NSRIT

AUTONOMOUS

ANSWER KEY & SCHEME OF EVALUATION

Semester V Regular
2020 Admitted

ACADEMIC REGULATION 2020

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NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	ECE			Academic Year	2022 - 2023
Course Code	20EC501	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Analog and Digital Communications						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define amplitude modulation and draw the single tone AM signal waveform.	20EC501.1	L1
2	Define frequency deviation in FM.	20EC501.2	L1
3	What is quantization error?	20EC501.3	L1
4	Draw the block diagram of Coherent detection of FSK.	20EC501.4	L2
5	Define entropy.	20EC501.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain how the amplitude modulation can be expressed in time domain and frequency domain	6M	20EC501.1	L2
6 (b)	Explain the generation of SSB by using the phase discrimination method.	6M	20EC501.1	L2
	OR			
7 (a)	Explain the generation technique of an AM wave using the square law modulator.	7M	20EC501.1	L2
7 (b)	Compare different types of AM systems in terms of its merits and demerits.	5M	20EC501.1	L3
8	Define modulation index in FM. Discuss the spectra of NBFM and WBFM for various modulation indices.	12M	20EC501.2	L2
	OR			
9 (a)	Draw the block diagram of Super-heterodyne Receiver and explain each block.	6M	20EC501.2	L2
9 (b)	Show that Narrow band FM is equivalent to AM with respect to transmission bandwidth.	6M	20EC501.2	L3
10 (a)	Explain the term Quantization and its types.	4M	20EC501.3	L2
10 (b)	Derive the output signal power to quantization noise in a PCM system.	8M	20EC501.3	L3
	OR			
11	Explain the operation of delta modulation in detail and its drawbacks along with necessary expressions	12M	20EC501.3	L3
12 (a)	Draw and explain the power spectral density (PSD) and geometrical representation of BPSK.	8M	20EC501.4	L2
12 (b)	Write any two comparisons among ASK and FSK.	4M	20EC501.4	L2
	OR			
13 (a)	Derive an expression for signal to noise ratio for integrator and dump filter.	7M	20EC501.4	L3
13 (b)	Derive an expression for error probability of ASK.	5M	20EC501.4	L3

14 (a)	Define information and explain its properties	6M	20EC501.5	L2
14 (b)	Define channel capacity and explain channel capacity for discrete channels.	6M	20EC501.5	L2
	OR			
15 (a)	Explain about the trade–off between bandwidth and S/N ratio.	8M	20EC501.5	L2
15 (b)	What are the advantages of source coding techniques? Explain any one of the source coding technique steps.	4M	20EC501.5	L2

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SCHEME OF EVALUATION

Part –A

1. Amplitude modulation Definition & single tone AM signal waveform............ 1M+1M
2. frequency deviation in FM Definition 2M
3. Quantization error 2M
4. Block diagram of Coherent detection of FSK 2M
5. Definition of entropy 2M

Part-B

6. (a) Explanation for how the amplitude modulation can be expressed in time domain and frequency domain...3M+3M

 (b) Generation of SSB signal using phase discrimination method 6M

 Diagram.......................................1M
 Theory..1M
 Expressions.................................2M
 Spectrum....................................2M

OR

7. (a) Generation technique of an AM wave using the square law modulator 7M

 Diagram.......................................1M
 Theory..2M
 Expressions.................................2M
 Spectrum....................................2M

 (b) Comparisons of AM, D.S.B-SC and S.S.B-SC transmission 5M

8. modulation index in FM Definition Explanation about the spectra of NBFM and WBFM...................2M+10M

 For NBFM 4M
 For WBFM 6M

OR

9. (a) Block diagram of super-heterodyne receiver and explanation of each block......6M

 Diagram.......................................2M
 Theory..4M

 (b) Narrow band FM is equivalent to AM with respect to transmission bandwidth 6M

10. (a) Quantization & its types ....2M+2M

 (b) Derivation for the output signal to quantization noise ratio in PCM..........8M

(OR)

11. Delta Modulation Diagram....................................1M
Principle and operation..........................2M
Theory..3M
Expressions..................................3M
Drawbacks..................................3M

12. (a) Explanation the power spectral density (PSD) and geometrical representation of BPSK...4M+4M
(b) two comparisons among ASK and FSK..4M

(OR)

13. (a) Expression for Signal to Noise Ratio for Integrator and dump filter..............7M
(b) Derivation for the probability error of ASK system...................................5M

14. (a) Definition of Information...............................4M
Information Properties................................2M
(b) Definition for channel capacity & Explanation for channel capacity for discrete channels....2M+4M

(OR)

15. (a) Explanation for the trade–off between bandwidth and S/N ratio.............8M
Theory...........1M
Derivation........6M
Conclusion.......1M
(b) Advantages of source coding 2M
Steps for any one of source coding technique...............2M

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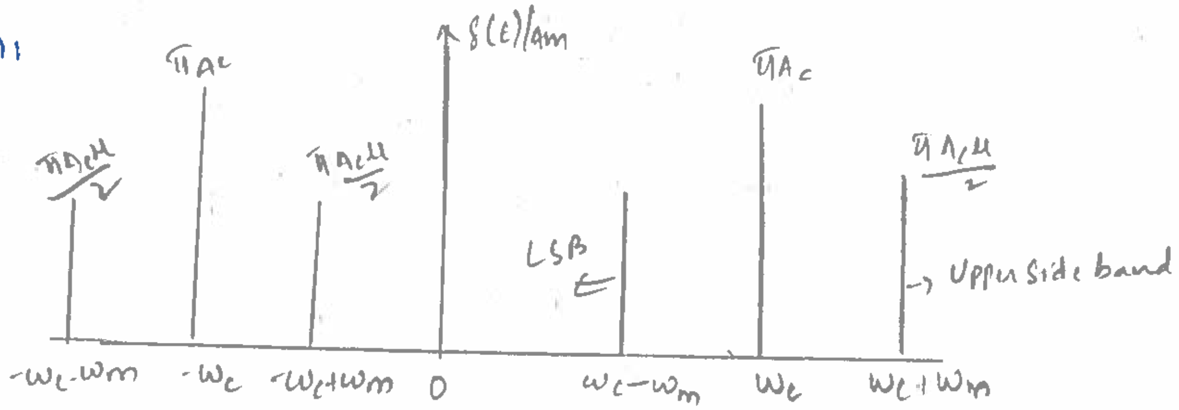
ANSWER KEY AND SCHEME OF EVALUATION

PART-A (Short Answers)

1. Define Amplitude Modulation and draw the Single tone AM Signal waveform.

A. It is a process of Amplitude of carrier Varies in accordance with the message Signal to convert Low frequency Signal to high frequency Signal. It is called as Amplitude Modulation.

Waveform:



2. Define frequency deviation in FM.

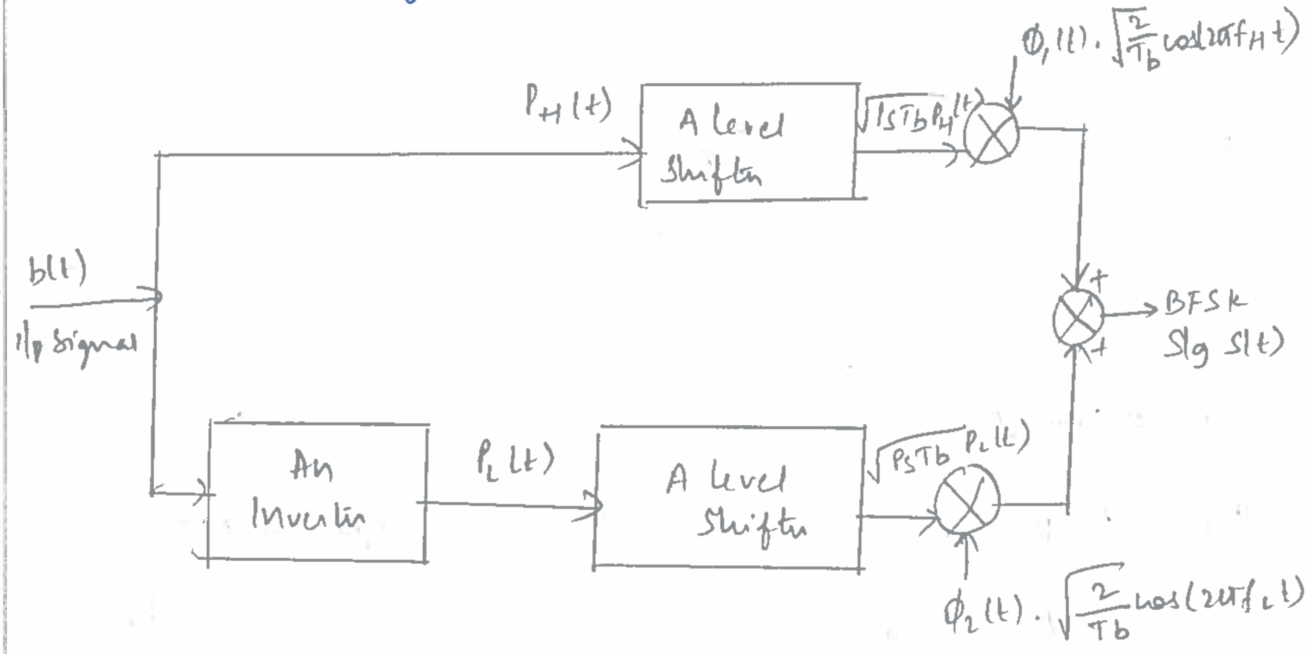
A. It is defined as carrier frequency subtraction from instantaneous frequency. It is denoted as $\Delta w(t)$

$$\Delta w(t) = w_i(t - w_c)$$

3. What is quantization Error?

A. Quantization Error is the difference b/w the analog signal and the closest available digital value at each Sampling instant from the A/D converter. Quantization Error also introduces noise, called quantization noise to the Sample signal.

4. Draw the block diagram of coherent detection of FSK.



5. Define Entropy.

A. The average amount of Information carried by each message is called Entropy $H(x)$

$$\therefore H(x) = \frac{I_{total}}{L} = \sum_{i=1}^{m} P_i \log \frac{1}{P_i} \text{ bits/symbol.}$$

$$\boxed{\begin{aligned} H(x) &= \sum_{i=1}^{m} P(x_i)\log_2 \frac{1}{P(x_i)} \\ H(x) &= -\sum_{i=1}^{m} P(x_i)\log_2 P(x_i) \end{aligned}}$$ bits/symbol

PART-B (Long Answers)

6(a). Explain how the amplitude modulation can be Expressed in time domain and frequency domain.

A. Time Domain Analyses

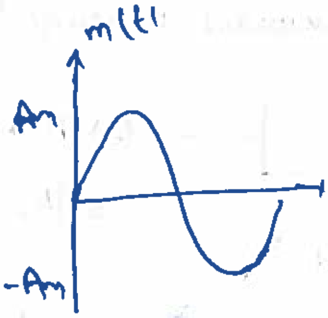
Case i : $\mu < 1$

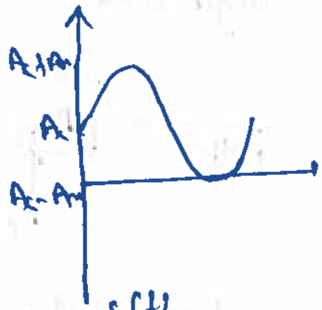
$\frac{A_m}{A_c} < 1$

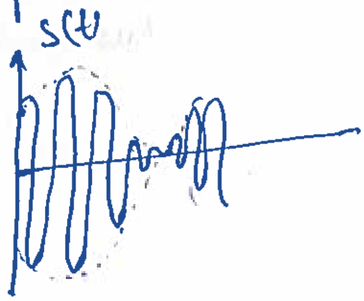
$A_m < A_c$

$A_m - A_c < 0$

$A_c - A_m > 0$







Case ii : $\mu = 1$

$\frac{A_m}{A_c} = 1$

$A_c = A_m$

$A_c - A_m = 0$

$s(t) = A(t) \cos w_c(t)$

$= [A_c + m(t)] \cos w_c t$

Case iii : $\mu > 1$

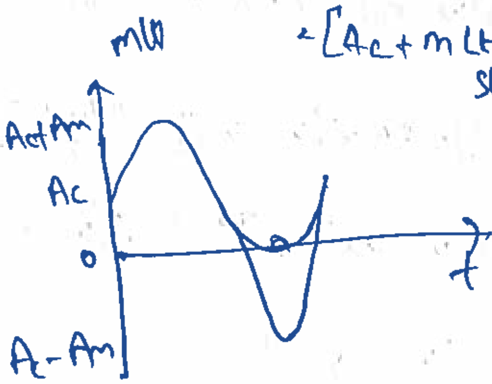
$\frac{A_m}{A_c} > 1$

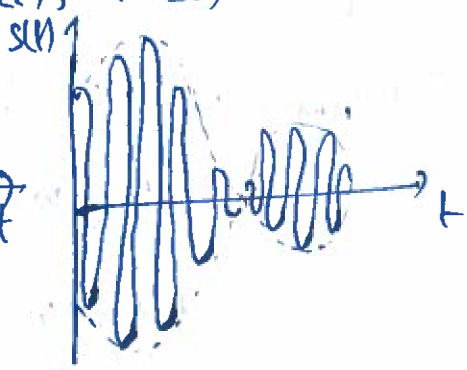
$A_c < A_m$

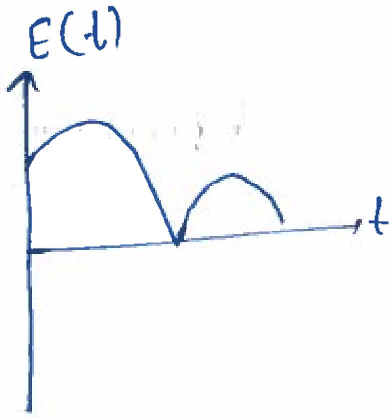
$A_c - A_m < 0$

$s(t) = A(t) \cos w_c(t)$

$= [A_c + m(t)] \cos w_c(t)$





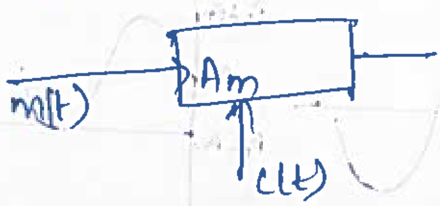


$S(t) \rightarrow$ Envelope

$$E(t) = |A(t)|$$

Time and Frequency Analysis of message Signal

Case i : Single tone Sinusoidal message Signal $m(t)$



$$S(t) = A(t) \cos \omega_c (t)$$

$$= [A_c + m(t)] \cos \omega_c t$$

message sig $m(t) = A_m \cos \omega_m t$

$$S(t) = [A_c + A_m \cos \omega_m t] \cos \omega_c t$$

$$A_c \left[1 + \frac{A_m}{A_c} \cos \omega_m t\right] \cos \omega_c t$$

$$A_c [1 + k_a A_m \cos \omega_m t] \cos \omega_c t$$

$$S(t)_{AM} = A_c [1 + \mu \cos \omega_m t] \cos \omega_c t$$

$$= A_c \cos \omega_c t + A_c \mu \cos \omega_c t . \cos \omega_m t$$

$$A_c \cos \omega_c t + \frac{A_c \mu}{2} . 2 \cos \omega_c t \cos \omega_m t$$

$$A_c \cos \omega_c t + \frac{A_c \mu}{2} [\cos (\omega_c + \omega_m) t + \cos (\omega_c - \omega_m) t]$$

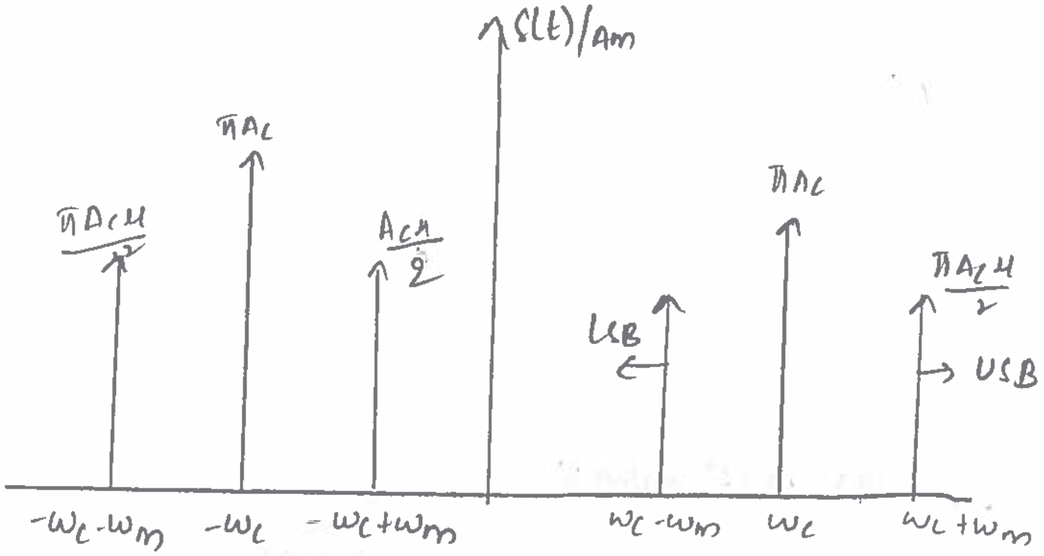
$$S(t)_{AM} = A_c \cos \omega_c t + \frac{A_c \mu}{2} \cos (\omega_c + \omega_m) t + \frac{A_c \mu}{2} \cos (\omega_c - \omega_m) t$$

$$\therefore \; FT [A \cos \omega_0 t] = \pi A \, \delta(\omega - \omega_0) + \pi A \, \delta(\omega + \omega_0)$$

In freq Analysis $\quad S(t)/_{AM} = \pi A_c \delta[(\omega - \omega_c)] + \pi A_c \delta[\omega + \omega_c] + \frac{\pi A_c \mu}{2} \delta[\omega - (\omega_c + \omega_m)]$

$$+ \frac{\pi A_c \mu}{2} \delta[\omega + (\omega_c + \omega_m)] + \frac{\pi A_c \mu}{2} \delta[\omega - (\omega_c - \omega_m)]$$

$$+ \frac{\pi A_c \mu}{2} \delta[\omega + (\omega_c - \omega_m)]$$



frequency components

For message sig m(t) - Am

carrier sig. c(t) - Ac

modulated sig. s(t) - fc

$= f_c \pm f_m$ → $f_c + f_m$ → USB, $f_c - f_m$ → LSB

→ Bandwidth = highest possible frequency - lowest possible frequency

$f_c + f_m - (f_c - f_m)$

$f_c + f_m - f_c + f_m$

$= 2f_m$

= 2 × maximum freq of msg sig.

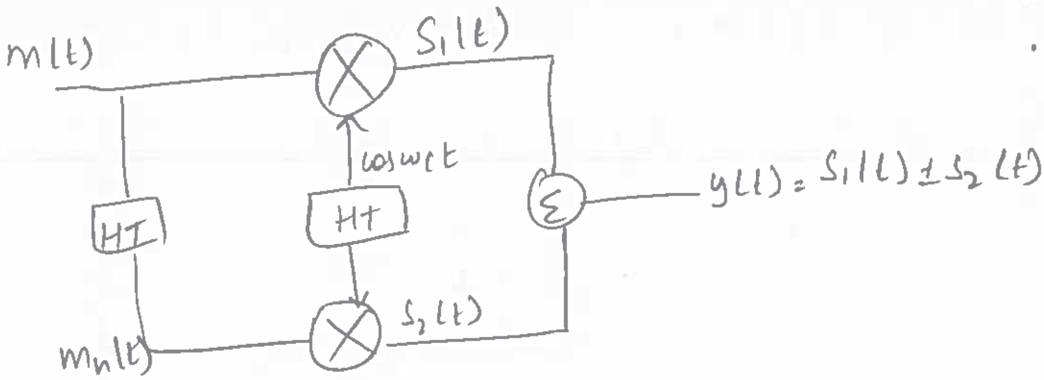
6(b) Explain the generation of SSB by using the phase discrimination method.

A. For high bandwidth require high transmitter for convience manner we require low transmitter power then require low band width.

Phase discrimination method.

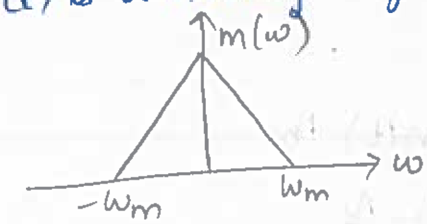
$y(t) = S_1(t) \pm S_2(t)$

$= m(t)\cos w_c t + m_h(t)\sin w_c t$



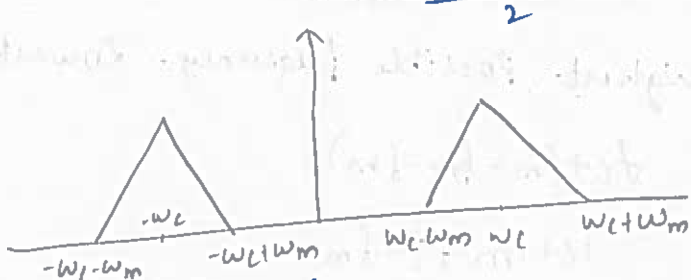
→ $s_1(t) = m(t)\cos\omega_c t$

where m(t) is a message signal



$$s_1(t) \xleftrightarrow{FT} S_1(\omega) \to FT[m(t)\cos\omega_c(t)]$$

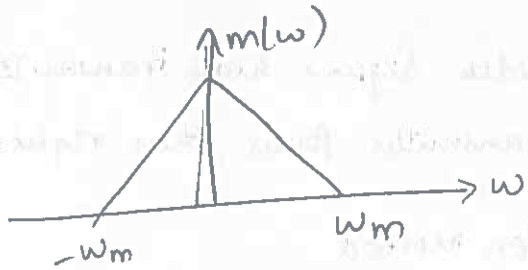
$$= \frac{m(\omega-\omega_c)}{2} + \frac{m(\omega+\omega_c)}{2}$$

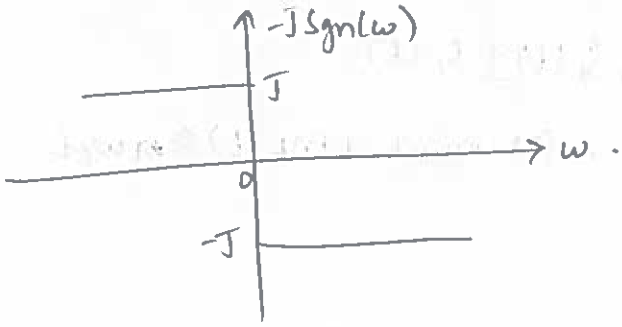


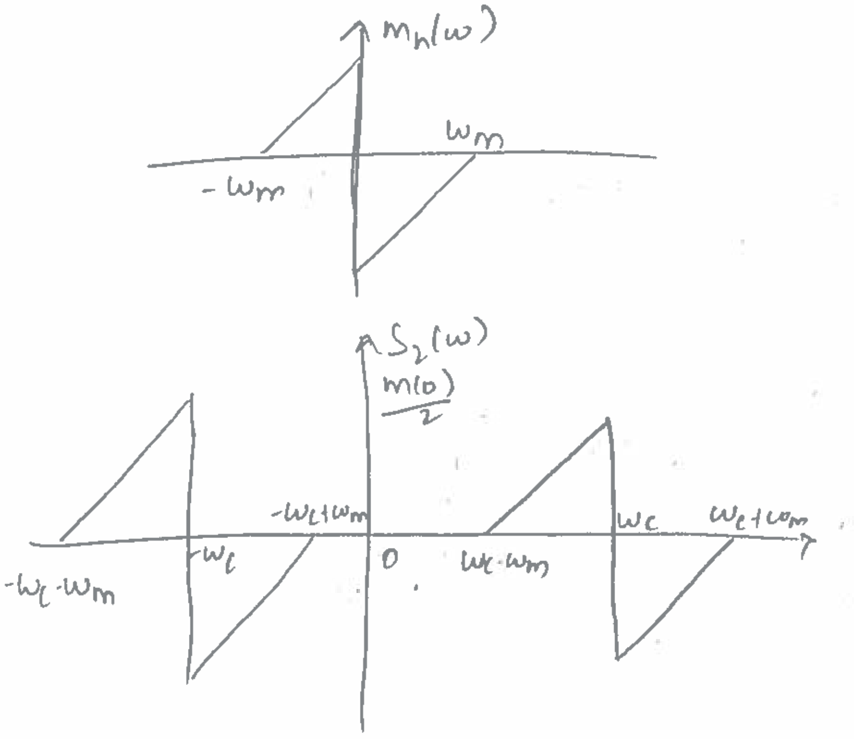
→ $s_2(t) = m_h(t)\sin\omega_c t$

$$M_h(\omega) \to HT[m_h(t)] \to HT\left[m(t) * \frac{1}{\pi t}\right]$$

$$M_h(\omega) \to M(\omega) \cdot -j\,sgn(\omega).$$



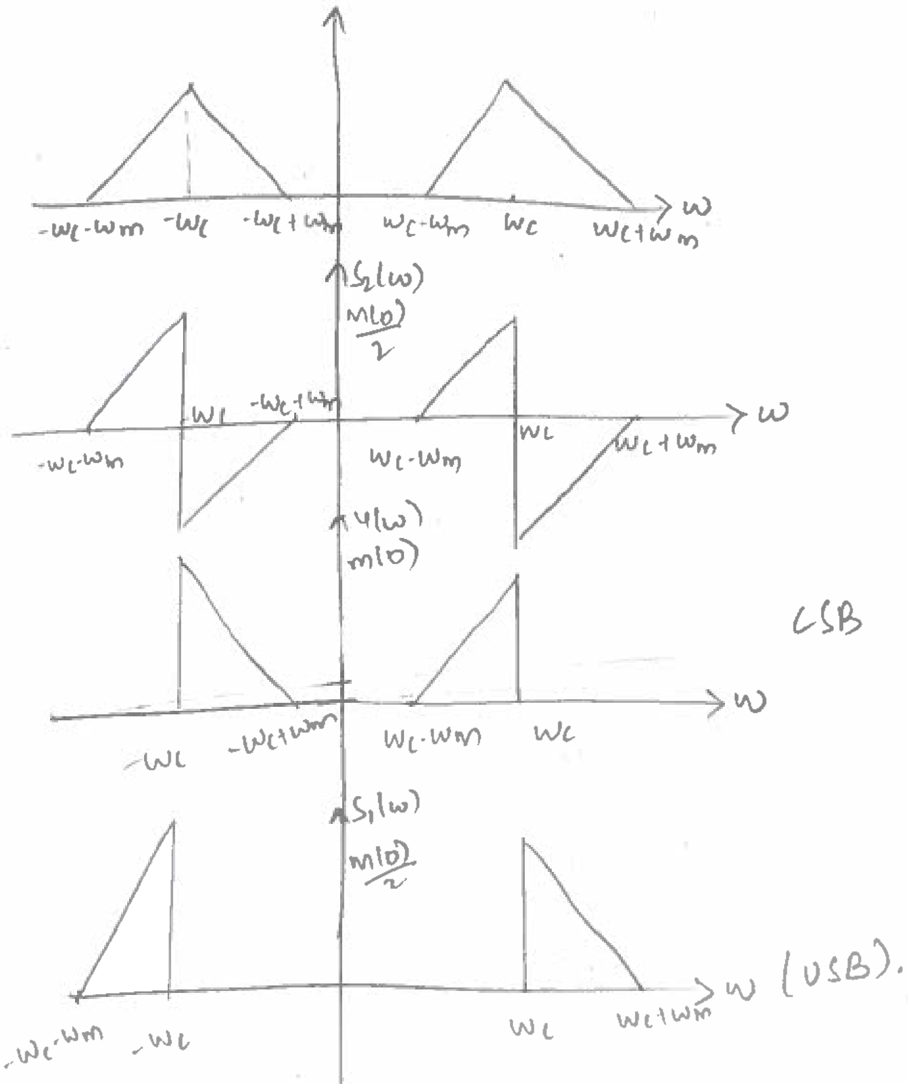




$y(t) = S_1(t) + S_2(t)$

$\downarrow FT$

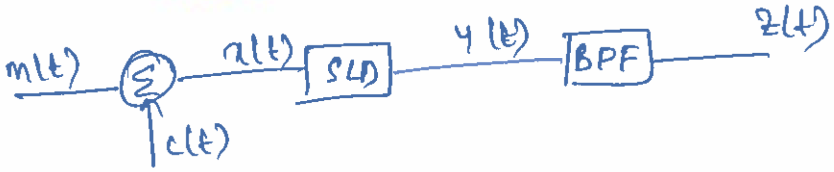
$Y(\omega) = S_1(\omega) + S_2(\omega)$



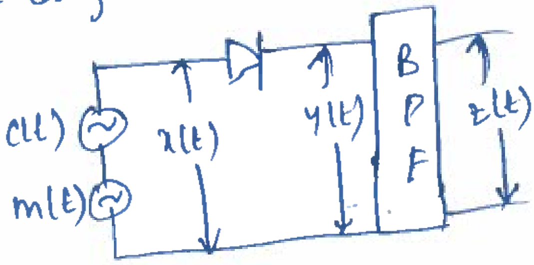
$Y(\omega) = S_1(\omega) - S_2(\omega)$

7(a) Explain the generation technique of an AM wave using the Square law modulator.

A.



Circuit diagram



$x(t) = c(t) + m(t)$

$c(t) = A_c \cos w_c t \, , \; m(t) = A_m \cos w_m t$

$x(t) = A_c \cos w_c t + m(t)$

$y(t) = a_0 x(t) + a_1 x^2(t)$

$= a_0(A_c \cos w_c t + m(t)) + a_1 (A_c \cos w_c t + m(t))^2$

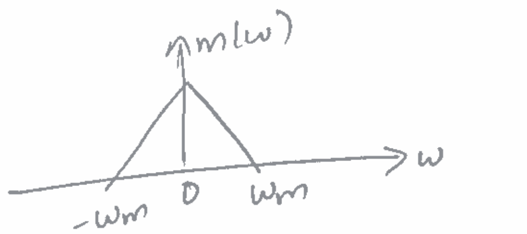
$y(t) = a_0 A_c \cos w_c t + a_0 m(t) + a_1 A_c^2 \cos^2 w_c t + a_1 m^2(t) + 2a_1 A_c m(t) \cos w_c t$

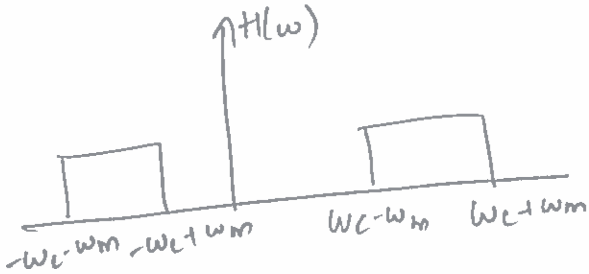
$y(t) = a_0 A_c \cos w_c t + a_0 m(t) + \frac{a_1 A_c^2}{2} + \frac{a_1 A_c^2 \cos 2w_c t}{2} + a_1 m^2(t) + a_1 2A_c m(t) \cos w_c t$

Let message sig, $m(t)$

$m(t) \xleftrightarrow{FT} M(\omega)$

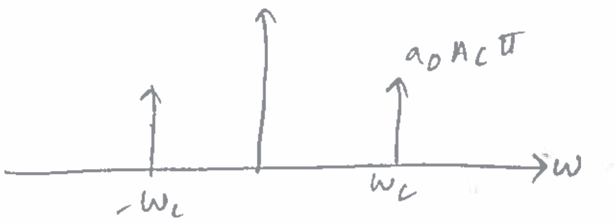
BPF $h(t) \xleftrightarrow{FT} H(\omega)$





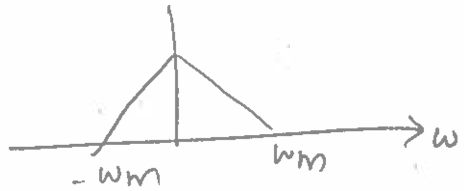
5

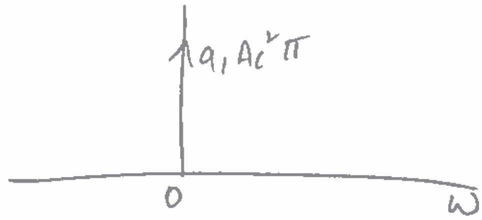
(i) $a_0 A_c \cos w_c(t) \xleftrightarrow{FT} a_0 A_c \pi\delta(w-w_0) + a_0 A_c \pi\delta(w+w_0)$



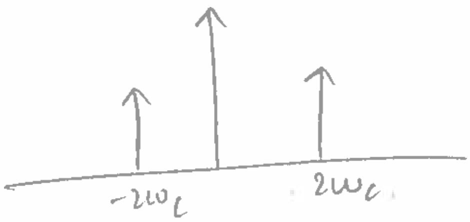
(ii) $a_0 m(t) \longleftrightarrow a_0 m(w)$

(iii) $\frac{a_1 A_c^2}{2} \longleftrightarrow a_1 A_c^2 \pi\delta(w-0)$



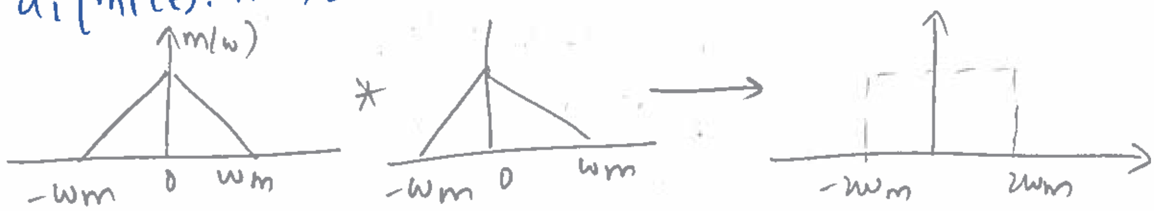


(iv) $a_1 A_c^2 \cos 2w_c t$

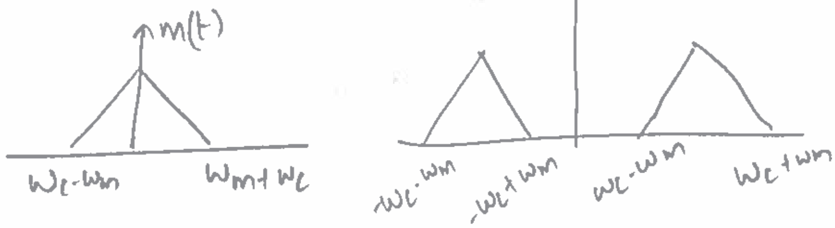


(v) $a_1 m^2(t)$

$a_1[m_1(t).m(t)] \longrightarrow a_1(m(w) * m(w))$



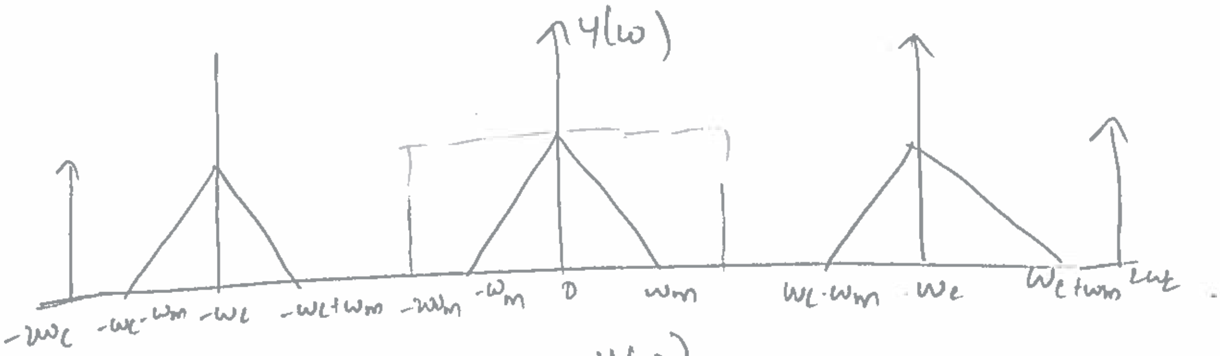
vi $2a_1 A_c m(t) \cos w_c t$

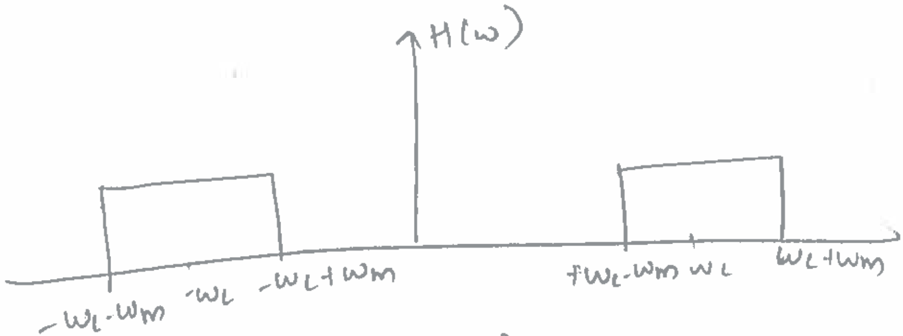


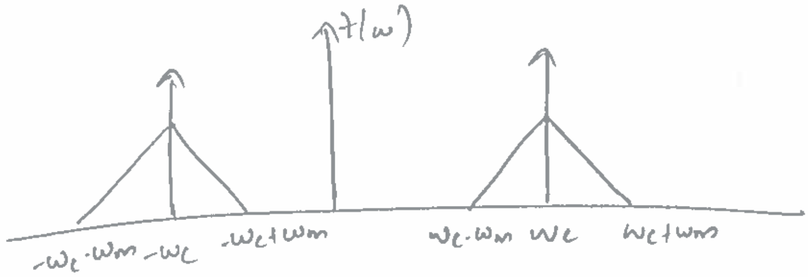
$z(t) = y(t) * h(t)$

$\downarrow F.T$

$Z(w) = Y(w)\, h(w)$







$$Z(\omega) = a_0 A_c \cos\omega_c t + 2a_1 A_c m(t) \cos\omega_c t$$

$$= a_0 A_c \left[1 + \frac{2a_1}{a_0} m(t)\right] \cos\omega_c t$$

$$= A_c' \left[1 + k_a m(t)\right] \cos\omega_c t$$

$$P_c = \frac{(A_c')^2}{2} = \frac{a_0^2 A_c^2}{2}$$

$$k_a = \frac{2a_1}{a_0}$$

$$\mu = k_a m(t)\big|_{max}$$

$$\mu = \frac{2a_1}{a_0} m(t)\big|_{max}.$$

7(b) Compare different types of AM systems in terms of its merits and demerits. 6

SNo	Parameter of Comparison	DSB FC (Standard AM)	DSBSC	SSB
1.	Carrier suppression	N.A	Fully	Fully
2.	Sideband suppression	N.A	N.A	One S.B Completely,
3.	Band width	$2f_m$	$2f_m$	f_m
4.	Transmission Efficiency	Minimum	Moderate	Maximum
5.	No. of Modulating inputs	1	1	1
6.	Application	Radio Broadcasting	Radio broadcasting	Point to point mobile Communication

8. Define modulation index in FM. Discuss the spectra of NBFM and WBFM for various modulation indices.

A. The modulation index of an FM wave is defined as

$$m_f = \frac{\text{Frequency deviation}}{\text{Modulating freq.}}$$

(or) $$m_f = \frac{\Delta f}{f_m}.$$

The modulation index (m_f) is very important in FM because.

It decides the B.W of the FM wave. The modulation index also decides the no. of sidebands having significant amplitudes.

Narrow band FM :

$$S_{FM}(t) = A_c \cos[\omega_c t + \beta \sin \omega_m t]$$

It contains carrier information in two side bands.

$$S_{FM}(t) = A_c \cos\omega_c t \cos\beta \sin\omega_m t - A_c \sin\omega_c t \sin(\beta \sin\omega_m t)$$

here $\beta << 1$ (very less)

Let $\theta = \beta \sin \omega_m t$

θ is very small because of $\beta << 1$

$$S_{FM}(t) = A_c[\cos[\omega_c t \cos\theta] - A_c(\sin\omega_c t \cdot \sin\theta)$$

Let for calculation purpose $\cos\theta = 1$ & $\sin\theta = \theta$

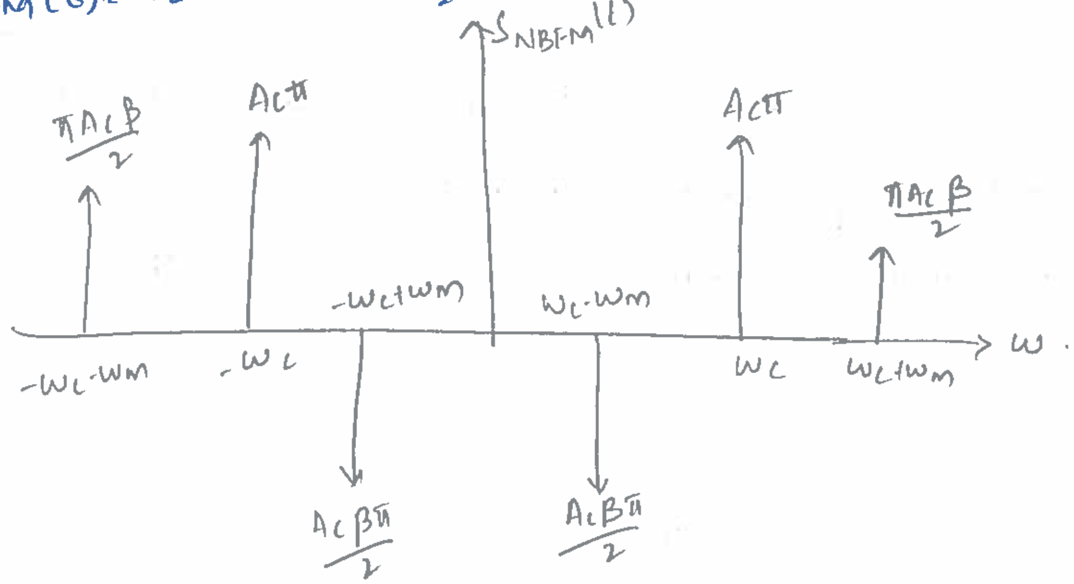
$$S_{FM}(t) = A_c \cos\omega_c t - A_c \sin\omega_c t\ \theta.$$

$$S_{FM}(t) = A_c \cos\omega_c t - A_c\beta \sin\omega_c t \sin\omega_m t$$

$$= A_c \cos\omega_c t - \frac{A_c\beta}{2} 2\sin\omega_c t \sin\omega_m t$$

$$= A_c \cos\omega_c t - \frac{A_c\beta}{2}\left[\cos(\omega_c - \omega_m)t - \cos(\omega_c + \omega_m)t\right]$$

$$S_{NBFM}(t) = A_c \cos\omega_c t + \frac{A_c\beta}{2}\cos(\omega_c + \omega_m t) - \frac{A_c\beta}{2}\cos(\omega_c + \omega_m)t]$$



$BW = \omega_c + \omega_m - \omega_c + \omega_m$

$= 2\omega_m$

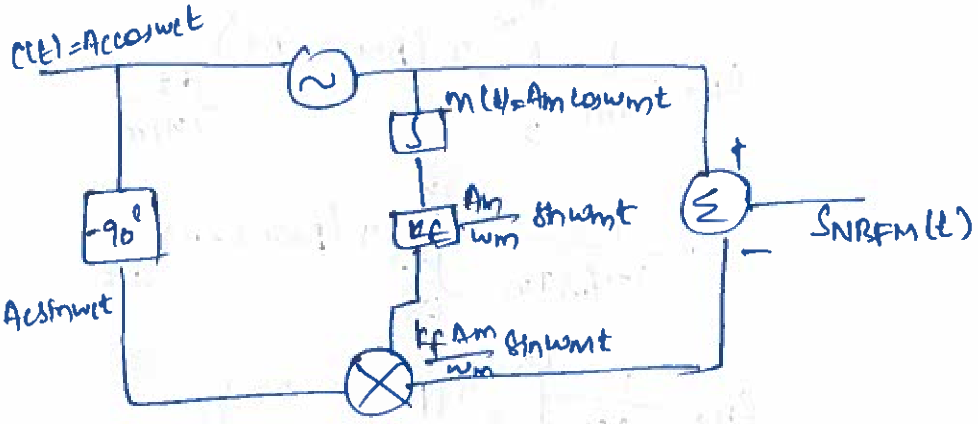
$= 2 \times$ max. freq of msg sig (or) modulated signal.

Block diagram

$$S_{FM} = A_c \cos(w_c t + \beta \sin w_m t)$$

$$S_{NBFM} = A_c \cos(w_c t) - A_c \beta \sin w_c t \sin w_m t$$

$$= A_c \cos(w_c t) - A_c \frac{k_f A_m}{w_m} \sin w_c t \sin w_m t$$



Wide band frequency modulation :

→ when $\beta > 1$, frequency modulation acts as one wide band freq. modulation

It consisting of one carrier sig and more than one side band

$$S_{FM}(t) = A_c \cos(w_c t + \beta \sin w_m t)$$

$$m(t) = A_m \cos w_m t$$

$$= A_m \cos 2\pi f_m t$$

$$c(t) = A_c \cos w_c t$$

$$A_c \cos 2\pi f_c t$$

$$S_{FM}(t) = A_c \cos[2\pi f_c t + \beta \sin 2\pi f_m t]$$

In above Expression consider real terms only

$$S_{FM}(t) = A_c \, Re\{ e^{j(2\pi f_c t + \beta \sin 2\pi f_m t)} \}$$

$$S_{WBFM}(t) = A_c Re\{ e^{j2\pi f_c t} \cdot e^{j\beta \sin 2\pi f_m t} \}$$

Let $x(t) = e^{j\beta \sin 2\pi f_m t}$

$$x(t) = e^{j\beta \sin w_m t}$$

Complex Exponential terms are always Periodic Signals

Now calculate F.S for $x(t)$

$$EFS[x(t)] = \sum_{n=-\infty}^{\infty} C_n e^{jn\omega_m t_m}$$

$$C_n = \frac{1}{T_m}\int_0^{T_m} x(t) e^{-jn\omega_m T_m} dt$$

$$C_n = \frac{1}{T_m}\int_0^{T_m} e^{j(\beta \sin z - nz)} \frac{dz}{2\pi f_m}$$

$$= \frac{1}{2\pi f_m T_m}\int_0^{2\pi} e^{j(\beta \sin z - nz)} dz$$

$$C_n = \frac{1}{2\pi}\int_0^{2\pi} e^{j(\beta \sin z - nz)} dz$$

It is nothing but a bessel's function of n^{th} order

$\Rightarrow C_n = J_n(\beta)$

C_n values substitute in EFS $x(t)$ eqn.

$$EFS[x(t)] = \sum_{n=-\infty}^{\infty} J_n(\beta) e^{jn\omega_m F_m}.$$

$$EFS\left[e^{j\beta \sin \omega_m t}\right] = \sum_{n=-\infty}^{\infty} J_n(\beta) e^{jn\omega_m t_m} \longrightarrow ①.$$

$$\rightarrow S_{WBFM}(t) = A_c R_e\left\{e^{j2\pi f_c t} \cdot e^{j\beta \sin \omega_m t}\right\}$$

$$= R_e\left\{\sum_{n=-\infty}^{\infty} A_c J_n(\beta) e^{j[2\pi(f_c + nf_m)]t}\right\}$$

$$S_{WBFM}(t) = \sum_{n=-\infty}^{\infty} A_c J_n(\beta) \cos\left[2\pi(f_c + nf_m)t\right]$$

$\Rightarrow$ Properties of Bessel's function:

1. $\sum_{n=-\infty}^{\infty} J_n^2(\beta) = 1$
2. $J_{-n}(\beta) = (-1)^n J_n(\beta)$
3. $J_{-n}^2(\beta) = J_n^2(\beta)$
4. $\sum_{n=-\infty}^{\infty} J_n^2(\beta) = 1$
5. when $n \uparrow$ then $J_n(\beta)$

$$\sum_{n=-\infty}^{\infty} J_n^2(\beta) = + \ldots + J_{-3}^2(\beta) + J_{-2}^2(\beta) + J_1^2(\beta) + J_0^2(\beta) + J_1^2(\beta) + J_2^2(\beta) + J_3^2(\beta) + \ldots = 1$$

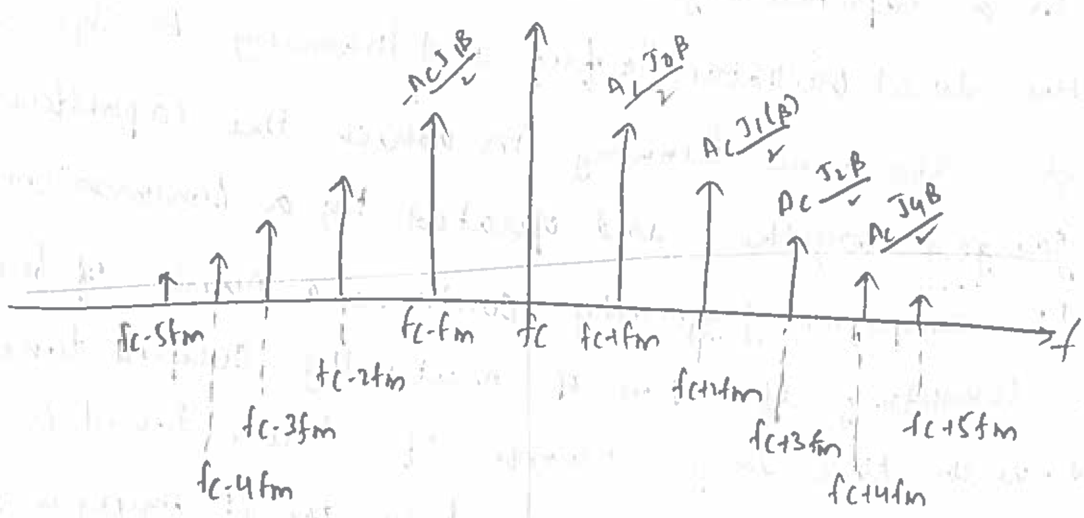
8

$$= J_0^2(\beta) + 2\left[J_1^2(\beta) + J_2^2(\beta) + J_3^2(\beta) + \ldots\right] = 1$$

from Eq (1)

$$S_{WBFM} = + \ldots + A_c\,(-J_3\beta)\cos[2\pi(f_c - 3f_m)]t + A_c J_2 \beta\cos[2\pi(f_c - 2f_m)t] - A_c J_1 \beta \cos[2\pi(f_c - f_m)]t + A_c J_0(\beta)\cos[2\pi f_c t] + A_c J_1(\beta)\cos[2\pi(f_c + f_m)t] + A_c J_2(\beta)\cos[2\pi(f_c + 2f_m)t] + A_c J_3(\beta)\cos[2\pi(f_c + 3f_m)t]$$

Frequency Spectrum

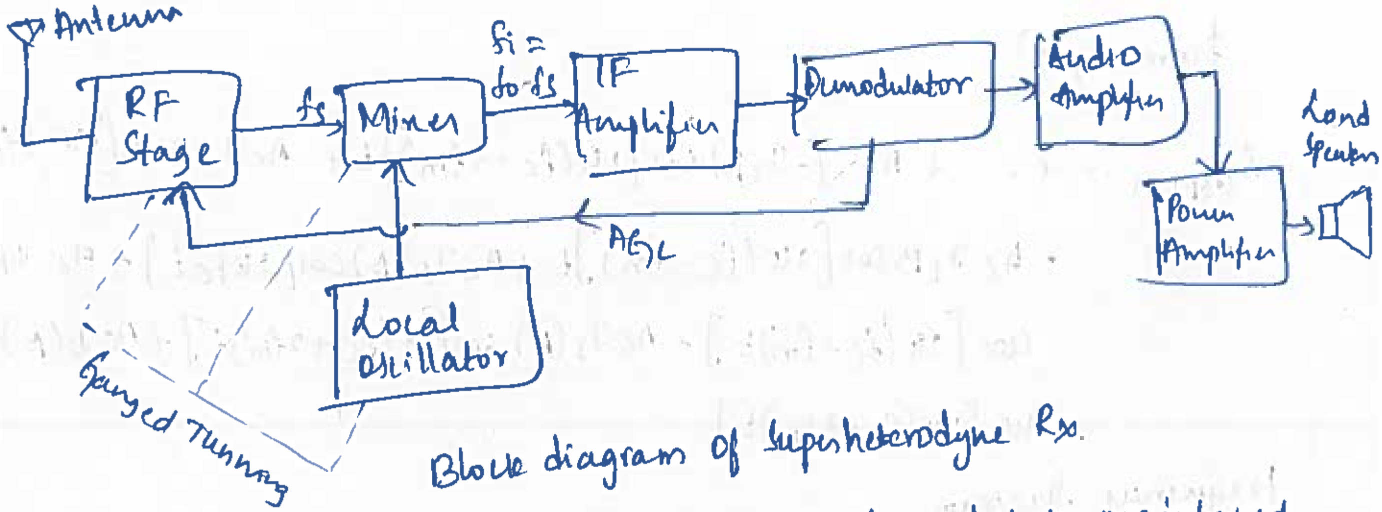


7(a). Draw the block diagram of Superheterodyne Receiver and Explain block diagram.

A. All the drawbacks in TRF Rx has been removed in a Super heterodyne receiver. The basic superheterodyne receiver is most widely used. The Super heterodyne Rx is used in all types of Rx like T.V Rx, radar Rx etc.,

In a Superheterodyne Rx, the incoming RF S/g freq is combined with the local oscillator S/g freq through a mixer and is converted into a signal of lower fixed freq.

the lower fixed freq. is known as intermediate frequency. However, the intermediate freq. slg contains the same modulation as the original slg. The intermediate freq. is now amplified and demodulated to reproduce the original slg.



Block diagram of superheterodyne Rx.

In a superheterodyne Rx, a constant freq diff is maintained b/w the local oscillator slg freq and incoming RF slg's freq through capacitance tunning in which the capacitances are ganged together and operated by a common control knob.

The IF amplifier generally contains a number of transformers each consisting of a pair of mutually coupled tuned circuits. Thus, with this large number of double-tuned circuits, operating at a specially chosen freq, the IF Amplifier provides most of the gain (sensitivity) and band width (selectivity) requirements of the receiver.

Since, the characteristics of the IF Amplifier are independent of the incoming freq to which the receiver is tuned, the selectivity and sensitivity of the superheterodyne Rx are quite uniform throughout its tunning range and not subject to the variations like a TRF Rx. IF Amplifier works at a fixed I.F freq.

After IF Amplifier, the s/g is applied at the i/p of demodulator which Extracts the Original modulating Signal. This audio s/g is amplified by an audio Amplifier to get a particular voltage level. This amplified audio signal is further amplified by power Amplifier to get a specified power level so that it may activate the loudspeaker. The loudspeaker is a transducer which converts this audio Electrical s/g into audio sound s/g and thus the original signal is reproduced.

Advantages.

(i) No variation in bandwidth, The BW remains constant over the Entire operating range

(ii) High Sensitivity & selectivity

(iii) High adjacent channel rejection.

Frequency Parameters of AM receiver.

(i) Two freq. bands: Medium wave band & Short wave band

(ii) RF carrier range (MW): 535 kHz to 1650 kHz (SW band): 5 to 15 MHz

(iii) Intermediate frequency IF: 455 kHz

(iv) IF bandwidth B: 10 kHz.

9(b). Show that Narrow band FM is Equivalent to AM with respect to transmission bandwidth.

4. Narrow band FM:

$$S_{FM}(t) = A_c \cos[\omega_c t + \beta \sin \omega_m t]$$

It contains carrier Information in two sidebands

$$S_{FM}(t) = A_c \cos\omega_c t \cos\beta \sin\omega_m t - A_c \sin\omega_c t \sin(\beta \sin\omega_m t)$$

here $\beta << 1$ (very less)

Let $\theta = \beta \sin\omega_m t$

θ is very small because of $\beta << 1$

$S_{FM}(t) = A_c[\cos w_c t \cos\theta] - A_c[\sin w_c t \sin\theta]$

Let for calculation purpose $\cos\theta = 1$ & $\sin\theta = \theta$

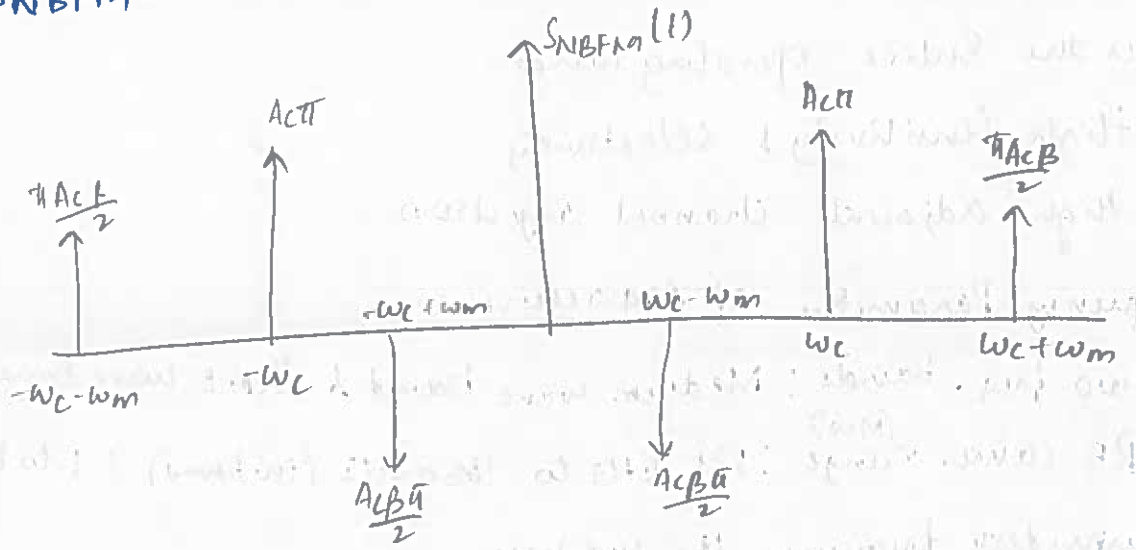
$S_{FM}(t) = A_c \cos w_c t - A_c \sin w_c t \, \theta$

$S_{FM}(t) = A_c \cos w_c t - A_c \beta \sin w_c t \sin w_m t$

$= A_c \cos w_c t - \frac{A_c \beta}{2} 2\sin w_c t \sin w_m t.$

$= A_c \cos w_c t - \frac{A_c \beta}{2}[\cos(w_c - w_m)t - \cos(w_c + w_m)t]$

$S_{NBFM}(t) = A_c \cos w_c t \frac{A_c \beta}{2} \cos(w_c + w_m)t - \frac{A_c \beta}{2} \cos(w - w_m)t$

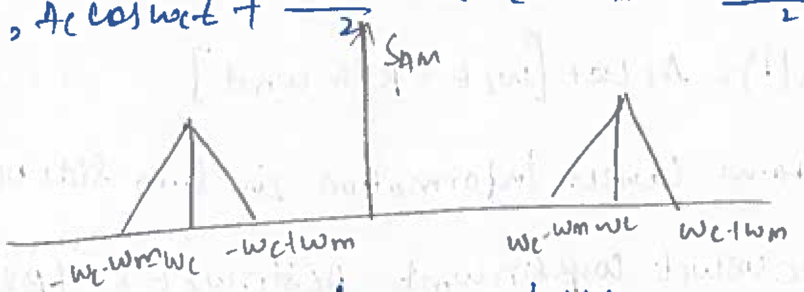


$Bw = \cancel{w_c} + w_m - \cancel{w_c} + w_m$

$Bw = 2w_m \longrightarrow ①$

For Am,

$S_{AM} = A_c \cos w_c t + \frac{A_c \mu}{2} \cos(w_c + w_m)t + \frac{A_c \mu}{2} \cos(w_c - w_m)t$



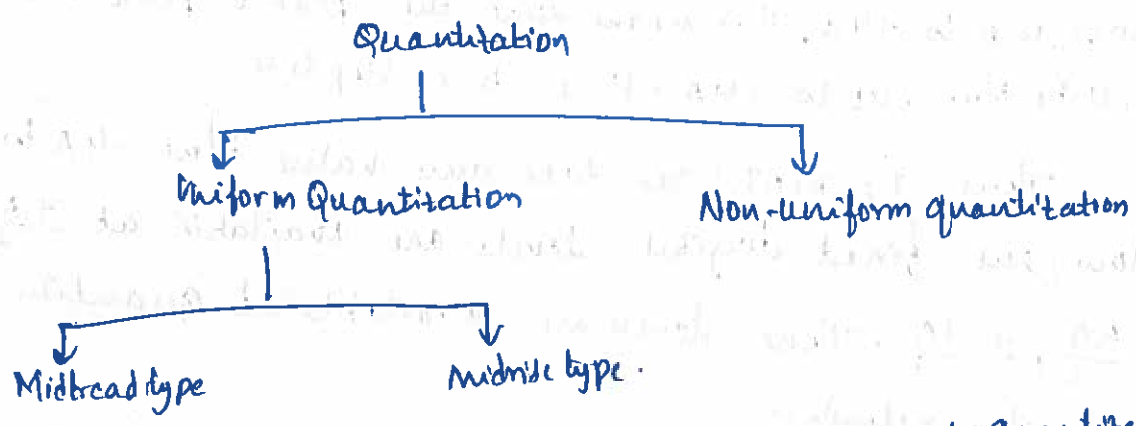
$Bw = \cancel{w_c} + w_m - \cancel{w_c} + w_m$

$Bw = 2w_m \longrightarrow ②$

By comparing eq ① & eq ②, we can say that Narrow Band FM is Equivalent to AM w.r.t transmission bandwidth.

10(a) Explain the term quantitation & its types 10

Quantisation Process may be classified as



(i) Uniform Quantizer: A uniform Quantizer is that type of quantizer in which step size remains same throughout the I/p range

(ii) Non uniform Quantizer: It is a type of quantizer in which 'Step size' Varies according to the I/p sig Value.

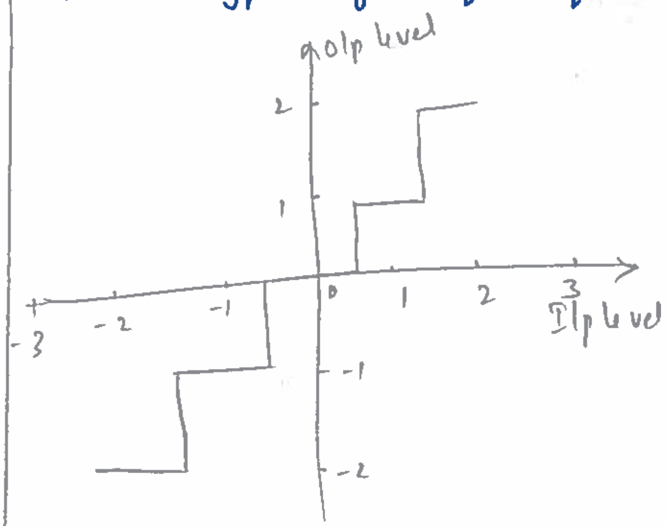
Uniform Quantizer:

There are 2 types of Uniform Quantizer as Under:

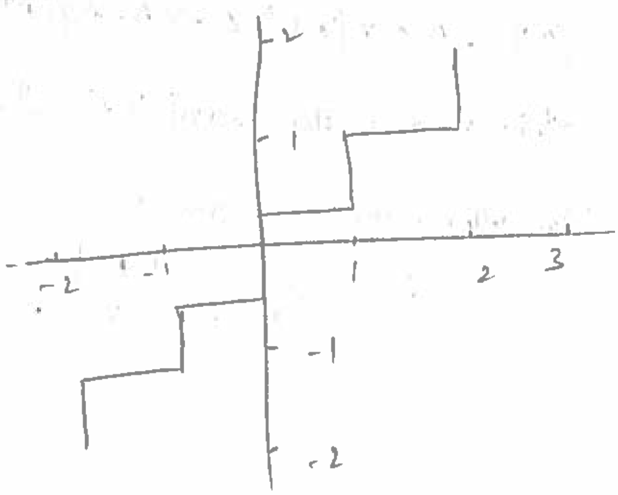
(i) Symmetric quantizer of the midtread type

(ii) Symmetric Quantizer of the mid rise type.

In a uniform quantizer, the representation levels are Uniformly spaced. The quantizer characteristic can also be a mid tread or midrise type. In midtread type, the Origin lies in ~~the~~ the middle of a tread of the Staircase like graph. In midrise type, the Origin lies in the middle of a rising Part of the Staircase like graph. Both the midtread and mid rise types of Uniform quantizers are Symmetric about the origin.



(a) midtread



(b) mid rise.

working principle of quantizer.

In fig(1), let us assume that the I/p of quantizer $x(nT_s)$ varies from -4Δ to $+4\Delta$. This means that the peak to peak value of $x(nT_s)$ will be b/w -4Δ to $+4\Delta$. Here, 'Δ' is step size.

Thus, I/p $x(nT_s)$ can take any value b/w -4Δ to $+4\Delta$. Now, the fixed digital levels are available at $\pm\frac{\Delta}{2}, \pm\frac{3\Delta}{2}, \pm\frac{5\Delta}{2}, \pm\frac{7\Delta}{2}$. These levels are available at quantizer because of its characteristic.

$x(nT_s) = 4\Delta$, then $x_q(nT_s) = \frac{7}{2}\Delta$ (from fig(1))

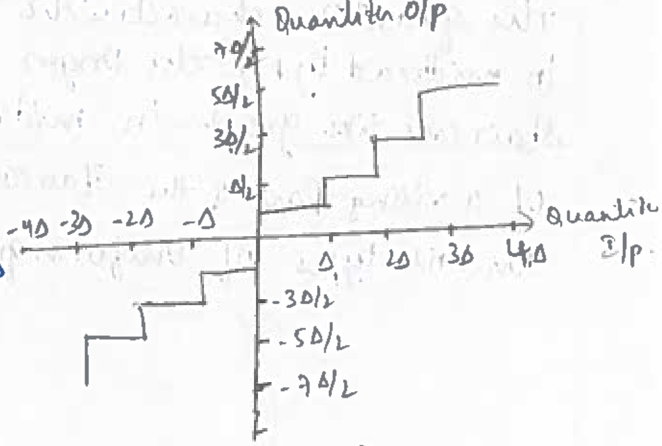
and if $x(nT_s) = -4\Delta$, then $x_q(nT_s) = \frac{-7}{2}\Delta$

from fig(2), max. quantization Error must be $\pm\frac{\Delta}{2}$

From above, we conclude the quantization Error as.

$$\mathcal{E} = x_q(nT_s) - x(nT_s)$$

Now, when $x(nT_s) = 0$, Quantizer will assign any one of the nearest binary ~~val~~ levels i.e., $\frac{\Delta}{2}$ (or) $-\frac{\Delta}{2}$. If $\frac{\Delta}{2}$ is assigned, the $\mathcal{E}$ will be.



fig(1)

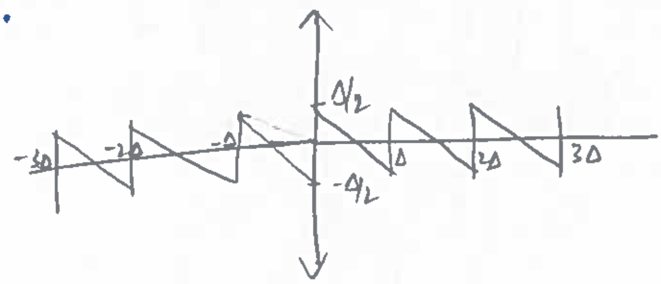
$\mathcal{E} = x_q(nT_s) - x(nT_s) = \frac{\Delta}{2} - 0 = \frac{\Delta}{2}$

for $\Delta < x(nT_s) < 2\Delta$, $x_q(nT_s) = \frac{3}{2}\Delta$

(or) $-\Delta < x(nT_s) < -2\Delta$, $x_q(nT_s) = \frac{-3}{2}\Delta$

this means the max($\mathcal{E}$) is $\pm\frac{\Delta}{2}$

In otherwords, max $\mathcal{E}$ is given by $\mathcal{E}_{max} = \left|\frac{\Delta}{2}\right|$.



fig(2)

10(b) Derive the output signal power to quantization noise in a PCM system.

A. The Quantization Error is given as

$$\varepsilon = x_q(nT_s) - x(nT_s)$$

Let us assume that the I/p $x(nT_s)$ to a linear or Uniform Quantizer has continuous amplitude in the range of $-x_{max}$ to $+x_{max}$

$$\text{Total amplitude range} = x_{max} - (-x_{max}) = 2x_{max}$$

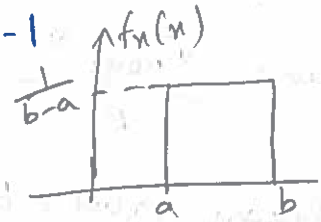
Now, if the amplitude range is divided into 'q' levels of Quantizer, then step size 'Δ' will be,

$$\text{Step size } \Delta = \frac{x_{max} - (-x_{max})}{q} = \frac{2x_{max}}{q}$$

If sig $x(t)$ is normalized to min and max values Equal to 1, then $x_{max} = 1$, $-x_{max} = -1$

then step size would be

$$\Delta = \frac{2}{q}$$

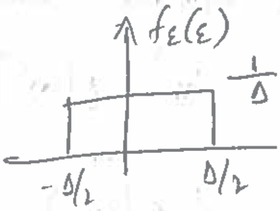


fig(1)

The maximum quantization Error is given as

$$\varepsilon_{max} = \left|\frac{\Delta}{2}\right|$$

i.e., $$-\frac{\Delta}{2} \le \varepsilon_{max} \le \frac{\Delta}{2}$$



$\frac{1}{\Delta} = \frac{1}{\Delta/2 - (-\Delta/2)}$

fig(2)

Hence, over the interval $\left(-\frac{\Delta}{2}, \frac{\Delta}{2}\right)$, ε may be assumed as an Uniformly distributed random Variable.

Fig(1) shows an Uniformly distributed random Variable 'X' over an interval (a,b).

$$f_X(x) = \begin{cases} 0 & \text{for } x < a \\ \frac{1}{b-a} & \text{for } a < x \le b \\ 0 & \text{for } x > b \end{cases} = \frac{\Delta^2}{12}, \text{ for linear quantization}$$

The PDF for quantization error is defined as

$$f_\varepsilon(\varepsilon) = \begin{cases} 0 & \text{for } \varepsilon \le -\frac{\Delta}{2} \\ \frac{1}{\Delta} & \text{for } -\frac{\Delta}{2} < \varepsilon \le \frac{\Delta}{2} \\ 0 & \text{for } \varepsilon > \frac{\Delta}{2} \end{cases} \longrightarrow Eq(a)$$

from fig(2), ε has zero average value, the mean m_ε of the quantization error is 0. The sig to quantization noise ratio of a quantizer is defined as

$$\frac{S}{N} = \frac{\text{Signal power (normalised)}}{\text{Noise power (normalised)}}$$

If $x(t)$ is known, then it is possible to calculate sig power. The noise power is defined as

$$\text{Noise power} = \frac{V^2_{noise}}{R} \xrightarrow{Eq(b)} [V^2_{noise} = \text{mean square value of noise voltage}]$$

$$\text{Mean square value} = E[\varepsilon^2] = \overline{\varepsilon^2} = V^2_{noise} \longrightarrow Eq(c)$$

The mean square value of a random variable 'x' is defined as

$$\overline{X^2} = E[x^2] = \int_{-\infty}^{\infty} x^2 f_x(x)\,dx$$

$$E[\varepsilon^2] = \int_{-\infty}^{\infty} \varepsilon^2 f_\varepsilon(\varepsilon)\,d\varepsilon$$

from eq(a), the above eqn is defined as

$$E[\varepsilon^2] = \int_{-\Delta/2}^{\Delta/2} \varepsilon^2 \times \frac{1}{\Delta}\,d\varepsilon = \frac{1}{\Delta}\left[\frac{\varepsilon^3}{3}\right]_{-\Delta/2}^{\Delta/2} = \frac{1}{\Delta}\left[\frac{(\Delta/2)^3}{3} + \frac{(\Delta/2)^3}{3}\right] = \frac{1}{3\Delta}\left[\frac{\Delta^3}{8} + \frac{\Delta^3}{8}\right]$$

$$E[\varepsilon^2] = \frac{\Delta^2}{12}$$

$$\text{from } Eq(c) = V^2_{noise} = \frac{\Delta^2}{12}$$

if $R = 1\,\text{Ohm}$, the noise power is normalized i.e.,

$$\text{Noise power} = \frac{V^2_{noise}}{1} = \frac{\Delta^2/12}{1} = \frac{\Delta^2}{12} \quad (\text{from } Eq(b))$$

Hence, finally we write

Quantization Error (in terms of power) $= \frac{\Delta^2}{12}$

→ Signal to Quantization Noise ratio for Linear Quantization

In, PCM System, it is defined as

$$\frac{S}{N} = \frac{\text{Normalized s/g power}}{\text{Normalized noise power}}$$

Normalized noise power is calculated as $\frac{\Delta^2}{12}$

$$\frac{S}{N} = \frac{\text{Normalized Signal power}}{\Delta^2/12} \longrightarrow Eq(1)$$

The no. of bits 'v' and quantization levels are related as

$$q = 2^v$$

Total Amplitude range $= 2x_{max}$

Step size will be $\Delta = \frac{2x_{max}}{q}$

$$\Rightarrow \Delta = \frac{2x_{max}}{2^v}$$

from Eq(1) we get,

$$\frac{S}{N} = \frac{\text{Normalized s/g power}}{\left(\frac{2x_{max}}{2^v}\right)^2 \cdot \frac{1}{12}}$$

$$\frac{S}{N} = \frac{P}{\frac{4x^2_{max}}{2^{2v}} \times \frac{1}{12}} = \frac{3P}{x^2_{max}} 2^{2v}.$$

$\therefore P =$ Normalized s/g power.

Hence, Signal to quantization noise ration :

$$\frac{S}{N} = \frac{3P}{x^2_{max}} \cdot 2^{2v}.$$

This Expression shows that s/g to noise power ratio of quantizer increases Exponentially with increasing bits/sample.

If we assume $x(t)$ is normalized.

$$x_{max} = 1$$

∴ sig to quantization noise ration will be

$$\frac{S}{N} = 3 \times 2^{2v} \times P$$

If the destination sig. power 'P' is normalized i.e.,

$$P \leq 1$$

then SNR will be given as

$$\frac{S}{N} \leq 3 \times 2^{2v} \longrightarrow Eq(2)$$

Because $x_{max} = 1$ and $P \leq 1$, the SNR given by Eq(2) is said to be normalized,

SNR in decibels, we get.

$$\left(\frac{S}{N}\right)dB = 10\log_{10}\left(\frac{S}{N}\right)dB \leq 10\log_{10}\left[3 \times 2^{2v}\right) \leq (4.8 + 6v)dB$$

Thus, the Signal to quantization noise ration for normalized values of power 'p' and amplitude of input $x(t)$ will be

$$\left(\frac{S}{N}\right)dB \leq (4.8 + 6v)dB.$$

11) Explain the operation of delta modulation in detail and its drawbacks along with necessary Expressions. 13

A. Delta modulation transmits only one bit / sample. Here, the present sample value is compared with the previous value and this results whether the amplitude is increased or decreased if transmitted.

→ I/p s/g x(t) is approximated to step signal by the delta modulator. This step size is kept fixed.

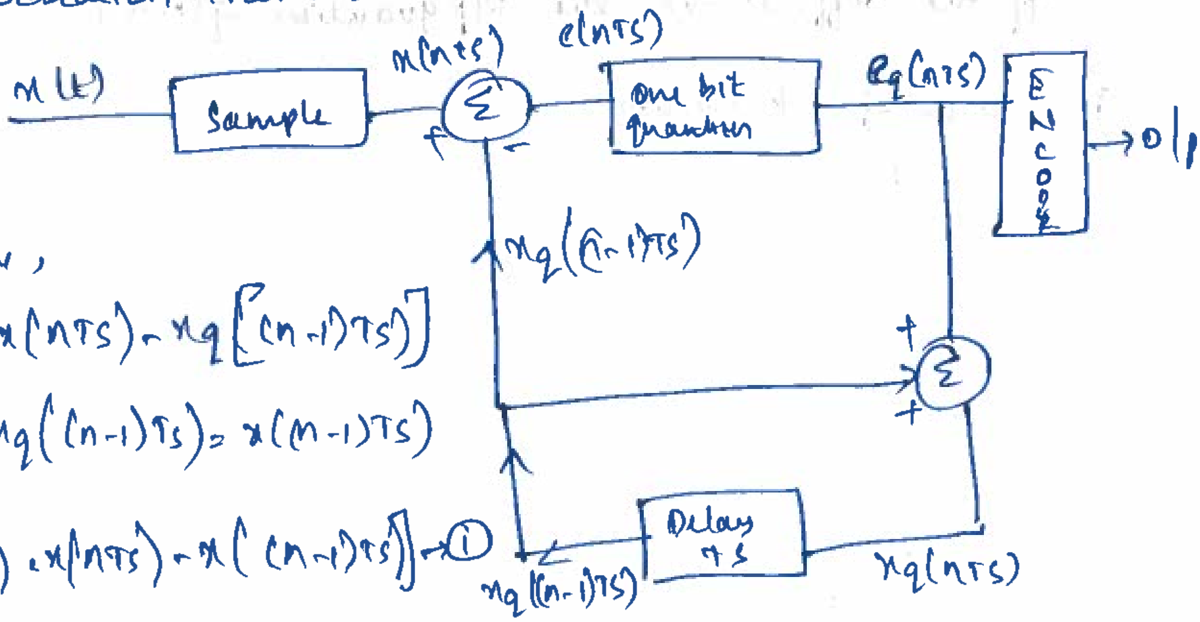
→ The difference b/w the I/p s/g x(t) and staircase approximated s/g is confined to two levels +Δ and -Δ

→ Now, if the difference is +ve, then approximated s/g is increased by step size 'Δ'., if the diff. is -ve, the step size is reduced by 'Δ'.

→ If step size is reduced '0' is transmitted & if step size is increased '1' is transmitted.

→ Hence, for each sample, only one bit is transmitted.

Delta modulation Transmitter



from figure,

$$e(nTs) = x(nTs) - x_q[(n-1)Ts]$$

where $x_q((n-1)Ts) = x((n-1)Ts)$

$$\therefore e(nTs) = x(nTs) - x[(n-1)Ts] \rightarrow (1)$$

$x(t) \rightarrow$ s/g I/p

$X(nTs) \rightarrow$ sample version of s/g I/p

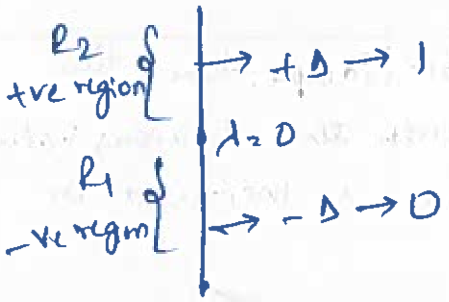
Quantizer $e(nTs) \longrightarrow \boxed{Q} \longrightarrow e_q(nTs)$

$\lambda = 0$

↑ threshold value

↳ o/p of quantization

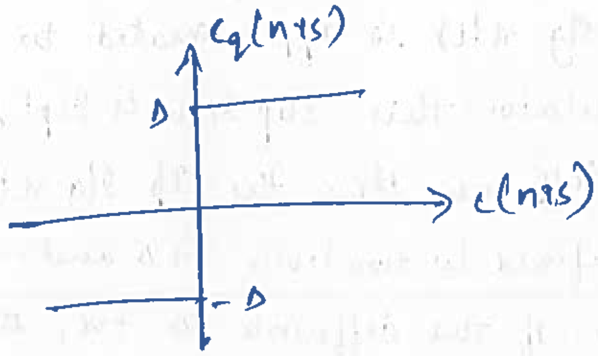
(i) $e(nT_s)|_{max}$



Δ is a value of quantized value

(ii) $e(nT_s)|_{min}$

$$e_q(nT_s) = \begin{cases} +\Delta, & e(nT_s) > 0 \\ -\Delta; & e(nT_s) < 0 \end{cases}$$



(iii) max quantization error : $Q_{e\,max} = \pm\Delta$

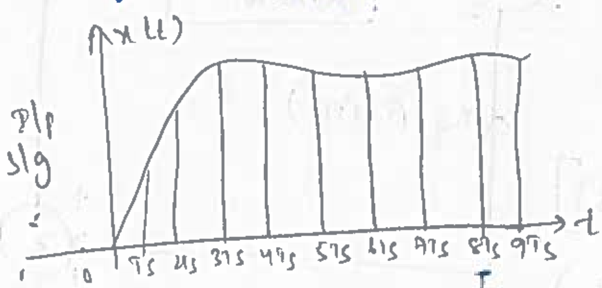
from eq (1) $e(nT_s) = x(nT_s) - x((n-1)T_s)$

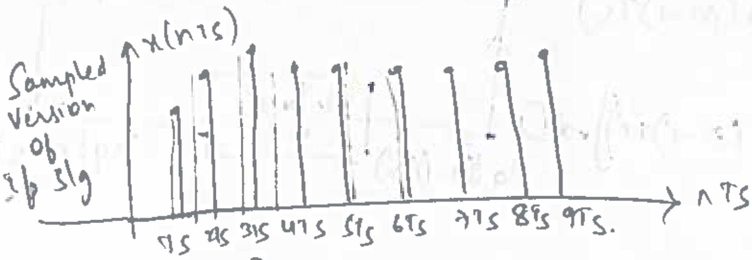
diff b/w present sample ~~value~~ & previous sample

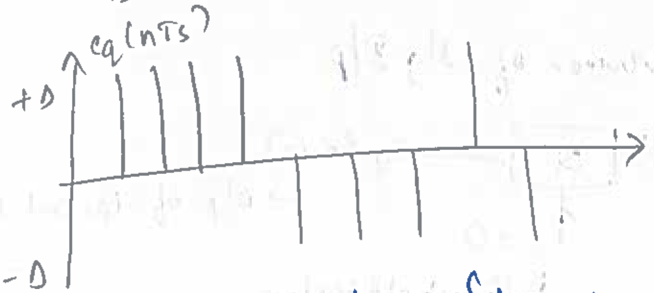
If the diff. is +ve the o/p quantizer $e_q(nT_s) = +\Delta$

If the diff. is -ve the o/p quantizer $e_q(nT_s) = -\Delta$

generation of bit stream.





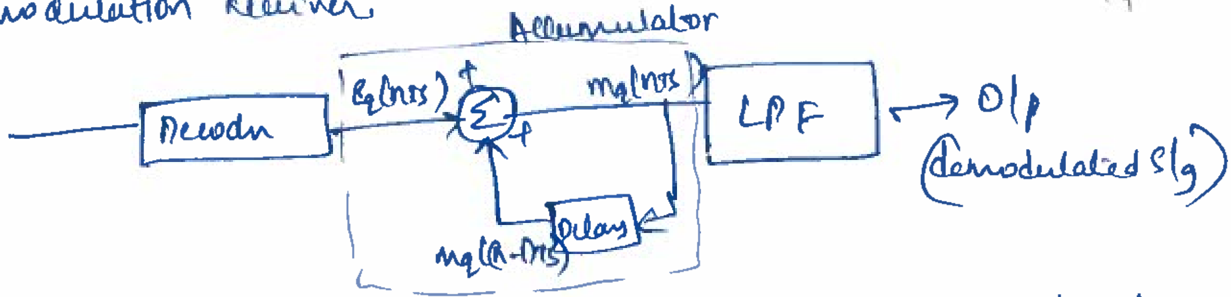


$+\Delta \rightarrow 1$ bit

$-\Delta \rightarrow 0$ bit.

$\therefore$ bit stream $b_k = \{1, 1, 1, 1, 0, 0, 0, 0, 1, 0\}$

Delta modulation Receiver



The Accumulator generates the staircase approximated s/g o/p and is delayed by one sampling period T_s.

from fig, $m_q(nT_s) = m_q[(n-1)T_s] + e_q[nT_s]$

$n=0$; $m_q(0) = m_q(-T_s) + e_q[0]$

here, $m_q[-T_s] = 0$

$m_q(0) = e_q(0)$

$n=1$; $m_q(T_s) = m_q(0) + e_q(T_s)$

$m_q(T_s) = e_q(0) + e_q[T_s]$ $[m_q[0] = e_q(0)]$

$n=2$; $m_q(2T_s) = m_q[T_s] + e_q[2T_s]$

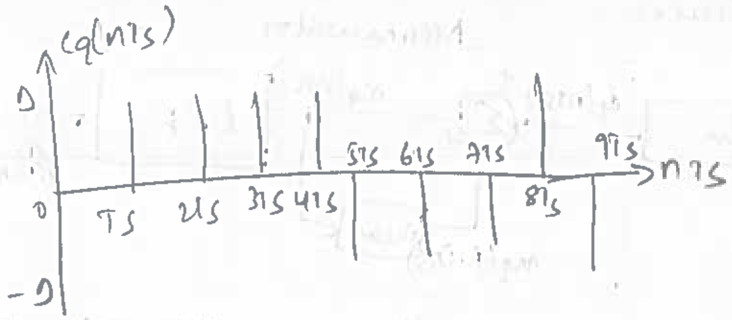
$= e_q[0] + e_q[T_s] + e_q[nT_s]$

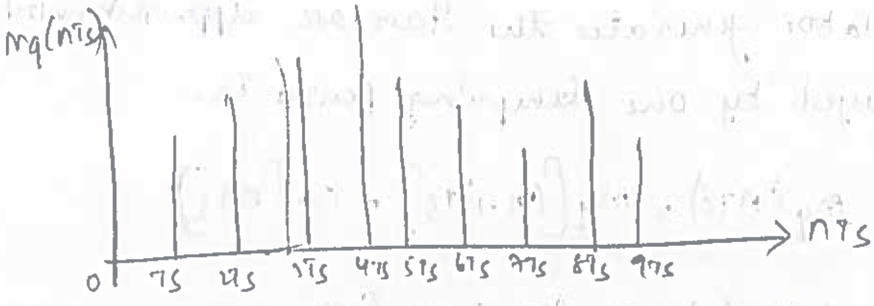
From this, we can write

$$\boxed{m_q[nT_s] = \sum_{k=0}^{n} e_q(kT_s)} \rightarrow \text{Accumulator System}$$

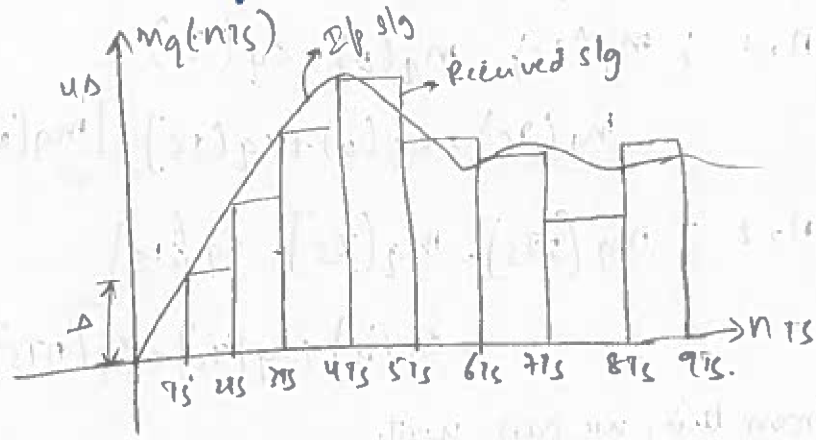
→ ~~It~~ is then added to the I/p s/g. If I/p is binary '1' then it adds Δ step to the previous o/p (which is delayed).

→ If I/p is binary '0' then one step 'Δ' is subtracted from the delayed signal.





If, the low pass filter has the cut off freq. Equal to the highest freq. in $x(t)$ this LPF smoothens the staircase sig to reconstruct original msg sig $x(t)$.



Received Signal : Staircase approximation of transmitted msg sig $x(t)$

Advantages:

→ Since, the DM transmits only one bit for one sample, the signalling rate and Transmission channel BW is quite small for DM when compared to PCM.

→ The Tx and Rx implementation is very much simple for DM. there is no analog to digital converter required in DM.

Drawbacks of DM

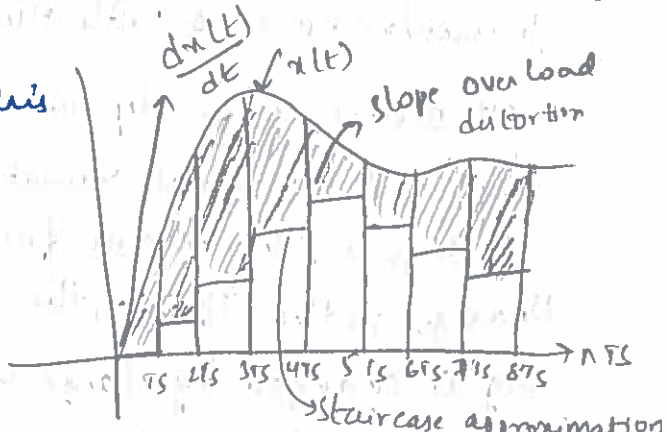
There are two major drawbacks.

(i) Slope Overload distortion

(ii) Granular (or) idle noise.

Slope overload distortion:

→ If step size is very very small, this distortion arises because of large dynamic range of the I/p signal



→ As shown in fig, the rate of rise of I/p sig $x(t)$ is so high that the staircase sig cannot approximate it, the Δ becomes too small for staircase $m_q(t)$ to follow the step segment $x(t)$. hence there is a large error b/w the staircase approximated sig and the original signal $x(t)$. this error/noise is called slope overload distortion.

$$\boxed{\left.\frac{dx(t)}{dt}\right|_{max} >> \frac{\Delta}{T_S}}$$ → Condition for occurrence of slop overload Noise.

Condition to not occur slope overload distortion:

$$\left.\frac{dx(t)}{dt}\right|_{max} \leq \frac{\Delta}{T_S}$$

let $x(t) = A_m \cos w_m t$

$$\frac{dx(t)}{dt} = A_m \cdot w_m \cos w_m t$$

$$\left.\frac{dx(t)}{dt}\right|_{max} = A_m w_m \quad (\cos\theta = 1 \text{ for } \theta = max)$$

$$\therefore A_m w_m \leq \frac{\Delta_{min}}{T_S}$$

$$A_m w_m T_s \leq \Delta_{min}$$

$$\Delta_{min} \geq \frac{A_m w_m}{f_s}$$

$$\boxed{\Delta_{min} \geq \frac{A_m(2\pi f_m)}{f_s}}$$ → Condition for not occurrence of SOD.

granular noise (or) idle noise.

→ It occurs when step size is too large compared to small variations in the I/p signal. This means for very small change in the I/p sig, the staircase sig is changed by large amount (Δ) because of large step size.

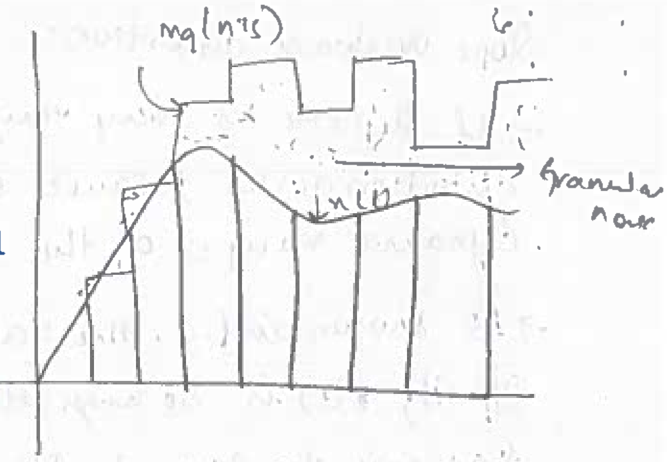


fig. shows that when the I/p sig is almost flat, the staircase mq(nTs) keeps on oscillating by $\pm\Delta$ around the sig.

→ The error b/w the I/p and approximated sig is called granular noise. The solution is to make step size small.

Maximum O/p signal to noise ration for DM:

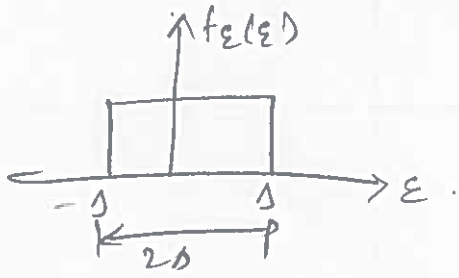
→ First we require to find Quantization Noise power

→ the Quantization Error in DM is Equally likely to lie ~~also~~ in the interval $(-\Delta, \Delta)$

→ max. Quantization Error $\varepsilon_{max} = \pm\Delta$

this Error can be uniformly distributed as shown in fig

$$f_\varepsilon(\varepsilon) = \begin{cases} \frac{1}{2\Delta} ; & -\Delta < \varepsilon < \Delta \\ 0 ; & \text{OW} \end{cases}$$



The mean square value or the variance of quantization noise is given by

$$\overline{\varepsilon^2} = \int_{-\Delta}^{\Delta} \varepsilon^2 f_\varepsilon(\varepsilon) d\varepsilon$$

$$\int_{-\Delta}^{\Delta} \varepsilon^2 \frac{1}{2\Delta} d\varepsilon$$

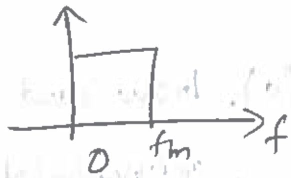
$$\frac{1}{2\Delta}\left[\frac{\varepsilon^3}{3}\right]_{-\Delta}^{\Delta} = \frac{1}{2\Delta}\left[\frac{\Delta^3}{3} + \frac{\Delta^3}{3}\right]$$

now $\overline{\varepsilon}^2 = \frac{2\Delta^3}{3}\cdot\frac{1}{2\Delta}$

$$\frac{\overline{\varepsilon}^2}{1} = \frac{\Delta^2}{3} = N_q$$

$$N_q = \frac{\Delta^2}{3}.$$

The DM sig passed through a reconstruction LPF at the o/p of DM Rx. the B.W of this LPF is f_M such that



$$\boxed{f_M \ge f_m \text{ and } f_M << f_s}$$

Now, assuming, N_q is distributed uniformly over frequency band upto f_s, the o/p quantization Noise power within the B.W f_M is given by

Normalized noise power at o/p, $N_q' = \frac{\Delta^2}{3}\times\frac{f_M}{f_s}$

Signal power (S):

let us consider $x(t) = A\cos\omega_m t$

$$\Delta_{min} \ge \frac{A_m \omega_m}{f_s}$$

$$\frac{\Delta f_s}{\omega_m} \ge A_m$$

$$A_m \le \frac{\Delta f_s}{\omega_m}$$

$$A_m \le \frac{\Delta}{2\pi}\left(\frac{f_s}{f_m}\right)$$

for sig power (S) $= \frac{A_m^2}{2}$

$$\frac{1}{2}\left(\frac{\Delta^2 f_s^2}{4\pi^2 f_m^2}\right)$$

$\therefore$ Signal to Noise power (S/N) $= \frac{S}{N_i}$

$$= \frac{\Delta^2 f_s^2}{8\pi^2 f_m^2} \times \frac{3 \times f_s}{\Delta^2 \times f_m}$$

$\left[\because f_s = \frac{1}{T_s}\right]$

SNR for DM $\boxed{\left[\frac{S}{N_0}\right] = \frac{3}{8\pi^2 f_M f_m^2 T_s^3}}$

12(a). Draw and Explain power spectral density (PSD) and geometrical representation of BPSK.

A. In BPSK, phase of the sinusoidal carrier is changed according to the data bit to be transmitted. Also, a bipolar NRZ signal is used to represent the digital data coming from the digital source

Spectrum of BPSK signals.

We know that the waveform $b(t)$ is a NRZ binary waveform. In this, there are rectangular pulses of Amplitude $\pm V_b$. If we assume that Each pulse is $\pm\frac{T_b}{2}$ around its centre then it becomes easy to find F.T of Each pulse. The F.T of this type of pulse is given as

$$X(f) = V_b T_b \frac{\sin(\pi f T_b)}{(\pi f T_b)} \rightarrow (1)$$

For Large no. of Such positive and negative pulses, the power spectral density $S(f)$ is Expressed as

$$S(f) = \frac{|X(f)|^2}{T_s} \rightarrow (2)$$

Substituting eq (1) in Eq (2), we get

$$S(f) = \frac{V_b^2 T_b^2}{T_s}\left[\frac{\sin(\pi f T_b)}{\pi f T_b}\right]^2.$$

For BPSK, only one bit is transmitted at a time, symbol and bit durations are same $T_b = T_S$, then,

$$S(f) = V_b^2 T_b \left[\frac{\sin(\pi f T_b)}{\pi f T_b}\right)^2 \rightarrow (3)$$

The BPSK slg is generated by modulating a carrier with the baseband signal $b(t)$. Due to the modulation of the carrier of freq. f_c, the spectral components are translated from f to f_c+f and f_c-f, then Eq (3) can be written as

$$S_{BPSK}(f) = V_b^2 T_b \left\{ \frac{1}{2}\left[\frac{\sin\pi(f_c-f)T_b}{\pi(f_c-f)}\right] + \frac{1}{2}\left[\frac{\sin\pi(f_c+f)T_b}{\pi(f_c+f)T_b}\right]^2\right\}$$

Let us assume the value of $\pm V_b = \sqrt{P}$. This means NRZ has amplitude of $+\sqrt{P}$ & $-\sqrt{P}$. Then the above Eqn is written as

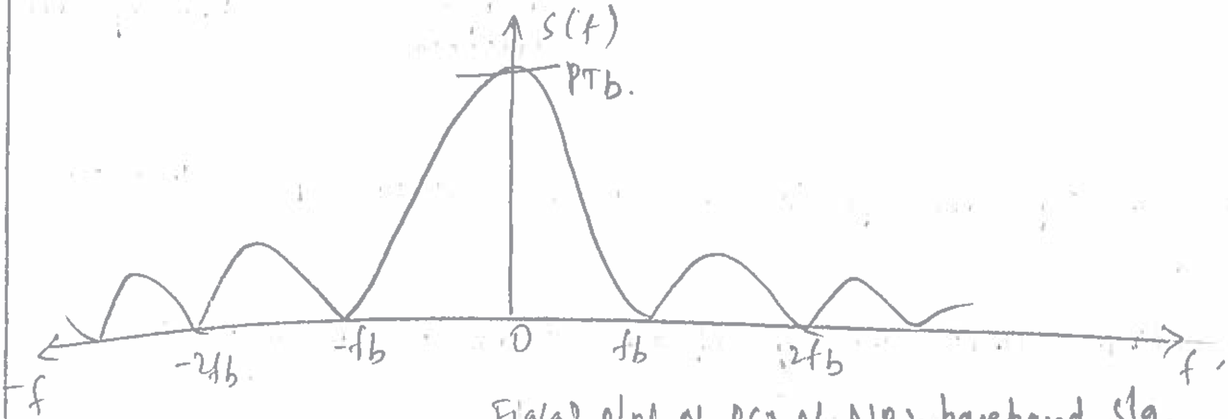
$$S_{BPSK}(f) = \frac{PT_b}{2}\left\{\left[\frac{\sin\pi(f-f_c)T_b}{\pi(f-f_c)T_b}\right]^2 + \frac{1}{2}\left[\frac{\sin\pi(f_c+f)T_b}{\pi(f_c+f)T_b}\right]^2\right\} \rightarrow (4)$$

This Eqn gives power spectral density of BPSK slg for modulating slg $b(t)$ having amplitude equal to $\pm\sqrt{P}$

$$S(t) = \pm\sqrt{2P}\cos(2\pi f_c t)$$

If $b(t) = \pm\sqrt{P}$, then the carrier becomes,

$$\phi(t) = \sqrt{2}\cos(2\pi f_c t)$$

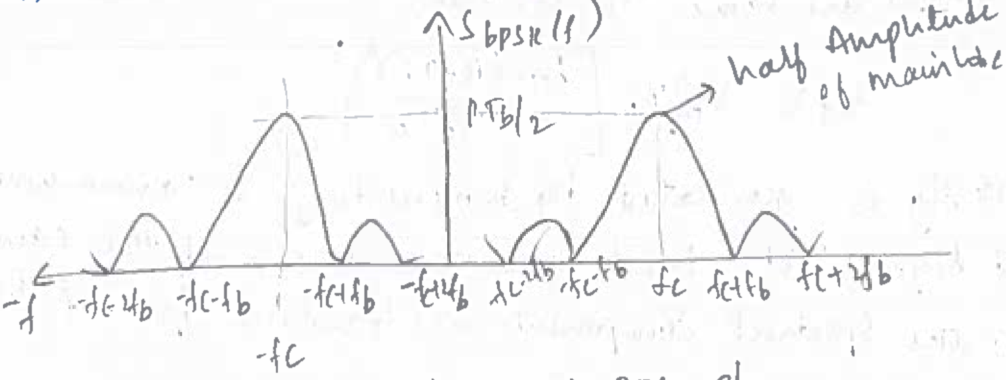


Fig(a) plot of PSD of NRZ baseband slg.

In fig(a), the main lobe ranges from $-f_b$ to $+f_b$. because we have taken $\pm V_b = \pm\sqrt{P}$ in Eq (3), therefore the peak value of the main lobe is PT_b. Now let us consider the PSD of BPSK slg

Expressed by Eq (4)

Fig(b) shows the plot of Eq (4)



fig(b): Plot of psd of BPSK sig.

12(b) Write any two comparisions among ASK and FSK.

A.

S.NO	Parameter of comparision	Binary ASK	FSK
1.	Variable characteristic	Amplitude	frequency
2.	Bandwidth (Hz) (Spectral Efficiency)	$2f_b$	$4f_b$
3.	Noise immunity	low	high
4.	Probability of Error	high	low.
5.	Performance in presence of noise	Poor	better than ASK
6.	System complexity	Simple	Moderately Complex
7.	Bit rate / data rate	Suitable upto 100 bits/sec	Suitable upto about 1200 bits/sec
8.	Demodulation method	Envelope detection	Envelope detection

13(a) Derive an Expression for signal to noise ratio for integrator and dump filter

A. We know that the o/p of the integrator may be written as

$$y(t) = \frac{1}{RC}\int_0^T [x(t)+n(t)]dt$$

here, the integration is performed over one bit period i.e., from 0 to T

$$y(t) = \frac{1}{RC}\int_0^T x(t)dt + \frac{1}{RC}\int_0^T n(t)dt$$

$y(t) = x_o(t) + n_o(t)$

$\therefore$ $x_o(t) \rightarrow$ o/p signal voltage

$n_o(t) \rightarrow$ o/p noise voltage

Let us consider $x_o(t) = \frac{1}{RC}\int_0^T x(t)dt$

because of the value of $x(t) = A$ from 0 to T)

$$x_o(t) = \frac{1}{RC}\int_0^T A\,dt$$

$$\frac{A}{RC}\int_0^T 1\,dt$$

$$x_o(t) = \frac{AT}{RC}$$

Let the time constant $RC = \tau$

$$\text{hence } x_o(t) = \frac{AT}{\tau}$$

the normalised sig power in standard resistance of 1Ω would be,

$$\text{O/p sig power} = \frac{x_o^2(t)}{1\Omega}$$

$$\therefore \text{O/p sig power} = \left(AT/\tau\right)^2$$

$$\boxed{\text{O/p sig power} = \frac{A^2T^2}{\tau^2}} \rightarrow (1)$$

Let us find the Noise power, we know the transfer function of Integrator is $\frac{1}{j\omega RC}$.

Now since a delay of $t = T$ in time domain is Equivalent to $e^{-j\omega T}$ in freq. domain, the n/w performing Integration over the Period of T can be represented by the T.F

$$H(f) = \frac{1 - e^{-j\omega T}}{j\omega RC}$$

we know $\omega = 2\pi f$, $RC = \tau$

$$H(f) = \frac{1 - e^{-j2\pi fT}}{j2\pi fT}$$

$[e^{j\theta} = \cos\theta + j\sin\theta]$

$$H(f) = \frac{1-[\cos(2\pi fT) - J\sin(2\pi fT)]}{J2\pi fT}$$

$$H(f) = \frac{1-\cos(2\pi fT)}{J2\pi fT} + \frac{J\sin(2\pi fT)}{J2\pi fT}$$

$$H(f) = \frac{\sin(2\pi fT)}{2\pi fT} - J\frac{1-\cos(2\pi fT)}{2\pi fT}$$

the magnitude of above Eqn is

$$H(f) = \sqrt{\frac{\sin^2(2\pi fT)}{(2\pi fT)^2} + \frac{(1-\cos 2\pi fT)^2}{(2\pi fT)^2}}$$

$$|H(f)| = \sqrt{\frac{\sin^2(2\pi fT) + 1 + \cos^2(2\pi fT) - 2\cos(2\pi fT)}{(2\pi fT)^2}}$$

S.O.B.S

$$|H(f)|^2 = \frac{\sin^2(2\pi fT) + \cos^2(2\pi fT) + 1 - \cos(2\pi fT)}{(2\pi fT)^2}$$

$$= \frac{2[2\sin^2(\pi fT)]}{4(\pi fT)^2}$$

finally $\boxed{|H(f)|^2 = \frac{\sin^2(\pi fT)}{(\pi fT)^2}}$

Now, the avg. power of the o/p Noise Sig $n_0(t)$ may be obtained by Integrating its P.D.S

$$\text{Power, } P = \int_{-\infty}^{\infty} S_{n_0}(f)\,df$$

Noise power $P = \frac{\overline{n_0^2(t)}}{R}$ for standard resistance $R = 1\Omega$

Noise power $= P = \overline{n_0^2(t)}$

$$\therefore \overline{n_0^2(t)} = \int_{-\infty}^{\infty} S_{n_0}(f)\,df \longrightarrow (1)$$

The I/p & O/p power spectral densities are related as

$$S_{n_0}(f) = |H(f)|^2 S_{ni}(f) \longrightarrow (3)$$

Substitute Eq (3) in Eq (1)

$$\overline{n_0^2(t)} = \int_{-\infty}^{\infty} |H(f)|^2 S_{ni}(f)\, df$$

$$= \int_{-\infty}^{\infty} \frac{\sin^2(\pi f T)}{(\pi f T)^2} \frac{N_0}{2}\, df$$

let $\pi f T = x$

$\pi f\, df = dx$ and $f = \frac{x}{\pi T}$

$df = \frac{dx}{\pi \tau}$ $\qquad \pi f T = \frac{\pi T x}{\pi \tau} = \frac{x T}{\tau}$

$$\overline{n_0^2(t)} = \int_{-\infty}^{\infty} \frac{\sin^2\left(\frac{xT}{\tau}\right)}{x^2} \frac{N_0}{2} \cdot \frac{dx}{\pi \tau}$$

$$\overline{n_0^2(t)} = \frac{N_0 T^2}{2\pi \tau^3} \int_{-\infty}^{\infty} \frac{\sin^2\left(\frac{xT}{\tau}\right)}{\left(\frac{xT}{\tau}\right)^2}\, dx.$$

let $\frac{xT}{\tau} = u$

$dx \frac{T}{\tau} = du \Rightarrow dx = du \frac{\tau}{T}$.

$$= \frac{N_0 T^2}{2\pi \tau^3} \int_{-\infty}^{\infty} \frac{\sin^2 u}{u^2}\, du \cdot \frac{\tau}{T}$$

$$= \frac{N_0 T}{2\pi \tau^2} \int_{-\infty}^{\infty} \left(\frac{\sin u}{u}\right)^2 du.$$

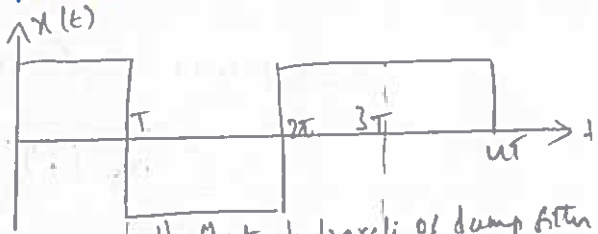
If the function $\frac{\sin u}{u}$ is squared then the above Equation is written as

$$\overline{n_0^2(t)} = \frac{N_0 T}{2\pi \tau^2}\, 2\int_{0}^{\infty} \left(\frac{\sin u}{u}\right)^2 du$$

$$= \frac{N_0 T}{2\pi \tau^2} \cdot 2\left(\frac{\pi}{2}\right)$$

$$\overline{n_o^2(t)} = \frac{N_0 T}{2\tau^2} \rightarrow (4)$$

This relation describes the Noise power at the O/p.

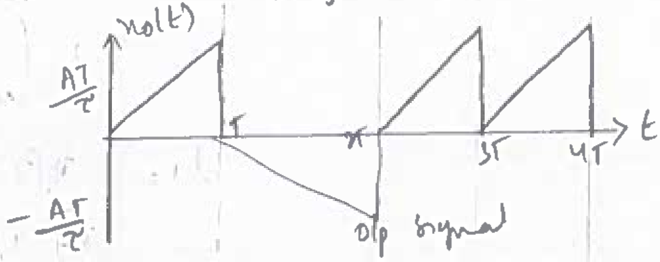


I/p sig to Integrate of dump filter

The SNR $\left(\frac{S}{N}\right)_0 = \frac{\text{Signal Power}}{\text{Noise Power}}$

Substitute Eq(1) in Eq(4)

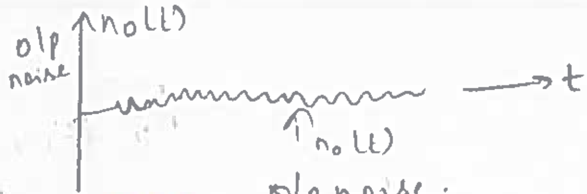
$$\left(\frac{S}{N}\right)_0 = \frac{\frac{A^2T^2}{\tau^2}}{\frac{N_0 T}{2\tau^2}} = \frac{2A^2T}{N_0}.$$



O/p signal

Hence, we conclude SNR of Integrate and dump receiver:

$$\left(\frac{S}{N}\right)_0 = \frac{2A^2T}{N_0} \rightarrow (5).$$



O/p noise.

Signal to Noise ratio is known as 'Figure of merit'.

13(b) Derive an Expression for Error probability of ASK.

A: We know that the ASK signal is represented as

Binary 1 : $x_1(t) = A\cos w_c t$

Binary 0 : $x_2(t) = 0$

We know that Average normalized Sig power

$$P_s = \frac{A^2}{2}$$

i.e., $A = \sqrt{2P_s}$

$\therefore x_1(t) = \sqrt{2P_s}\cos w_c t \; ; 0 < t < T$

$x_2(t) = 0.$

Let us obtain the expression for probability of Error using the matched filter.

The Expression of Error probability with optimum filter is

$$P_e = \frac{1}{2}\,\text{erfc}\left[\frac{x_{01}(T) - x_{02}(T)}{2\sqrt{2}\,\sigma}\right] \rightarrow (1).$$

The Expression for max. o/p SNR of an Optimum filter is

$$\left[\frac{x_{01}(T)-x_{02}(T)}{\sigma}\right]^2_{max} = \int_{-\infty}^{\infty} \frac{|X(f)|^2}{S_{ni}(f)}\, df \rightarrow (1)$$

The value of PSD for a white Noise i/p is $S_{ni}(f) = \frac{N_0}{2} \rightarrow (2)$.

Substitute Eq (2) in Eq (1).

$$\left[\frac{x_{01}(T)-x_{02}(T)}{\sigma}\right]_{max} = \frac{2}{N_0}\int_{-\infty}^{\infty} |X(f)|^2 df \rightarrow (3)$$

According to rayleigh's Energy theorem.

$$\int_{-\infty}^{\infty} |X(f)|^2 df = E = \int_{-\infty}^{\infty} x^2(t)\, dt \rightarrow (4)$$

Substitute (4) in (3), we get

$$\left[\frac{x_{01}(T)-x_{02}(T)}{\sigma}\right]^2_{max} = \frac{2}{N_0}\int_{-\infty}^{\infty} x^2(t) dt$$

$$\left[\frac{x_{01}(T)-x_{02}(T)}{\sigma}\right]^2 = \frac{2}{N_0}\int_0^T \left(\sqrt{2P_s}\cos\omega_c t\right)^2 dt$$

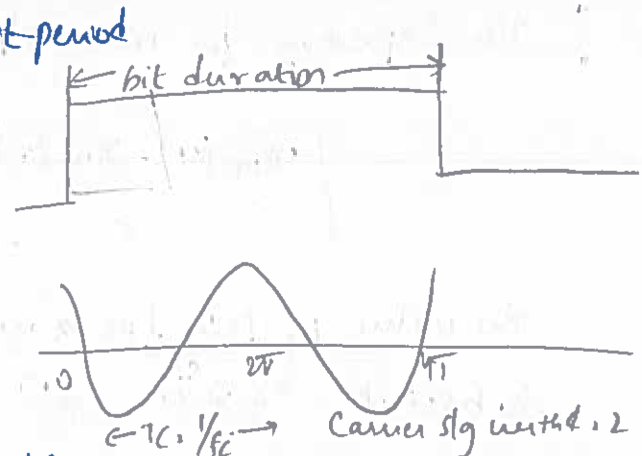
$$= \frac{4P_s}{N_0}\int_0^T \cos^2\omega_c t\, dt$$

$$= \frac{4P_s}{2N_0}\int_0^T 1.dt + \frac{4P_s}{2N_0}\int_0^T \cos 2\omega_c t\, dt$$

$$= \frac{4P_sT}{2N_0} + \frac{4P_s}{2N_0}\cdot\frac{\sin 2\omega_c T}{2\omega_c}.$$

we assume that the freq of the carrier signal (f_c) is selected such that there are I no. of complete cycles of the carrier during the bit duration (T), as shown below.

from fig(a), it shows that one bit period



fig(a). Relation b/w bit Period (T) of carrier frequency.

$$T = 2T_C = 2/f_C$$

as $T_C = 1/f_C$

$$f_C \cdot T = 2$$

i.e., $f_C \cdot T = k$.

$\therefore \sin 2\omega_c T = \sin 4\pi k$, where $k = 1, 2, \ldots$

$= \sin 4\pi, \sin 8\pi, \ldots$

$\sin 2\omega_c T = 0$ for all values of k

$$\therefore \left[\frac{x_{01}(t) - x_{02}(t)}{\sigma}\right]^2 = \frac{4P_sT}{2N_0} + 0$$

$$= \frac{2P_sT}{N_0}$$

Apply square root on both sides

$$\left[\frac{x_{01}(T) - x_{02}(T)}{\sigma}\right]_{max} = \sqrt{\frac{2P_sT}{N_0}} \quad \text{--- (5)}$$

Substitute Eq (5) in Eq (1)

we get error probability of the ASK as under

$$\therefore P_e = \frac{1}{2} \operatorname{erfc}\left[\frac{\sqrt{\frac{2P_sT}{N_0}}}{2\sqrt{2}}\right]$$

$$P_e = \frac{1}{2} \operatorname{erfc}\left[\sqrt{\frac{P_sT}{4N_0}}\right]$$

This is the Expression for the bit error probability by P_B.

$$\therefore P_B = \frac{1}{2} \operatorname{erfc}\left[\sqrt{\frac{P_sT}{4N_0}}\right]$$

But we know that $P_sT = E$: Energy of the signal [Energy/bit]

$$\therefore P_B = \frac{1}{2} \operatorname{erfc}\left[\sqrt{\frac{E}{4N_0}}\right]$$

This is the Expression for bit error probability of ASK using the matched filter

→ The complementary error function $erfc(x)$ is a monotonically decreasing function of x. This means that as x increases, the value of $erfc(x)$ decreases. hence, the bit error probability P_B decreases with increase in the ratio $\frac{E}{4N_0}$. Thus the error probability depends only on the signal energy and not its shape (or) any other parameter.

P_B expressed in terms of a functions as under

we know $Q(x) = \frac{1}{2} erfc\left(x/\sqrt{2}\right)$

$$2Q(\sqrt{2}\,x) = erfc(x)$$

$$\therefore erfc\left(\sqrt{\frac{E}{4N_0}}\right) = 2Q\left(\sqrt{2}\cdot\sqrt{\frac{E}{4N_0}}\right)$$

$$= 2Q\left(\sqrt{\frac{2E}{4N_0}}\right)$$

$$= 2Q\left(\sqrt{\frac{E}{2N_0}}\right)$$

bit error probability : $P_B = \frac{1}{2}\left[2Q\sqrt{\frac{E}{2N_0}}\right]$

$$P_B = Q\left(\sqrt{\frac{E}{2N_0}}\right)$$

from constellation diagram we know that $d = \sqrt{E}$

$d^2 = E$

error probability in terms of distance (d)

$$\therefore P_B = \frac{1}{2} erfc\left[\sqrt{\frac{d^2}{4N_0}}\right]$$

(or)

$$P_B = Q\left[\sqrt{\frac{d^2}{2N_0}}\right]$$

14(a) Define Information and Explain its properties.

A. → The Amount of Information received from the knowledge of occurrence of an Event may be related to the likelihood (or) Probability of Occurrence of that Event

→ The amount of Information contained in an Event is closely related to its Uncertainity.

→ Messages containing knowledge of low probability of occurrence convey relatively more information

i.e., Information is Inversely proportional to probability

$$I \propto \frac{1}{p}$$

$p=1,\ 1-p=0\ ;\ I=0.$

$p=0,\ 1-p=1\ ;\ I=\infty$

$p \to$ Probability, $I \to$ Information.

Let us consider a discrete memory less source (DMS) denoted by X and having alphabet $(x_1, x_2 \dots x_m)$ the Information Content by a Symbol x_i, denoted by $I(x_i)$ is defined as.

$$I(x_i) = \log_b \frac{1}{P(x_i)}$$

$$I(x_i) = -\log_b P(x_i)$$

$\therefore P(x_i) \to$ Probability of occurrence of symbol x_i

Let us consider messages m_1, m_2 with their respective Probabilities P_1, P_2

i.e., $m_1 \to P_1$
$m_2 \to P_2$

Information of first msg $\therefore I_1 \propto 1/P_1$

$$I_2 \propto \frac{1}{P_2}$$

$$I_{12} \propto \frac{1}{P_1} \cdot \frac{1}{P_2}$$

$$I_T = \log\left(\frac{1}{P_1} \cdot \frac{1}{P_2}\right)$$

$$= \log\frac{1}{P_1} + \log\frac{1}{P_2}$$

$I = \log_b \frac{1}{P}$.

Units of Information : I

$$I = \log_b \frac{1}{P}$$

If $b=2$; Units of $I =$ bit

If $b=10$; units of $I =$ Hartley

If $b=e$; Units of $I =$ nat (natural unit)

It is standard to use $b=2$

$$I = \log_2 1/P$$

Properties of Information $I(x_i)$:

(i) $I(x_i) = 0$ for $P(x_i) = 1$

we know, $I(x_i) = +\log_2 \frac{1}{P(x_i)}$

given $P(x_i) = 1$

$$I(x_i) = \log_2 \frac{1}{1} = \log_2 1$$

$$= \frac{\log 1}{\log 2}$$

$$I(x_i) = 0.$$

$\therefore$ If $P(x_i) = 1$ then $I(x_i) = 0$.

(ii) $I(x_i) \geq 0$.

The Information content $I(x_i)$ must be a non-negative quantity since each message contains some information in the worst case $I(x_i)$ can be equal to zero.

(iii) $I(x_i) > I(x_3)$ if $P(x_i) < P(x_3)$

we know that Information contain is Inversely proportional to probability $I(x_i) \propto \frac{1}{P(x_i)}$

Ex: In a binary PCM if '0' occur with probability 1/4 and '1' occur with probability Equal to 3/4. then amount of Information carried by each bits

for bit '0' has $P(x_i) = \frac{1}{4}$

for information of bit has $P(x_j) = 3/4$

$$I(x_i) = \log_2 \frac{1}{P(x_i)}$$

with $P(x_i) = 1/4$

$$I(x_i) = \log_2 \frac{1}{(1/4)}$$

$$= \log_2 4$$

$$\log_2 2^2 \Rightarrow 2\log_2 2$$

$$I(x_i) = 2 \text{bits}$$

with $P(x_3) = 3/4$

$$I(x_j) = \log_2 \frac{1}{P(x_3)}$$

$$\log_2 \frac{1}{3/4}$$

$$\log_2 4/3$$

$$= \frac{\log_{10}(4/3)}{\log_{10}(2)}$$

$$I(x_j) = 0.415 \text{ bits}$$

(iv) $I(x_i, x_j) = I(x_i) + I(x_j)$ if x_i and x_j are independant

we know that $I(x_i) = \log_2 \frac{1}{P(x_i)}$ &

$$I(x_j) = \log_2 \frac{1}{P(x_j)}$$

$$I(x_1, x_2) = \log_2 \frac{1}{P(x_1) \cdot P(x_2)}$$

$$\log_2 \left[\frac{1}{P(x_1)} \cdot \frac{1}{P(x_2)}\right]$$

$$\log_2 \left[\frac{1}{P(x_1)}\right] + \log_2 \left[\frac{1}{P(x_2)}\right]$$

$I(x_1, x_2) = I(x_1) + I(x_2)$ Hence proved.

14(b) Define channel capacity and Explain Channel Capability for discrete channels. 23

A. → The Maximum amount of Information that can be conveyed by channel is called 'channel capacity'.

The channel capacity/symbol of a discrete memoryless channel (DMC) is defined as

$$C_c = \max[I(x,y)]$$

(or)

$$= \max[H(x) - H(x/y)]$$

$$= \max[H(y) - H(y/x)]$$

bits/symbol (or) msg.

→ The channel capacity (C_c) is a function of only the channel transition probabilities which define the channel

→ Generally $H(x/y) \geq 0$, so capacity of any channel is always less than (or) Equal ($\leq$) to max. Entropy of source.

Channel capacity rate [C_r]:

If 'r' symbols are being transmitted/second, then the maximum rate of transmission of information/sec is rC_c

$$\boxed{\therefore C_r = rC_c} \text{ bps.}$$

→ channel capacity for some discrete channel:

(1) Loseless channel: By knowing the receiver, we can find the transmitter, that channel is called as loseless channel

i.e., $H(x/y) = 0$

$$\therefore C = \max[I(x,y)]$$

$$= \max[H(x) - H(x/y)]$$

$$C = \max[H(x)]$$

we know maximum Entropy $H(x) = \log_2 m$,

$\therefore C = \log_2 m$ where $m \rightarrow$ no of symbols in 'x'

(ii) Deterministic channel: By knowing the transmitter, we can observe the receiver, that channel is called the Deterministic channel

$$i.e., \; H(Y/X) = 0$$

$$C = \max[I(x,y)]$$

$$= \max[H(Y) - H(Y/X)]$$

$$= \max[H(Y)]$$

$$C = \log_2 n$$ where $n \rightarrow$ no. of symbols in 'y'.

(iii) Noiseless channel: Since a noiseless channel is both lossless and deterministic we have $H(Y/X) = H(X/Y)$

$$\therefore C = \log_2(m) = \log_2 n.$$

(iv) Binary symmetric channel:

for bsc for the mutual information

$$I(x,y) = H(Y) + p\log_2 p + (1-p)\log_2(1-p)$$

and channel capacity for BSC

$$C = 1 + p\log_2 p + (1-p)\log_2(1-p).$$

15(a) Explain about the trade-off between bandwidth and S/N ratio.

A. For a continuous source Entropy is maximum, if the probability density function associated with o/p of the source is gaussian

$$i.e., \; f(x) = \frac{1}{\sqrt{2\pi\sigma^2}} e^{-x^2/2\sigma^2} \quad (\text{mean} = 0)$$

then $\frac{1}{f(x)} = \sqrt{2\pi\sigma^2} \cdot e^{x^2/2\sigma^2}$

Apply log on both sides

$$\log\left(\frac{1}{f(x)}\right) = \log\left(\sqrt{2\pi\sigma^2}\, e^{x^2/2\sigma^2}\right)$$

we know $H(x) = -\int_{-\infty}^{\infty} f(x) \log f(x)\, dx$

$$\therefore H(x) = +\int_{-\infty}^{\infty} f(x) \left[\log \sqrt{2\pi\sigma^2}\, e^{x^2/2\sigma^2}\right] dx$$

$$\therefore \frac{\log}{2} 2\pi\sigma^2 + \frac{\log e}{2\sigma^2} \cdot \sigma^2$$

$$= \frac{1}{2} \log 2\pi\sigma^2 + \log_2 e \cdot$$

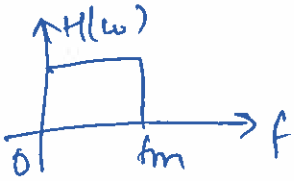
$$H(x) = \frac{1}{2} \log\left(2\pi\sigma^2 e\right) \text{ b/msg}$$

for Entropy of a continous source also given by

$$H(x) = 2f_m \left[\frac{1}{2} \log\left(2\pi\sigma^2 e\right)\right]$$

$$H(x)\big|_{max} = f_m \log\left(2\pi e\sigma^2\right)$$

for LPF



$\Rightarrow BW(B) = f_m - 0$

$B = f_m$.

$$\boxed{\therefore H(x)\big|_{max} = B \log 2\pi e\sigma^2} \text{ b/msg.}$$

Let communication channel to be band limited to B Hz.
the o/p of the channel y(t) will also be band limited to B Hz

In a additive white Gaussian Noise (AWGN) channel, the channel o/p y is given by

$$y = x + n.$$

Channel capacity $C = \max[I(x,y)]$

$$= \max[H(y) - H(y/x)]$$

$$H(y/x) = -\sum_i \sum_j P(x_i, y_j) \log p(y_j/x_i)$$

$$= \int_{-\infty}^{\infty} \int_{-\infty}^{\infty} f(x,y) \log f(y/x) \, dx \, dy$$

$$= -\int_{-\infty}^{\infty} f(x) dx \int_{-\infty}^{\infty} f(y/x) \log f(y/x) \, dy$$

$$H(y/x) = -\int_{-\infty}^{\infty} f(y/x) \log f(y/x) \, dx$$

$$H(y/x) = -\int_{-\infty}^{\infty} f(x) \log f(x) \, dx.$$

$$H(y/x) = H(n).$$

Channel capacity $C = \max[H(y) - H(y/x)]$

$$= \max[H(y) - H(n)]$$

$$H(n) = \underset{\substack{\downarrow \\ \text{Channel BW}}}{B} \log 2\pi e \underset{\substack{\downarrow \\ \text{Noise variance.}}}{\sigma_N^2} \text{ bps}$$

Entropy R_x will have its max. Entropy if the density function associated with its I/p is gaussian and the max. Entropy is given by $B \log 2\pi e \sigma_y^2$.

$$C = B \log 2\pi e \sigma_y^2 - B \log 2\pi e \sigma_N^2.$$

From the knowledge on variance

$$\text{Var}(x,y) = \text{Var}(x) + \text{Var}(y) + 2\,\text{conv}(x,y)$$

$$y = x + n.$$

$$Var(y) = Var(x) + Var(n) + \underbrace{2Cov(x,n)}_{\text{independent} = 0}.$$

$$\therefore Var(y) = Var(x) + Var(n)$$

$$\sigma_y^2 = \sigma_x^2 + \sigma_n^2.$$

$$\sigma_y^2 = E(y^2) - \underbrace{[E(y)]^2}_{\text{mean} = 0}$$

$$\sigma_y^2 = E[y^2]$$

$$\sigma_y^2 = S + N$$

$\therefore$ channel capacity $= B\log 2\pi e(S+N) - B\log(2\pi e N)$

$$= B\log\left[\frac{2\pi e(S+N)}{2\pi e N}\right]$$

$$B\log\left[\frac{S+N}{N}\right]$$

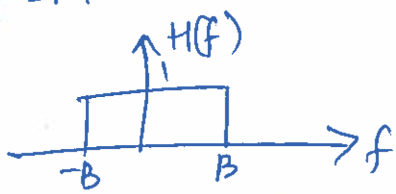
$$\boxed{C = B\log_2\left[1+\frac{S}{N}\right] \text{ bps.}}$$

$\hookrightarrow$ this Expression is known as Shannon's law

Trade off b/w Band width & S/N ration :

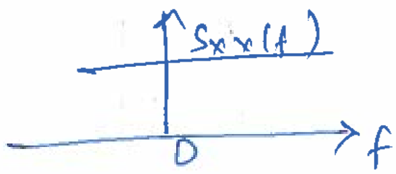
Let common channel be an ideal LPF

$$\therefore H(f) = 1 \; ; \; |f| \leq B$$
$$0 \; ; \; \text{OW}.$$



and I/p power spectral density (PSD) is constant its called a white Noise

$$S_{xx}(f) = \frac{\eta}{2}$$



from the relation b/w PSD of O/p & I/p

$$S_{yy}(f) = |H(f)|^2\, S_{xx}(f)$$

$$S_{yy}(f) = 1(\eta/2) \; ; \; -B < f < B.$$

O/p Noise power $N = \int_{-B}^{B} S_{YY}(f)\,df$

$$= \frac{\eta}{2}(f)\Big|_{-B}^{B} = 2B(\eta/2)$$

$$= \eta B.$$

we know $C = B\log\left(1+\frac{S}{\eta B}\right)$ bps

$$\boxed{\lim_{B\to\infty} C = \lim_{B\to\infty} B\log\left(1+\frac{S}{\eta B}\right)}\ \text{bps.}$$

in any communication channel noise $\propto$ BW

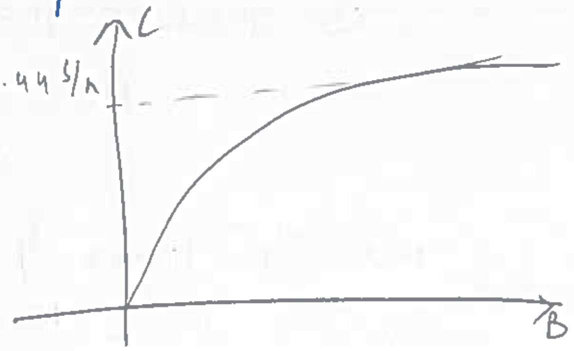
$$\lim_{B\to\infty} C = \lim_{B\to\infty} B\cdot\frac{\eta}{S}\cdot\frac{S}{\eta}\log\left(1+\frac{S}{\eta B}\right)$$

$$\frac{S}{\eta}\lim_{B\to\infty}\frac{\eta B}{S}\log\left(1+\frac{S}{\eta B}\right)$$

$$\lim_{x\to\infty} x\log_2\left(1+\frac{1}{x}\right) = \log_2 e$$

$$\therefore \lim_{B\to\infty} C = \log_2 e\,\frac{S}{\eta}$$

$$\lim_{B\to\infty} C = 1.44\frac{S}{\eta}\ \text{bps}$$



if $C = B\log\left(1+\frac{S}{N}\right)$

$C = 20$ kbps ; $B = 4$ kHz $\Rightarrow \frac{S}{N} = 31$
$C = 20$ kbps ; $B = 5$ kHz $\Rightarrow \frac{S}{N} = 15$
$C = 20$ kbps ; $B = 3.3$ kHz $= \frac{S}{N} = 63$.

from this, we can say $\frac{S}{N} \propto \frac{1}{B}$

→ Same channel can be obtained with different combination of Band width & SNR.

→ Both need not to be increased (or) decreased simultaneously

This is called 'Exchange b/w B.W and S/N ratio (or) Trade off b/w Band width and S/N ratio'.

15(b). What are the advantages of Source coding techniques? Explain any one of the source coding technique steps. 26.

4. If redundancy ~~increases~~ reduces, Efficiency increases.

Shannon-Fano coding.

Algorithm:

1. List the source symbols in order of decreasing probability
2. Partition the set into two sets that are as close to equiprobables as possible, and assign 0 to the upper set 1 to the lower set
3. Continue this process, each time partioning the sets with as nearly equal probabilities as possible until further partioning is not possible.

NSRIT

Semester End Examination, Nov./Dec., 2022

Degree	B. Tech. (U. G.)	Program	CSE	Academic Year	2022 - 2023
Course Code	20CS501	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Java Programming				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Why is Java known as platform independent?	20CS501.1	L1
2	What is the use of super keyword?	20CS501.2	L1
3	What are the advantages of multithreading?	20CS501.3	L1
4	How do applets differ from application program?	20CS501.4	L1
5	Draw AWT Class Hierarchy.	20CS501.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the concept of Type Conversion and Type Casting in Java with a Suitable Java Program?	6M	20CS501.1	L2
6 (b)	What is Recursion and Write a Java Program to display a Factorial of a number using Recursion?	6M	20CS501.1	L3
	OR			
7 (a)	Discuss various loop statements and branching statements available in Java? Show their Syntax.	6M	20CS501.1	L2
7 (b)	Discuss about java Buzzwords?	6M	20CS501.1	L2
8 (a)	Differentiate Method Overloading and Method Overriding.	6M	20CS501.2	L2
8 (b)	What is Inheritance? Explain the different types of Inheritance Mechanisms in Java with an example for each?	6M	20CS501.2	L2
	OR			
9 (a)	Define Constructor and Explain the different types of Constructors in Java with a suitable example	6M	20CS501.2	L2
9 (b)	With suitable code segments illustrate various uses of 'final' keyword.	6M	20CS501.2	L3
10 (a)	Explain Multithreading with an example.	6M	20CS501.3	L2
10 (b)	Write a Java Program that demonstrates the use of Abstract Class and Interface?	6M	20CS501.3	L3
	OR			
11 (a)	Explain the thread life cycle with neat diagram	6M	20CS501.3	L2
11 (b)	Explain thread priorities in Java. Write a java program using thread priorities	6M	20CS501.3	L2
12 (a)	Define Applet. What are the uses of applet?	6M	20CS501.4	L1
12 (b)	Write the HTML Applet Tag and explain each part of it.	6M	20CS501.4	L2
	OR			
13 (a)	Write a java program to read and count the characters from files.	6M	20CS501.4	L3
13 (b)	Distinguish between i.Input Stream and Reader classes ii.Output Stream and Writer Classes	6M	20CS501.4	L2
14 (a)	Explain the concept of AWT Button and AWT Text Field with suitable example	6M	20CS501.5	L2
14 (b)	Difference between Adapter classes and Inner classes.	6M	20CS501.5	L2
	OR			
15 (a)	List out various layout managers in Java. Write a java program for Border Layout class in Java.	6M	20CS501.5	L3
15 (b)	Demonstrate the mouse related events with java program	6M	20CS501.5	L3

NSRIT

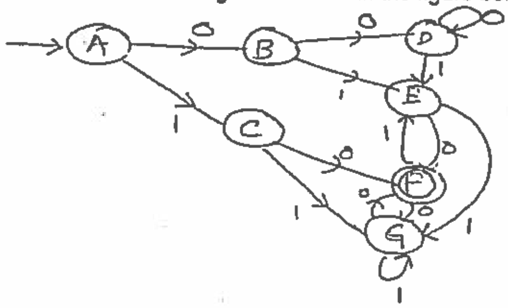
Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE (AI & ML) & CSE (DS)			Academic Year	2022 – 2023
Course Code	20CS405	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Theory of Computation						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Give the formal definition of Moore machine	20CS405.1	L1
2	List any two applications of PDA	20CS405.2	L1
3	What is undecidability?	20CS405.3	L1
4	List the two types of Parsers	20CS405.4	L1
5	Translate the following expression into Three address code a*(b+c)	20CS405.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Let $\Sigma = \{a, b\}$, Design DFA that accepts any string with "aababb" as a substring	6M	20CS405.1	L3
6 (b)	Design DFA to accept strings of a's and b's having even number of a's and b's	6M	20CS405.1	L3
	OR			
7	Minimize the following DFA shown in the figure below 	12M	20CS405.1	L3
8 (a)	Given CFG G = ({S, A}, {a, b}, P, S) where P consists of $S \rightarrow aAS \mid a$ $\quad A \rightarrow SbA \mid SS \mid ba$ Give the LMD, RMD and parse tree for the string "aabbaa"	6M	20CS405.2	L3
8 (b)	Design a PDA for the language $L = \{ww^R / w \in \{0, 1\}^*$, R stands for Reverse}	6M	20CS405.2	L3
	OR			
9	Convert the following grammar to Chomsky Normal Form S-> ABA $\quad$ A-> aA | Є $\quad$ B-> bB | Є and simplify the grammar	12M	20CS405.2	L3
10	Design TM for performing proper subtraction of two numbers	12M	20CS405.3	L3
	OR			
11 (a)	Explain Various Types of Turing Machines	6M	20CS405.3	L2
11 (b)	Explain Church Turing Thesis	6M	20CS405.3	L2
12 (a)	Describe the role performed by lexical analysis of the compiler	6M	20CS405.4	L2
12 (b)	Write the steps to perform FIRST and FOLLOW function	6M	20CS405.4	L2
	OR			
13	Define Compiler, Explain the Phases of compiler, and how the following statement will be translated into every phase Position:=initial+rate*60	12M	20CS405.4	L2
14	Define a DAG. Construct a DAG and write the sequence of instructions for the following expression a+a*(b-c)+(b-c)*d	12M	20CS405.5	L3
	OR			
15	Write short notes on a. Peephole optimization technique b. Global data flow analysis	12M	20CS405.5	L2

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Semester End Exams — NOV, DEC - 2022

Key of Theory of computation

1. Give the formal definition of Moore Machine.

A. $M = (Q, \Sigma, \Delta, \delta, \lambda, q_0)$

where Q = set of non empty states
Σ = set of i/p symbols
δ = Transition fn. $\delta : Q \times \Sigma \rightarrow Q$
Δ = set of o/p symbols.
λ = o/p fn $\lambda : Q \rightarrow \Delta$
q_0 = Initial State. — 2M

2. List any 2 applications of PDA.

A. • Context free language.
• Top down parsing
• Timed Automata model
• Tower of Hanoi
etc.., — 2M

3. What is undecidability?

A. If the language L of all yes instances to P is not decidable or a language is undecidable if it is not decidable.

(or)

The pb's for which we can't construct an algorithm that can answer the problem correctly in finite time are termed as undecidable pb's. — 2M

4. List the 2 types of Parsers.

A. 1. Topdown Parser
2. Bottom up Parser

— 2M

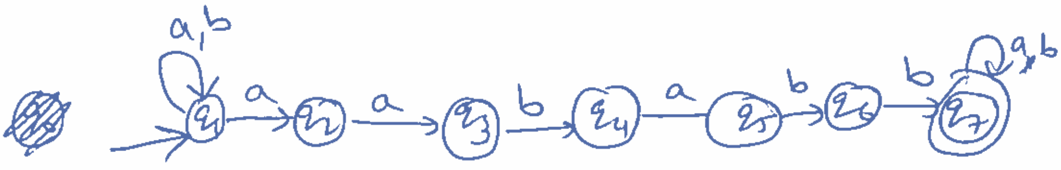
5. Translate the following exp into 3 addr code.

$a * (b+c)$

A. $t1 := b+c$

$t2 := t1 * a$ → 2M

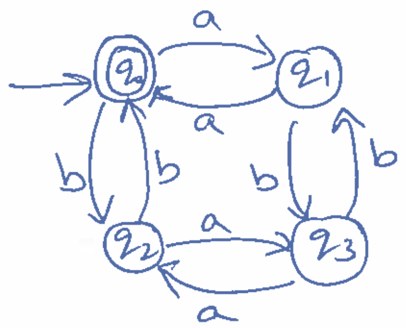
6.
(a)



L = {aaababbaba, baababbb, . . . }

→ 6M

6
(b)



L = { (aabb), (baba), (abab), . . . }

→ 6M

7

on 0: {A, B, C, D, E, G} {F}

A → B, B → D, C → F, D → D, E → F, G → F

→ 4M

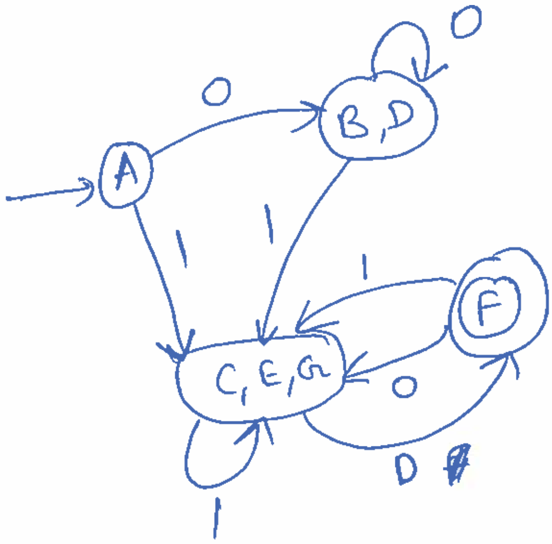
on 1: {A, B, D} {C, E, G} {F}

A → C, B → E, D → E; C → G, E → G, G → G

→ 4M

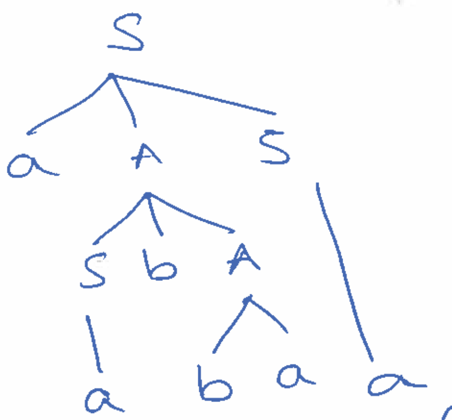
{A} {B, D} {C, E, G} {F}

→ 2M



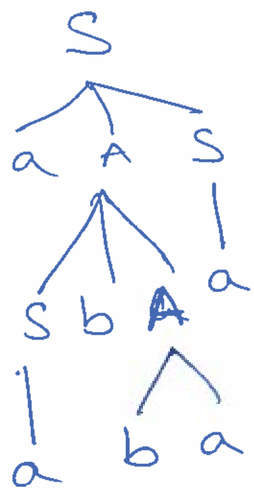
2M

8 (a) LMD: $S \rightarrow aAS$
$\rightarrow aSbAS$
$\rightarrow aabAS$
$\rightarrow aabbaS$
$\rightarrow aabbaa$ — 2M



1M

RMD: $S \rightarrow aAS$
$\rightarrow aAa$
$\rightarrow aSbAa$
$\rightarrow aSbbaa$
$\rightarrow aabbaa$ — 2M



1M

8 (b) $\delta(q_0, 0, Z_0) = (q_0, 0Z_0)$

$\delta(q_0, 0, 0) = (q_0, 00)$

$\delta(q_0, 1, q_0) = (q_0, 1Z_0)$

$\delta(q_0, 1, 1) = (q_0 11)$

$\delta(q_0, 0, 1) = (q_0, 01)$

$\delta(q_0, 1, 0) = (q_0 10)$

$\delta(q_0, c, Z_0) = (q_1, Z_0)$

$\delta(q_0, c, 0) = (q_1, 0)$

$\delta(q_0, c, 1) = (q_1, 1)$

$\delta(q_1, 0, 0) = (q_1, \epsilon)$

$\delta(q_1, 1, 1) = (q_1, \epsilon)$

$\delta(q_1, \epsilon, Z_0) = (q_f, \epsilon)$

3M

$(q_0, 011c110, z_0) \vdash (q_0, 11c110, 0z_0)$

$\vdash (q_0, 1c110, 10z_0)$

$\vdash (q_0, c110, 110z_0)$

$\vdash (q_1, 110, 110z_0)$

$\vdash (q_1, 10, 10z_0)$

$\vdash (q_1, 0, 0z_0)$

$\vdash (q_1, \epsilon, z_0)$

$\vdash (q_f, \epsilon)$. Accepted.

3M

9. Chomsky normal form: (CNF):

$S \to a$

$S \to AB$.

4M

Step 1:- $S \to ABA$

Remove start symbol in R.H.S. Here no starting symbol in RHS.

$\therefore S \to ABA$

Step 2:- Remove null productions, unit production, useless symbols

Step 3:- Replace terminals on RHS with non terminals & Construct new production.

4M

Finally we get, $S \to ABA \Rightarrow S \to XA$
$X \to AB$

$A \to aA/\epsilon \Rightarrow A \to YA$
$Y \to a$
$A \to a$

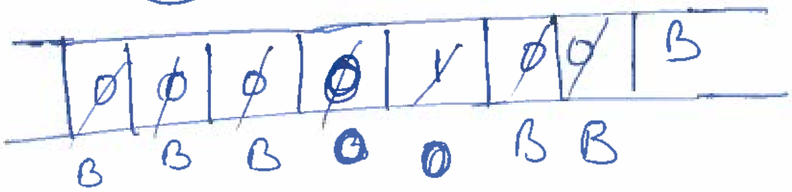
$B \to bB/\epsilon \Rightarrow B \to ZB$
$Z \to b$
$B \to b$

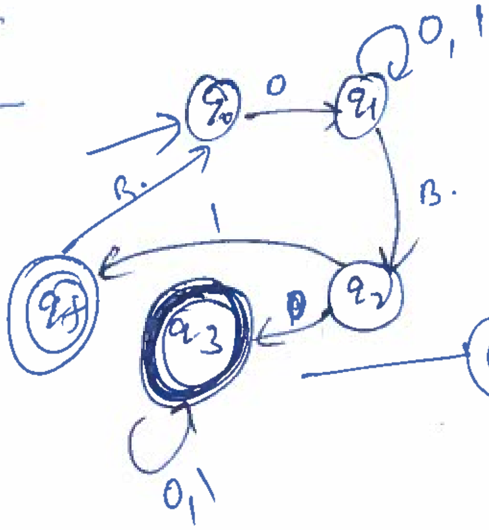
4M

10. (10) $m=4, n=2$

	0	1	B
$\rightarrow q_0$	(q_1, B, R)		
q_1	$(q_1, 0, R)$	$(q_1, 1, R)$	(q_2, B, L)
q_2	(q_3, B, L)	$(q_f, 0, R)$	
q_3	$(q_3, 0, L)$	$(q_3, 1, L)$	(q_0, B, R)
q_f			

6M





6M

11(a)

1. Multitape TM
2. Multitrack TM
3. Non - Deterministic TM
4. Two way infinite TM

If they write anything related to turing machine 6M

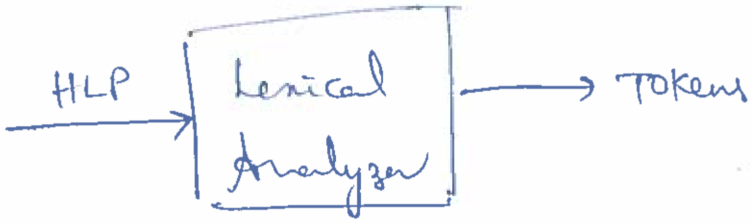
11(b) Church Turing Thesis:

"The assumption that the intuitive notion of computable functions can be ~~identifi~~ identified with partial recursive functions".

In 1930, this statement was first formulated by Alonzo Church & is usually referred to as Church's thesis, or the Church turing thesis.

6M

12 (a)



(2M)

- It converts HLP into no. of tokens.
- Lexical Analyzer is responsible for removing the white spaces & comments from the source program.
- It corresponds to the error messages with the source program.
- It helps to identify the tokens.
- The i/p characters are read by the lexical Analyzer from the source code. (4M)

12 (b)

FIRST :-

- If the terminal symbol is 'a' then FIRST{a} = {a}
- If there is a rule $X \to \epsilon$ then add ϵ to first of (X)
  i.e.. $FIRST(X) = \{\epsilon\}$
- If $X \to a\alpha$ then $FIRST(X) = a$.

(3M)

Any Example also they can write

FOLLOW :-

- For starting symbol S, place \$ in FOLLOW(S)
- If there is a production $A \to \alpha B \beta$ then everything in $FIRST(\beta)$ without ϵ is to be placed in FOLLOW(B)
- If there is a production $A \to \alpha B \beta$ or $A \to \alpha B$ & $FIRST(\beta) = \{\epsilon\}$ then $FOLLOW(A) = FOLLOW(B)$

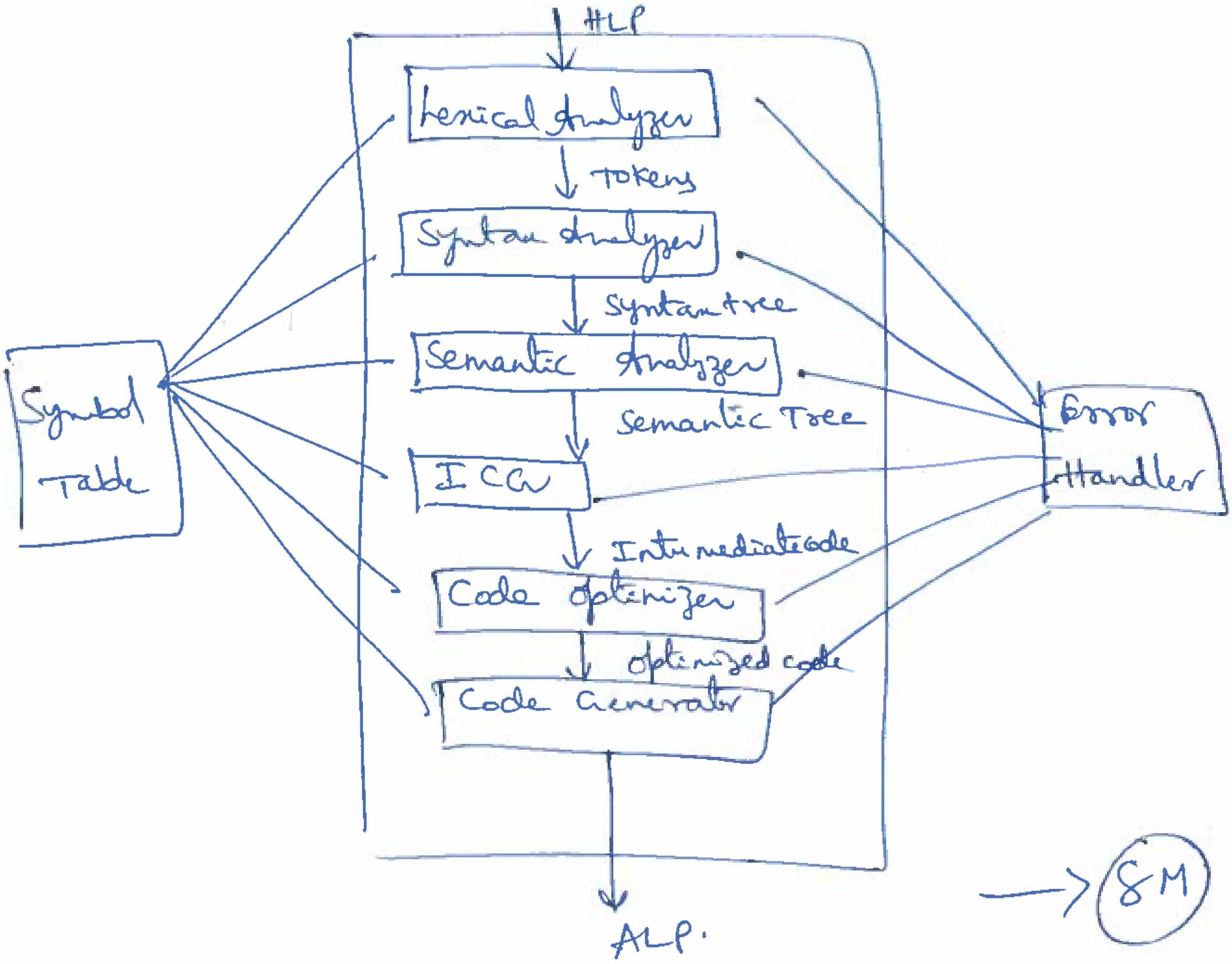
(3M)

Any example also they can write

13.

$HLP \rightarrow \boxed{\text{Compiler}} \rightarrow ALP.$

- Compiler converts HLP into ALP.



→ 8M

Ex:- Position := initial + rate * 60

↓

LA

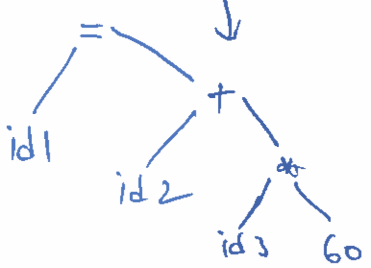
↓

id1 := id2 + id3 * 60

↓

SA

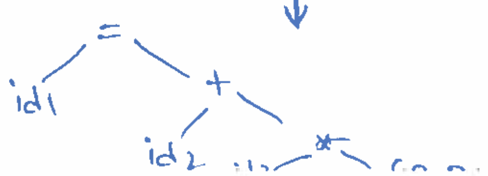
↓



↓

Semantic Analyzer

↓



↓

Intermediate code generator

↓

t1 := inttofloat(60)

t2 := id3 * t1

t3 := id2 + t2

id1 := t3.

↓

Code optimizer

t1 := id3 * 60.0

t2 := id2 + t1

id1 := t2

↓

Code generator

↓

MOV R1, #60.0

MUL R2, R1

MOV R3, id2

ADD R3, R2

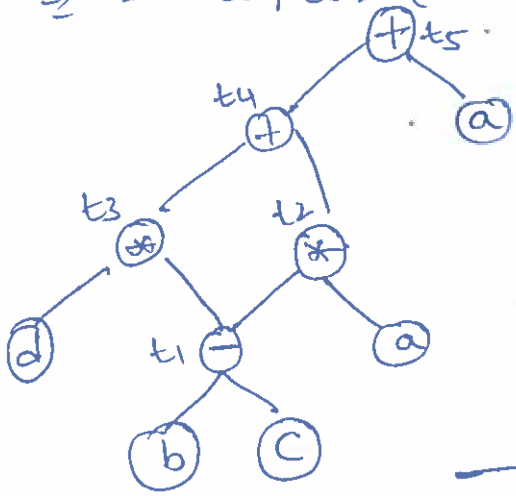
MOV id1, R3.

→ 4M

(14) **DAG :-** (Directed Acyclic Graph)

It is a directed graph with no directed cycles.

Ex :- $a + a * (b-c) + (b-c) * d$



DAG

6M

$t1 := b - c$
$t2 := a * t1$
$t3 := d * t1$
$t4 := t2 + t3.$
$t5 := a + t4$

6M

Three Address code

15(a) **Peephole optimization :-**

- It is performed on a small part of the code.
- The objective of peephole optimization is
  - To improve performance
  - To reduce memory footprint
  - To reduce code size
- Peephole optimization Techniques :-

(1) Redundant load & Store elimination

ex :-

$y = x + 5$
$i = y.$
$z = i$
$w = z * 3.$

→

$y = x + 5$
$i = y$
$w = y * 3$

2M

② Constant Folding :- The code can be simplified.

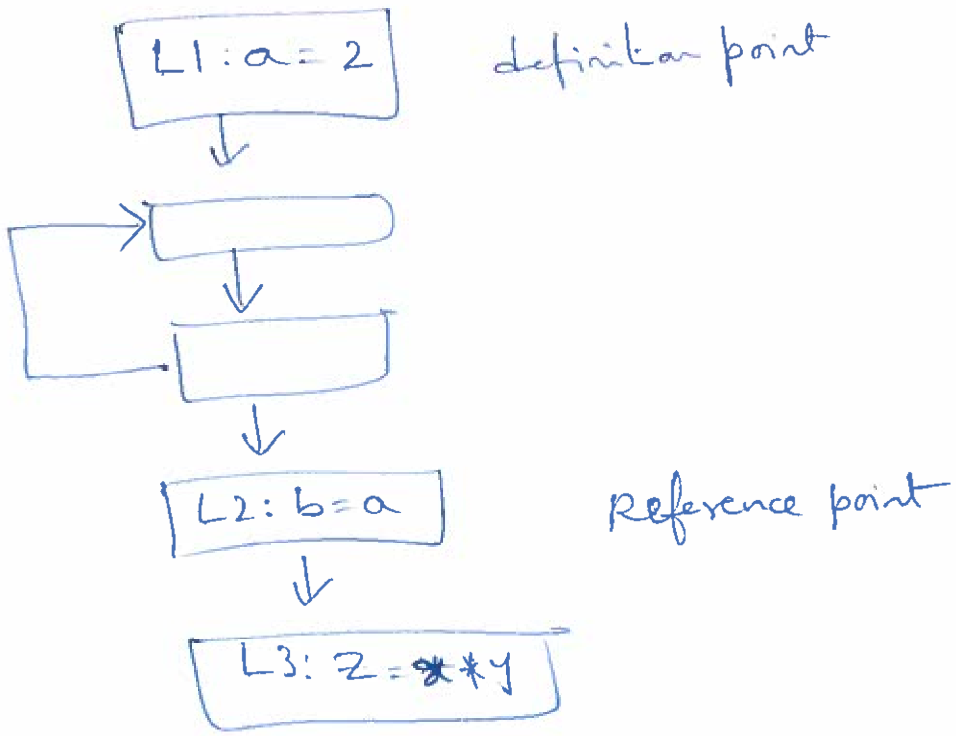
Ex: $x = 2*3 \rightarrow x = 6.$ 2M

③ Strength reduction :- The operators that consumes higher execution time are replaced by the operators consuming less execution time.

$y = x*2 \longrightarrow y = x + x.$ 2M

5 ⓑ Global data flow Analysis:

It is the analysis of flow of data in control flow graph i.e.. the analysis that determines the information regarding the definition & use of data in program. 2M



Evaluation point. 2M

- Definition pt :- A point in a program containing some definition
- Reference point :- A point in a program containing a reference to a data item
- Evaluation pt :- A point in a program containing evaluation of expression.

2M

SCHEME OF EVALUATION

Degree: B.Tech Program: civil Academic year: 2012-23

Course code: 20CE502 Test duration: 3hrs Max. Marks: 70

SEM: V, Course: DRCE (Design of Reinforced concrete elements)

PART A (5×2 = 10M)

1. 4-Assumptions → 4 × 1/2 = 2M
2. Bending moment written ⇒ 2M
3. Max & Min → 2M
4. Reasoning → 2M
5. Types (any 4) → 2M

10M

PART-B (Each 12M)

7 → Data notations → 2M
Procedure → 6M
Results + Diat Calculations → 4M

12M

6 (a) → Assumption + Diagram → 4M

(b) → Procedure → 4M
formulae → 4M

12M

8 Data + notations → 2M
Procedure → 6M
Results + Calculation → 4M

12M

9. Data + notations → 2M
   Procedure + Dia → 6M
   Results + calulation → 4M
   12M

10. Data + notations → 2M
    Procedure + Dia → 6M
    Results + calulation → 4M
    12M

11. Data + notations → 2M
    Procedure + Dia → 6M
    Results + calulations → 4M
    12M

12. Data + notation → 2M
    Procedure + Dia → 6M
    Results + calulations → 4M
    12M

13. Data + notations → 2M
    Procedure + sketch → 6M
    Results + calutions → 4M
    12M

14. Principle → 6M
    Assumptions → 6M
    12M

15. Design Principle → 2M
    Procedure → 6M
    Diagram → 4M
    12M

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ANSWER KEY AND SCHEME OF EVALUATION

SEMESTER END REGULAR EXAMINATION, Nov/Dec, 2022

Degree: B.Tech Program: Civil Engineering Academic Year: 2022-23

CODE: 20CE502 Test Duration: 3 hrs Semester: V

Course: Design of Reinforced Concrete elements

PART - A (5×2 = 10M)

1 Ans

(1) The plain sections remains plane before and after bending; these assumptions ensure that the strain diagram is linear.

(2) At the time of failure, the maximum strain in concrete at the topmost fiber will be 0.0035

(3) The stress strain curve for the concrete is parabolic up to strain of 0.002, and stress will be constant up to strain of 0.0035

(4) The tensile strength of concrete below the neutral axis is ignored; says tensile strength of concrete is ignored

2 Ans: The stiff elements restrain the slab and cause additional bending moments at the exterior corners. Corner reinforcemt must be provided in top and bottom of the slab to resist these bending moments

3 Ans: As per IS-456-200 the cross sectional area of longitudinal reinforcemt in columns shall not be less than 0.8% nor more than 6% of the gross cross-sectional area of the column

4 Ans: → Combined footing is used while construction of two (or) more columns when they are close to each other and their foundations overlap.

→ Main purpose of using combined footing is to distribute uniform pressure under the footing

5 Ans: → useful when soil bearing capacity low which causes overlapping of adjacent isolated footing

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ANSWER KEY AND SCHEME OF EVALUATION

5 Ans: Types of water tanks:

(i) Based on water tank location

(a) underground tanks
(b) Tanks resting on ground
(c) Overhead tanks

(ii) Based on water tank shape.

(a) Rectangular tank
(b) circular tank
(c) spherical tank
(d) circular tank with conical bottom

PART-B (5x12 = 60M)

6(a) limit state collapse

→ Plane section normal to the axis remains plane after bending

→ The maximum strain in the concrete at the outermost compression fiber can be take 0.0035 in bending

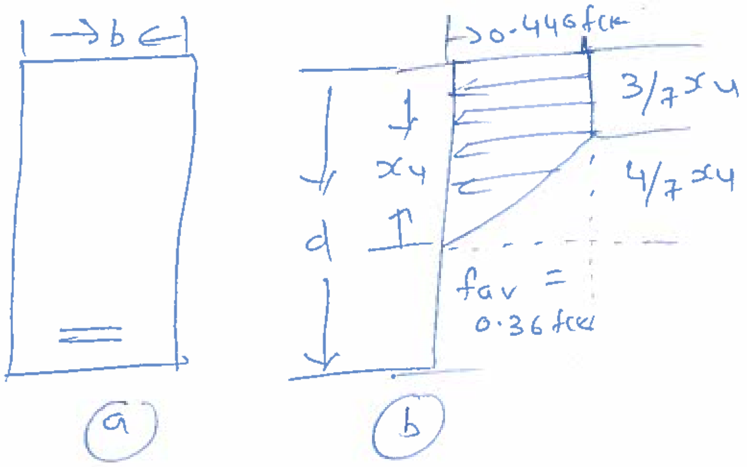
→ Tensile strength of concrete ignored

STRESS BLOCK PARAMETERS

→ Diagram show distribution compressive strength in concrete across depth x_u of the section termed as "Stress block"

→ Since strain diagram linear over this depth x_u, then shape of stress block is same

→ Varies parabolically upto height $\frac{4}{7}x_u$ i.e $\left(\frac{0.002}{0.0035}x_u\right)$ ϵ_p has constant value equal to design stress $0.446 f_{ck}$ i.e $\left(0.67 \times \frac{1}{1.5} f_{ck}\right)$ for the $\frac{3}{7}x_u$ i.e $\left(\frac{0.0035-0.002}{0.0035}\right)x_u$



Area A of stress block = Rect Portion + Area Parabolic

$= 0.446 f_{ck} \times \frac{3}{7}x_u \times \frac{2}{3} \times 0.446 f_{ck} \times \frac{4}{7}x_u$

$= 0.361 f_{ck} x_u = 0.36 f_{ck} x_u$

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ANSWER KEY AND SCHEME OF EVALUATION

Area stress block = $f_{av}\, x_u$

Compressive force = $C = b x_u f_{av}$

$= b x_u\, 0.36 f_{ck} = k_1 f_{ck} b x_u$

$k_1 = 0.36$ is defined as stress block parameter

$C\bar{x} = C_1 x_1 + C_2 x_2$

$C = 0.36 f_{ck} b x_u$, $A_1 = 0.446 f_{ck}\, b \frac{3}{7} x_u$

$x_1 = \frac{1}{2} \times \frac{3}{7} x_u = \frac{3}{14} x_u$, $A_2 = \frac{2}{3} \times 0.446 f_{ck} \times b \times \frac{4}{7} x_u$

$x_2 = \frac{3}{7} x_u + \frac{3}{8}\left(\frac{4}{7} x_u\right)$

Distance of parabolic depth 'a' $\frac{3}{8} a$

$= \frac{3}{7}(1+0.5) x_u = \frac{4.5}{7} x_u$

$0.36 f_{ck} b x_u \bar{x} = 0.446 f_{ck} \times b \frac{3}{7} x_u \times \frac{3}{14} x_u +$

$\frac{2}{3} \times 0.446 f_{ck} \times b \frac{4}{7} x_u \times \frac{4.5}{7} x_u$

$\bar{x} = 0.417 x_u = 0.42 x_u$

$= k_2 x_u$

$k_2 = 0.42$

(6b) Design of Singly Reinforced section:

(i) Take depth as $\frac{1}{12}$th – $\frac{1}{15}$th span (nearest 50mm)

(ii) Take breadth $\frac{1}{3}$rd – $\frac{1}{2}$ the depth (min 200mm)
(Round off nearest multiple 50mm)
(230 permitted)

(iii) Determine Mu lim

(iv) If Mu lim > Mu , step (v) : otherwise increase the depth

(v) Equating moment of resistance to B.M ,
determine area of steel

$$C = T \text{ , gives}$$

$$0.36 f_{ck}\, b\, x_u = 0.87 f_y A_{st}$$

$$x_u = \frac{0.87 f_y A_{st}}{0.36 f_{ck}\, b}$$

(or)

Moment of resistance = Design moment

$$T \times LA = M_u$$

$$0.87 f_y A_{st} (d - 0.42 x_u) = M_u$$

$$M_u = 0.87 f_y A_{st} \left(d - \frac{0.42 \times 0.87 f_y A_{st}}{0.36 f_{ck}\, b} \right)$$

$$M_u = 0.87 f_y A_{st}\, d \left(1 - 1.05 \frac{A_{st}}{bd} \frac{f_y}{f_{ck}} \right)$$

$$M_u = 0.87 f_y A_{st}\, d \left(1 - \frac{A_{st}}{bd} \frac{f_y}{f_{ck}} \right)$$

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ANSWER KEY AND SCHEME OF EVALUATION

~~8 Ans~~

9 Ans Slab dimensions $l_y = 8m$, $l_x = 6m$

floor finish + L.L = 3 kN/m², $f_{ck} = 20N/m^2$, $f_y = 415$

STEP-1 $\frac{L_y}{L_x} = \frac{8}{6} < 2$

$\frac{l_y}{l_x} < 2$ slab – two way

STEP-2 fixing depth of slab

$\frac{L}{d} = 20$ (S.S case)

$\Rightarrow \frac{L_x}{d} = 20 \Rightarrow \frac{6000}{d} = 20 \Rightarrow \boxed{d = 300\text{ mm}}$

Eff cover 30mm (d')

$\therefore D = d + d' = 300 + 30 = 330mm$

$D = 330mm$

STEP-3 Eff depth calulation

(I) Eff span for short span

(i) clear span + eff depth = 6 + 0.3 = 6.3m

A/

(ii) c/c distance b/w supports $= 0.15 + 6 + 0.15$

$= 6.3m$

Assume (300mm wall) $\Rightarrow$

Select least

$\Rightarrow$ $\boxed{l_x = 6.3m}$

(II) Eff span for long span

(i) clear span + eff depth $= 0.3 + 8 = 8.3m$

(ii) c/c dist b/w supports $= 0.15 + 8 + 0.15$

$= 8.3m$

least

$\boxed{\therefore l_y = 8.3m}$

STEP-4 Load calculations

(1) Dead load $(w_D) = D \times b \times \gamma$

$= 0.3 \times 1 \times 25 = 7.5 KN/m^2$

(2) $L.L = 3 KN/m^2$

Total load $= 7.5 + 3 = 10.5 KN/m^2$

ultimated load $= (w_u) = 1.5 \times w = 1.5 \times 10.5$

$\boxed{w_u = 15.75 KN/m^2}$

STEP-5 Calculations of BM

IS-456

$M_x = \alpha_x w_u l_x^2$

$M_y = \alpha_y w_u l_x^2$

Pg 91 : 456

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ANSWER KEY AND SCHEME OF EVALUATION

α_x & α_y B.M coeff

$$\boxed{\frac{l_y}{l_x} = 1.33}$$

$$\frac{l_y}{l_x} = 1.3 \Rightarrow \alpha_x = 0.093,$$

$$\frac{l_y}{l_x} = 1.4 \Rightarrow \alpha_x = 0.099$$

Interpolating for $\frac{l_y}{l_x} = 1.33$

$$\left(\frac{l_y}{l_x}\right)_{1.33} = 0.093 + \left(\frac{0.099 - 0.093}{1.4 - 1.3}\right)(1.33 - 1.3)$$

$$= 0.093 + (0.06) \times 0.03 = 0.0948$$

$$\left(\frac{l_y}{l_x}\right)_{1.33} \Rightarrow \boxed{\alpha_x = 0.0948}$$

for, $\frac{l_y}{l_x} = 1.3 \Rightarrow \alpha_y = 0.055,$

$$\frac{l_y}{l_x} = 1.4 \Rightarrow \alpha_y = 0.051$$

for $\left(\frac{l_y}{l_x}\right)_{@ 1.33}$ Interpolating

$$= 0.055 - \left(\frac{0.055 - 0.051}{1.4 - 1.3}\right)(1.33 - 1.3)$$

$$= 0.055 - (0.04 \times 0.03) = \underline{0.0538}$$

for $\left(\frac{l_y}{l_x}\right)_{1.33}$ $\boxed{\alpha_y = 0.0538}$

Now B.M short span

$$M_x = \alpha_x W_u l_x^2$$

$$M_x = 0.0948 \times 15.75 \times (6.3)^2$$

$$M_x = 59.26 \text{ KN}$$

B.M long span

$$M_y = \alpha_y W_u l_x$$

$$M_y = 0.0538 \times 15.75 \times (\cancel{0.0538}\ 6.3)^2$$

$$M_y = 33.63 \text{ KN}$$

STEP-6

Depth req

$$d_{req} = \sqrt{\frac{M_{ux}}{0.138 f_{ck} b}} = \sqrt{\frac{59.26 \times 10^6}{0.138 \times 20 \times 1000}}$$

$$d_{req} = 146 \text{ mm}$$

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ANSWER KEY AND SCHEME OF EVALUATION

Depth provided = 300 > Depth req

∴ under reinforced section

Can be reduced to 120 mm

STEP - 7 Calculation of Ast

$$(A_{st})_x = \frac{0.5 f_{ck}}{f_y}\left(1 - \sqrt{1 - \frac{4.6 M_{ux}}{f_{ck}\, b\, d_x}}\right) b\, d_x$$

$$(A_{st})_x = \frac{0.5 \times 20}{415}\left(1 - \sqrt{1 - \frac{4.6 \times 59.26 \times 10^6}{20 \times 1010 \times (130)^2}}\right) \begin{matrix}1000 \times \\ 130\end{matrix}$$

$$(A_{st})_x = 0.024\left(1 - \sqrt{1 - \cancel{45.93}\ 0.15}\right) 300{,}000$$

$$(A_{st})_x = 0.024 \times 0.92 \times 300{,}000 = \cancel{6624 \text{ mm}^2}\ 334.71 \text{mm}^2$$

$$\cancel{A_{st\,x} = 6624 \text{ mm}^2}\ \ 6632.438$$

$$\boxed{A_{st\,x} = 334.71 \text{mm}^2}$$

Ast long span

$$d_y = 150 - \left(\frac{8}{2} + \frac{8}{2}\right) = 142 \text{ mm}$$

$$(A_{st})_y = \frac{0.5 f_{ck}}{f_y}\left(1 - \sqrt{1 - \frac{4.6 M_{uy}}{f_{ck}\, b (d_y)^2}}\right) b \cdot d_y$$

$$= \frac{0.5 \times 20}{415}\left(1 - \sqrt{1 - \frac{4.6 \times \cancel{9.04}\ 33.63}{20 \times 1000 \times (142)^2}}\right) 1000 \times 142$$

$$= 0.024 \left(1 - 0.99\right) 292{,}000 = 70.08 \text{ mm}^2$$

$$\boxed{(A_{st})_y = \cancel{70.08 \text{ mm}^2}\ 233.31 \text{ mm}^2}$$

$$A_{st\,min} = 0.12\,\%\, b D = 0.12\,\% \times 1000 \times 150$$

$$\boxed{A_{st\,min} = \cancel{396 \text{ mm}^2}\ 180 \text{ mm}^2}$$

Now for 8 mm bars

$$a_{st} = \frac{\pi}{4} \times (8)^2 = 50.26 \text{ m}^2$$

STEP-8

Spacing for short span

$$S_x = \frac{a_{st}}{(A_{st})_x} \times 1000 = \frac{50.26}{\cancel{364.74}\ 344.71} \times 1000$$

$$\boxed{S_x = 145.8 \text{ mm c/c}}$$

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ANSWER KEY AND SCHEME OF EVALUATION

$\therefore$ Provide S_x = 140mm c/c along short span

Spacing for long span

$$S_y = \frac{ast}{(Ast)_y} \times 1000 = \frac{50.26}{233.81} \times 1000$$

$$\boxed{S_y = 214\,mm\ c/c}$$

Provide S_y = 200mm c/c long span

Now actual $(Ast)_x$ Provided $= \frac{50.26}{140} \times 1000$

$$\boxed{(Ast_x)_{pro} = 359\,mm^2}$$

actual (Ast_y) Prov

$$(Ast_y)_{pro} = \frac{50.26}{200} \times 1000 = 251.3\,mm^2$$

$$\boxed{(Ast_y)_{prov} = 251.3\,mm^2}$$

Alternate bars of short span can be bent 0.15 l_x from centre of support

$\therefore$ $0.15\ l_x = 0.15 \times (4.12)$

$= 0.618 \simeq 0.62$ m from centre of support

Alternate bars of long span to be bent 0.15 l_y from centre of support

$\therefore$ $0.15 \times l_y = 0.15 \times 5.12 = 0.768$ m from centre of support & remaining alternate bars will continue to support

x

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ANSWER KEY AND SCHEME OF EVALUATION

11 Ans

Dia = 450mm, P_u = 1200KN, M_u = ?

f_{ck} = 20 N/mm², f_y = 415 N/mm2

Assume d' = 50mm, Now $\frac{d'}{D} = \frac{50}{450} = 0.11$

using chart on Sp-16

Non dimensional parameter

$$\frac{P_u}{f_{ck}D^2} = \frac{1200 \times 10^3}{20 \times (450)^2} = 0.29$$

See chart in SP-16

Now we get $\frac{P}{f_{ck}} = 0.06$

$\therefore P = f_{ck} \times 0.06 = 20 \times 0.6 \Rightarrow P = 1.2\%$

Now $A_{st} = \frac{P\pi D^2}{400} = \frac{1.2 \times \pi \times 500^2}{400}$

$= 2356$ mm2

Now using 20mm ϕ bars $ast = \frac{\pi}{4}(20)^2 = 314.8 mm^2$

$\therefore$ No of bars $= \frac{Ast}{ast} = \frac{2356}{314.16} = 7.49$

$\therefore$ Provide 8 - 20mm ϕ

Now, Design of lateral ties (Pg 49 456-2000)

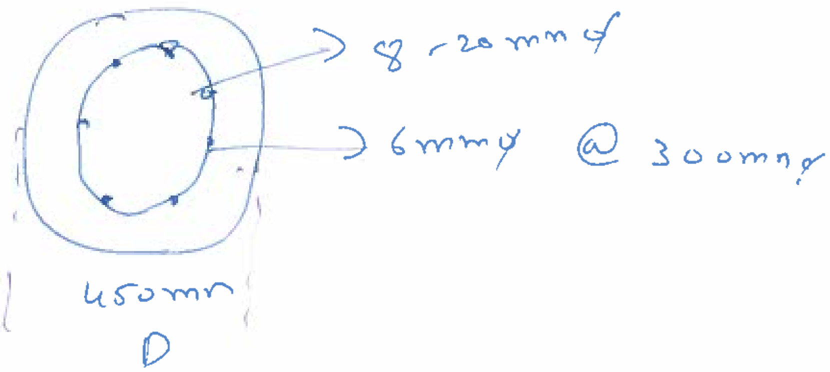
dia of ties $> \frac{1}{4}(20) = 5mm$

Now spacing

(1) least lateral dim = 500mm
(2) $16\phi = 16 \times 20 = 320mm$
(3) 300mm

Select the least value

$\therefore$ Provide 6mm ϕ @ 300mm lateral tie



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ANSWER KEY AND SCHEME OF EVALUATION

12 Ans

$L \cdot L = 1000 kN$, Size = 300 mm × 300 mm

$SBC = 160\ kN/m^2$, $f_{ck} = 25\ N/mm^2$, $f_y = 500$

STEP-1 Assume self weight of footing = 10% of super imposed load

$W_D = 10\% \times 1000\ kN = 0.1 \times 1000$

$$\boxed{W_D = 100\ kN}$$

Now Total load = $w = W_L + W_D = 1000 + 100$

$$\boxed{w = 1100\ kN}$$

So, req area of footing 'A' $= \dfrac{w}{SBC}$

$$= \frac{1100}{160} = 6.875 m^2$$

$$\boxed{A = 6.875 m^2}$$

Now for square footing $B = \sqrt{A}$

$$B = \sqrt{6.875} = 2.62\ m$$

$$\boxed{B = 2.62 m}$$

$\therefore$ Provide square footing of size 2.7 m × 2.7 m

STEP-2 Net upward pressure

$$P_o = \frac{W}{B^2} = \frac{1100}{(2.7)^2} = \frac{1100}{7.344} = 149.78 \text{ kN/m}^2$$

$$\boxed{P_o = 149.78 \text{ kN/m}^2}$$

STEP-3: Depth of footing on basis of B.M

Max B.M at face of column given by

$$M = P_o \frac{B}{8} (B-b)^2$$

$$M = 149.78 \times \frac{(2.7)}{8} (2.7 - 0.3)^2$$

$$M = 149.78 \times 0.33 \times 5.76$$

$$\boxed{M = 284.70 \text{ kN.m}}$$

ultimate moment (Mu)

$$M_u = 1.5 M \Rightarrow 1.5 \times 284.70$$

$$\boxed{M_u = 427.05 \text{ kNm}}$$

Dreq

$$M_u = 0.138 f_{ck} \cdot B d^2$$

$$d^2 = \frac{M_u}{0.138 f_{ck} B} \Rightarrow d = \sqrt{\frac{M_u}{0.138 f_{ck} B}}$$

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ANSWER KEY AND SCHEME OF EVALUATION

$$d = \sqrt{\frac{427.05 \times 10^6}{0.138\ f_{ck}\ B}} = \sqrt{\frac{427.05 \times 10^6}{0.138 \times 25 \times 2700}}$$

$$\boxed{d = 214.11\ mm} = 215\ mm$$

Provide 50mm cover i.e d'

$$\therefore D = 215 + 50 = 265\ mm$$

$$\boxed{D = 265\ mm}$$

Due to shear consideration adopt higher eff depth

$\therefore$ adopt 400mm as eff depth (d) & 50mm cover

$$\therefore d = 400\ mm, \quad D = 400 + 50$$

$$\boxed{D = 450\ mm}$$

STEP-4 Area of Reinforcement

$$A_{st} = \frac{0.5\ f_{ck}}{f_y}\left(1 - \sqrt{1 - \frac{4.6\ M_u}{f_{ck}\ Bd^2}}\right) Bd$$

$$A_{st} = \frac{0.5 \times 25}{500}\left(1 - \sqrt{1 - \frac{4.6 \times 427.05 \times 10^6}{20 \times 2700 \times 400^2}}\right)$$

$$\times 2700 \times 400$$

$$\boxed{\begin{array}{l} A_{st} = (0.025)(0.121)(1080000 \\ A_{st} = \cancel{14121\ mm^2}\ 3267 mm^2 \end{array}}$$

$$P = \frac{A_{st}}{Bd} \times 100 = \frac{1\,3121}{2700 \times 400} \times 100 = 3 \times 10^{-3}$$

$$\boxed{P = 1.30\%} \qquad P = 0.0\cancel{0}3\%$$

~~Min is 0.12% therefore OK~~

using 12mm ϕ bars in both directions

$$a_{st} = \frac{\pi}{4}(12)^2 = 113 mm^2$$

$$\therefore \text{No. of bars} = \frac{A_{st}}{a_{st}} = \frac{3267}{113} = 26 \text{ bars}$$

Provide 12 mm bars in both direction

$$\therefore \text{Spacing} = \frac{B}{\text{no of bars}} = \frac{2700}{26} = 103 mm c/c$$

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ANSWER KEY AND SCHEME OF EVALUATION

13 Ans) Column size = 300x500 mm, $P = 1500 kN$

S.B.C = 200 kN/m², $f_{ck} = 20$, $f_y = 415$

Area of footing

$P = 1500 kN$

Selfweight = 200 kN

Total load = 2200 kN

SBC = 200 kN/m² $\Rightarrow A = \frac{2200}{150} = 14.667 m^2$

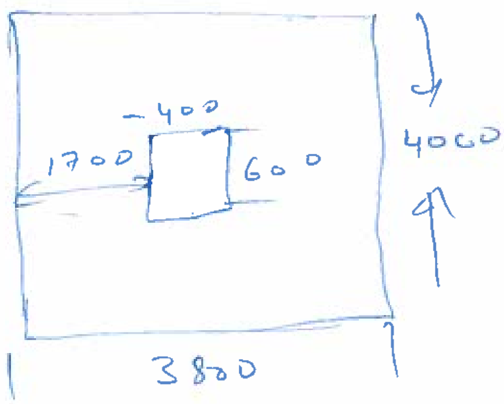
Providing 3.8 m x 4.0 footing

Area provided = 3.8 x 4 = 15.2 m²

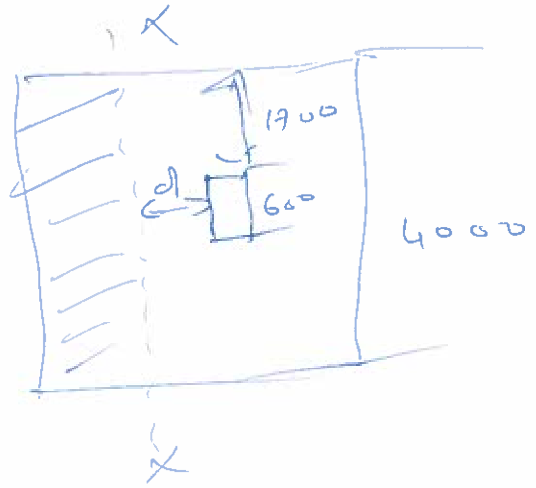
Soil press for design

$q_u = 1.5 \times \frac{2000}{15.2} = 197.4 kN/m^2$

Cantilever Projection $l = \frac{3.8 - 0.4}{2} = 1.7 m$



Depth of footing



Total shear force across section x-x

$V_u = q_u \times$ shaded area $= 0.1974 \times 4000\,(1700 - d)$

Assuming $P_t = 0.2$, τ_c for $M_{20} = 0.32\,N/mm^2$

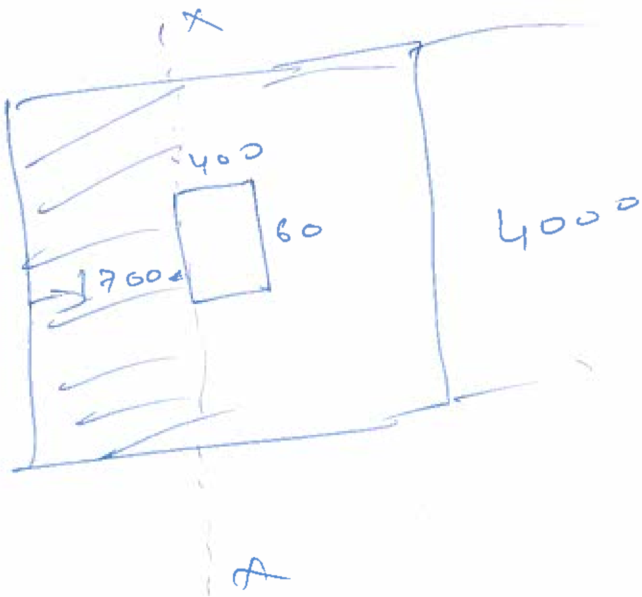
let depth of footing be "d".

$0.1974 \times 4000\,(1700 - d) = 0.32 \times 4000\,d$

$d = 648\,mm$

Provide, $d = 650\,mm$ & $D = 725\,mm$

Check for bending



M_u for 4000 mm

$M_u = 0.1974 \times 4000 \times 1700 \times \frac{1700}{2}$

$M_u = 1140.972 \times 10^6\,N\text{-}mm$

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$x_{u\,lim} = 0.48 \times 650 = 312$

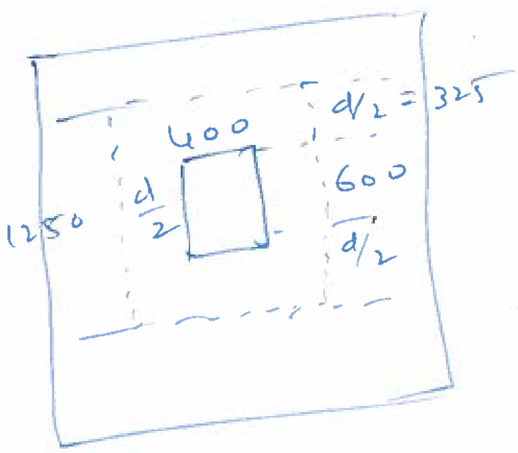
$M_{u\,lim} = 0.36 \times 20 \times 4000 \times 312 \left(650 - 0.42 \times 312\right)$

$= 4663.16 \times 10^{6}$ N-mm $> M_u$

∴ depth selected sufficient to design footing as singly reinforced section

Check for depth from the consideration of 2-way shear

$\frac{d}{2} = \frac{650}{2} = 325$ mm



Perimeter $= 2(400 + 650 + 600 + 650) = 2(1050 + 1250)$

$= 4600$ mm

$\therefore$ Area of concrete resisting 2-way shear

$= 4600 \times 650 = 2990000 mm^2$

Punching shear $V_u = 0.1974 (4000 \times 3800 - 1050 \times 1250)$

$= 2741392 N$

Equating to resisting shear

$2741392 = 2990000 \tau_c$

$\tau_c' = 0.917 N/mm^2$

2-way shear stress permitted (clause 31.6.3.1)

$= k_s \tau_c$

where $k_s = 0.5 + \beta_c$

$\beta_c = \frac{3.8}{4} = 0.95, \quad k_s = 1$

$\tau_c = 0.25 \sqrt{f_{ck}} = 0.25 \sqrt{20} = 1.118 N/mm^2$

But actual $\tau_c = 0.917 N/mm^2$

$\therefore$ footing safe in 2-way shear

Depth $= 650 mm$

Reinforcement : $M_u = 0.87 f_y A_{st} d \left[1 - \frac{A_{st} f_y}{bd f_{ck}}\right]$

$1140.972 \times 10^6 = 0.87 \times 415 \times A_{st} \times 650$

$\left(1 - \frac{A_{st}}{4000 \times 650} \times \frac{415}{20}\right)$

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$$4861.8 = A_{st}\left(1 - \frac{A_{st}}{125301}\right)$$

$$\therefore A_{st}^2 - 125301 A_{st} + 4861.8 \times 125301 = 0$$

$$A_{st} = 5068 \text{ mm}^2$$

using 16 m bars $S = \frac{\pi/4 \times 16^2}{5068} = 158.1$ mm

Provide 16 mm @ 150 mm c/c both direction

$$P_t = \frac{\pi/4 \times 16^2}{150 \times 650} \times 100 = 0.206 > 0.2$$ assumed for taking τ_c

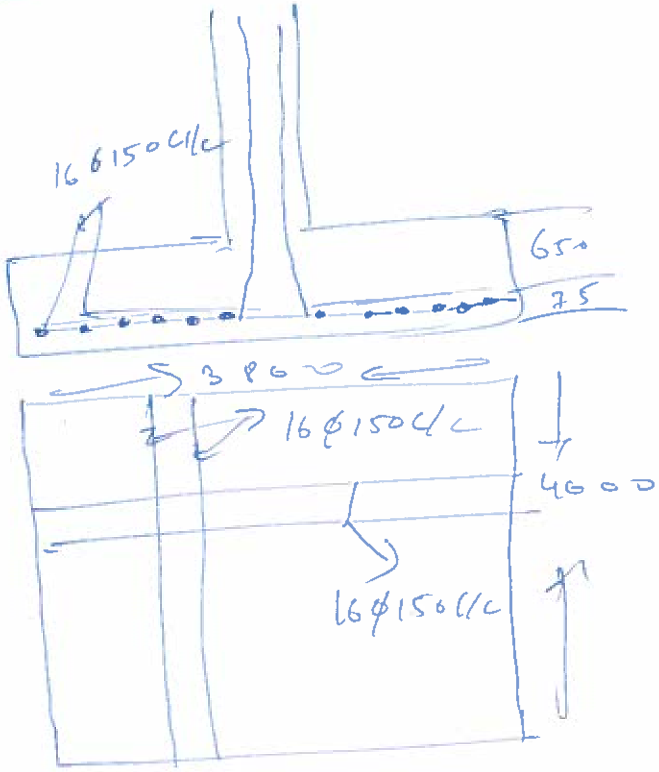
Hence ok

Check for development length

$L_d = \frac{0.87 f_y \phi}{4 \tau_{bd}}$, τ_{bd} for M_{20} & Fe 415 $= 1.92$ N/mm² (IS-456 Clause 26.2.1.1)

$$L_d = \frac{0.87 \times 415 \times 16}{4 \times 1.92} = 752 \text{ mm} < 1700 \text{ mm}$$

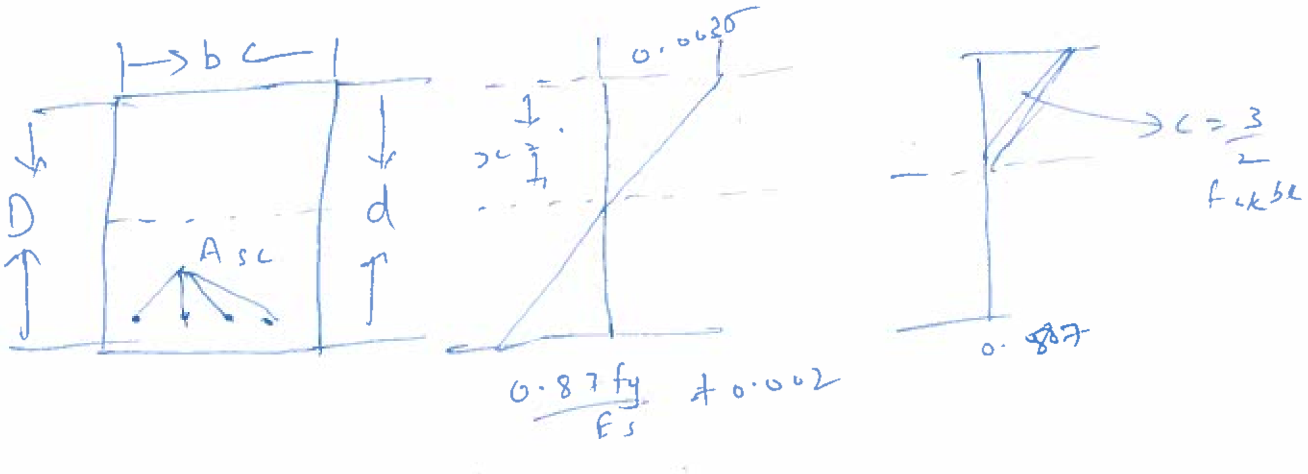
Reinforcement details



14 Ans

working stress design method & its assumptions with limitations

working stress design method used for the reinforced concrete design where concrete is assumed as elastic, steel & concrete act together elastically where the relationship between loads & stresses is linear



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ANSWER KEY AND SCHEME OF EVALUATION

Assumptions of working stress method

(i) Plane section before bending will remain plane after bending

(ii) Bond b/w steel & concrete is perfect with in elastic limit of steel.

(iii) The steel and concrete behaves as linear elastic material

(iv) All tensile stress are taken by reinforcement & none by concrete

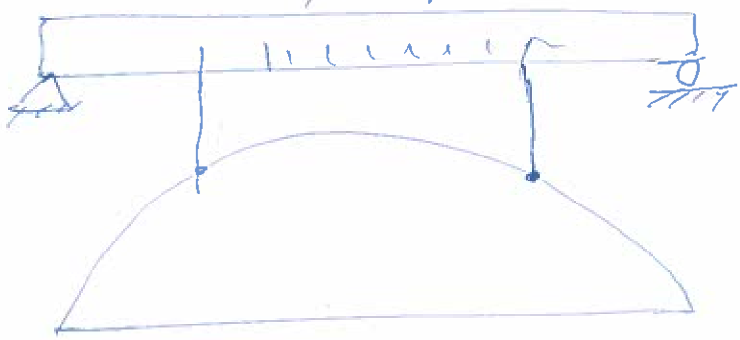
(v) Stresses in steel & concrete are related by factor "modular ratio"

(vi) Stress - strain relationship of steel & concrete is a straight line under working load

Principle WSM

→ Traditional method used in RCC & Steel, timber.

→ Obeys Hook's law, Stress directly proportional to strain upto point of collapse



→ Main principle structural material behaves in linear elastic manner and also ensure adequate safety by suitability restricting the stress in materials induced

$$\text{Permissible stress} = \frac{\text{ultimate (or) yield stress}}{F.O.S}$$

$$\text{working stress} \leq \text{Permissible stress}$$

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ANSWER KEY AND SCHEME OF EVALUATION

(15) Design Principle cantilever retaining wall

Design Parameters (i) Height of earth to be retained, h (m)

(ii) Surcharge pressure on backfill, w_s (kN/m²)

(iii) Soil properties

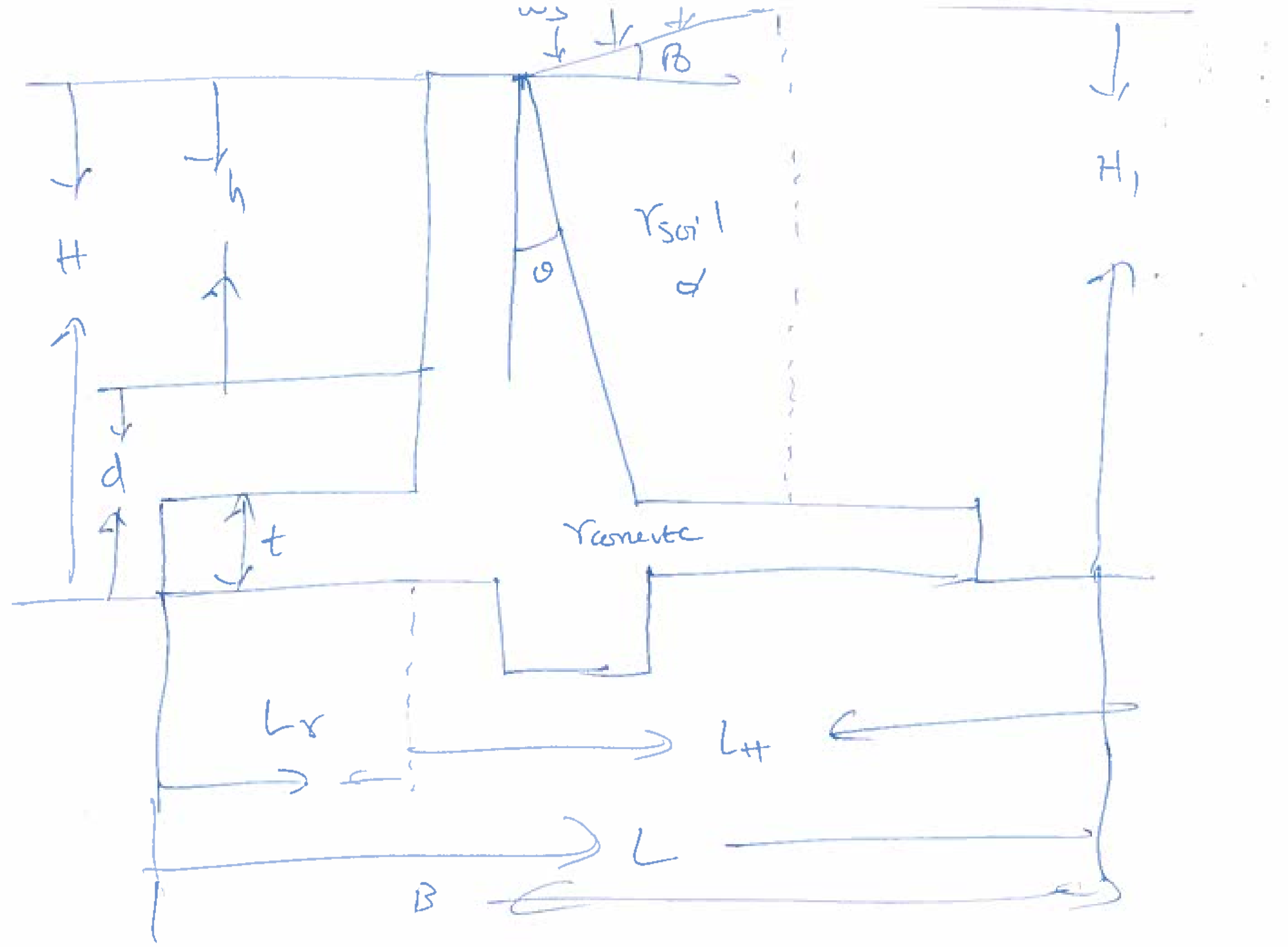
→ Angle of internal friction of soil, (ϕ°)

→ slope of back fill (β°) with horizontal

→ unit weight of soil

→ Soil – wall interface friction, δ (°)

→ concrete density in kN/m²

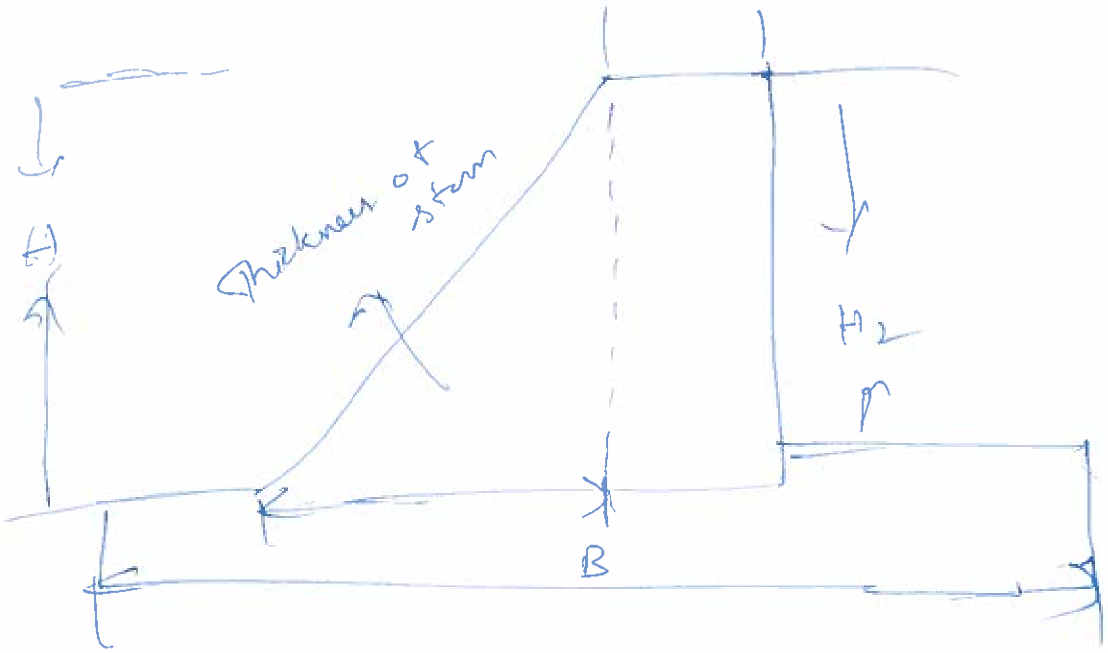


B = Base width $= 0.48H - 0.56H$

Toe projection $= 0.3b$

Thickness of base slab = Thickness of stem $= \frac{H}{12}$

width of stem at top, Min = 200 mm



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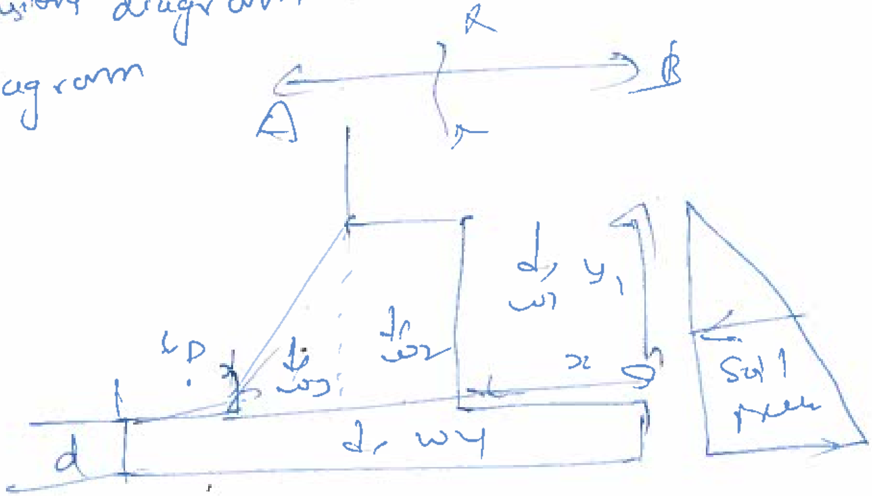
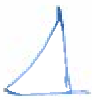
ANSWER KEY AND SCHEME OF EVALUATION

fig 1 : → General dimensions diagram

fig 2 : → Pressure diagram



Check for stability

Component	weight	x in m	M in kNm
① weight of Back fill	$w_1 = x_1 \times y_1 \times \gamma$	$A = B - \frac{x_1}{2}$	$w_1 \times D_1 = M_1$
② Rectangular portion in stem	$w_2 = (T_s - 0.2) \times y_1$	$D_2 = (T_D + D_M \ \text{of}\ w_3) - w\ 2/2$	$w_2 \times D_2 = M_2$
③ Triangular portion in stem 	$w_3 = \frac{1}{2} \times b \times h \times$ Ut of RC	$D_3 = T_D + \frac{2}{3} h$	$w_3 \times D_3 = M_3$
④ Base slab	$w_4 = B \times d \times$ Ut w of RCC	$D_4 = B/2$	$w_4 \times D_4 = M_4$

Horizontal pressure

$$P_H = \frac{1}{2} K_a \gamma H^2$$

K_a = Active earth pressure

γ = Density of earth

H = overall height of wall upto base

Overturning moment

$$M_o = P_h \frac{H}{3}$$

As per IS: 456-2000 factor of safety against overturning is

$$F_1 = \frac{0.9 \times M_S \rightarrow \text{Moment of stability}}{M_o \rightarrow \text{overturning moment}} > 1.4$$

$$F_2 = \frac{0.9\, \mu\, \Sigma W \rightarrow \text{Sum of self weight}}{P_H \rightarrow \text{Horizontal pressure}}$$

Distance of the point of application of the resultant force from the heel end

$$Z = \frac{\Sigma M}{\Sigma W}$$

Eccentricity $e = Z - b/2$

we need to check

$$e \leq \frac{b}{6}$$

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ANSWER KEY AND SCHEME OF EVALUATION

Extreme pressure intensity at G_u

$$= \frac{W}{b}\left(1 + \frac{6e}{b}\right)$$

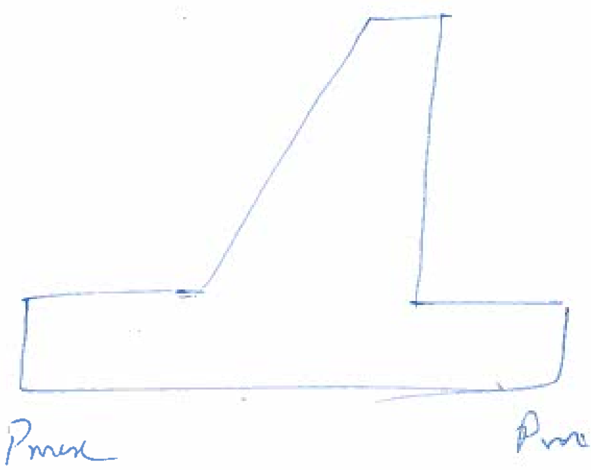
$$P_{max} = \frac{W}{b}\left(1 + \frac{6e}{b}\right)$$

$$P_{min} = \frac{W}{b}\left(1 - \frac{6e}{b}\right)$$

To get value of g & f
we need to do interpolation

P_{min}	x_1	y_1
g	x	y
P_{max}	x_2	y_2

$$\frac{x - x_1}{x_2 - x_1} = \frac{y - y_1}{y_2 - y_1}$$



Design of stem

$$\text{Max B.M } M = \frac{wh^3}{6}$$

$$\text{Ultimate moment} = M \times 1.5 = M_u$$

Effective depth = total width of the stem → 40 mm

$$\frac{M_u}{b^2} \Rightarrow b = 1000 \text{ mm as it is unit width}$$

$$\% \text{ steel} = P_t = 50 \left(\frac{1 - \sqrt{1 - \frac{4.6 M_u}{b d^2} \Big/ f_{ck}}}{f_y / f_{ck}} \right)$$

$$P_t = \frac{100 \times A_{st}}{bd}$$

$$A_{st} = \frac{P_t \times b d}{100}$$

spacing, dia & total bars

Design of Toe slab

① upward pressure due to   magnitude   distance   moment

cdj f

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ANSWER KEY AND SCHEME OF EVALUATION

8 Ans' Depth of slab $\frac{Span}{30}$ =

Given = 6.5 m x 13.5 m slab
Beam width = 230 mm
L.L = 2 kN/m², floor finish 1.5 kN/m²
M20, Fe415

Design Procedure

(i) Assume a depth of $\frac{1}{36}$th spann

(ii) Effective span shall be found as explained in Art 6.3 (clause 22.2 IS-456)

(iii) find design moment & shear force

(iv) Design for moment

(v) check for shear

(vi) Check for deflection control

(vii) Design distribution steel

(viii) Sketch reinforcement details

Depth of Slab

Assuming Span 3.5 m

$$d = \frac{Span}{30} = \frac{3500}{30} = 116.67\ m \simeq 120\ mm$$

(Cover = 30 mm)

$$\boxed{D = d + cover}$$

$\Rightarrow D = 150\ mm$

Effective Span

width of support = 230 mm

clear span = 3500 - 230 = 3270 mm

$\therefore \frac{l}{12}$ th clear span = $\frac{3270}{12}$ = 272.5 mm

Thus width of support < $\frac{l}{12}$ clr span

$\therefore$ Eff Span = 3270 + 12 = 3390 mm

= 3.39 m

STEP-2

Design Moment & shear for 1m wide slab

Self wt of Slab = $0.15 \times 1 \times 1 \times 25 = 3.75\ kN/m^2$

finishing load = $1.5\ kN/m^2$

Total DL = $1.5 + 3.75 = 4.75\ kN/m^2$

Super imposed load = $2\ kN/m^2$

Max moment occurs at support next to the end support is given by $M_{max} = \frac{w_d l^2}{10} + \frac{w_L l^2}{9}$

$$M_u = \frac{1.5 \times 4.75 \times 3.39^2}{10} + \frac{1.5 \times 2 \times 3.39^2}{9}$$

$8.18 + 3.83$ = 12.01 kN-m

Max shear force $V_u = (0.6\, w_d + 0.6\, w_L)$

$V_u = 0.6 \times 1.5 \times 4.75 \times 3.39 + 0.6 \times 1.5 \times 2 \times 3.39$

$= 20.59 \times 10^3$ kN

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ANSWER KEY AND SCHEME OF EVALUATION

Design of main Reinforcemnt:

$= 0.48 \times d$

$x_u\, lim = 0.48 \times 120 = 57.6\, mm$

$Mu\, lim = 0.36\, f_{ck}\, b\, x_u\, lim\, (d - 0.42\, x_u\, lim)$

$Mu\, lim = 0.36 \times 20 \times 1000 \times 57.6\, (120 - 0.42 \times 57.6)$

$= 39.733 \times 10^6\, N\text{-}mm$

$\therefore\ Mu < Mu\, lim$

Singly reinforced section can be designed

$$Mu = 0.87\, f_y\, A_{st}\, d \left(1 - \frac{A_{st}}{bd} \frac{f_y}{f_{ck}}\right)$$

$$\Rightarrow 12.01\, \cancel{KN\text{-}m} \times 10^6 = 0.87 \times 415 \times A_{st} \times 120 \left(1 - \frac{A_{st}}{1000 \times 120} \times \frac{415}{20}\right)$$

$$12.01 \times 10^6 = 43326\, A_{st} \left(\frac{1}{1} - \frac{A_{st}}{1.72 \times 10^{-4}}\right)$$

$$12010000 = \frac{745207200\, A_{st} - A_{st}}{1.72 \times 10^{-4}}$$

$$12.01 \times 10^6 = 0.87 \times 415 \times A_{st} \times 120 \left(1 - \frac{A_{st}}{1000 \times 120} \times \frac{415}{20}\right)$$

$$12.01 \times 10^6 = 43326 A_{st} \left(1 - A_{st}(1.72 \times 10^{-4})\right)$$

$$12.01 \times 10^6 = 43326 A_{st} - 7.452 A_{st}^2$$

$$7.452 A_{st}^2 - 43326 A_{st} + 12.01 \times 10^6 = 0$$

$$A_{st}\left(7.452 A_{st} - 4326 + 12.01 \times 10^6\right) = 0$$

$$A_{st}^2 - 5814.00 A_{st} + 1611647.88$$

$$A_{st} = 340.7 \ mm^2$$

using 10mm bars $S = \frac{\pi/4 \times 10^2}{3} \times 1000 = 230$ mm

Provide 10mm bars at 225 mm c/c

$\therefore$ Ast Provided $= \frac{\pi/4 \times 10^2}{225} \times 1000 = 349 \ mm^2$ per m width

Max spacing permitted is 3×120 or 300, which does

Shear $p_t = \frac{\pi/4 \times 10^2}{225 \times 120} \times 100 = 0.291$

Table 19 IS-456 $\tau_c = 0.38$

slab thickness < 150mm, $k_s = 1.3$ (clause 40.2.1.1)

$\tau_c = 1.3 \times 0.38$

$= 0.494 \ N/mm^2$

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ANSWER KEY AND SCHEME OF EVALUATION

$\tau_{c\,max} = 0.5 \times 2.8 = 1.4 N/mm^2$

$\tau_v < \tau_c$ & $< \tau_{c\,max}$

∴ Shear reinforcement not req

Check for deflection control

$$\frac{l}{d}\ prov = \frac{3.39 \times 1000}{120} = 28.25$$

for continuous slab ⇒ $\frac{l}{d} = 26$

$$P_t = 0.291 \quad \& \quad f_s = 0.58 \times 415 \times \frac{341.7}{349} = 235.7 N/mm^2$$

from fig 4 IS-456, $F_1 = 1.55$

∴ $\frac{l}{d}\ max = 26 \times 1.55 = 40.3$

$\frac{l}{d}\ prov < \frac{l}{d}\ max$

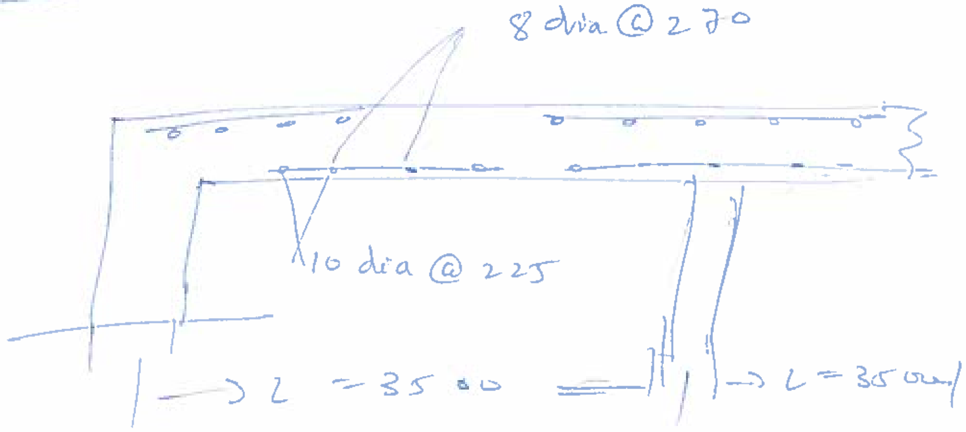
∴ Deflection control Satisfactory

Disturbuted steel $A_s = \frac{0.12 \times 1000 \times 150}{100} = 180\,mm^2$

Using 8 mm bars $S_v = \frac{\pi/4 \times 8^2}{180} \times 1000 = 279\,mm$

Provide 8 dia @ 270 mm c/c

Reinforcement detail



10 Ans!

STEP 1 Assume % of reinforcement (P)

STEP-2 Assume eff cover & calculate $\frac{d'}{D}$ & d'/B

STEP-3 Calculate $\frac{P_u}{f_{ck}\, b D}$

STEP-4 Choose graph from SP-16 find $\frac{M_{ux_1}}{f_{ck}\, b D^2}$ (chart 27-62)

STEP-5 Choose graph for SP-16 find $\frac{M_{uy_1}}{f_{ck}\, b D^2}$

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STEP-6 Calulate P_{uz}

M_{ux_1} & M_{uy_1} = max moment capacity from axial load about x & y axis

STEP-7 find $\frac{P_u}{P_{uz}}$ & $\frac{M_{ux}}{M_{ux_1}}$ & $\frac{M_{uy}}{M_{uy_1}}$

STEP-8 find $\left(\frac{M_{ux}}{M_{ux_1}}\right)$ permissible from chart-64

STEP-9 check $\left(\frac{M_{ux}}{M_{ux_1}}\right)$ permissible > $\frac{M_{ux}}{M_{ux_1}}$

STEP-10 calculate A_{sc} & Provide main & lateral reinf of Column

7 Ans Given ⇒ width of flange = 2400mm = b_f

Breadth of beam = 300mm = b_w

D = Total depth = 450mm, Thickness of slab = 130mm = D_f

d + cover

ultimate moment = 850 kN.m = Cover

M20, Fe415 materials

Eff Span = Clr span + d

= If cover 50mm then d = 400mm

(a) flange width = $b_f = \frac{l_o}{6} + b_w + 6D_f$

(b) $b_f = 0.5(L_1 + L_2) + b_w$, i.e c/c adjacent slabs

= 3.5 m = 3500mm

b_f = 2408

Design moment (Mu) & Shear force (Vu)

Self weight of Slab

weight of floor finish

live load =

total load

Self weight of rib

width of rib

Depth of rib

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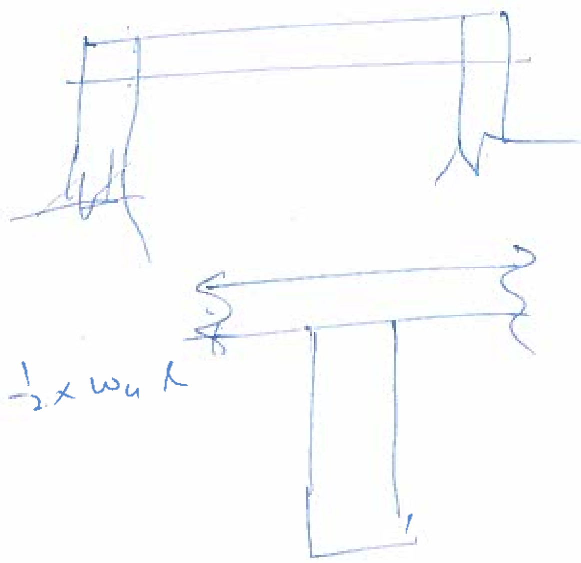
ANSWER KEY AND SCHEME OF EVALUATION



Self weight of rib

weight of plaster to rib

Total load on beam

factored load w_u

$$M_u = \frac{w_u l^2}{8} \quad \& \quad V_u = \frac{1}{2} \times w_u l$$

Design of long bars

$$x_{u\,lim} = 0.48\, d$$

$$M_{u\,lim} = 0.446 f_{ck}\, b_f\, D_f\, (d - 0.5 D_f)$$

$\therefore M_u < M_{u\,lim}$

Assuming neutral axis $x_u = D_f = \cancel{125mm}$

$$M_u' = 0.36 f_{ck}\, b_f\, D_f\, (d - 0.42 D_f)$$

$M_u < M_u'$ i.e. N-A is width in flange

Equating moment to moment of resistance

A_{st} provided

Design of shear reinforcement

find P_t, based on this τ_c

from table 19 in IS-456 $\tau_v = \frac{V_u}{bd}$

$$\tau_c < \tau_v < \tau_{c\,max}$$

Shear reinforcement designed

$$V_{us} = V_u - \tau_c b d$$

$$V_{us} = \frac{0.87 f_y A_{sv} d}{S_v}$$

Check for deflection

$$\frac{L}{d} = 20, \quad P_t = !$$

A_{st} req, A_{st} Prov

$f_s =$

from fig 4, IS-456, $F_1 = 1.0$

find $\frac{b_w}{b_f}$

from fig-6 IS-456 F_3

$\frac{L}{d}$ max $= F_1 F_2 F_3$ basic value

(M. Sai Kiran)

Sig: HoD (Civil)

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Mechanical Engineering	Academic Year	2022 - 2023
Course Code	20ME502	Test Duration	3 Hrs. | Max. Marks 70	Semester	V
Course	Design of Machine Elements-I				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the various phases of design?	20ME502.1	L1
2	What is meant by fatigue stress concentration factor?	20ME502.2	L1
3	List any 4 advantages of bolted joints over welded joints.	20ME502.3	L1
4	What is the use of universal coupling?	20ME502.4	L1
5	Recall the stresses in Helical Springs of circular wire.	20ME502.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw the stress strain diagrams of Ductile materials and Brittle materials.	6M	20ME502.1	L2
6 (b)	A steel saw blade 1 mm thick is bent into an arc of a circle of 50 cm radius. Determine the flexural stresses induced and the bending moment required bending the blade which is 15 mm wide. Take E = 2.1X10^5 N/mm^2.	6M	20ME502.1	L3
	OR			
7 (a)	Explain the manufacturing considerations in design.	6M	20ME502.1	L2
7 (b)	A cast iron pulley transmits 10 KW at 400 rpm. The diameter of the pulley is 1.2meter and it has four straight arms of elliptical cross section. In which the major axis is twice the minor axis. Determine the dimensions of the arm if the allowable bending stress is 15 MPa.	6M	20ME502.1	L3
8 (a)	Define fatigue, endurance limit, stress concentration and notch sensitivity.	6M	20ME502.2	L1
8 (b)	Determine the thickness of a 120 mm wide a uniform plate for safe continuous operation if the plate is to be subjected to a tensile load that has a maximum value of 250 kN and a minimum value of 100 kN. The properties of the plate material are as follows: Endurance limit stress 225 MPa, and yield point stress 300 MPa. The factor of safety based on yield point may be as 1.5	6M	20ME502.2	L3
	OR			
9 (a)	Explain the factors that affect the fatigue strength.	6M	20ME502.2	L2
9 (b)	A simply supported beam has a concentrated load at the center, which fluctuates from a value of P to 4 P. The span of the beam is 0.5 m and its cross-section is circular with a diameter of 0.06 m. Taking for the beam material an ultimate stress of 700 MPa, a yield stress of 500 MPa, endurance limit of 330 MPa for reversed bending, and a factor of safety of 1.3, calculate the maximum value of P. Take a size factor of 0.85 and a surface finish factor of 0.9.	6M	20ME502.2	L3
10 (a)	What do you understand by the terms riveted joint and welded joints?	6M	20ME502.3	L1
10 (b)	Two plates 16 mm thick are joined by a double riveted lap joint. The pitch of each row of rivets is 90 mm. The rivets are 25 mm in diameter. The permissible stresses are 140 MPa in tension, 80 MPa in shear and 160 MPa in crushing. Find the efficiency of the joint.	6M	20ME502.3	L3
	OR			

11 (a)	List any 6 applications of sleeve and cotter joints.	6M	20ME502.3	L1
11 (b)	Design a sleeve and cotter joint to resist a tensile load of 60 Km all parts of the joints are made of the same material with the following allowable stresses: σ_t = 60 MPa, τ = 70 MPa, σ_c = 125 MPa.	6M	20ME502.3	L3
12 (a)	Write about the uses of internal and external circlips with neat sketch.	6M	20ME502.4	L2
12 (b)	Compare weight, strength and stiffness of two shafts of same material subjected to same torque. One being solid, other being hollow with inner diameter to outer diameter ratio 0.5.	6M	20ME502.4	L2
	OR			
13	Write about the rigid couplings and its type with neat diagram.	12M	20ME502.4	L2
14 (a)	Explain about the co-axial springs.	4M	20ME502.5	L2
14 (b)	A helical spring is made from a wire of 6 mm diameter and has outside diameter of 75 mm. If the permissible shear stress is 350 MPa and modulus of rigidity 84 Kn/MM^2, find the axial load which the spring can carry and the deflection per active turn.	8M	20ME502.5	L3
	OR			
15 (a)	Define spring. What is the purpose of mechanical springs?	6M	20ME502.5	L1
15 (b)	A truck spring has 12 number of leaves, two of which are full length leaves. The spring supports are 1.05 m apart and the central band is 85 mm wide. The central load is to be 5.4 KN with a permissible stress of 280 MPa. Determine the thickness and width of the steel spring leaves. The ratio of the total depth to the width of the spring is 3. Also determine the deflection of the spring.	6M	20ME502.5	L3

NSRIT

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Design of Machine Elements-I ~~Part-A~~ Mechanical Engineering

B.Tech. - 20ME502 : course code Academic Year - 2022-2023

PART-A

1. Various phases of design – 2M
2. Definition fatigue stress concentration factor – 2M
3. 4 Advantages of bolted joints over welded joints – 2M
4. Use of universal coupling - 2M
5. Stresses in helical springs of circular wire – 2M

PART-B

6 a) Diagram - Stress strain for Ductile materials and Brittle materials – 6M

b) Given data – 1M, Formula – 2M, Substitution Calculation – 3M

or

7 (a) Manufacturing consideration in design – 6M

(b) Given data - 1M, Formula – 2M, Substitution & calculation – 3M

8(a) Definition for fatigue – 1.5M, endurance limit – 1.5M, stress concentration – 1.5M, notch sensitivity – 1.5M

(b) Given data – 1M, Formula – 2M, Substitution & calculation – 3M

or

9(a) Factors that affect the fatigue strength – 6M.

(b) Given data – 1M, Formula – 2M, Substitution & calculation – 3M

10 (a) Riveted joint – 3M, welded joints – 3M

b) Given data – 1M, Formula – 2M, Substitution & calculation – 3M

11 a) 6 applications of Sleeve and cotter joints – 6 M

b) Given data – 1 M, formula – 2 M, substitution & calculation – 3 M

12 a) Uses of internal circlips – diagram – 3 M

Uses of External circlips – diagram – 3 M

b) Compare weight – 2 M, Strength – 2 M, Stiffness – 2 M

or

13. Rigid couplings and its type with neat diagram.

14 a) Co-axial springs with diagram – 6 M

b) Given data – 1 M, formula – 2 M, substitution & Calculation – 3 M

or

15 a) Definition spring – 3 M, Purpose of mechanical springs – 3 M

b) Given data – 1 M, formula – 2 M, substitution & calculation – 3 M

PART-A (Short Answer questions).

1. what are the various phases of design?

1. Recognition of need. 2. Synthesis (Mechanisms)
3. Analysis of forces 4. Material selection
5. Design of elements (Size and stresses) 6. Modification.
7. Detailed drawing 8. Production.

2. what is meant by fatigue stress concentration factor?

$$K_f = \frac{\text{Endurance limit without stress concentration}}{\text{Endurance limit with stress concentration}}$$

when a machine member is subjected to cyclic or fatigue loading the value of fatigue stress concentration factor shall be applied instead of theoretical stress concentration factor.

3. List any 4 advantages of bolted joints over welded joints!

1. Simple design 2. Easy operation 3. Low cost
4. Less skilled or Semi skilled Labour required 5. Availability.
6. Replacement is easy 7. No risk. 8. Less time required
9. Non permanent joint 10. Less man power required.

4. what is the use of universal coupling.

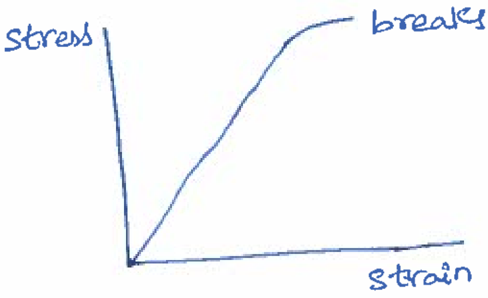
A universal or Hooke's coupling is used to connect two shafts whose axes intersect at a small angle. The main application of the universal or Hooke's coupling is found in the transmission from the gear box to the differential or back axle of the automobiles. A Hooke's coupling is also used for transmission of power to different spindles of multiple drilling machine. It is used as a knee joint in milling machine.

5. Recall the stresses in helical springs of circular wire.

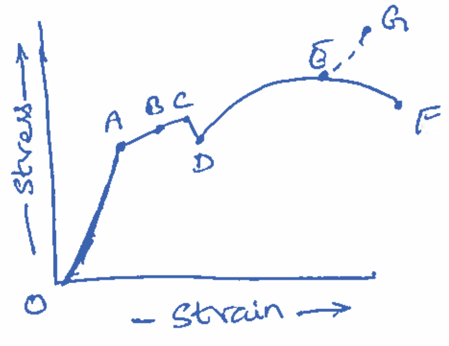
1. Torsional shear stress.
2. Direct shear stress.
3. Stress due to curvature of wire

PART-B (Long Answer Questions 5X12 = 60 marks)

6a) Draw the stress strain diagram of ductile materials and Brittle materials.



Brittle material



Ductile material

The various properties of the material are

1. proportional limit (from point O to A)
2. Elastic limit (up to the point B)
3. Yield point (the points C and D are called upper and lower Yield points respectively)
4. Ultimate stress (point E)
5. Breaking stress (point F)

6 (b) A steel saw blade 1mm thick is bent into an arc of a circle of 50 cm radius. Determine the flexural stresses induced and the bending moment required bending the blade which is 15mm wide. Take $E = 2.1 \times 10^5 N/mm^2$.

$b = 15mm$ wide

thickness $(t) = 1cm$, radius $(r) = 50cm$, By using bending equation

$E = 2.1 \times 10^5 N/mm^2$,

$$\frac{M}{I} = \frac{\sigma}{Y} = \frac{E}{R}, \quad \frac{\sigma}{Y} = \frac{E}{R}, \quad \frac{t}{2} = \frac{1}{2} = 0.5mm$$

$$\sigma = \frac{E}{R} Y = \frac{2.1 \times 10^5 \times 0.5}{200} = 210 N/mm^2$$

$$\frac{M}{I} = \frac{\sigma}{Y}, \quad M = \frac{\sigma}{Y} \times I = \frac{210 \times 1.24}{0.5} = 525 N\text{-}mm$$

$M = 525$ N-mm

$$I = \frac{bd^3}{12} = \frac{15 \times (1)^3}{12} = 1.24 mm^4.$$

(OR)

7(a) Explain the manufacturing considerations in design.

1. Minimum total number of parts in a product
2. Minimum variety of parts
3. Use standard parts
4. Use modular design
5. Design parts to be multifunctional
6. Design parts for multiple use
7. Select least costly material
8. Design parts for ease of manufacture
9. Shape the parts for minimizing the operations.

7(b) A cast iron pulley transmits 10 KW at 400 rpm. The diameter of the pulley is 1.2 meter and it has four straight arms of elliptical cross section. In which the major axis is twice the minor axis, Determine the dimensions of the arm if the allowable bending stress is 15MPa.

Given $P = 10\,KW = 10\times10^3\,W$, $N = 400\,rpm$, $D = 1.2\,m = 1200\,mm$ or $R = 600\,mm$, $\sigma_b = 15\,MPa = 15\,N/mm^2$

Let T = Torque transmitted by the pulley, we know that the power transmitted by the pulley $P = \frac{2\pi NT}{60}$

$$10\times10^3 = \frac{2\pi NT}{60} = \frac{2\pi\times400\times T}{60} = 42T$$

$$T = \frac{10\times10^3}{42} = 238\,N\text{-}m = 238\times10^3\,N\text{-}mm$$

∴ Tangential load acting on the pulley $= \frac{T}{R} = \frac{238\times10^3}{600} = 396.7\,N$

∴ Tangential load on each arm $w = \frac{396.7}{4} = 99.2\,N$

maximum bending moment on the arm $M = w\times R = 99.2\times600$

$M = 59520\,N\text{-}mm$

Let $2b$ = Minor axis in mm and $2a$ = Major axis in mm $= 2\times2b = 4b$

∴ Section modulus for an elliptical cross-section

$$Z = \frac{\pi}{4}\times a^2 b = \frac{\pi}{4}(2b)^2\times b = \pi b^3\,mm^3.$$

We know that bending stress $(\sigma_b) = \frac{M}{Z}$

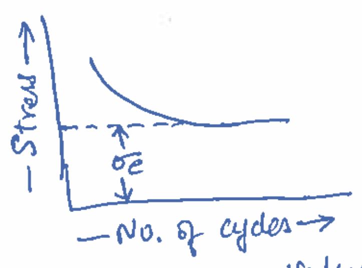
$$15 = \frac{M}{Z} = \frac{59520}{\pi b^3} = \frac{18943}{b^3}, \quad b^3 = 1263 \;\&\; b = 10.8\,mm$$

∴ Minor axis $2b = 2\times10.8\,mm = 21.6\,mm$

and major axis $2a = 2\times2b = 4\times10.8 = 43.2\,mm$.

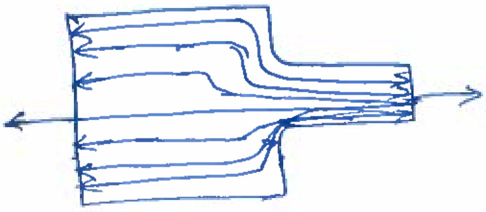
8 a) Define fatigue, endurance limit, stress concentration and notch sensitivity.

When a material is subjected to repeated stresses, it fails at stresses below the yield point stresses. Such type of failure of a material is known as fatigue.



A little consideration will show that if the stress is kept below a certain value as shown in dotted line in Fig, the material will not fail whatever may be the number of cycles. This stress as represented by dotted line is known as endurance or fatigue limit (σ_e). It is defined as maximum value of the completely reversed bending stress which a polished standard specimen can withstand without failure for infinite number of cycles (usually 10^7 cycles).

Stress concentration : whenever a machine component changes the shape of its cross-section, the simple stress distribution no longer holds good and the neighbourhood of the discontinuity is different. This irregularity in the stress distribution caused by abrupt changes of form is called stress concentration.



Notch sensitivity: It may defined as the degree to which the theoretical effect of stress concentration is actually reached. The stress gradient depends mainly on the radius of the notch, hole or fillet and on the grain size of the material.

$$q = \frac{K_f - 1}{K_t - 1}$$

& $K_f = 1 + q(K_t - 1)$ for tensile or bending stress

$K_{fs} = 1 + q(K_{ts} - 1)$ for shear stress.

K_f = fatigue stress concentration factor

K_t = Theoretical or form stress concentration factor for axial or bending loading,

K_{ts} = theoretical stress concentration factor for torsional or shear loading.

8b) Determine the thickness of a 125mm wide a uniform plate for safe continuous operation. If the plate is to be subjected to a tensile load that has a maximum value of 250KN and a minimum value of 100KN. The properties of the plate material are as follows! Endurance limit stress 225MPa, and Yield point stress 300 MPa. The factor of safety based on yield point may be as 1.5

Given $b = 120$ mm, $W_{max} = 250$KN, $W_{min} = 100$KN, $\sigma_e = 225$MPa $= 225$N/mm², $\sigma_y = 300$MPa $= 300$ N/mm², $FS = 1.5$

Let t = thickness of the plate in mm.

$\therefore$ Area $A = b \times t = 120t$ mm².

We know that mean or Average load $W_m = \frac{W_{max} + W_{min}}{2}$

$$W_m = \frac{250 + 100}{2} = 175\text{KN} = 175 \times 10^3 \text{N}$$

$$\therefore \text{Mean stress } \sigma_m = \frac{W_m}{A} = \frac{175 \times 10^3}{120t} \text{ N/mm}^2.$$

$$\text{Variable Load } W_v = \frac{W_{max} - W_{min}}{2} = \frac{250 - 100}{2} = 75\text{KN} = 75 \times 10^3 \text{N}$$

$$\therefore \text{Variable stress } \sigma_v = \frac{W_v}{A} = \frac{75 \times 10^3}{120t} \text{ N/mm}^2$$

According to Soderberg's formula

$$\frac{1}{FS} = \frac{\sigma_m}{\sigma_y} + \frac{\sigma_v}{\sigma_e}, \quad \frac{1}{1.5} = \frac{175 \times 10^3}{120t \times 300} + \frac{75 \times 10^3}{120t \times 225}$$

$$\frac{4.86}{t} + \frac{2.78}{t} = \frac{7.64}{t}, \quad t = 7.64 \times 1.5 = 11.46 \text{ say } 11.5\text{mm}$$

ok

9 (a) Explain the factors that affect the fatigue strength.

1. Effect of stress concentration 2. The influence of size factor. 3. The influence of surface processing state 4. Influence on loading experience 5. the effect of chemical composition 6. Effect of heat treatment and microstructure 7) The effect of inclusions 8. Influence of surface properties and residual stress.

~~9. Effect of ...~~

9 (b) A simply supported beam has a concentrated load at the center, which fluctuates from a value of P to 4P. The span of the beam is 0.5m and its cross-section is circular with a diameter of 0.06 m. Taking for the beam material an ultimate stress of 700MPa, a yield stress of 500MPa, endurance limit of 330MPa for reversed bending, and a factor of safety of 1.3, calculate the maximum value of P. Take a size factor of 0.85 and a surface finish factor of 0.9

Given $W_{min} = P$, $W_{max} = 4P$, $L = 500\,mm$, $d = 60\,mm$, $\sigma_u = 700\,MPa = 700\,N/mm^2$, $\sigma_y = 500\,MPa = 500\,N/mm^2$, $\sigma_e = 330\,MPa = 330\,N/mm^2$,

$F.S = 1.3$, $K_{sz} = 0.85$, $K_{sur} = 0.9$

we know that maximum bending moment

$$M_{max} = \frac{W_{max} \times L}{4} = \frac{4P \times 500}{4} = 500P\ N\text{-}mm$$

minimum bending moment $M_{min} = \frac{W_{min} \times L}{4} = \frac{P \times 500}{4} = 125P\ N\text{-}mm$

∴ mean or Average bending moment

$$M_m = \frac{M_{max} + M_{min}}{2} = \frac{500P + 125P}{2} = 312.5P\ N\text{-}mm$$

and variable bending moment $M_v = \frac{M_{max} - M_{min}}{2} = \frac{500P - 125P}{2} = 187.5P\ N\text{-}mm$

Section Modulus $Z = \frac{\pi}{32} d^3 = \frac{\pi}{32}(60)^3 = 21.21 \times 10^3\ mm^3$

∴ Mean bending stress

$$\sigma_m = \frac{M_m}{Z} = \frac{312.5P}{21.21 \times 10^3} = 0.0147P\ N/mm^2$$

and variable bending stress

$$\sigma_v = \frac{M_v}{Z} = \frac{187.5P}{21.21 \times 10^3} = 0.0088P\ N/mm^2$$

we know that according to Goodman's formula

$$\frac{1}{FS} = \frac{\sigma_m}{\sigma_u} + \frac{\sigma_v \times K_f}{\sigma_e \times K_{sur} \times K_{sz}}$$

(Taking $K_f = 1$)

$$\frac{1}{1.3} = \frac{0.0147P}{700} + \frac{0.0088P \times 1}{330 \times 0.9 \times 0.85}$$

$$= \frac{21P}{10^6} + \frac{34.8P}{10^6} = \frac{55.8P}{10^6}$$

$$P = \frac{1}{1.3} \times \frac{10^6}{55.8} = 13785\ N = 13.785\ kN$$

and according to Soderberg's formula.

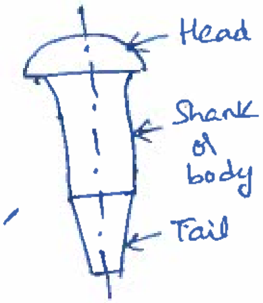
$$\frac{1}{F.S} = \frac{\sigma_m}{\sigma_y} + \frac{\sigma_v \times K_f}{\sigma_e \times K_{sur} \times K_{sz}}$$

$$\frac{1}{1.3} = \frac{0.0167P}{500} + \frac{0.0088P \times 1}{330 \times 0.9 \times 0.85} = \frac{29.4P}{10^6} + \frac{34.8P}{10^6} = \frac{64.2P}{10^6}$$

$$P = \frac{1}{1.3} \times \frac{10^6}{64.2} = 11982N = 11.982kN$$

From the above we find that maximum value of $P = 13.785kN$.

10 (a) what do you understand by the terms riveted joint and welded joints?



A rivet is a short cylindrical bar with a head integral to it. The cylindrical portion of the rivet is called shank or body and lower portion of shank is known as tail. the rivets are used to make permanent fastening between the plates such as structural work, ship building, bridges, tanks and boiler shells. the riveted joints are widely used for joining light metals.

The fastening (joints) may be classified into the following groups:

1. Permanent fastenings :- The permanent fastenings are those fastenings which can not be disassembled without destroying the connecting components. The examples of permanent fastenings in order of strength are soldered, brazed, welded and riveted joints.
2. Temporary or detachable fastenings :- the temporary or detachable fastenings are those fastenings which can be ~~disasse~~ disassembled without destroying the connecting components. The examples of temporary fastenings are screwed, keys, cotters, pins and splined joints.

A welded joint is a permanent joint which is obtained by the fusion of the edges of the two parts to be joined together, with or without the application of pressure and a filler material. The heat required for the fusion of the material may be obtained by burning of gas (in case of gas welding). or by an electric arc (in case of electric arc welding). The latter method is extensively used because of greater speed of welding.

welding is extensively used in fabrication as an alternative method for casting or forging and as a replacement for bolted and riveted joints. It is also used as a repair medium. e.g. to reunite metal of a crack. to build up a small part that has broken off such as gear tooth or to repair a worn surface such as a bearing surface.

10 (b). Two plates 16 mm thick are joined by a double riveted lap joint. The pitch of each row of rivets is 90 mm. The rivets are 25 mm in diameter. The permissible stresses are 140 MPa in tension, 80 MPa in shear and 160 MPa in crushing. Find the efficiency of the joint.

Given $t = 16$ mm, $\sigma_t = 140$ MPa $= 140$ N/mm², $\tau = 80$ MPa $= 80$ N/mm², $\sigma_c = 160$ MPa $= 160$ N/mm². $d = 25$ mm. pitch $p = 90$ mm.

i) Tearing resistance of the plate

$$P_t = (p-d)t \times \sigma_t = (90-25)16 \times 140 = 145600\ N$$

(ii) Shearing resistance of the rivet

$$P_s = n \times \frac{\pi}{4} \times d^2 \times \tau = 2 \times \frac{\pi}{4} \times 25^2 \times 80 = 78539.816\ N$$

(iii) Crushing resistance of the rivet

$$P_c = n \times d \times t \times \sigma_c = 2 \times 25 \times 16 \times 160 = 128000\ N$$

∴ strength of the joint = Least of P_t, P_s and P_c = ~~78539.816~~ 78539.816 N

we know that the strength of the unriveted & solid plate

$$P = p \times t \times \sigma_t = 90 \times 16 \times 140 = 201600\ N$$

∴ efficiency of the joint $\eta = \frac{\text{Least of } P_t, P_s \text{ and } P_c}{P} = \frac{78539.816}{201600}$

$$\eta = 38.95\%$$

11 a) List any 6 applications of sleeve and cotter joints

1. Cotter joint is used to connect two rods subjected to axial tensile or compressive loads. Cotter joint is widely used to connect the piston rod and crosshead of the steam engine.
2. It is used in a bicycle to connect the paddle to the sprocket wheel.
3. Use a wet air pump to join a tail rod with the piston rod.
4. Arrangement of cotter and dowel to join two parts of a flywheel.
5. Cotter joints are used between the slide spindles and the fork of the valve mechanism
6. It is used to connect two rods of equal diameter subjected to axial forces

b) Design a sleeve and cotter joint to resist a tensile load of 60kN all parts of the joints are made of the same material with the following allowable stresses $\sigma_t = 60$ MPa, $\tau = 70$ MPa, $\sigma_c = 125$ MPa

Given $P = 60$ kN $= 60 \times 10^3$ N, $\sigma_t = 60$ MPa $= 60$ N/mm², $\tau = 70$ MPa $= 70$ N/mm², $\sigma_c = 125$ MPa $= 125$ N/mm²

1. Diameter of the rods

Let d = Diameter of the rods

Considering the failure of the rods in tension

$P = \frac{\pi}{4} \times d^2 \times \sigma_t$, $60 \times 10^3 = \frac{\pi}{4} \times d^2 \times 60 = 47.13 d^2$

$d = 60 \times 10^3 / 47.13 = 1273$ or $d = 35.7$ say 36 mm

2. Diameter of enlarged end of rod and thickness of cotter

Let d_2 = Diameter of enlarged end of rod and

t = thickness of cotter, It may be taken as $d_2/4$.

Considering the failure of the rod in tension across the weakest section (ie slot). We know that load P

$60 \times 10^3 = \left[\frac{\pi}{4}(d_2)^2 - d_2 \times t\right]\sigma_t = \left[\frac{\pi}{4}(d_2)^2 - d_2 \times \frac{d_2}{4}\right] 60$

$= 32.13 (d_2)^2$, $\therefore d_2^2 = \frac{60 \times 10^3}{32.13} = 1867$.

$d_2 = 43.2$ Say 44 mm, $t = \frac{44}{4} = \frac{d_2}{4} = 11$ mm.

Let us now check the induced crushing stress in the rod or cotter. We know that load (P)

$60 \times 10^3 = d_2 \times t \times \sigma_c = 44 \times 11 \times \sigma_c = 484 \sigma_c$

$\therefore \sigma_c = \frac{60 \times 10^3}{484} = 124$ N/mm²

Since the induced crushing stress is less than the given value of 125 N/mm², therefore the dimensions of d_2 and t are within safe limits.

3. Outside diameter of sleeve.

Let d_1 = outside diameter of sleeve

Considering the failure of sleeve in tension across the slot.

We know that load (P)

$60 \times 10^3 = \left[\frac{\pi}{4}\left[(d_1)^2 - (d_2)^2\right] - (d_1 - d_2)t\right]\sigma_t$

$= \left[\frac{\pi}{4}\left[(d_1)^2 - (44)^2\right] - (d_1 - 44)11\right] 60$

$\therefore 60 \times 10^3 / 60 = 0.7854 (d_1)^2 - 1520.7 - 11 d_1 + 484$

or $(d_1)^2 - 14 d_1 - 2593 = 0$

$\therefore d_1 = \frac{14 \pm \sqrt{(14)^2 + 4 \times 2593}}{2} = \frac{14 \pm 102.8}{2}$

$= 58.4$ Say 60 mm.

4. Width of cotter

Let b = Width of cotter.

Considering the failure of cotter in shear. Since the cotter is in double shear, therefore load (P)

$60 \times 10^3 = 2b \times t \times \tau = 2 \times b \times 11 \times 70 = 1540b$

$b = 60 \times 10^3 / 1540 = 38.96$ Say 40mm.

5. Diameter of the rod from the beginning to the cotter hole. (Inside the sleeve end)

let a = Required distance.

Considering the failure of the rod end in shear. Since the rod end is in double shear, therefore load(P)

$60 \times 10^3 = 2a \times d_2 \times \tau = 2a \times 44 \times 70 = 6160a$

$a = \frac{60 \times 10^3}{6160} = 9.74$ Say 10 mm

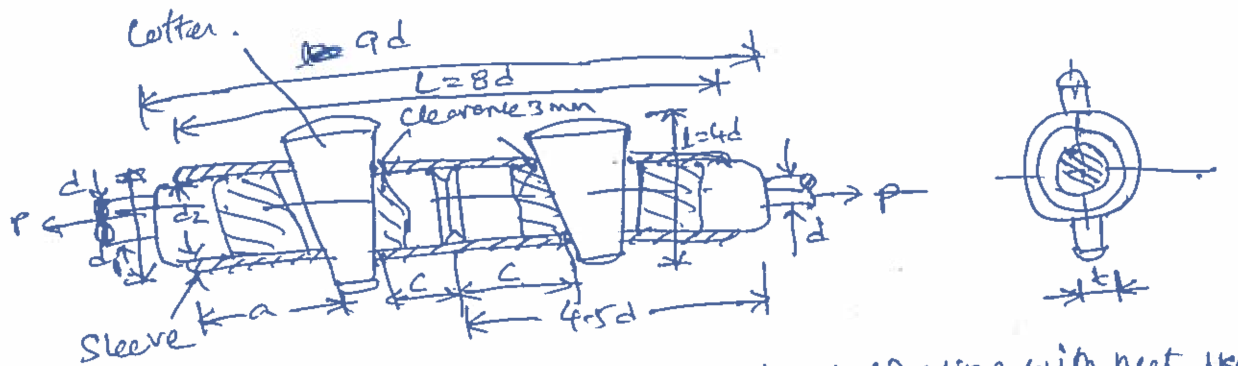
6. Diameter of the rod end from its end to the cotter hole

let c = Required distance.

Considering the failure of the sleeve end in shear. therefore load(P) Since the sleeve end is in double shear

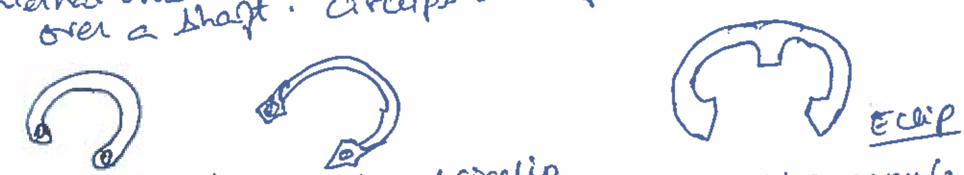
$60 \times 10^3 = 2(d_1 - d_2) c \times \tau = 2(60-44) c \times 70 = 2240c$

$c = \frac{60 \times 10^3}{2240} = 26.78$ Say 28 mm



12(a) Write about the uses of internal and external circlips with neat sketch.

A circlip also known as a C-clip, Rotor clip, snap ring or Jesus clip. is a type of fastener or retaining ring consisting of a semi-flexible metal ring with open ends which can be snapped into place into a machined groove on a dowel pin or other part to permit rotation but to prevent axial movement. There are two basic types: Internal and external, referring to whether they are fitted into a bore or over a shaft. Circlips are often used to secure pinned connection



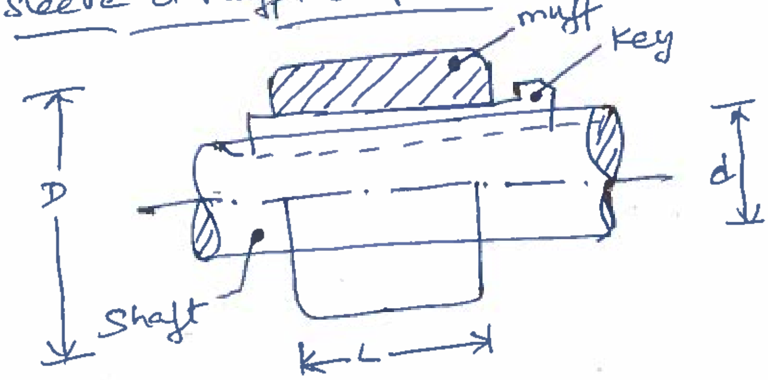
Internal Circlip   External Circlip   E clip

Internal circlips are used in heavy machine manufacturing, mining, railways, construction, heavy machine manufacturing, turbine ~~manufacturing~~ manufacturing, energy sector and port, harbor and ship repairing. Circlips are generally used in motors, pistons and turbines, clutches. Internal circlips are used for bores while external circlips are used for shafts.

13. Write about the rigid couplings and its type with neat diagram.

1. Rigid coupling: It is used to connect two shafts which are perfectly aligned. Following types of rigid coupling are important from the subject point of view:
   a) Sleeve or muff ~~case~~ coupling
   b) Clamp or split-muff or compressive coupling
   c) Flange coupling.
2. Flexible coupling: It is used to connect two shafts having both lateral and angular misalignment. Following types of flexible couplings are important from the subject point of view.
   a) Bushed pin type coupling. b) Universal coupling c) Oldham coupling.

a) Sleeve or muff-coupling



$D = 2d + 13\,mm$, $L = 3.5d$

$T = \frac{\pi}{16} \times \tau_c \left(\frac{D^4 - d^4}{D}\right)$

$= \frac{\pi}{16} \times \tau_c \times D^3 (1 - k^4)$

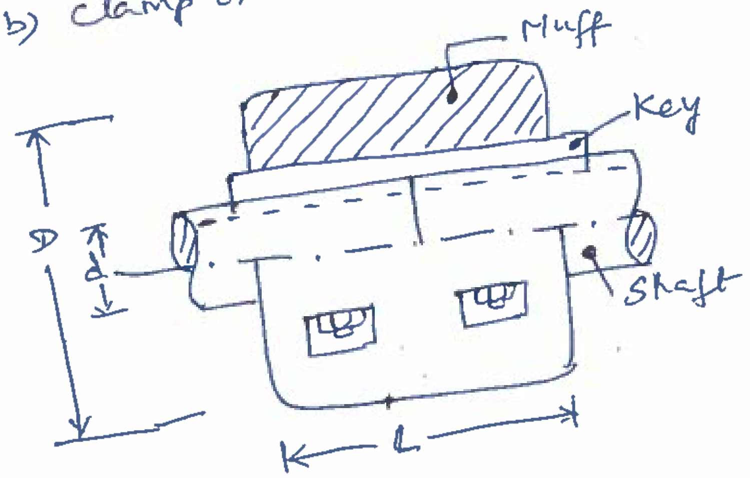
$k = d/D$

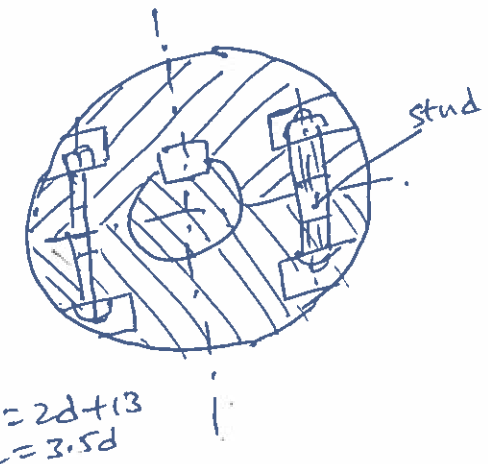
$l = \frac{L}{2} = \frac{3.5d}{2}$

$T = l \times w \times \tau \times \frac{d}{2}$ — shearing of key

$= l \times \frac{t}{2} \times \sigma_c \times d/2$ — crushing of key

b) Clamp or compression coupling.

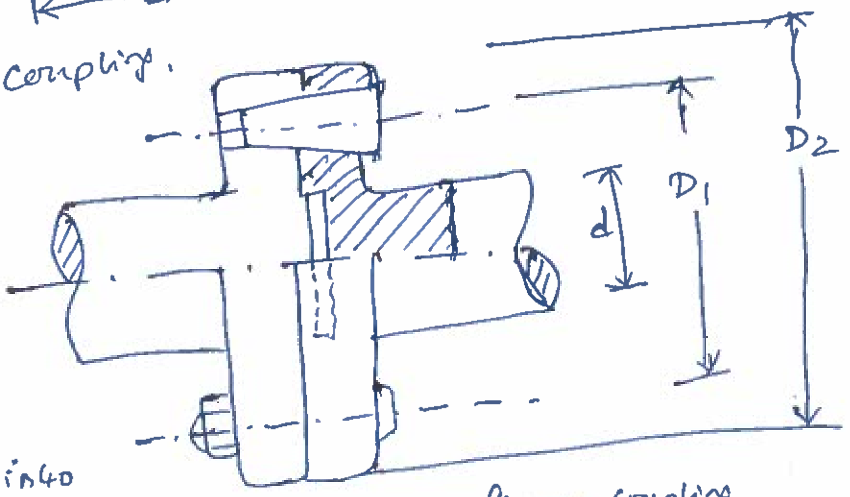




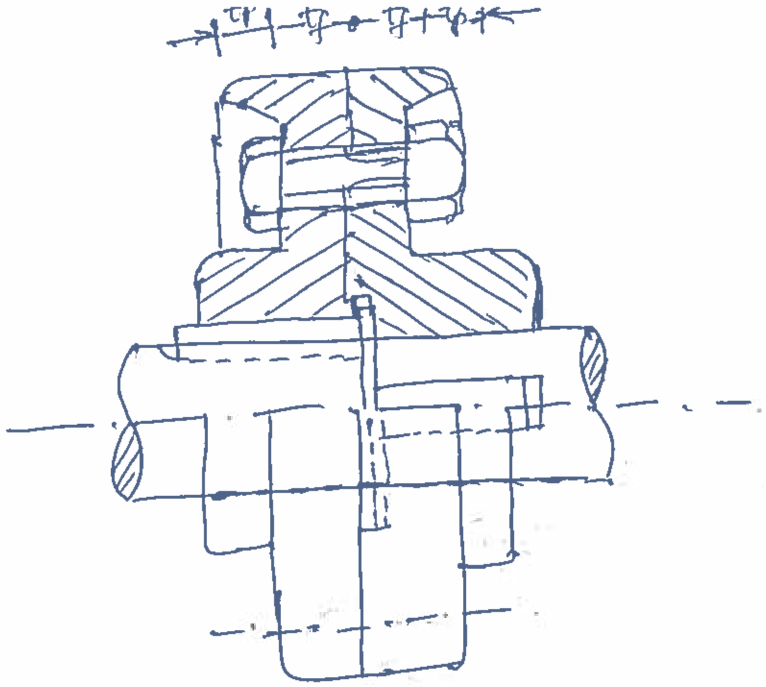
$D = 2d + 13$
$L = 3.5d$

c) Flange coupling.

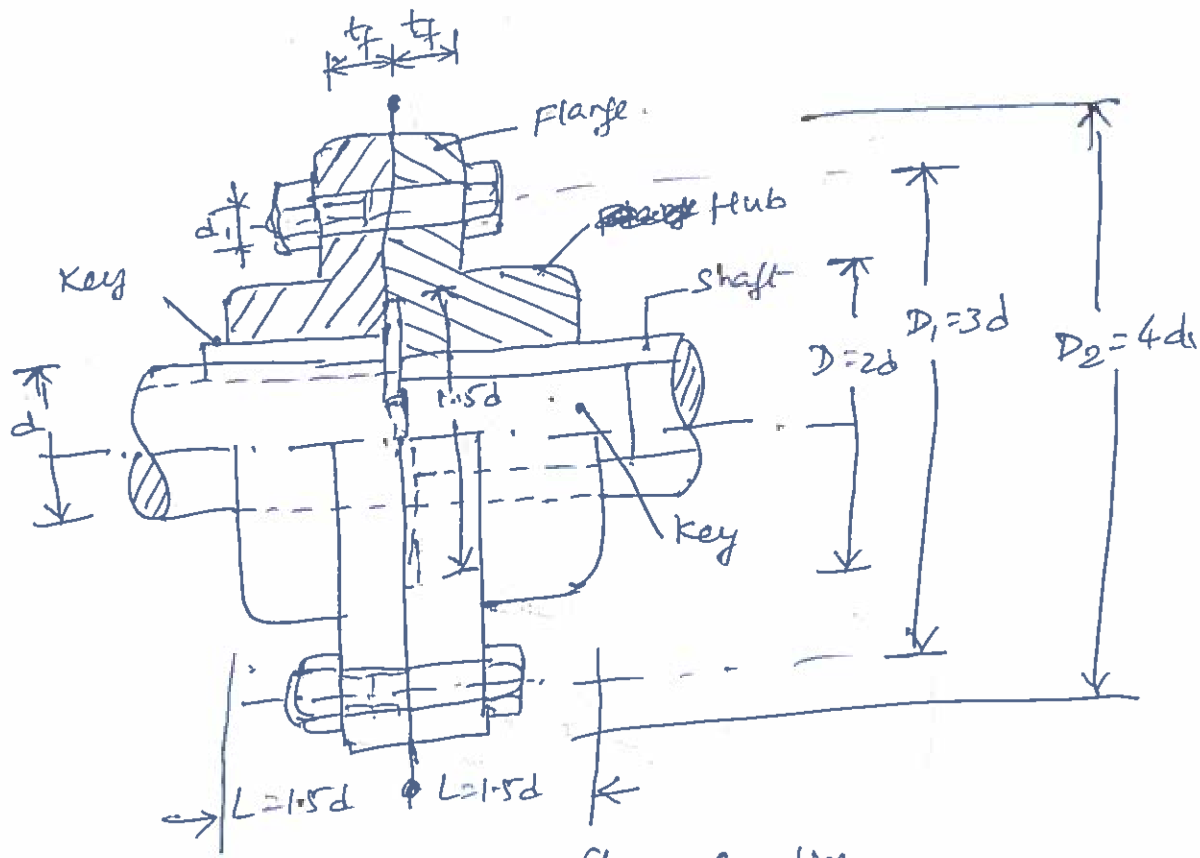
$D_1 = 1.6d$
$D_2 = 2.2d$
Thickness of flange $= d/3$
Taper of bolt $=$ 1 in 20 to 1 in 40



Marine type flange coupling.



Protective type flange coupling.



Unprotected type flange coupling.

outside diameter of hub $D = 2d$

Length of hub $L = 1.5d$

pitch circle diameter of bolts $D_1 = 3d$

outside diameter of flange $D_2 = D_1 + (D_1 - d) = 2D_1 - D = 4d$

thickness of flange $t_f = 0.5d$

Number of bolts $= 3$ for d upto 40 mm

$= 4$ for d upto 100 mm

$= 6$ for d upto 180 mm

12(b) Compare weight, strength and stiffness of two shafts of same material subjected to same torque. One being solid, other being hollow with inner diameter to outer diameter ratio 0.5

Given $d_o = d$, $d_i = d_o/2$ or $K = d_i/d_o = 1/2 = 0.5$

Comparison of weight

we know that weight of a hollow shaft

we know that weight = Cross-sectional area × Length × Density.

W_H = Cross-sectional area × Length × Density. → ①

$= \frac{\pi}{4}[(d_o)^2 - (d_i)^2] \times$ Length × Density

and weight of the solid shaft

$W_s = \frac{\pi}{4} \times d^2 \times$ Length × Density. → ②

Since both the shafts have the same material and length, therefore by dividing equation ① by equation ② we get

$$\frac{W_H}{W_S} = \frac{(d_o)^2 - (d_i)^2}{d^2} = \frac{(d_o)^2 - (d_i)^2}{(d_o)^2} \quad (\because d = d_o)$$

$$= 1 - \frac{(d_i)^2}{(d_o)^2} = 1 - K^2 = 1 - (0.5)^2 = 0.75$$

Comparison of strength.

we know that strength of the hollow shaft

$$T_H = \frac{\pi}{16} \times \tau \times (d_o)^3 (1 - K^4) \rightarrow ③$$

and strength of the solid shaft

$$T_s = \frac{\pi}{16} \times \tau \times d^3 \rightarrow ④$$

∴ Dividing equation ③ by equation ④ we get.

$$\frac{T_H}{T_S} = \frac{(d_o)^3(1-K^4)}{d^3} = \frac{(d_o)^3(1-K^4)}{(d_o)^3} = 1 - K^4$$

$$= 1 - (0.5)^4 = 0.9375$$

Comparison of stiffness.

we know that stiffness $= \frac{T}{\theta} = \frac{G \times J}{L}$

∴ Stiffness of a hollow shaft $S_H = \frac{G}{L} \times \frac{\pi}{32}[(d_o)^4 - (d_i)^4] \rightarrow ⑤$

and stiffness of a solid shaft $S_s = \frac{G}{L} \times \frac{\pi}{32} \times d^4 \rightarrow ⑥$

Dividing equation ⑤ by equation ⑥ we get

$$\frac{S_H}{S_s} = \frac{(d_o)^4 - (d_i)^4}{d^4} = \frac{(d_o)^4 - (d_i)^4}{(d_o)^4} = 1 - \frac{(d_i)^4}{(d_o)^4} = 1 - K^4 \quad (\because d = d_o)$$

$$= 1 - (0.5)^4 = 0.9375$$

A helical spring is made from a wire of 6mm diameter and has outside diameter of 75mm. If the permissible shear stress is 350 MPa and modulus of rigidity 84 kN/mm², Find the axial load which the spring can carry and the deflection per active turn.

14(b) Given $d = 6mm$, $D_o = 75mm$, $\tau = 350 MPa = 350 N/mm^2$, $G = 84 kN/mm^2$ $= 84 \times 10^3 N/mm^2$

We ~~have~~ know that mean diameter of the spring,

$$D = D_o - d = 75 - 6 = 69 \text{ mm}$$

$\therefore$ Spring index $c = \frac{D}{d} = \frac{69}{6} = 11.5$

Let W = Axial load and δ/n = Deflection per active turn.

1. Neglecting the effect of curvature.

we know that the shear stress factor

$$K_s = 1 + \frac{1}{2C} = 1 + \frac{1}{2 \times 11.5} = 1.043$$

and maximum shear stress induced in the wire (τ)

$$350 = K_s \times \frac{8WD}{\pi d^3} = 1.043 \times \frac{8 W \times 69}{\pi \times 6^3} = 0.848 W$$

$\therefore$ $W = 350/0.848 = 412.7 N$.

We know that deflection of the spring $\delta = \frac{8 W D^3 n}{G \cdot d^4}$

$\therefore$ Deflection per active turn

$$\frac{\delta}{n} = \frac{8 W D^3}{G \cdot d^4} = \frac{8 \times 412.7 (69)^3}{84 \times 10^3 \times 6^4} = 9.96 \text{ mm}$$

2. Considering the effect of curvature.

we know that wahl's stress factor.

$$K = \frac{4C-1}{4C-4} + \frac{0.615}{C} = \frac{4 \times 11.5 - 1}{4 \times 11.5 - 4} + \frac{0.615}{11.5} = 1.123$$

We also know that the maximum shear stress induced in the wire (τ)

$$350 = K \times \frac{8WC}{\pi d^2} = 1.123 \times \frac{8 \times W \times 11.5}{\pi \times 6^2} = 0.913 W$$

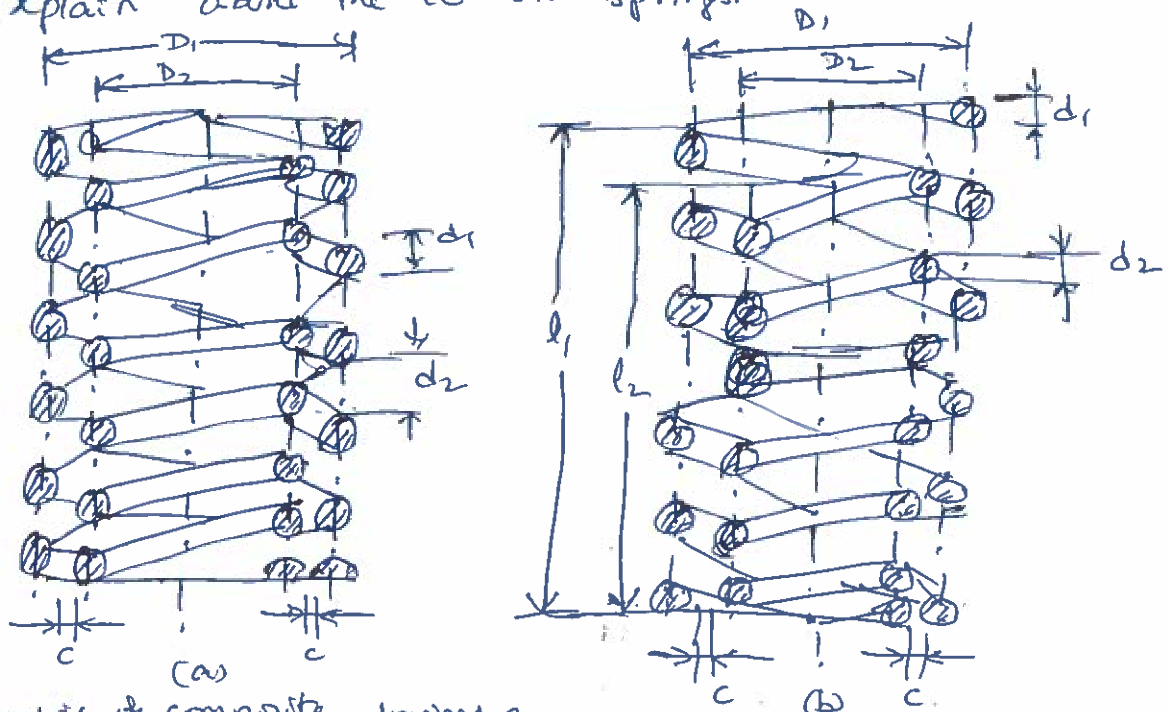
$\therefore$ $W = \frac{350}{0.913} = 383.4 N$.

and deflection of the spring $\delta = \frac{8 W D^3 \cdot n}{G d^4}$

$\therefore$ deflection per active turn

$$\frac{\delta}{n} = \frac{8 W D^3}{G \cdot d^4} = \frac{8 \times 383.4 (69)^3}{84 \times 10^3 \times 6^4} = 9.26 \text{ mm}.$$

14 (a) Explain about the co-axial springs.



Concentric or composite springs or co-axial springs

A concentric or composite spring is used for one of the following purposes;

1. To obtain greater spring force within a given space.
2. To ensure the operation of a mechanism in the event of failure of one of the springs.

The concentric springs for the above two purposes may have two or more springs and have the same free lengths ~~and~~ as shown in fig (a) and are ~~compressed~~ compressed equally. Such springs are used in automobile clutches, valve springs in aircraft, heavy duty diesel engines and rail-road car suspension systems. Sometimes concentric springs are used to obtain a spring force which does not increase in a direct relation to be deflection but increases faster. Such springs are made of different lengths as shown in fig (b). The shorter spring begins to act only after the longer spring is compressed to a certain amount. These springs are used in governors of variable speed engines to take care of the variable centrifugal force.

15(a) Define spring. What is the purpose of mechanical springs?

A spring is defined as an elastic body. whose function is to distort when loaded and to recover its original shape. when the load is removed. The various important applications of springs are as follows;

1. To cushion, absorb or control energy due to either shock or vibration as in car springs, railway buffers, aircraft landing gears, shock absorbers and vibration dampers.
2. To apply forces, as in brakes, clutches and spring loaded valves.
3. To control motion by mainting contact between two elements as in cams and followers.
4. To measure forces, as in spring balances and engine indicators
5. To store energy as in watches, toys. etc.

15 (b) A truck spring has 12 number of leaves, two of which are full length leaves. The spring supports are 1.05 m apart and the central band is 85 mm wide. The central load is to be 5.4 kN with a permissible stress of 280 MPa. Determine the thickness and width of the steel spring leaves. The ratio of the total depth to the width of the spring is 3. Determine the deflection of the spring.

15(b) A truck spring has 12 number of leaves, two of which are full length leaves. The spring supports are 1.05 m apart and the central band is 85 mm wide. The central load is to be 5.4 kN with a permissible stress of 280 MPa. Determine the thickness and width of the steel spring leaves. The ratio of the total depth to the width of the spring is 3. Also determine the deflection of the spring.

Given $n = 12$, $n_F = 2$, $2L_1 = 1.05\,m = 1050\,mm$, $\ell = 85\,mm$,

$2W = 5.4\,kN = 5400\,N$ or $W = 2700\,N$; $\sigma_F = 280\,MPa = 280\,N/mm^2$

Thickness of and width of the spring leaves.

Let t = Thickness of the leaves and b = width of the leaves.

Since it is given that the ratio of the total depth of the spring $(n \times t)$ and width of the spring (b) is 3, therefore

$$\frac{n \times t}{b} = 3 \quad \text{or} \quad b = \frac{n \times t}{3} = \frac{12 \times t}{3} = 4t$$

We know that the effective length of the spring.

$$2L = 2L_1 - \ell = 1050 - 85 = 965\,mm$$

$$L = \frac{965}{2} = 482.5\,mm.$$

and number of graduated leaves

$$n_G = n - n_F = 12 - 2 = 10$$

Assuming that the leaves are not initially stressed, therefore maximum stress or bending stress for full length ~~long~~ leaves (σ_F)

$$280 = \frac{18 W L}{b t^2 (2n_G + 3n_F)} = \frac{18 \times 2700 \times 482.5}{4t \times t^2 (2 \times 10 + 3 \times 2)}$$

$$280 = \frac{225476}{t^3}, \quad t^3 = 225476/280 = 805.3$$

or $t = 9.3$ say $10\,mm$ and $b = 4t = 4 \times 10 = 40\,mm$

and Deflection of the spring

We know that deflection of the spring.

$$\delta = \frac{12 W L^3}{E b t^3 (2n_G + 3n_F)}$$

$$= \frac{12 \times 2700 \times (482.5)^3}{210 \times 10^3 \times 40 \times 10^3 (2 \times 10 + 3 \times 2)}\ mm$$

$$= 16.7\,mm \quad (\text{Taking } E = 210 \times 10^3\,N/mm^2)$$

K. [signature]
30/11/2022

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30/11/2022

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	EEE			Academic Year	2022-2023
Course Code	20EE502	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Power Electronics						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Draw the output characteristics of power MOSFET.	20EE502.1	L1
2	At what conditions fully control rectifier will act as a line commutated inverter.	20EE502.2	L1
3	What are the control strategies for the regulation of output voltage in ac voltage controllers?	20EE502.3	L1
4	A step-up chopper is supplied from a 110V DC source. The voltage required by the load is 440V. if the switch is turned on for 0.25ms find the chopper frequency?	20EE502.4	L2
5	List any 4 methods for reduction of harmonics in the output voltage of inverters.	20EE502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	For a power diode reverse recovery time is 3.9 μs and the rate of diode current decay is 50 A/μs. For a softness factor of 0.3, calculate the peak inverse current and storage charge.	6M	20EE502.1	L2
6 (b)	List any 6 differences between SCR, IGBT & BJT.	6M	20EE502.1	L2
	OR			
7	Draw the switching characteristics of SCR, define delay time, rise time, spread time, reverse recovery time, gate recovery time and device turn off time	12M	20EE502.1	L2
8 (a)	Draw the circuit diagram of a single-phase half wave-controlled rectifier with RL load and derive the expression for r.m.s. output voltage.	6M	20EE405.2	L2
8 (b)	What is the effect of source inductance on the output waveform of 1-phase full controlled converter?	6M	20EE405.2	L2
	OR			
9	Describe the working principle of a single phase fully controlled bridge converter for rectifying and inverting modes. Draw the relevant load voltage waveforms for firing angle $\alpha = 45^\circ$ & 120°.	12M	20EE502.2	L3
10	A 3-phase full converter charges a battery from a three-phase supply of 230 V, 50 Hz.The battery emf is 200 V and its internal resistance is 0.5Ω. On account of inductance connected in series with the battery, charging current is constant at 20 A. Compute the firing angle delay and supply power factor.	12M	20EE502.3	L3
	OR			
11 (a)	Describe working of 1-Phase AC-AC regulators with R load only and draw the relevant waveforms.	6M	20EE405.3	L2
11 (b)	Discuss in detail the principle of operation of a circulating current mode dual converter. Compare circulating and circulating current free dual converter.	6M	20EE405.3	L2
12	Explain the operating principle of dc chopper with a suitable diagram. Draw the voltage and current waveforms of Buck chopper. Derive expressions for output voltage volt sec balance.	12M	20EE405.4	L2
	OR			
13	With the help of a neat circuit diagram and associated waveforms, discuss the operation of Buck-Boost converter and derive expressions for output voltage volt sec balance.	12M	20EE405.4	L2
14	Explain the working of a 1-phase half bridge, full bridge Inverter with RL load. Draw the relevant output waveforms.	12M	20EE405.5	L2
	OR			
15	With suitable circuit diagrams and wave forms explain the operation of a three phase inverter in 180^0 mode?	12M	20EE405.5	L2

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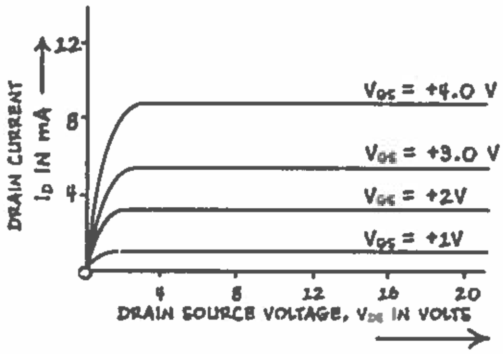
ANSWER KEY AND SCHEME OF EVALUATION

POWER ELECTRONICS

1. Draw the output characteristics of power MOSFET.



2. At what conditions fully control rectifier will act as a line commutated inverter.

   Ans: For firing angle greater than 90 degrees the fully control rectifier will act as a line commutated inverter

3. What are the control strategies for the regulation of output voltage in ac voltage controllers?

   There are two different types of thyristor control used in practice to control the ac power flow

   •On-Off control (Or) integral Cycle Control

   •Phase control

   In On-Off control technique Thyristors are used as switches to connect the load circuit to the ac supply (source) for a few cycles of the input ac supply and then to disconnect it for few input cycles. The Thyristors thus act as a high speed contactor (or high speed ac switch).

   In phase control the Thyristors are used as switches to connect the load circuit to the input ac supply, for a part of every input cycle. That is the ac supply voltage is chopped using Thyristors during a part of each input cycle. The thyristor switch is turned on for a part of every half cycle, so that input supply voltage appears across the load and then turned off during the remaining part of input half cycle to disconnect the ac supply from the load.

4. A step-up chopper is supplied from a 110V DC source. The voltage required by the load is 440V. if the switch is turned on for 0.25ms find the chopper frequency?

   Given Data:

   DC source =110V

   Output voltage = 440V

   Ton = 0.25ms

From Given Data

Step up chopper

Vo = Vs / (1-D)

440 = 110/ (1-D)

1-D = 110/440= 0.25

D= 0.75

0.75 = Ton/(Ton+ Toff)

0.75 = 0.25/ (0.25+ Toff)

(0.25+ Toff) = 0.25/ 0.75 = 0.33

Toff= 0.33-0.25 =0.083ms

Total Time period is

T = Ton + Toff

T = 0.25 ms+ 0.083ms = 0.33ms

and the switching frequency,

$$f_{switching} = \frac{1}{T}$$

= 1/0.33ms = 3 k Hz

5. List any 4 methods for reduction of harmonics in the output voltage of inverters.

Using Stepped Wave Inverter:

PWM Inverters

- Single Pulse Width Modulation(SPWM)
- Multiple Pulse Width Modulation(MPWM)
- Sinusoidal Pulse Width Modulation(SPWM)

For a power diode reverse recovery time is 3.9 μs and the rate of diode current decay is 50 A/μs. For a sofiness factor of 0.3, calculate the peak inverse current and storage charge.

6 (a) For a power diode reverse recovery time is 3.9 μs and the rate of diode current decay is 50 A/μs. For a softness factor of 0.3, calculate the peak inverse current and storage charge.

$$\begin{cases} t_{rr} = t_a + t_b = 3.9\mu s \\ S.F. = \frac{t_b}{t_a} = 0.3 \end{cases} \Rightarrow t_a = 3\mu s$$

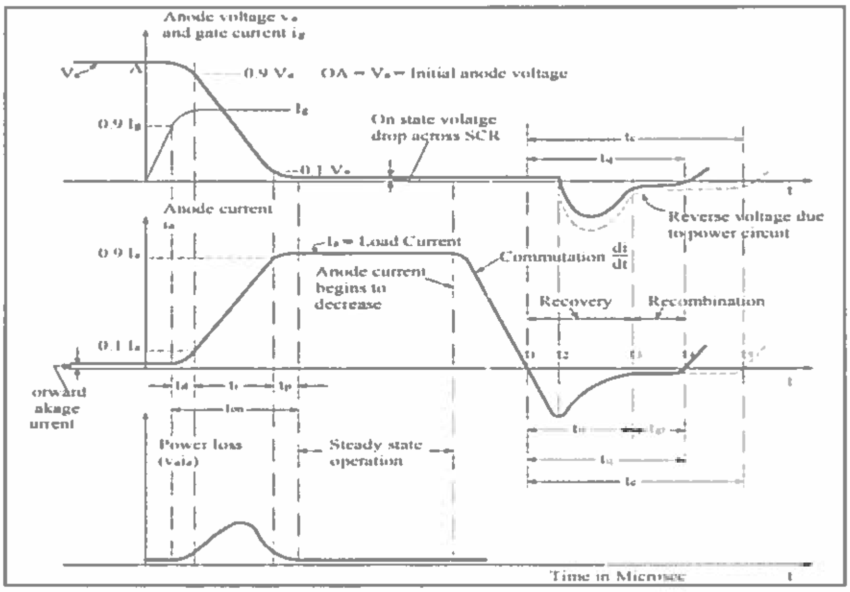
$$I_{RR} = t_a \frac{di}{dt} = 3 \times 50 = 150\ A$$

$$Q_{RR} = \frac{1}{2} I_{RR} t_{RR} = \frac{1}{2} \times 150 \times 3.9 = 292.5$$

6 (b) List any 6 differences between SCR, IGBT & BJT.

BJT	SCR	IGBT
Bipolar Junction Transistor	Silicon Controlled Rectifier	Insulated Gate Bipolar Junction Transistor
It has 3 Terminals Emitter, Base, Collector	It has 3 Terminals Anode, cathode, Gate	It has 3 Terminals Emitter, Gate, Collector
It is current controlled device	It is current controlled device	It is Voltage controlled device
Bidirectional current flow device	Unidirectional current flow device	Unidirectional current flow device
Bipolar voltage withstanding device	Bipolar voltage withstanding device	Bipolar voltage withstanding device
Suitable in High Current Withstanding applications	Suitable in high voltage with standing applications	Suitable in high switching frequency applications

7 Draw the switching characteristics of SCR, define delay time, rise time spread time, reverse recovery time, gate recovery time and device turn off time



Delay time (td):

1. Time taken by ig to reach 90% to 100% is called td
2. Time taken by Va to reach 100% to 90%is called td
3. Time taken by F.L.C to reach 10% of Ia is called td

Rise time (tr):

1. Time taken by Va to reach 90% to 10%is called tr
2. Time taken by Ia to reach 10% to 90% is called tr)

Spread time (tp):

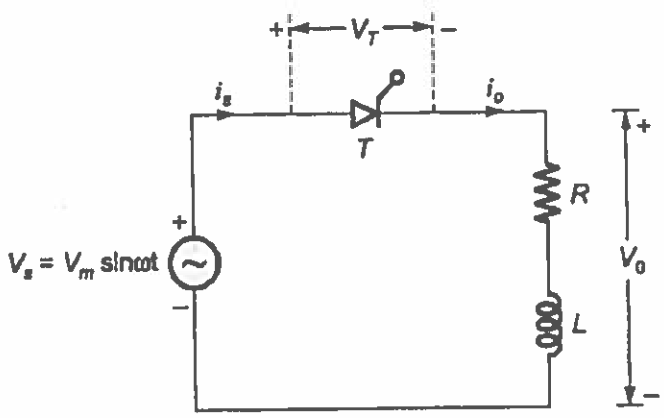
i. time taken by Va to reach 10% to min value is called tp
2. time taken by Ia to reach 90% to 100% is called tp

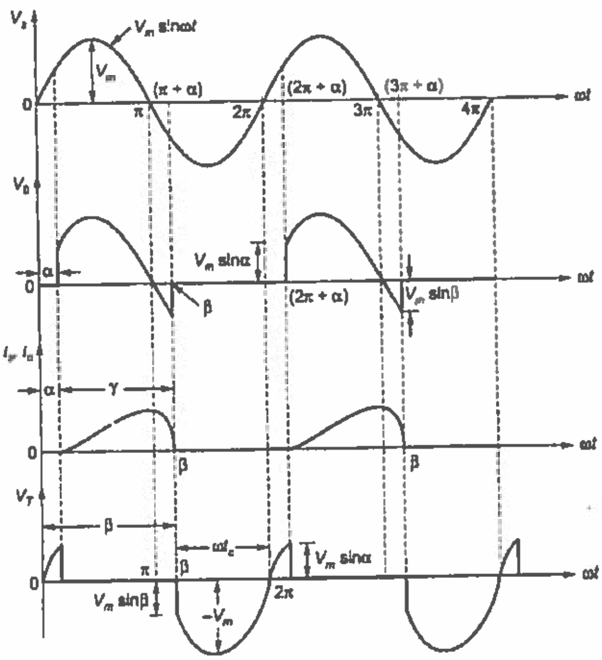
Reverse recovery time:

Gate recovery time

Device turn off time

8 (a) Draw the circuit diagram of a single-phase half wave-controlled rectifier with RL load and derive the expression for r.m.s. output voltage.



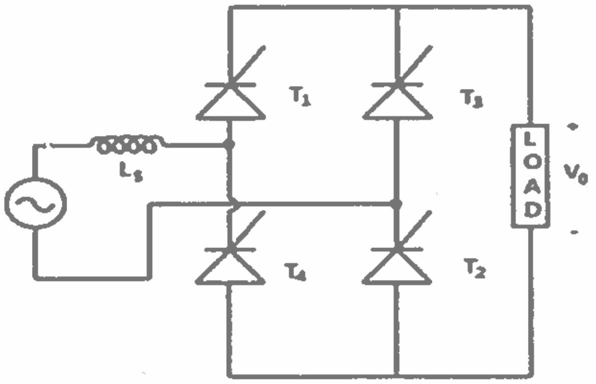


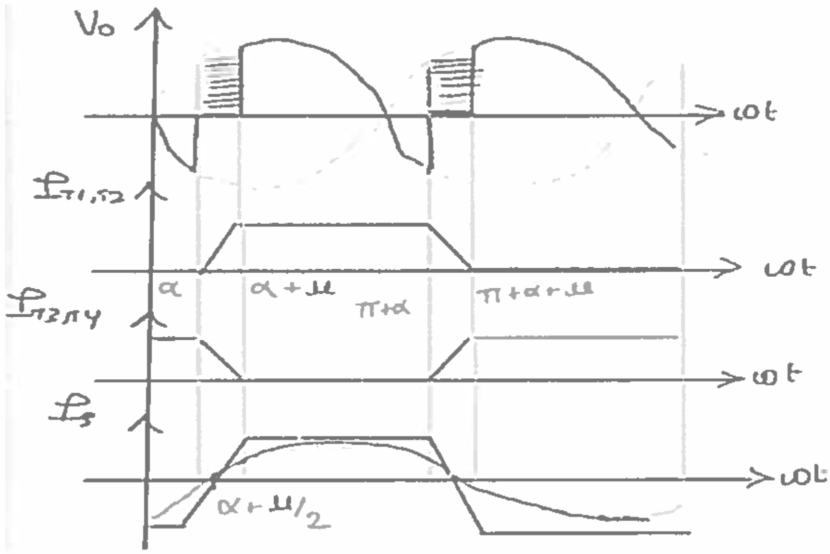
Expression for r.m.s. output voltage:

$$RMS\ Voltage, Vo = \sqrt{\frac{1}{2\pi}\int_{\alpha}^{\beta}(Vm\sin\omega t)^2\, d(\omega t)}$$

$$= \frac{Vm}{2\sqrt{\pi}}\sqrt{[(\beta-\alpha) - 1/2\{\sin 2\beta - \sin 2\alpha\}]}$$

8 (b) What is the effect of source inductance on the output waveform of 1-phase full controlled converter?

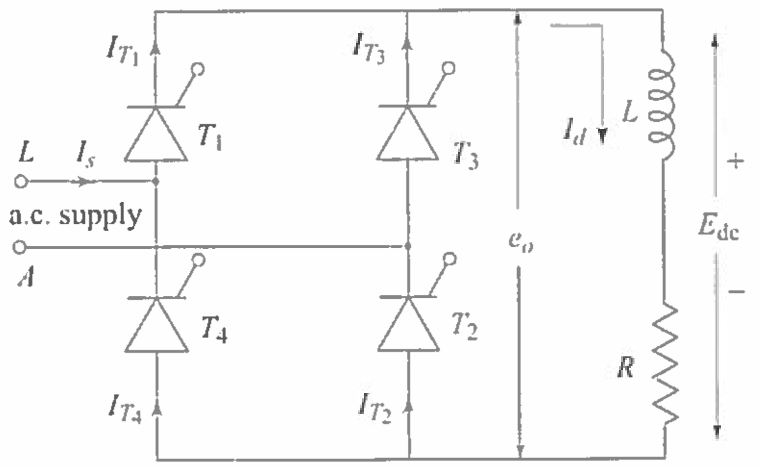


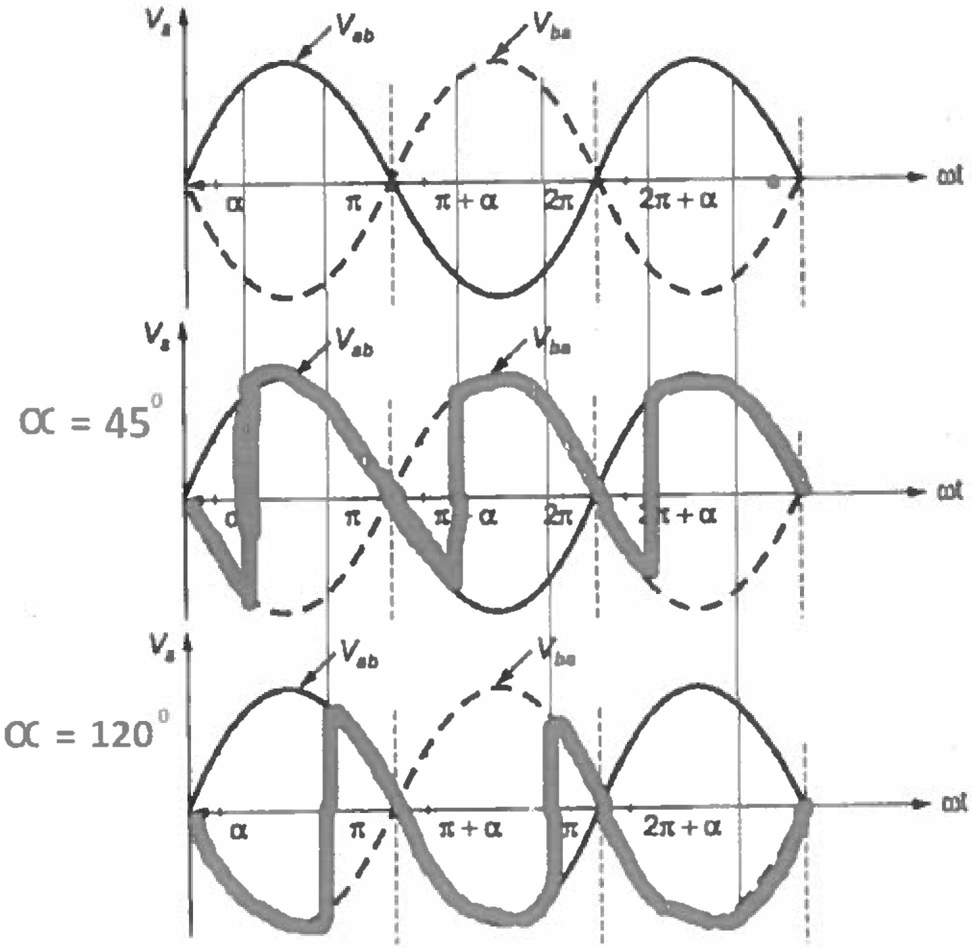


1. During overlap interval the load current freewheels through the thyristors and the output voltage is clamped to zero. On the other hand, the input current starts changing polarity as the current through T1 and T2 increases and T3 T4 current decreases. At the end of the overlap interval the current through T3 and T4 becomes zero and they commutate, T1 and T2 starts conducting the full load current

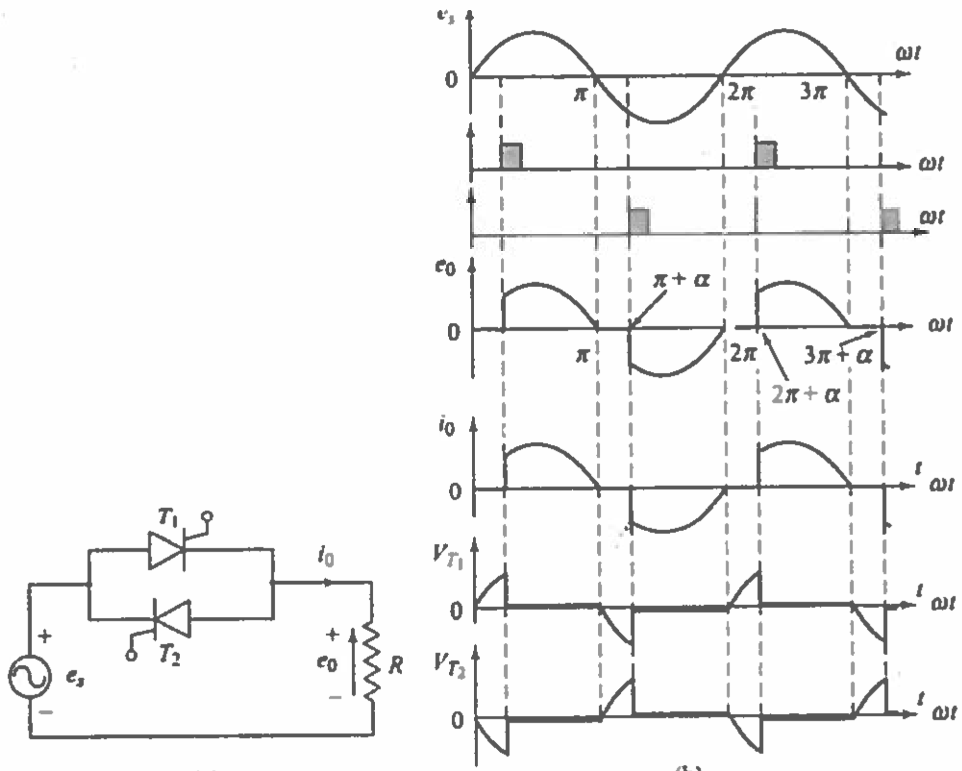
2. The same process repeats during commutation from T1 T2 to T3T4 at $\omega t = \pi + \alpha$. From Fig. 2(b) it is clear that, commutation overlap not only reduces average output dc voltage but also reduces the extinction angle γ which may cause commutation failure in the inverting mode of operation if α is very close to 180°.

9. Describe the working principle of a single phase fully controlled bridge converter for rectifying and inverting modes. Draw the relevant load voltage waveforms for firing angle α = 45° & 120°.





11 (a) Describe working of 1-Phase AC-AC regulators with R load only and draw the relevant waveforms ?



(a) (b)

Figure 1.a, shows the single-phase a.c. voltage controller with resistive load. As shown, two thyristors connected in antiparallel have been employed. Waveforms for source voltage e_s, gating pulses i_{g_1}, i_{g_2}, load-current i_o, load voltage e_0, voltage across T_1 as V_{T_1}, and voltage across T_2 as V_{T_2} are shown in Fig.1.b.

Thyristors T_1 and T_2 are forward biased during positive and negative half- cycle, respectively. During positive half-cycle, T_1 is triggered at a firing angle α. T_1 starts conducting and source voltage is applied to load from α to π. At π, both e_0, i_0 fall to zero. Just after π, T_1 is subjected to reverse bias and it is, therefore, turned-off. During negative half-cycle, T_2 is triggered at $(\pi + \alpha)$. T_2 conducts from $(\pi + \alpha)$ to 2π. Soon after 2π, T_2 is subjected to a reverse bias and it is, therefore, commutated. Load and source currents have the same waveform.

From zero to α, T_1 is forward biased, therefore $V_{T_1} = e_s$ as shown in Fig.1.b. From α to π, T_1 conducts, V_{T_1} is therefore about 1 V. After π, T_1 is reverse biased by source voltage, therefore, $V_{T_1} = e_s$ from π to $(\pi + \alpha)$. The voltage variation V_{T_1} across T_1 is shown in Fig.1.b. Similarly, the variation of voltage V_{T_2} across thyristor T_2 can be drawn. In Fig.1.b, voltage drop across thyristors T_1 and T_2 is purposely shown just to highlight the duration of reverse bias across T_1 and T_2. Examination of this figure reveals that for any value of α, each thyristor is reverse biased for π/ω seconds.

11 (b) Discuss in detail the principle of operation of a circulating current mode dual converter. Compare circulating and circulating current free dual converter.

Principle of operation of a circulating current mode dual converter:

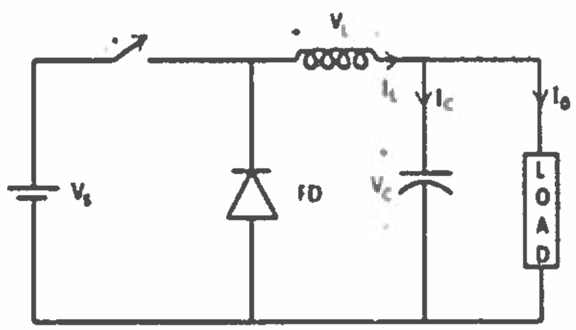
The firing angles are adjusted such that the firing angle of converter 1 (α_1) + firing angle of converter 2 (α_2) = 180°. Converter 1 performs as a controlled rectifier when firing angle be $0 < \alpha_1 < 90^\circ$ and Converter 2 performs as an inverter when the firing angle be $90^\circ < \alpha_2 < 180^\circ$.

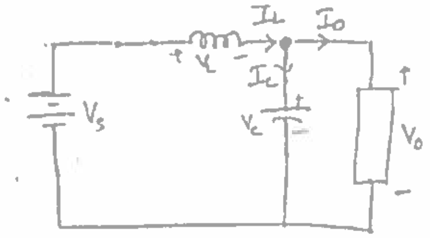
Sr. No	With circulating Current	Without circulating Current
1.	Reactors are required to limit the circulating current. These reactors may be costly and bulky and consume lot of power.	Reactors are not required for normal operation but sometimes may have to be used to make the load current continuous.
2.	Losses increase due to circulating current and hence efficiency decreases.	Efficiency is higher.
3.	The converter current is due to presence of circulating current.	Converter current can be discontinuous.
4.	Less response time, therefore fast response.	Response time is more so slow response.
5.	Transfer characteristics are linear.	Transfer characteristics are nonlinear due to discontinuous current.
6.	The fault circulating current, due to undesired firing of converters is limited by the current reactor.	Simultaneous firing of both the converter due to faulty signals will create a short circuit condition.
7.	The converter current is higher than the load current, due to circulating current.	The converter current is the same as output current.

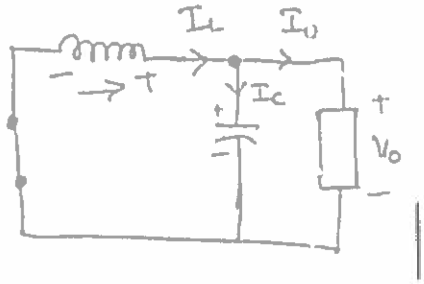
12. Explain the operating principle of dc chopper with a suitable diagram. Draw the voltage and current waveforms of Buck chopper. Derive expressions for output voltage volt sec balance.

Principle of operation of DC chopper:

DC chopper works on DC voltage. They work as **step up and step down transformers on DC voltage**. They can convert the steady constant DC voltage to a higher value or lower value based on their type. DC choppers are more efficient, fast, and optimized devices.





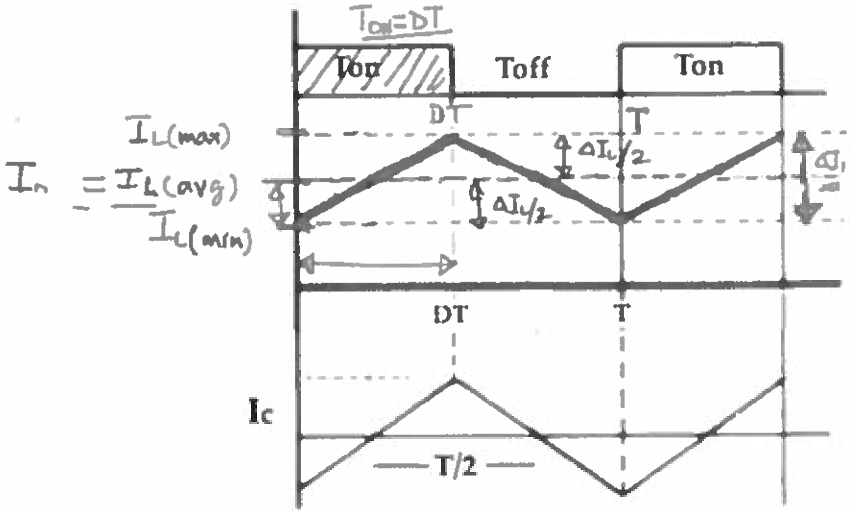


(1) $V_{L(ON)} = V_s - V_o$

$I_{c(ON)} = I_L - I_o$

(2) $V_{L(off)} + V_o = 0 \Rightarrow V_{L(off)} = -V_o$

$I_{c(off)} = I_L - I_o$



(3) Volt sec Balance

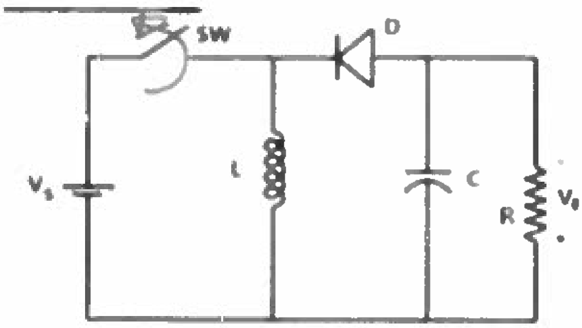
$V_{L(ON)} T_{ON} + V_{L(off)} T_{off} = 0$

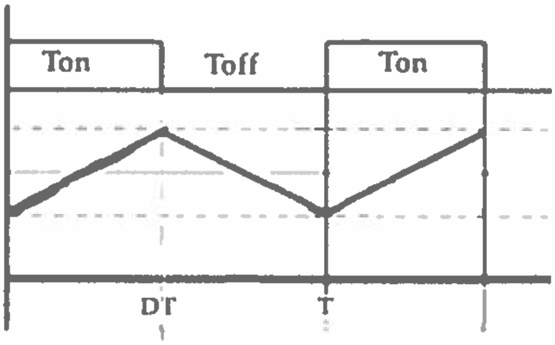
$(V_s - V_o) DT - V_o (1-D) T = 0$

$\boxed{V_o = DV_s}$ $\quad 0 \le D \le 1$

(Step down chopper)

13. With the help of a neat circuit diagram and associated waveforms, discuss the operation of Buck-Boost converter and derive expressions for output voltage volt sec balance.





① 0 to DT

$V_{L(ON)} = V_s$

$I_{c(ON)} = -I_o$

② DT to T

$V_L(off) - V_o = 0$

$V_{L(off)} = V_o$

$I_{c(off)} = -(I_L + I_o)$

③ Volt sec Balance

$V_s DT + V_o(1-D)T = 0$

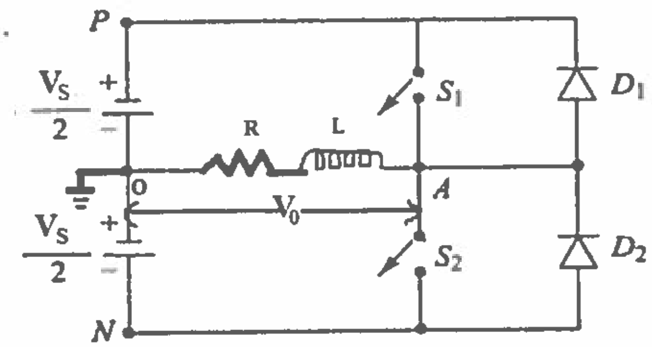
$$\boxed{V_o = \frac{-D}{(1-D)} V_s}$$

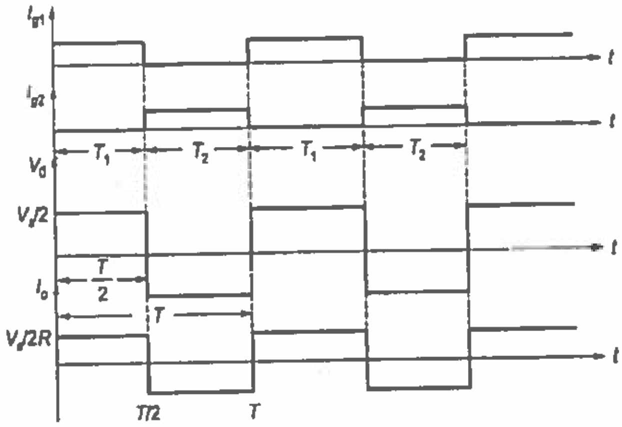
④ Amp sec Balance

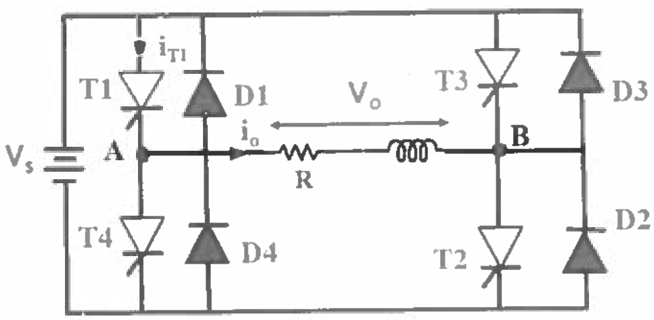
$-I_o DT - (I_L + I_o)(1-D)T = 0$

$$I_L = \frac{-I_o}{(1-D)}$$

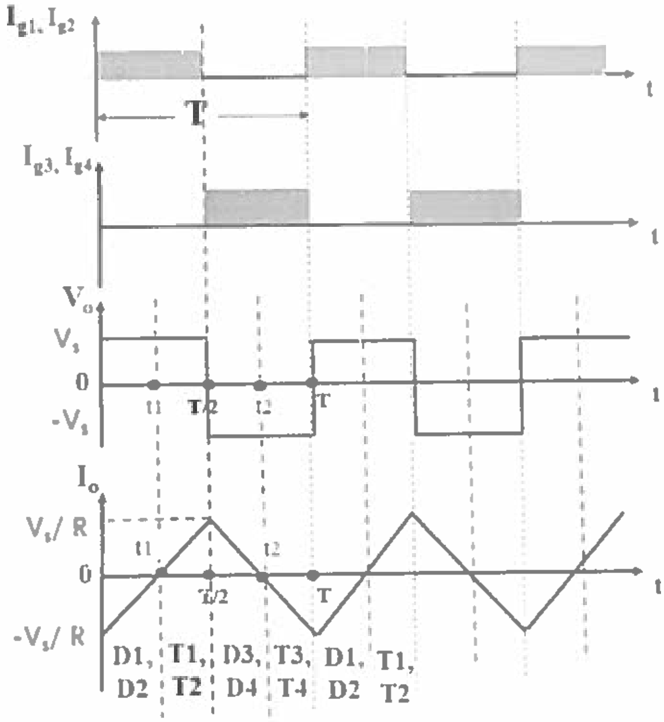
14. Explain the working of a 1-phase half bridge, full bridge Inverter with RL load. Draw the relevant output waveforms.



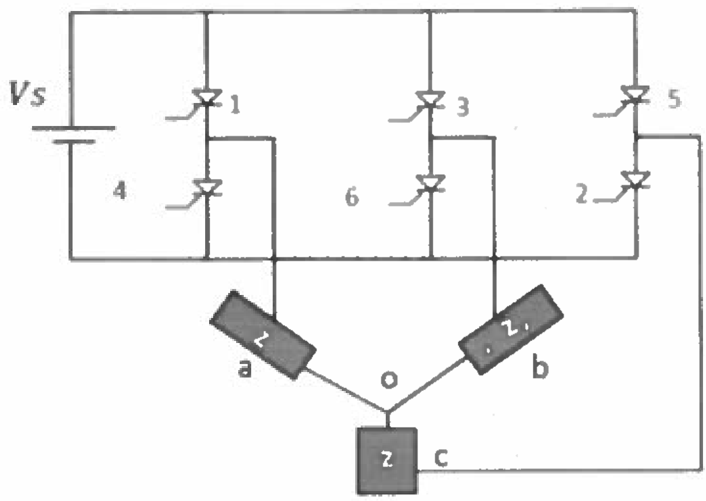


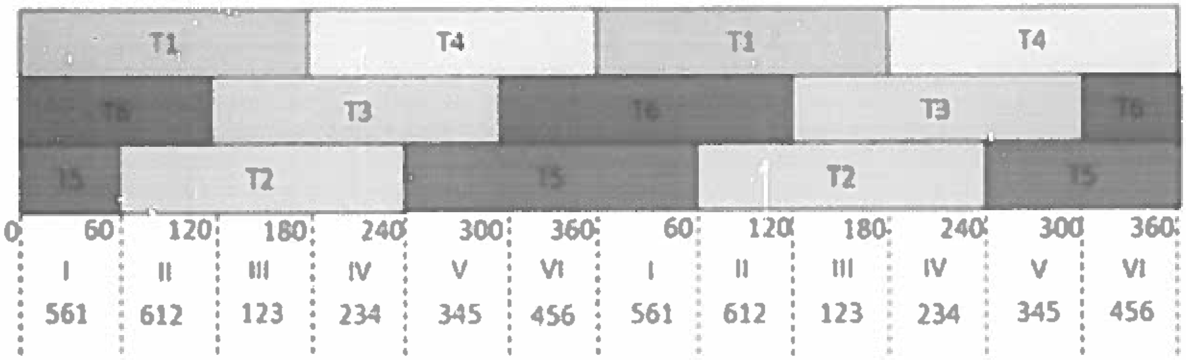


Circuit Diagram of Single Phase Full Bridge Inverter with RL Load

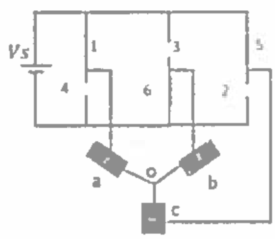


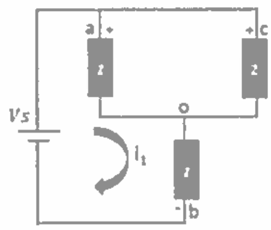
• 15. With suitable circuit diagrams and wave forms explain the operation of a three phase inverter in 180^0 mode?





STEP 1





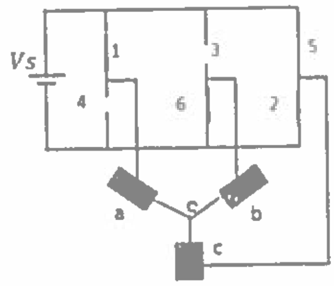
$$i_1 = \frac{V_s}{z + \frac{z}{2}} = \frac{2}{3}\cdot\frac{V_s}{z}$$

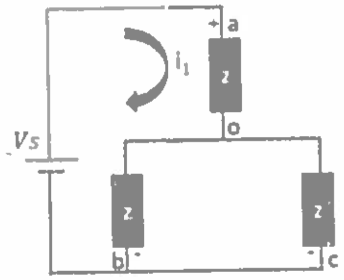
$$Vao = i_1 \cdot \frac{Z}{2} = \left[\frac{2}{3}\cdot\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

$$Vob = i_1 . Z = \left[\frac{2}{3}\cdot\frac{V_s}{Z}\right] Z = \frac{2Vs}{3}$$

$$Vco = i_1 . Z = \left[\frac{2}{3}\cdot\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

STEP 2





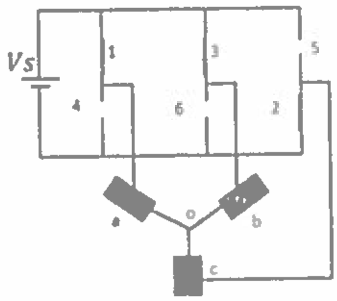
$$i_1 = \frac{V_s}{z + \frac{z}{2}} = \frac{2}{3}.\frac{V_s}{z}$$

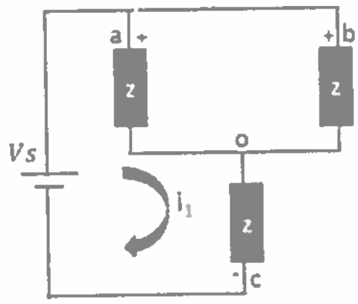
$$Vao = i_1.Z = \left[\frac{2}{3}.\frac{V_s}{Z}\right]Z = \frac{2Vs}{3}$$

$$Vob = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

$$Voc = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

STEP 3





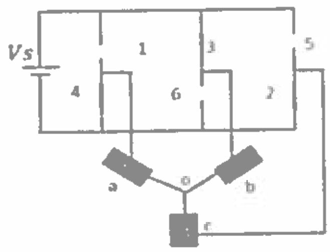
$$i_1 = \frac{V_s}{z + \frac{z}{2}} = \frac{2}{3}.\frac{V_s}{z}$$

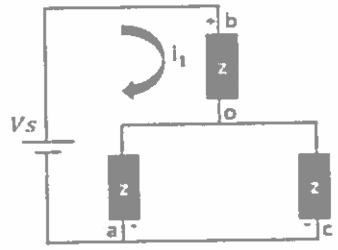
$$Vao = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{z}{2} = \frac{V_s}{3}$$

$$Vbo = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

$$Voc = i_1.Z = \left[\frac{2}{3}.\frac{V_s}{Z}\right]Z = \frac{2Vs}{3}$$

STEP 4



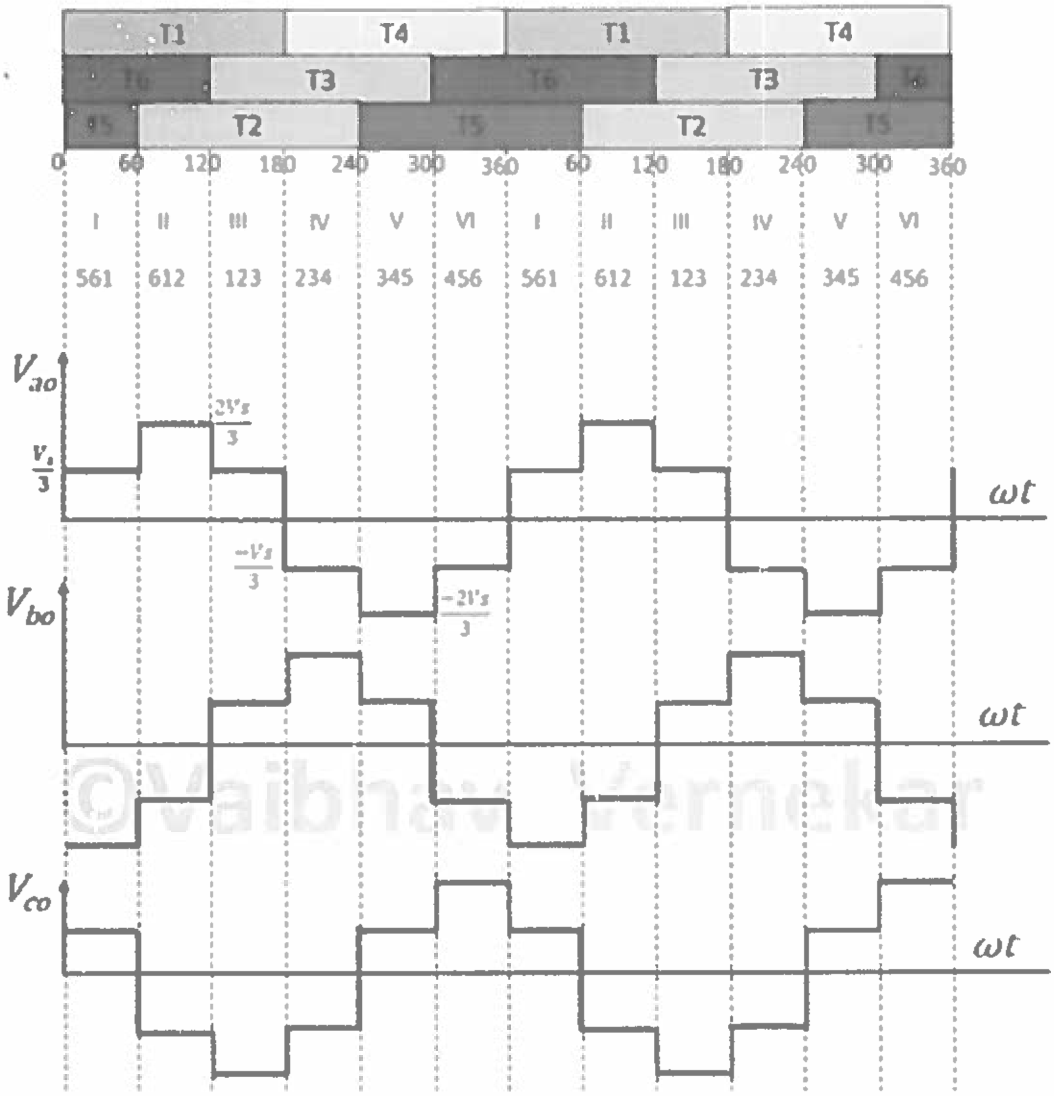


$$i_1 = \frac{V_s}{z + \frac{z}{2}} = \frac{2}{3}.\frac{V_s}{z}$$

$$Voa = i_1.Z = \left[\frac{2}{3}.\frac{V_s}{Z}\right]Z = \frac{2Vs}{3}$$

$$Vbo = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$

$$Voc = i_1.\frac{Z}{2} = \left[\frac{2}{3}.\frac{V_s}{Z}\right]\frac{Z}{2} = \frac{V_s}{3}$$



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NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	ECE			**Academic Year**	2022 - 2023
Course Code	20EC502	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	V
Course	**Linear and Digital IC Applications**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Sketch the Integrator and Differentiator circuit using IC 741.	20EC502.1	L1
2	List any two applications of VCO.	20EC502.2	L1
3	Differentiate R-2R and Inverted R-2R DAC methods.	20EC502.3	L2
4	Draw the circuit diagram of CMOS NAND circuit.	20EC502.4	L1
5	List any two IC's used for Adders.	20EC502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Draw the circuit diagram of Inverting Amplifier and Non Inverting Amplifier and explain its operation.	12M	20EC502.1	L2
	OR			
7	Explain the D.C. characteristics of an Op-Amp.	12M	20EC502.1	L2
8 (a)	Explain the working of an Astable Multivibrator using 555 timer with relevant circuit and Waveforms and derive the expression for frequency of operation and duty cycle.	8M	20EC502.2	L2
8 (b)	Design a 555 based Astable Multivibrator to generate an output signal with frequency 2.5KHz and duty cycle of 50%(C=0.01uF).	4M	20EC502.2	L3
	OR			
9	Explain the working of PLL and derive an expression for the lock-in range of a PLL.	12M	20EC502.2	L2
10	Explain Successive Approximation ADC with neat diagram.	12M	20EC502.3	L2
	OR			
11	Explain the binary weighted resistor and R-2R type DAC with necessary equations and write any 4 advantages and disadvantages.	12M	20EC502.3	L2
12	Explain the 2 input TTL NAND gate logic in detail with neat sketch and its truth table.	12M	20EC502.4	L2
	OR			
13	Realize a CMOS OR and NOR gate using CMOS Logic with function table	12M	20EC502.4	L3
14 (a)	Discuss the Magnitude Comparator (IC7485).	4M	20EC502.5	L2
14 (b)	Explain any two 74 Series combinational logic IC's with example.	8M	20EC502.5	L2
	OR			
15	Explain the concept of Universal Shift Register with necessary diagrams.	12M	20EC502.5	L2

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(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

B.TECH V Semester ECE.

Linear and Digital IC Applications (20EC502)

PART-A 10M

1. Sketch the Integrator & Differentiator circuit using IC741
   Integrator circuit – 1 M
   Differentiator circuit – 1 M

2. List any two applications of VCO
   Applications + diagram - 2M

3. Differentiate R-2R and Inverted R-2R DAC Methods
   Comparing the methods – 2M

4. Draw the circuit diagram of CMOS NAND circuit
   CMOS NAND circuit + Truth Table – 2M

5. List any two IC's used for adders
   IC NO & Application – 2M.

PART B

6. Draw the Circuit diagram of Inverting Amplifier and Non Inverting Amplifier and Explain its operation. 12M
   Diagrams – 4M
   Description – 4M
   Equations – 4M

7. Explain the DC characteristics of an Op-Amp
   Diagrams – 4M
   Definitions + Description – 8M

8 ⓐ Explain the working of an Astable Multivibrator using 555 timer with relevant circuit and waveform and derive the expression for frequency of operation and duty cycle. (8M)

Diagram – 2M

Description + Derivation – 4M

waveforms – 2M

ⓑ Design a 555 based Astable Multivibrator to generate an o/p signal with frequency 2.5KHz & duty cycle 50% (C=0.01 μF) (4M)

Time period calculation – 1M

RA, RB calculation – 1M

Diagram – 2M.

⑨ Explain the working of PLL and derive an expression for the lock-in-range of PLL. (12M)

Diagram – 3M

Derivation for lock in range – 2M+2M=4M

Description – 5M

⑩ Explain Successive Approximation ADC with neat diagram. (12M)

Diagram – 4M

Expressions + Description – 8M.

⑪ Explain Binary weighted resistor & R-2R type DAC with necessary equations & write any 4 advantages and Disadvantages 12M

Binary weighted – 4M

R-2R ladder method – 6M

Advantages, Disadvantages – 2M

(12) Explain the 2 i/p TTL NAND gate using CMOS logic with function table (12M)

TTL NAND gate circuit - 5M

Description - 7M

(13) Realize a CMOS OR and NOR gate using CMOS logic with function table (12M)

CMOS Implementation OR Gate - 4M

Function Table + Description - 2M

CMOS Implementation NOR Gate - 4M

Function Table + Description - 2M

(14) (a) Discuss the Magnitude Comparator (IC 7485) (4M)

Pin diagram - 2M

Description - 2M

(b) Explain any two 74 series Combinational logic IC's with example (8M)

74 IC pin diagrams - 4M

Example - 4M

(15) Explain the Concept of Universal shift Register with necessary diagrams (12M)

Diagram - 2M

Truth Table - 2M

Description + VHDL code - 8M

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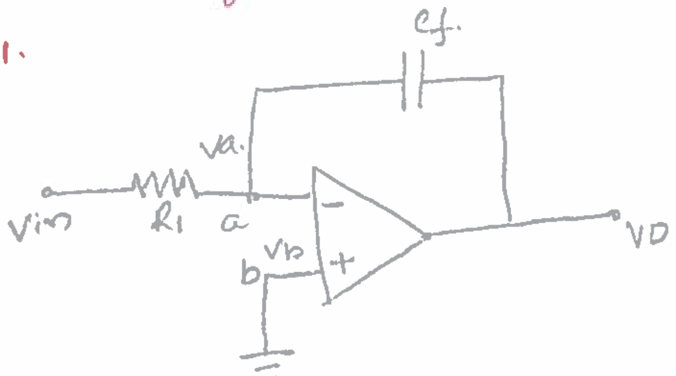
ANSWER KEY AND SCHEME OF EVALUATION

LINEAR AND DIGITAL IC APPLICATIONS (20EC502)

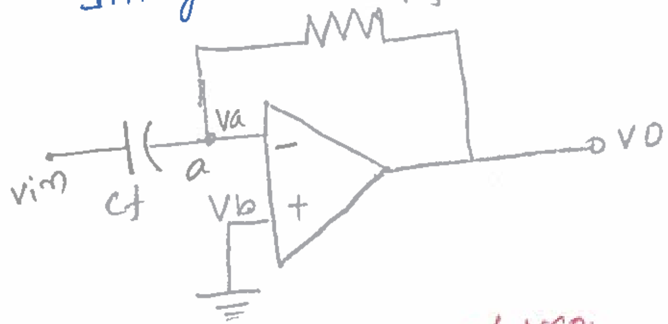
PART A.

1. Sketch the Integrator and Differentiator circuit using IC 741.



$$V_O = -\frac{1}{R_1 C_f} \int V_{in}(t)\, dt$$

Integrator.



$$V_O = -R_f C_1 \frac{dV_{in}}{dt}$$

2. List any two applications of VCO.

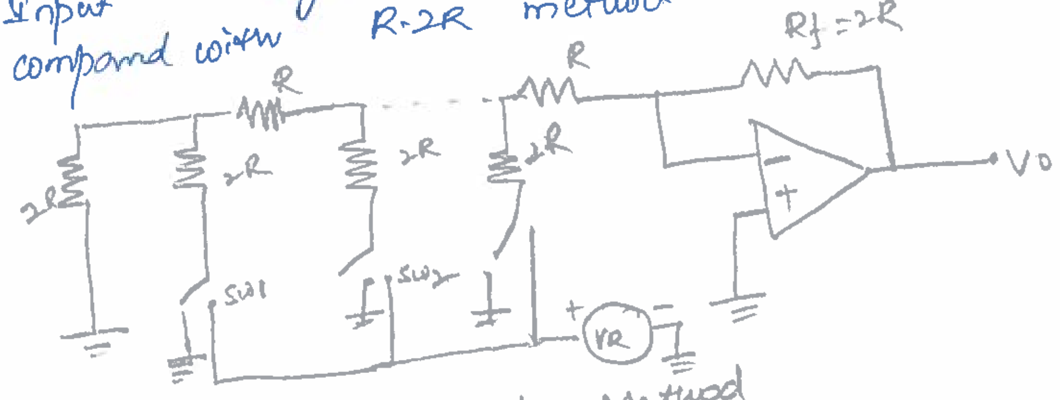
Applications:-

(1) FM demodulation
(2) Signal generation
(3) function Generator.
(4) FSK Demodulator
(5) frequency Multiplier.
(6) converting low frequency signals EEG & EKG into audio frequency range signals.
(7) Tone generation

③ Differentiate R-2R and Inverted R-2R DAC Methods.

In R-2R DAC current flowing through resistors changes as Input binary data changes. Power dissipation causes heating which creates non linearity in DAC.

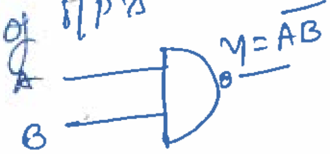
In Inverted R-2R DAC ladder node voltage remain constant even with change in Input binary word. Performance is good when compared with R-2R method DAC



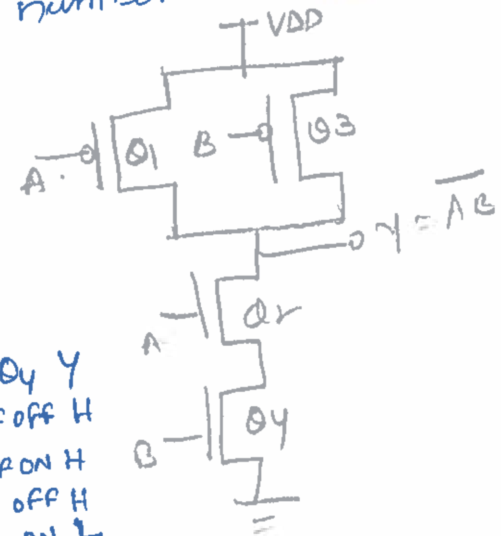
R-2R Ladder Method

④ Draw the circuit Diagram of CMOS NAND circuit

In CMOS implementation, for a NAND same as selected of number PMOS, NMOS circuit no. of I/P's



A	B	Y
0	0	1
0	1	1
1	0	1
1	1	0



A	B	Q_1	Q_3	Q_2	Q_4	Y
L	L	ON	ON	OFF	OFF	H
L	H	ON	OFF	OFF	ON	H
H	L	OFF	ON	ON	OFF	H
H	H	OFF	OFF	ON	ON	L

⑤ List any two IC's used for Adders.

IC7483 - 4 Bit binary full adder.

IC 74283 - 4 Bit Binary adder.

IC7486 + IC7408 - XOR Gate + AND Gate = Half adder.

6) Draw the Circuit diagram of Inverting Amplifier & Non Inverting Amplifier and explain its operation.

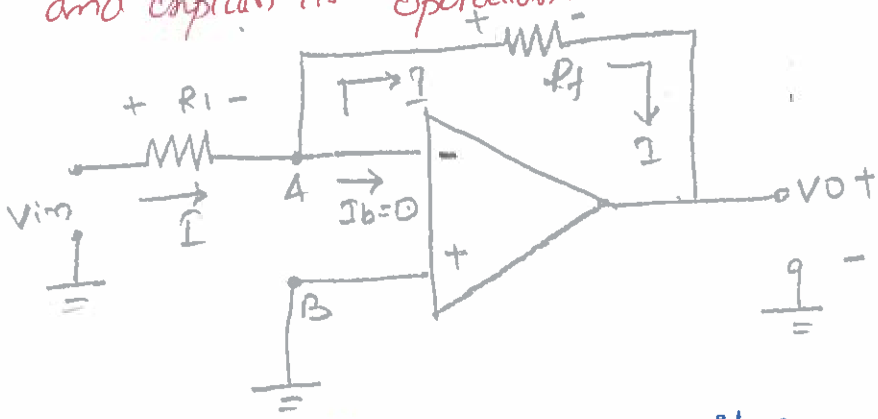


Fig: Inverting Amplifier.

The output of an amplifier is inverted as compared to the I/p signal. Inverted signal means o/p is having 180° phase shift when compared to the Input signal.

An amplifier which provides a phase shift of 180° between I/p & o/p signal is called "Inverting Amplifier".

<u>Ckt operation</u>:-

As node B is grounded, node A is also at ground potential from virtual ground concept $V_A = 0$.

$$I = \frac{V_{in} - V_A}{R_1}$$

$$I = \frac{V_{in}}{R_1}$$

from O/p side considering the the direction of current I

$$I = \frac{V_A - V_0}{R_f}$$

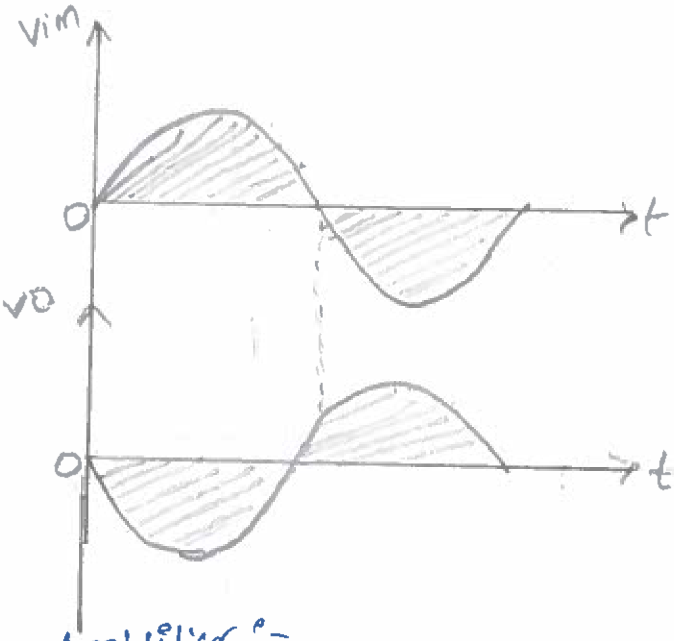
$$I = \frac{-V_0}{R_f}$$

Entire Current I passes through R_f as opamp current is zero

$$\frac{V_{in}}{R_1} = \frac{-V_0}{R_f}$$

$$\boxed{A_{vf} = \frac{V_0}{V_{in}} = -\frac{R_f}{R_1}}$$

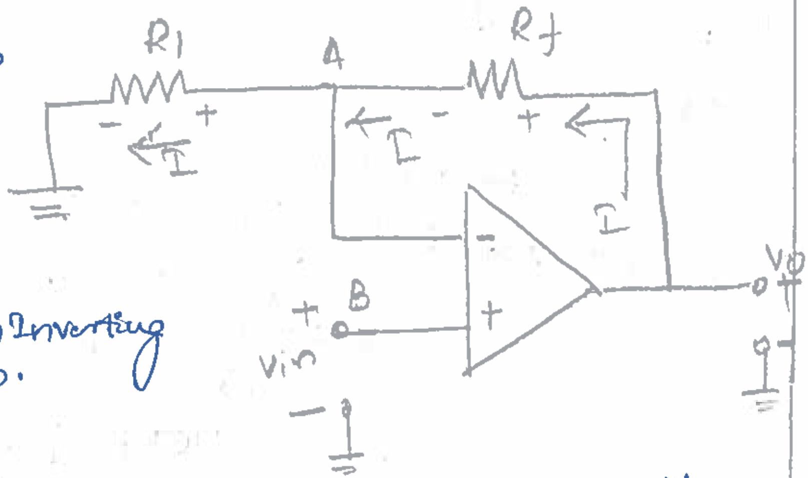
$\frac{R_f}{R_1}$ is gain of the amplifier, -ve sign indicates o/p polarity is opposite to i/p. Hence called as Inverting amplifier.



Non Inverting Amplifier :-

An amplifier which amplifies input without producing any phase shift b/w input and output is called Non Inverting Amplifier.

The input is applied to non Inverting terminal of the op-amp.



Non Inverting Amplifier

Ckt operation :-

Node B is at potential Vin, A is also at potential Vin

$$V_A = V_B = V_{in}$$

From o/p side

$$I = \frac{V_o - V_A}{R_f} = \frac{V_o - V_{in}}{R_f}$$

At Inverting terminal

$$I = \frac{V_A - 0}{R_1} = \frac{V_{in}}{R_1}.$$

Entire current passes through R_1 as I/p current of an op-amp is zero.

$$\frac{V_A}{R_1} = \frac{V_o - V_{in}}{R_f}$$

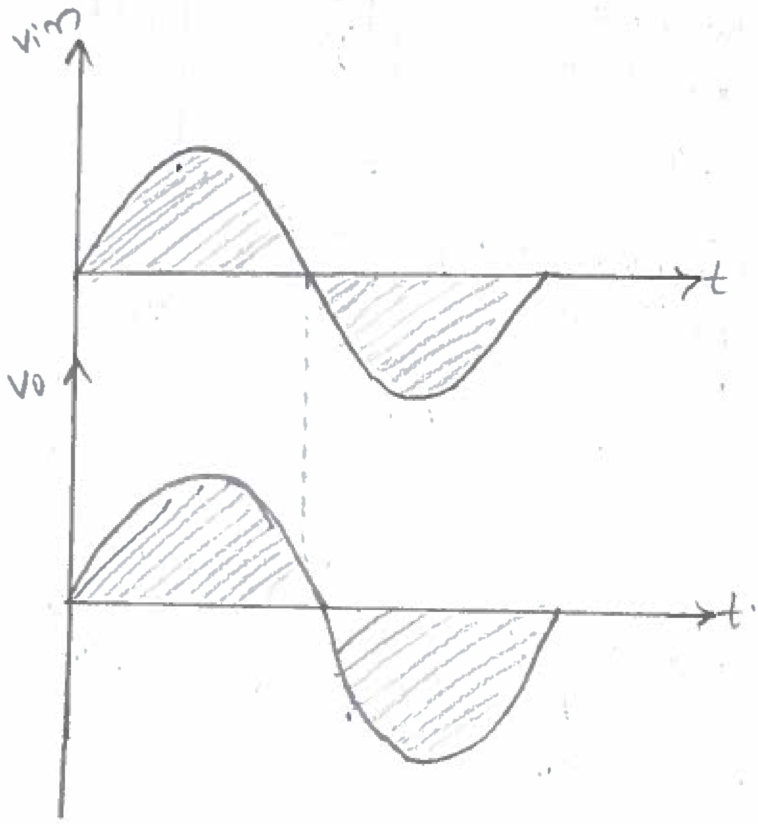
$$\frac{V_o}{R_f} = \frac{V_A}{R_1} + \frac{V_{in}}{R_f}$$

$$\frac{V_o}{R_f} = V_{in}\left[\frac{1}{R_1} + \frac{1}{R_f}\right]$$

$$\frac{V_o}{V_{in}} = R_f\left[\frac{R_1 + R_f}{R_1 R_f}\right]$$

$$\frac{V_o}{V_{in}} = \left[1 + \frac{R_f}{R_1}\right]$$

+ve sign. Indicates no phase shift between I/p & O/p



(7) Explain the DC characteristics of an op-amp

An ideal op-amp draws no current from source and its response is also independent of temperature.

A practical op-amp, current is taken from source into op-amp i/p's, i/p's respond differently to current & voltage due to mismatch in transistors. Real op-amp shifts its operation with temperature.

The non ideal dc characteristics that add error components to the dc o/p voltage are

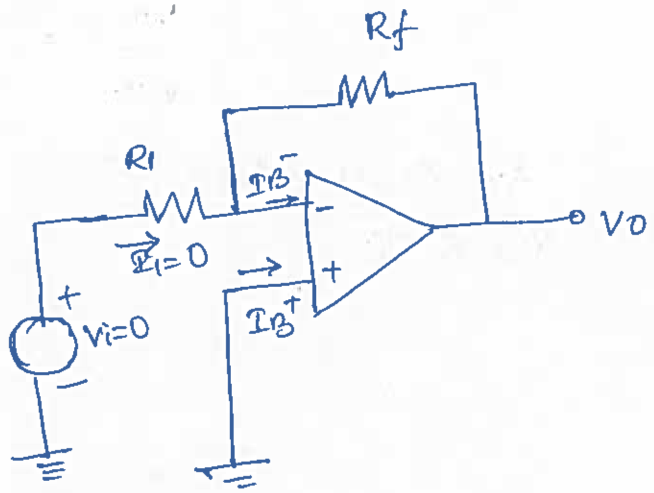
(i) Input bias current

(ii) Input offset current

(iii) Input offset voltage

(iv) Thermal Drift.

Input Bias Current :-



I_B^- & I_B^+ are base currents to the Inverting & Non Inverting terminals.

$$I_B = \frac{I_B^+ + I_B^-}{2}$$

Bias current of bipolar opamp = 500nA
FET I/P op-amp = 50pA

If Vi = 0, O/p voltage should be zero volts.

$V_o = I_B^- R_f$

If $R_f = 1M\Omega$

$V_o = 500nA \times 1M\Omega$

= 500mV.

If a signal is of mV ~~without~~ Bias currents unacceptable which can produce a mV voltage be compensated by Rcomp added b/w non Inverting terminal & ground.

$$R_{comp} = \frac{R_1 R_f}{R_1 + R_f}$$

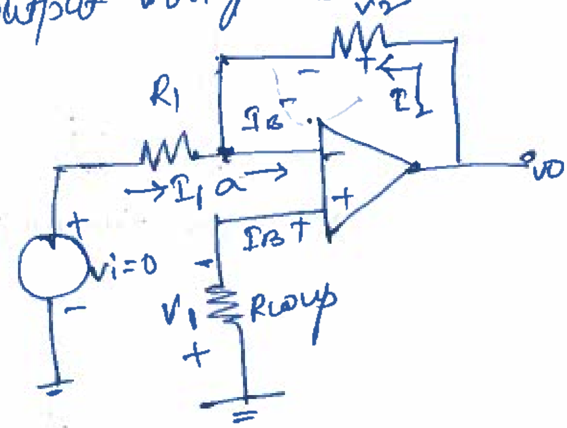
Input offset current :-

As the Input transistors can't be made Identical, there will be some difference b/w I_B^+ & I_B^- & is called as offset current I_{os} & can be written as

$$|I_{os}| = I_B^+ - I_B^-$$

I_{os} for BJT op-amp = 200nA
FET - 10pA

Even with Bias current compensation offset current produce an output voltage when $V_i = 0$



$$V_1 = I_B^+ R_{comp}$$

$$I_1 = \frac{V_1}{R_1}$$

KCL at node a

$$I_2 = (I_B^- - I_1)$$

$$= I_B^- - \left(\frac{I_B^+ R_{comp}}{R_1}\right)$$

$$V_0 = I_2 R_f - V_1$$

$$= I_2 R_f - I_B^+ R_{comp}$$

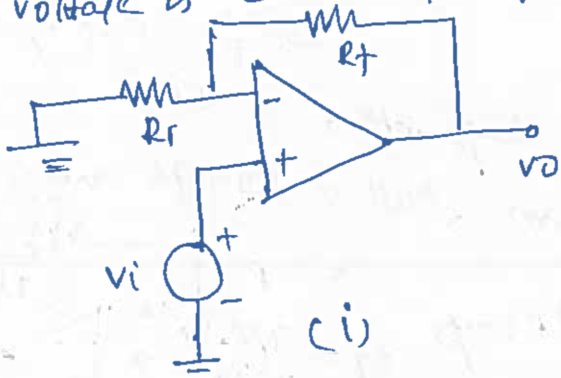
$$= \left(I_B^- - \frac{I_B^+ R_{comp}}{R_1}\right) R_f - I_B^+ R_{comp}$$

$$V_0 = R_f \left[I_B^- - I_B^+\right]$$

$$V_0 = R_f I_{os}$$

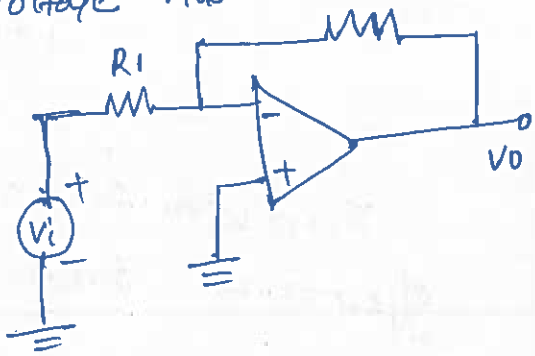
Input offset voltage :-

Inspite of compensation techniques, o/p voltage still may not be zero with zero I/p voltage. due to unavoidable imbalances inside the op-amp & small voltage applied at i/p terminals to make o/p voltage zero. This voltage is called Input offset voltage V_{ios}.

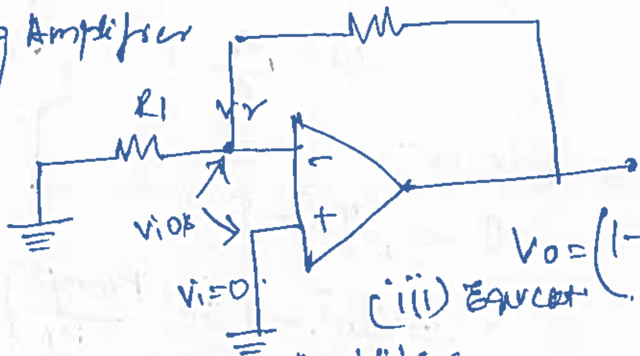


(i)

Non Inverting Amplifier



(ii) Inverting Amplifier.



(iii) Equcct

$$V_o = \left(1 + \frac{R_f}{R_1}\right) V_{ios}$$

For Non Inverting & Inverting Amplifier If $V_i = 0$, then the ckts will become as fig(iii)

voltage V_2 at Inverting terminal

$$V_2 = \left(\frac{R_1}{R_1 + R_f}\right) V_0$$

$$V_0 = \left(\frac{R_1 + R_f}{R_1}\right) V_2 = \left(\frac{1}{1} + \frac{R_f}{R_1}\right) V_2$$

$$V_{ios} = |V_i - V_2| \quad \text{and } V_i = 0$$

$$V_{ios} = |0 - V_2| = V_2$$

$$V_0 = \left(1 + \frac{R_f}{R_1}\right) V_{ios}$$

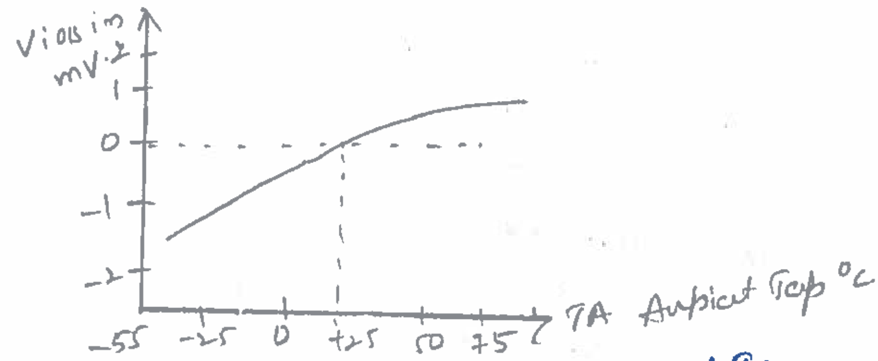
$$\text{Total o/p offset voltage} = \left(1 + \frac{R_f}{R_1}\right) V_{ios} + R_f I_B$$

$$= \left(1 + \frac{R_f}{R_1}\right) V_{ios} + R_f I_{os}$$

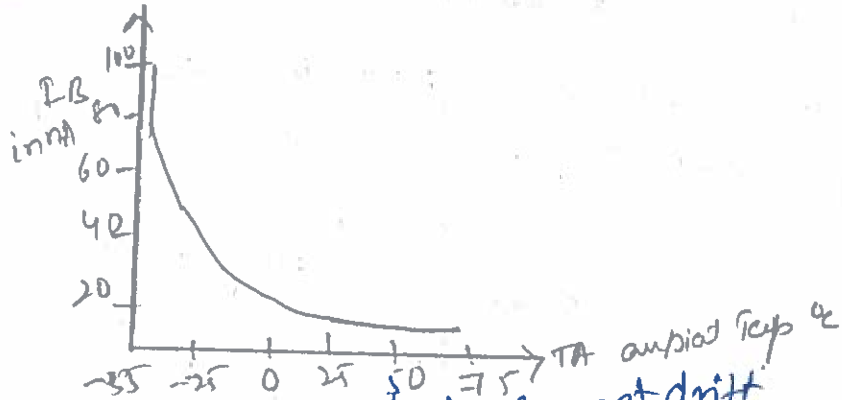
Thermal Drift:-

op-amp parameters Input offset Voltage V_{ios}, Input Bias current I_B & Input offset current I_{ios} are varying with Temperature, supply voltage & Time

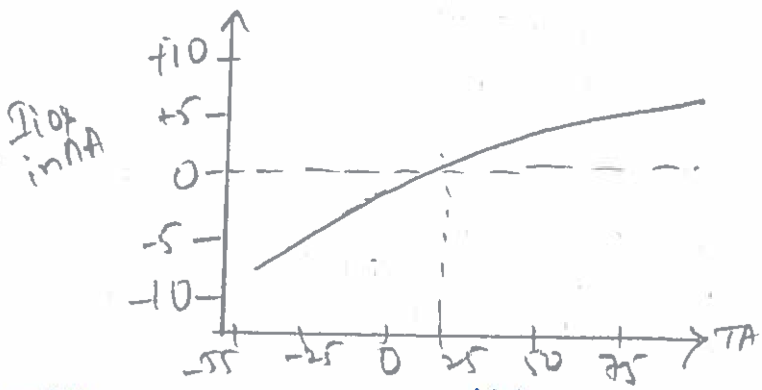
Input offset voltage drift $= \frac{\Delta V_{ios}}{\Delta T} = \frac{\text{Change in I/p offset voltage}}{\text{Change in temperature}}$



Thermal drift in Input Bias current $= \frac{\Delta I_B}{\Delta T}$

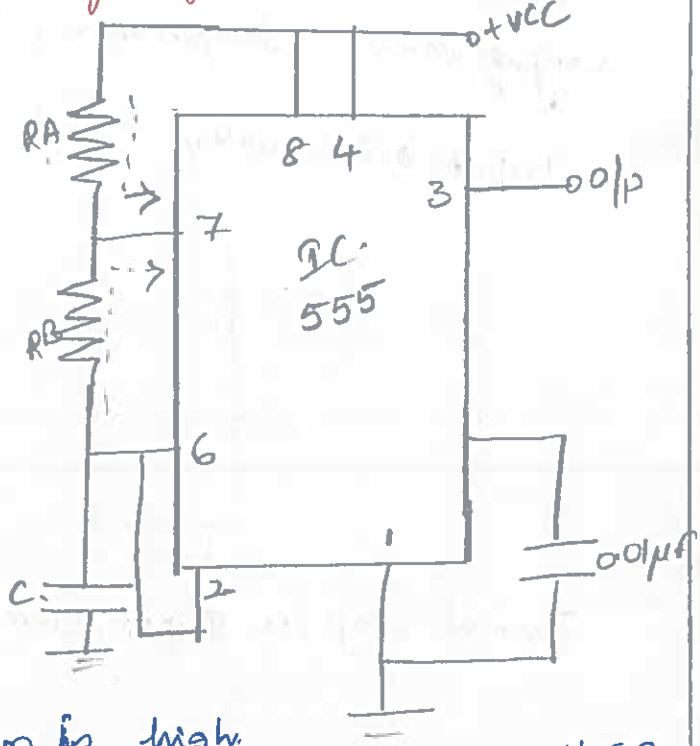


Input Bias Current drift



Input offset current drift $= \frac{\Delta I_{ios}}{\Delta T}$

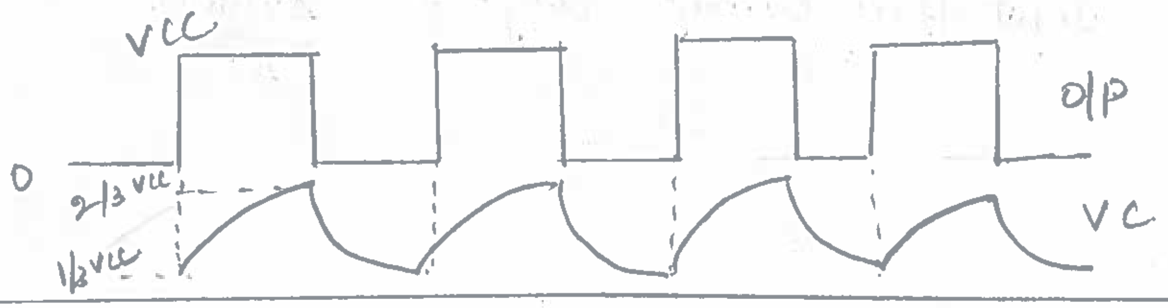
8 (a) Explain the working of an Astable Multivibrator using 555 timer with relevant circuit and waveforms and derive the expression for frequency of operation and duty cycle.



IC 555 timer connected for Astable operation in the fig. Timing Resistor is split into 2 sections RA & RB. Pin 7 of discharging transistor Q_1 is connected b/w RA & RB. When Vcc is connected external timing capacitor charges towards Vcc with time constant $(R_A+R_B)C$.

During this time o/p pin is high. $(S=1, R=0, Q=1, \bar{Q}=0)$ which unclamps the timing capacitor C. When capacitor voltage equals to $2/3$ Vcc the upper comparator triggers the flipflop & $Q=0$ & $\bar{Q}=1$ which inturn makes Q_1 ON & capacitor discharges towards ground through RB. RA & RB must be large to limit current & prevent damage to Q_1.

During discharge when C reaches Vcc/3 the lower comparator triggers $(S=1, R=0, Q=1, \bar{Q}=0)$ enclamps the timing capacitor C. C is charged b/w $2/3$ Vcc & $1/3$ Vcc & discharges



The capacitor voltage for a low pass circuit subjected to step i/p of vcc

$$v_c = V_{cc}(1-e^{-t/RC})$$

Time taken by the circuit to change from 0 to $2/3 V_{cc}$

$$2/3 V_{cc} = V_{cc}(1-e^{-t_1/RC})$$

$$t_1 = 1.09RC$$

The time t_2 taken to change from 0 to $1/3 V_{cc}$

$$\frac{1}{3} V_{cc} = V_{cc}(1-e^{-t_2/RC})$$

$$t_2 = 0.405RC$$

The time taken to charge from $1/3 V_{cc}$ to $2/3 V_{cc}$

$$t_{High} = t_1 - t_2$$

$$= 1.09RC - 0.405RC$$

$$t_{high} = 0.69RC = 0.69(R_A+R_B)C$$

The o/p is low when capacitor discharges from $2/3 V_{cc}$ to $1/3 V_{cc}$

$$\frac{1}{3} V_{cc} = \frac{2}{3} V_{cc} e^{-t/RC}$$

$$e^{-t/RC} = 1/2$$

$$T/RC = \ln(2)$$

$$T = RC\ln(2)$$

$$T = 0.69RC$$

$$T = 0.69R_B C$$

Total Time $T = T_{High} + T_{low}$

$$T = 0.69(R_A+R_B)C + 0.69R_B C$$

$$= 0.69(R_A+2R_B)C$$

Duty cycle % $= \frac{T_{High}}{T} \times 100$ or $\frac{T_{low}}{T} \times 100$

$$= \frac{0.69(R_A+R_B)C}{0.69(R_A+2R_B)C} \times 100 \text{ or } \frac{0.69R_B C}{0.69(R_A+2R_B)C} \times 100$$

Frequency $= 1/T = \frac{1}{0.69(R_A+2R_B)C} = \frac{1.45}{(R_A+2R_B)C}$

(8)(b) Design a 555 based Astable Multivibrator to generate an o/p signal with frequency 2.5kHz & duty cycle 50%. ($C = 0.01\mu F$)

$$\text{Timeperiod} = \frac{1}{f} = \frac{1}{2.5} = 0.4\,ms$$

Duty cycle = 50%

$T_H = T_L$

$T = T_H + T_L$

$0.4\,ms = 2T_L$

$T_L = 0.2\,ms$

$T_H = 0.2\,ms$

$$0.5 = \frac{T_H}{T} \times 100$$

$$0.5 = \frac{R_A + R_B}{R_A + 2R_B}$$

$$0.5R_A + \cancel{R_B} = R_A + \cancel{R_B}$$

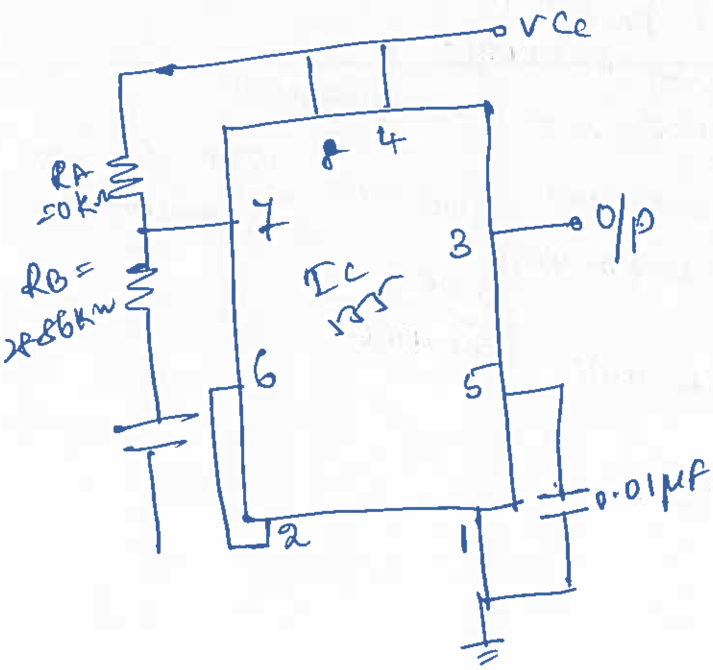
$T_H = 0.69(R_A + R_B)C$

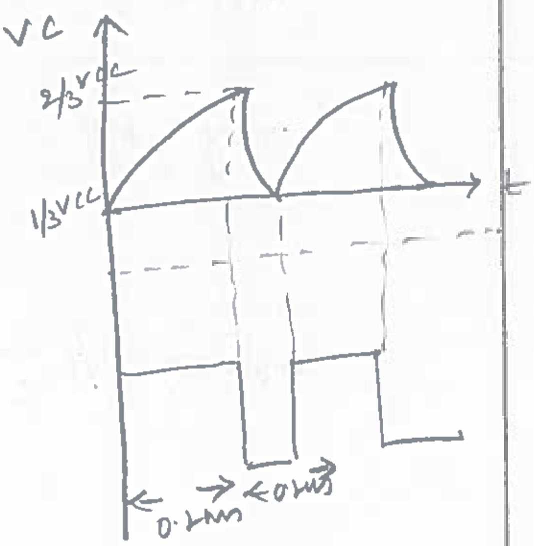
$0.2 \times 10^{-3} = 0.69(R_A + R_B)C$

$= 0.69(R_B) \times 0.01 \times 10^{-6}$

$$R_B = \frac{0.2 \times 10^{-3}}{0.69 \times 0.01 \times 10^{-6}}$$

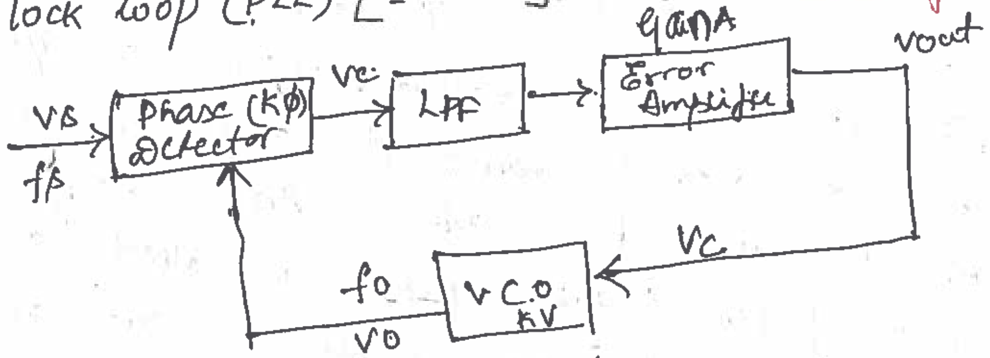
$R_B = 28.86\,k\Omega$





(9) Explain the working of PLL & derive an expression for lock-in range of PLL

Phase lock loop (PLL) [IC 565]



PLL consists of (i) Phase detector
(ii) Low pass filter
(iii) Error Amplifier
(iv) VCO (Voltage controlled oscillator)

Phase detector compares the i/p frequency with feed back frequency & generate an o/p signal which is a function of the difference b/w the phase of two i/p signals

$$V_e = K\phi(\theta_e - \pi/2) \qquad \theta_e = \text{phase error.}$$

The o/p voltage of phase detector is DC voltage, which is applied to low pass filter to remove high frequency noise components. The o/p of the filter is amplified by gain A, & then applied as control voltage V_c to VCO.

$$V_c = A \cdot V_e = K\phi A(\theta_e - \pi/2)$$

control voltage V_c will result in a shift in the VCO frequency from its center frequency f_o to a frequency

$$f = f_o + K_v \cdot V_c$$

when PLL is locked to the i/p signal frequency.

$$f = f_s = f_o + K_v \cdot V_c$$

Substitute V_c value.

$$f_s - f_o = K_v \cdot K\phi A(\theta_e - \pi/2)$$

($\because K_v$ = voltage to frequency transfer coefficient of VCO)

$$(\theta_e - \pi/2) = \frac{f_s - f_o}{K_v \cdot K\phi \cdot A}$$

$$\theta_e = \pi/2 + \frac{f_s - f_o}{K_v K\phi \cdot A}$$

Maximum o/p voltage magnitude available from phase detector occurs for $\theta = \pi$ & 0 radians.

$$\boxed{V_{c(max)} = \pm K\phi(\pi/2)}$$

The error or control voltage applied as i/p to VCO forces the VCO to change its O/P frequency in the direction to reduce the difference b/w i/p frequency & o/p frequency of VCO. & this process is called "capturing". This is done till o/p frequency of VCO is same as i/p signal frequency.

When the two frequencies are equal then the circuit is said to be locked.

$$V_{c\,max} = \pm K\phi(\pi/2)A$$

$$f = f_s = f_0 + K_V K\phi(\pi/2)A$$

$$= f_0 \pm \Delta f_L$$

$2\Delta f_L$ will be locking range

$$2\Delta f_L = K_V K\phi A \cdot \pi$$

where $\Delta f_L = K_V K\phi A\, \pi/2$

The locking range is symmetrically located w.r.t to VCO free running frequency f_0.

$$K_V = \frac{8f_0}{V}$$

$$V = +V_{CC} - (V_{CL})$$

$$K\phi = \frac{1.4}{\pi} \quad \& \quad A = 1.4$$

Substitute the values of (K_V, $K\phi$) in Δf_L equation

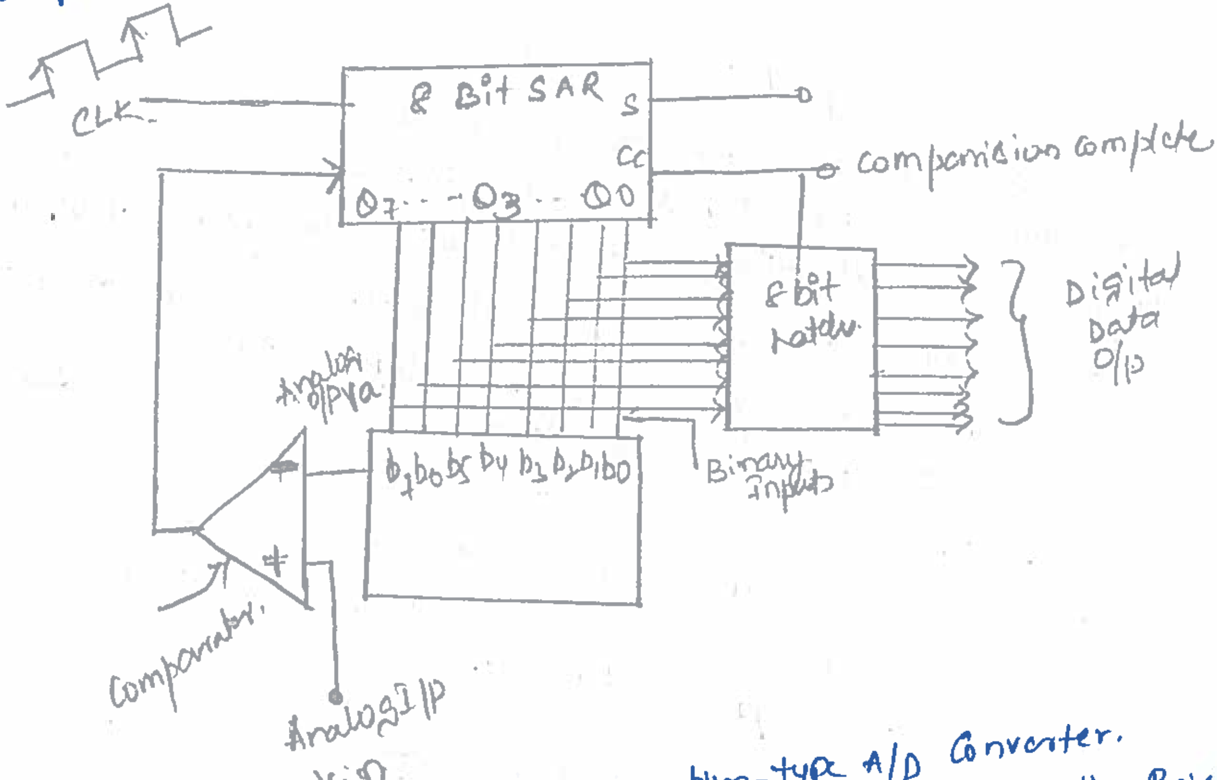
$$\Delta f_L = \pm \frac{8f_0}{V} \times \frac{1.4}{\pi} \times 1.4 \times \pi/2$$

$$\boxed{\Delta f_L = \pm \frac{7.84 f_0}{V}}$$

(10) Explain Successive Approximation ADC with neat diagram

Conversion time is maintained constant in successive Approximation type A/D converter & is proportional to the no. of bits in the digital o/p

The basic principle of this A/D converter is the unknown analog I/p voltage is approximated as n-bit digital value by trying one bit at a time, beginning with MSB.



Successive Approximation-type A/D Converter.

It consists of 8 bit SAR (successive Approximation Register) 8 bit D/A converter. The analog whose o/p is applied to then compared to an analog i/p o/p Va of D/A converter comparator. The o/p of the comparator signal (Vin) by the SAR. SAR adjusts its digital is 8 bit data i/p to equivalent to analog i/p Vin. The 8 bit o/p until it is end of the conversion holds the latch at the resultant digital data o/p

At the start of conversion cycle, SAR is reset by holding S signal high on the first clk pulse Q_7 of the SAR is set. The D/A converter generates an analog equivalent to the Q_7 bit which is compared with analog i/p V_{in}

If comparator o/p is Low D/A O/P > V_{in} SAR will clear its MSB at Q_7

If comparator o/p is high D/A O/P < V_{in} & SAR will keep MSB Q_7 set.

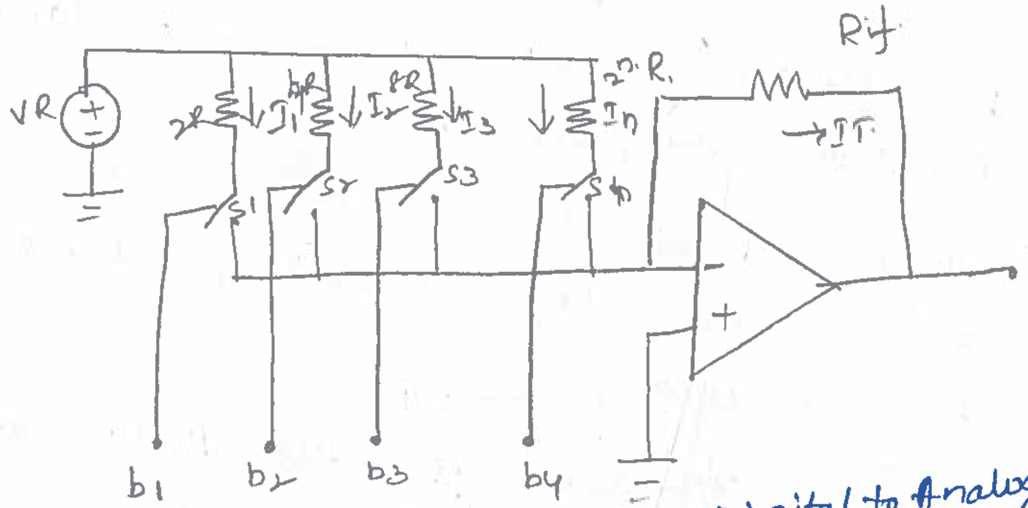
Consequently SAR tries all the bits. As soon as LSB Q_0 is tried SAR forces the conversion complete signal high to indicate the parallel o/p lines contain valid data. CC signal in turn enables the latch & digital data appears at the o/p of the latch

Resolution are the advantages highspeed & excellent

D/A converter o/p voltage get successively closer to analog i/p voltage V_i. For an 8 bit converter, it requires 8 pulses to complete the o/p

Corrected digital Representation	SAR o/p V_i at different stages of conversion	Comparator o/p
11010100	1000 0000	1
	1100 0000	1
	1110 0000	0
	1101 0000	1
	1101 1000	0
	1101 0100	1
	1101 0110	0
	1101 0101	0
	1101 0100	

(11) Explain the Binary weighted Resistor and R-2R type DAC with necessary equations & write any 4 advantages and disadvantages.



A Binary weighted Resistor Digital to Analog Converter has an op-amp to sum n binary weighted currents derived from a reference voltage VR via current scaling resistors 2R, 4R, 8R ...... $2^n R$. Switch positions are controlled by digital the I/p's. When digital I/p is logic 1, then it connects the corresponding resistance to the reference voltage. otherwise it leaves resistor open.

On switch $I = \frac{VR}{R}$

Off switch $I = 0$

Op-amp is used as a summing amplifier. Due to high I/p impedance of an op-amp total current flows through Rf

$I_T = I_1 + I_2 + I_3 + \dots I_n$

$V_0 = -I_T \cdot R_f$

$= -(I_1 + I_2 + I_3 + \dots I_n) R_f$

$= -\left(b_1 \frac{VR}{2R} + b_2 \frac{V_2}{4R} + \frac{b_3 VR}{8R} + \dots \frac{b_n VR}{2^n R} \right) R_f$

$= -\frac{VR}{R}(b_1 2^{-1} + b_2 2^{-2} + b_3 2^{-3} \dots + b_n 2^{-n}) R_f$

$R_f = R$ $\quad V_0 = -VR(b_1 2^{-1} + b_2 2^{-2} + b_3 2^{-3} \dots\dots + b_n 2^{-n})$

Analog o/p voltage is proportional to the i/p digital word.

Draw backs

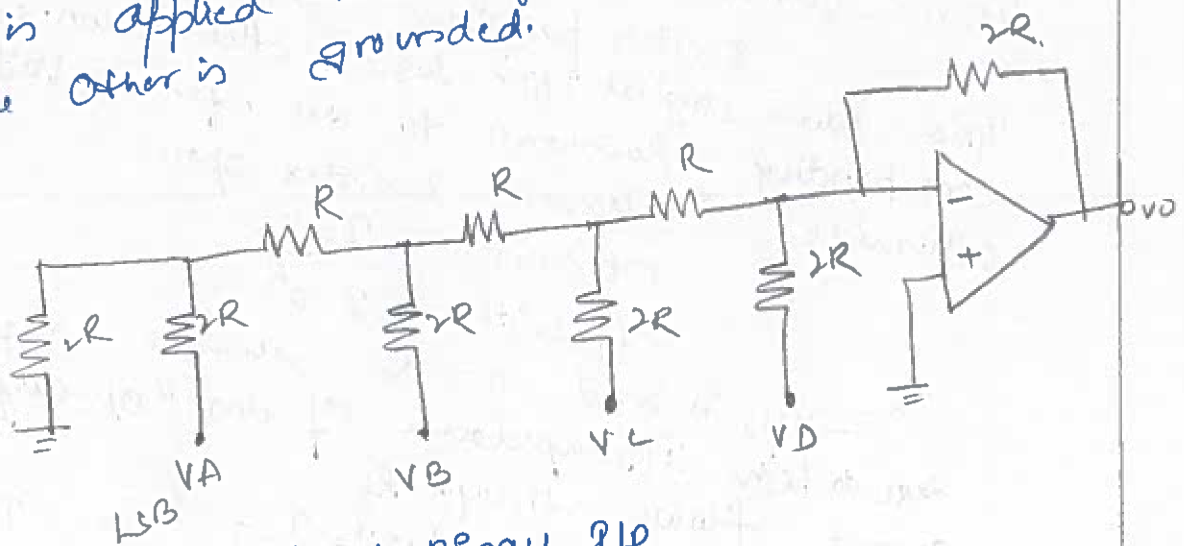
① wide range of Resistor values are required for 8-bit DAC, $2^1R, 2^2R \ldots 2^8 \cdot R$, largest Resistor is 128 times the small one

② Impractical to fabricate large Resistor values on IC & voltage drop across large Resistor affects the accuracy.

To over come this R-2R ladder Network is evolved.

R-2R Ladder Network :-

In R-2R ladder DAC method, Reference voltage is applied to one of the switch positions & the other is grounded.



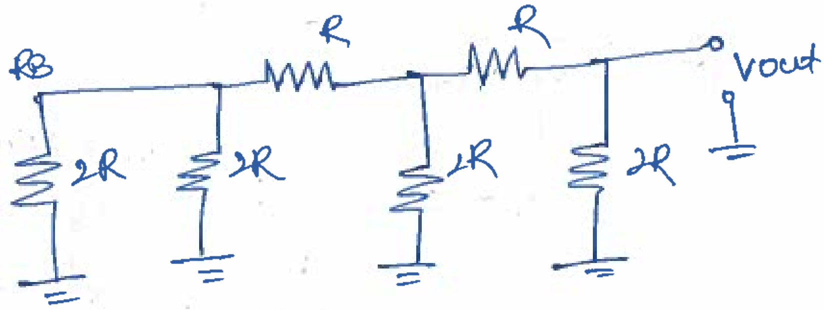
4 Bit Binary I/P

Principle of operation:-

Assume all the binary I/p's are Grounded at 0V. $V_A = V_B = V_C = V_D = 0V$. Binary Code corresponding to four i/p's "0000". Calculate the Equivalent Resistance by using parallel & series combination of Resistances 2R & R

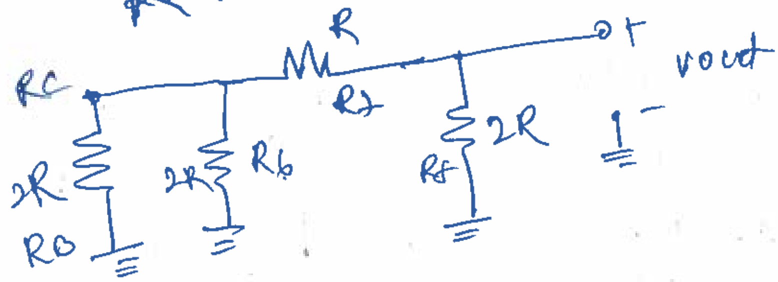
$$\frac{R_1 \times R_2}{R_1 + R_2} = \frac{R_1 R_2}{R_1 + R_2} = \frac{2R \times 2R}{2R + 2R} = R$$

R is in series with $R(R_3) =$ $R + R = 2R$.



$$R_B = \frac{2R \times 2R}{2R + 2R} + R5$$

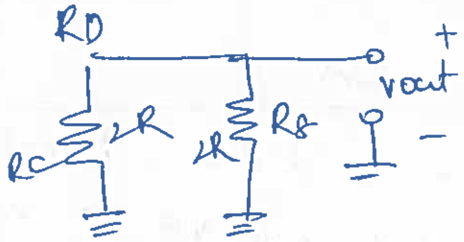
$$= \cancel{R}R + R = 2R.$$



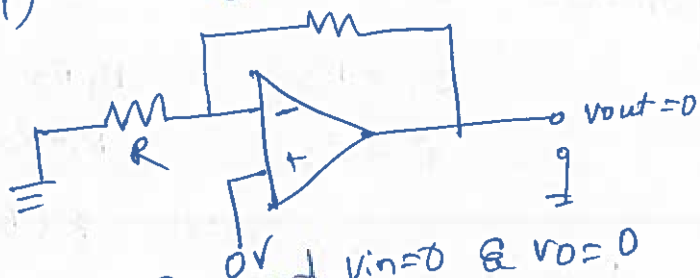
$$R_C = R7 + \frac{RB \times R6}{RB + R6}$$

$$= R + \frac{2R \times 2R}{2R + 2R}$$

$$= R + R = 2R$$

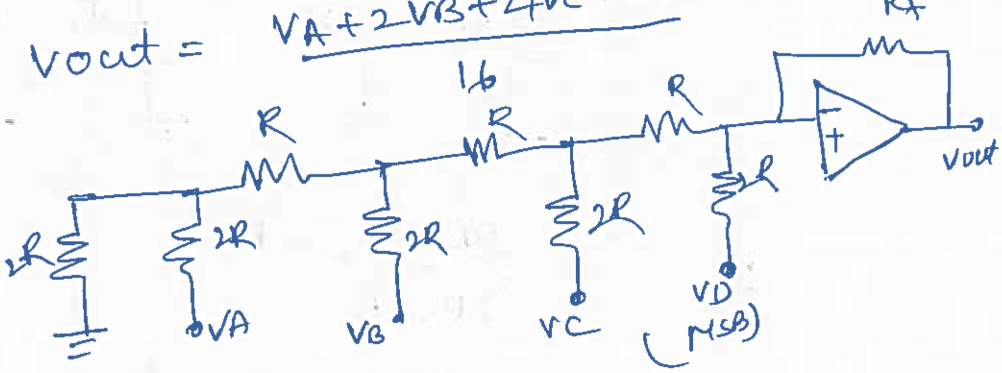


R-2R with "0000" I/P
O/P voltage for an Inverting op-amp is
(Rf/R1) Vin if Rf=R=1.



~~As~~ R terminated to Ground Vin=0 & Vo=0
For "0000" i/p voltage Vo=0

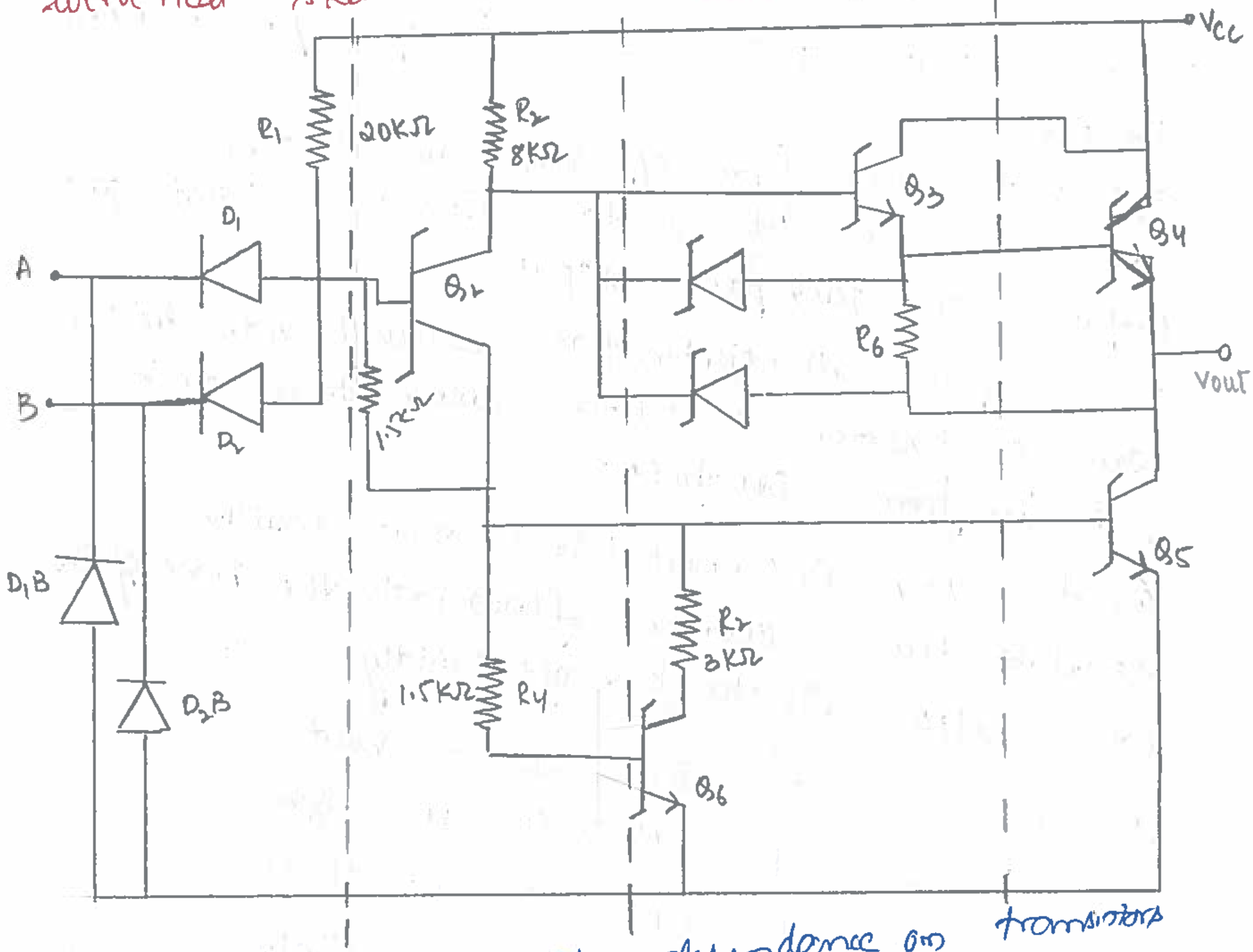
$$Vout = \frac{VA+2VB+4VC+8VD}{16}$$



Advantages
(1) Easier to build accurately due precision resistors R, 2R.
(2) No. of bits can be expanded by adding more sections of R/2R values of same.

Disadvantages
(1) Current through Resistors changes as I/p binary data changes
(2) Power dissipation causes heating
(3) Non linearity in DAC

(12) Explain the 2 input TTL NAND Gate logic in detail with neat sketch and its truthTable.



TTL logic is named as its dependence on transistors alone to perform logic operations.
The fig shows basic 2 i/p TTL NAND Gate (74LS00). NAND gate function is obtained by inverting the O/P of the AND gate.
There are 3 functional parts.

1. Diode gate & I/P protection
2. Phase Shifter.
3. O/P stage.

Circuit Operation :- D_1 & D_2 with 20kΩ Resistor forms 2 I/P AND Gate, D_2B & D_1B are used as protective diodes which protect the circuit from large Negative transients on Input lines.

Q_2 acts as an inverter along with surrounding transistors and resistors acts as phase shifter. Diode AND Gate & phase shifter finaly represents NAND function

Q_4 & Q_5 transistors form O/P stage in which only one transistor is ON at any time called as Totem pole output or push pull output

Q_3 & Q_4 forms Darlington pair which provides higher gain & shorter rise time when transistor is switching from OFF to ON.

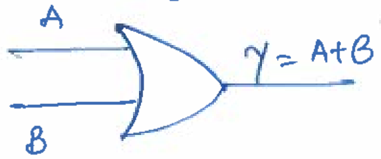
Q_6 transistor is connected in series which regulates the current flow into the base of Q_5 which helps Q_5 to turn off Rapidly.

A	B	Q_2	Q_3	Q_4	Q_6	Q_5	Vout
0	0	0	1	ON	0	Off	High
0	1	0	1	ON	0	off	High
1	0	0	1	ON	0	Off	High
1	1	1	0	Off	1	ON	Low.

Case i) & Case ii) & case iii) for these three cases $Q_2 = 0$
$A=0, B=0$, $A=0, B=1$, $A=1, B=0$ the high o/p at collector Q_2 is fed as I/p to Q_3 which makes Q_3 ON., inturn Q_4 is also turned off. so High o/p is observed at Vout
ON., Q_5 = OFF so AND o/p is High
case iv) when both i/p's $A=1$ & $B=1$ Q_2 = ON. Low o/p at collector of Q_2 is fed as I/p to Q_3 which makes it off & Q_4 OFF; Q_5 is ON so a Low o/p is observed at Vout.

13. Realize CMOS OR & NOR gate using CMOS logic with function table.

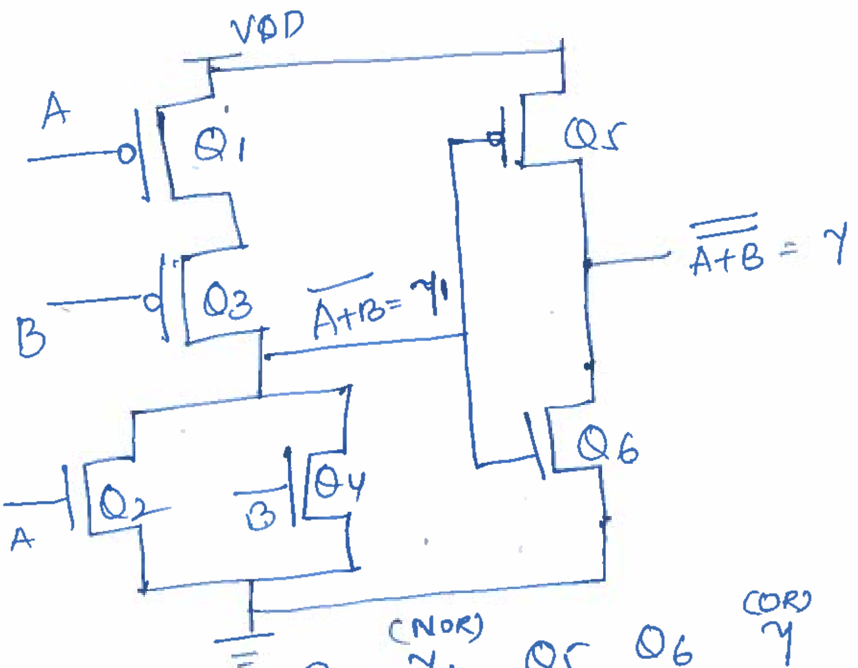
OR Gate :- In Digital logic Design OR Gate is one of the basic logic gate.



A	B	Y
0	0	0
0	1	1
1	0	1
1	1	1

OR gate output is logic high when one of I/p's is high.

OR gate is implemented using CMOS circuits. As the no of I/p's are 2 (A, B) we need 2NMOS, 2PMOS for design of the NOR Gate. OR gate again an inverter is Required which consists of 1NMOS & PMOS.



A	B	Q1	Q3	Q2	Q4	(NOR) Y_1	Q5	Q6	(OR) Y
L	L	ON	ON	OFF	OFF	H	OFF	ON	L
L	H	ON	OFF	OFF	ON	L	ON	OFF	H
H	L	OFF	ON	ON	OFF	L	ON	OFF	H
H	H	OFF	OFF	ON	ON	L	ON	OFF	H

Case(i) A = B = Low

Q_1, Q_3 ON, Q_2 & $Q_4 \rightarrow$ OFF.

when Both PMOS are ON there is a conducting path from VDD to O/P as NMOS (Q_2 & Q_4 are OFF) so they are considered as open circuits. supply voltage appears at the O/P terminal.

Case(ii) A = Low B = High

Q_1 ON Q_3 OFF

when Q_3 is OFF it is observed as open circuit then supply voltage can't appear at O/P terminal.

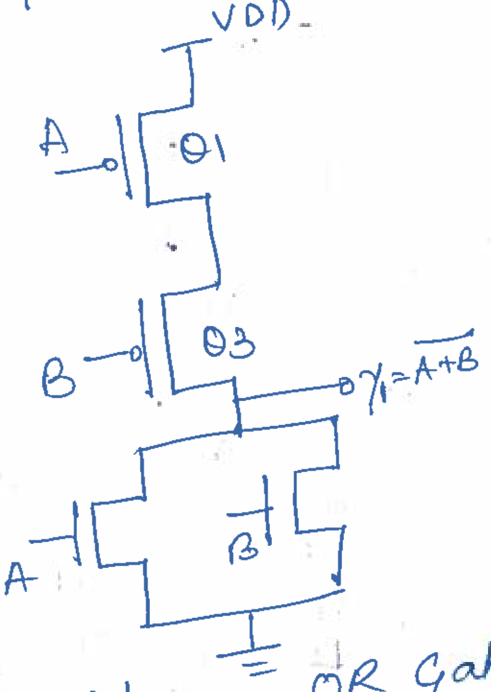
Similarly in case(iii) A = High B = Low

Q_1 OFF Q_3 ON. Q_1 is open ckt. &

Output = 0.

Case(iv) when A = High B = High

Q_1 OFF Q_3 OFF So Both transistors are observed as open circuits So there is no path from VDD to O/P terminal.



NOR Gate.

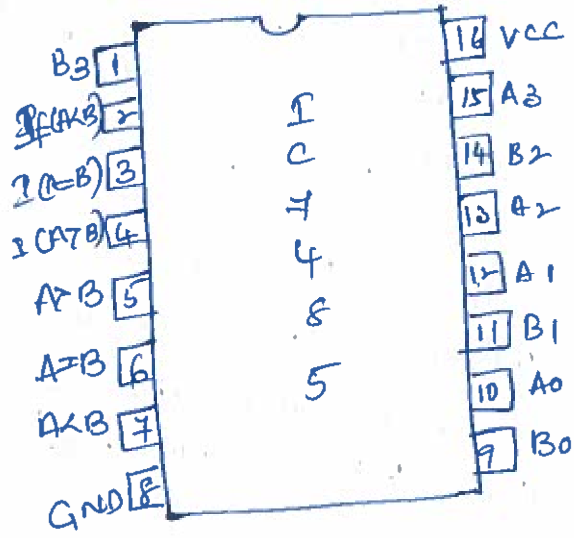
A	B	Q_1	Q_3	Q_2	Q_4	Y_1
L	L	ON	ON	OFF	OFF	H
L	H	ON	OFF	OFF	ON	L
H	L	OFF	ON	ON	OFF	L
H	H	OFF	OFF	ON	ON	L

To obtain OR Gate Inverter is connected at the O/P of the NOR Gate

14 a

Discuss the Magnitude Comparator (IC 7485)

IC 7485 is a magnitude comparator which compares Binary Numbers. of 4 bits.



$A_3 A_2 A_1 A_0$, $B_3 B_2 B_1 B_0$ are two binary inputs which provides 8 o/p's $A>B$, $A<B$, $A=B$.

VHDL code

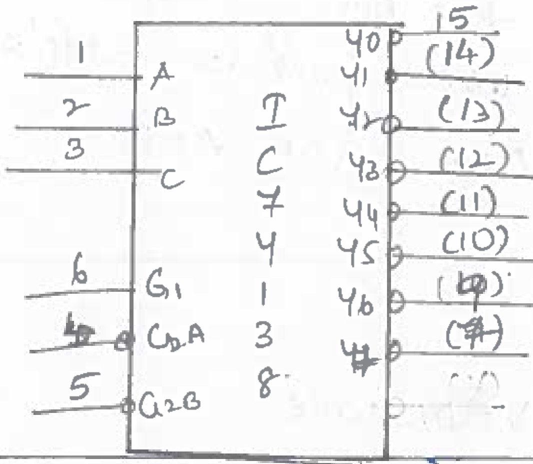
```
library IEEE;
Use IEEE.STD_logic_1164.all
Use. IEEE.std_logic_arith.all;
Use. IEEE std_logic_unsigned.all;

Entity comparator is
    port (a,b: in std_logic_vector (3 downto 0);
    Greater : out std_logic;
    smaller : out std_logic;
    equal   : out std_logic);
end comparator
Architecture behavioral of comparator is
begin
Greater <='1' when (a>b) else '0';
Smaller <='1' when (a<b) else '0';
equal <='1' when (a=b) else '0';
end behavioral;
```

(5) Explain any two 74 series Combinational logic IC's with example.

IC 74138 is a 3 to 8 decoder which accepts 3 binary Inputs & provides 8 active low O/p's. It has 2 active low enable pins (G_{2A} & G_{2B}), active high enable pin (G_1)

O/p is active low & is observed when $G_{2A}=0$, $G_{2B}=0$ & $G_1=1$. For every combination of I/p's One o/p is highlighted.

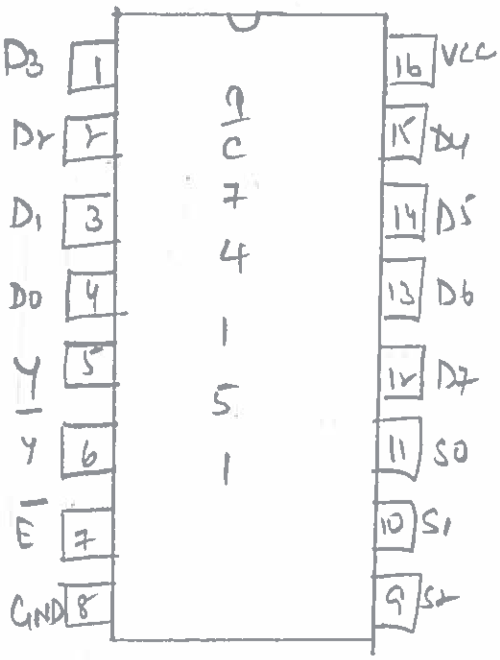


Enable			Inputs			O/p's							
G_{2B}	G_{2A}	G_1	C	B	A	$\overline{Y_7}$	$\overline{Y_6}$	$\overline{Y_5}$	$\overline{Y_4}$	$\overline{Y_3}$	$\overline{Y_2}$	$\overline{Y_1}$	$\overline{Y_0}$
1	X	X	X	X	X	1	1	1	1	1	1	1	1
X	1	X	X	X	X	1	1	1	1	1	1	1	1
X	X	0	X	X	X	1	1	1	1	1	1	1	1
0	0	1	0	0	0	1	1	1	1	1	1	1	0
0	0	1	0	0	1	1	1	1	1	1	1	0	1
0	0	1	0	1	0	1	1	1	1	1	0	1	1
0	0	1	0	1	1	1	1	1	1	0	1	1	1
0	0	1	1	0	0	1	1	1	0	1	1	1	1
0	0	1	1	0	1	1	1	0	1	1	1	1	1
0	0	1	1	1	0	1	0	1	1	1	1	1	1
0	0	1	1	1	1	0	1	1	1	1	1	1	1

Multiplexer (IC 74151) 8x1 MUX :-

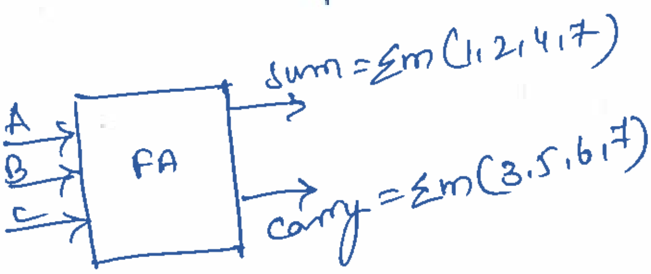
A Multiplexer is a combinational logic circuit which multiplexes the information from 2^n no. of Inputs on a single o/p line depending on n. no of selection lines.

E	S_2	S_1	S_0	Y	$\bar{Y}$
0	0	0	0	I_0	$\bar{I}_0$
0	0	0	1	I_1	$\bar{I}_1$
0	0	1	0	I_2	$\bar{I}_2$
0	0	1	1	I_3	$\bar{I}_3$
0	1	0	0	I_4	$\bar{I}_4$
0	1	0	1	I_5	$\bar{I}_5$
0	1	1	0	I_6	$\bar{I}_6$
0	1	1	1	I_7	$\bar{I}_7$
1	x	x	x	X	X



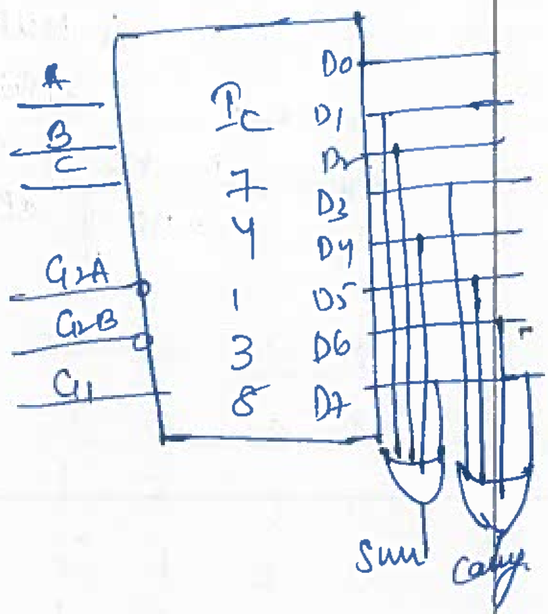
Implement full adder with mux (4x1)

Full adder is a combinational logic circuit which is used to perform addition of 3 binary bits and provides sum & carry as outputs.



A	B	C	Sum	Carry
0	0	0	0	0
0	0	1	1	0
0	1	0	1	0
0	1	1	0	1
1	0	0	1	0
1	0	1	0	1
1	1	0	0	1
1	1	1	1	1

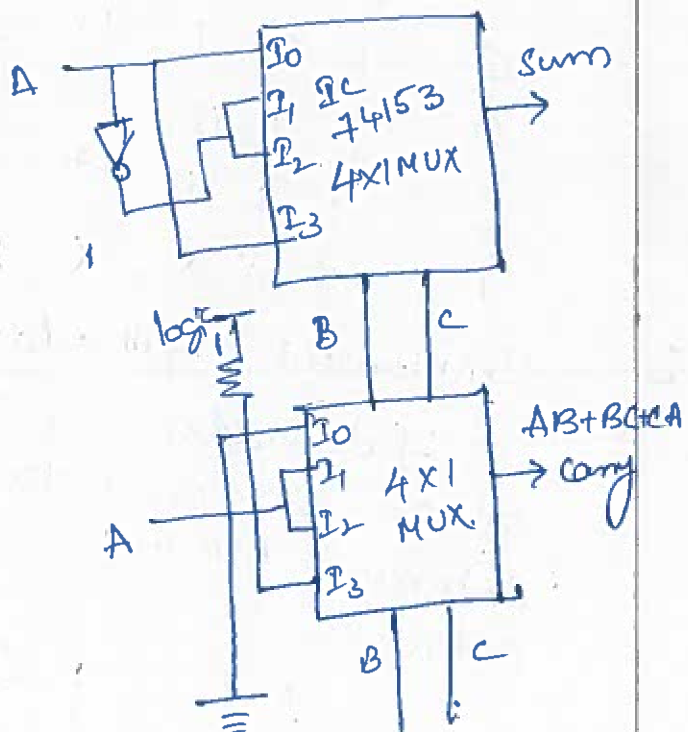
3 to 8 decoder



Sum

	I_0	I_1	I_2	I_3
$\bar{A}$	0	1	2	3
A	4	5	6	7
	A	$\bar{A}$	$\bar{A}$	A

$A \oplus B \oplus C$



Carry

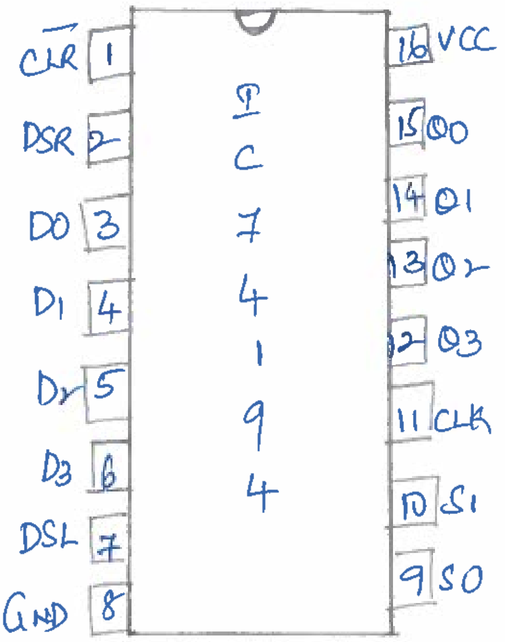
	I_0	I_1	I_2	I_3
$\bar{A}$	0	1	2	3
A	4	5	6	7
	0	A	A	1

15. Explain the concept of universal shift Register with necessary diagrams.

Universal shift Register (IC 74194)

A register which is capable of performing bidirectional shift along with parallel load capabilities is called as Universal shift Register.

Mode control S1	Mode control S0	Register operation
0	0	No change
0	1	Right shift
1	0	Left shift
1	1	Parallel load.



Pin diagram of Universal shift Register (IC 74194)

Function Table :-

Operation mode	CLK	$\overline{CLR}$	S1	S0	DSR	DSL	Dn	Q0	Q1	Q2	Q3
Reset	X	0	X	X	X	X	X	0	0	0	0
Shift left	↑	1	1	0	X	0	X	Q1	Q2	Q3	0
	↑	1	1	0	X	1	X	Q1	Q2	Q3	1
Shift Right	↑	1	0	1	0	X	X	0	Q0	Q1	Q2
	↑	1	0	1	1	X	X	1	Q0	Q1	Q2
parallel load	↑	1	1	1	X	X	Dn	D0	D1	D2	D3
Hold	↑	1	1	0	X	X	X	Q0	Q1	Q2	Q3

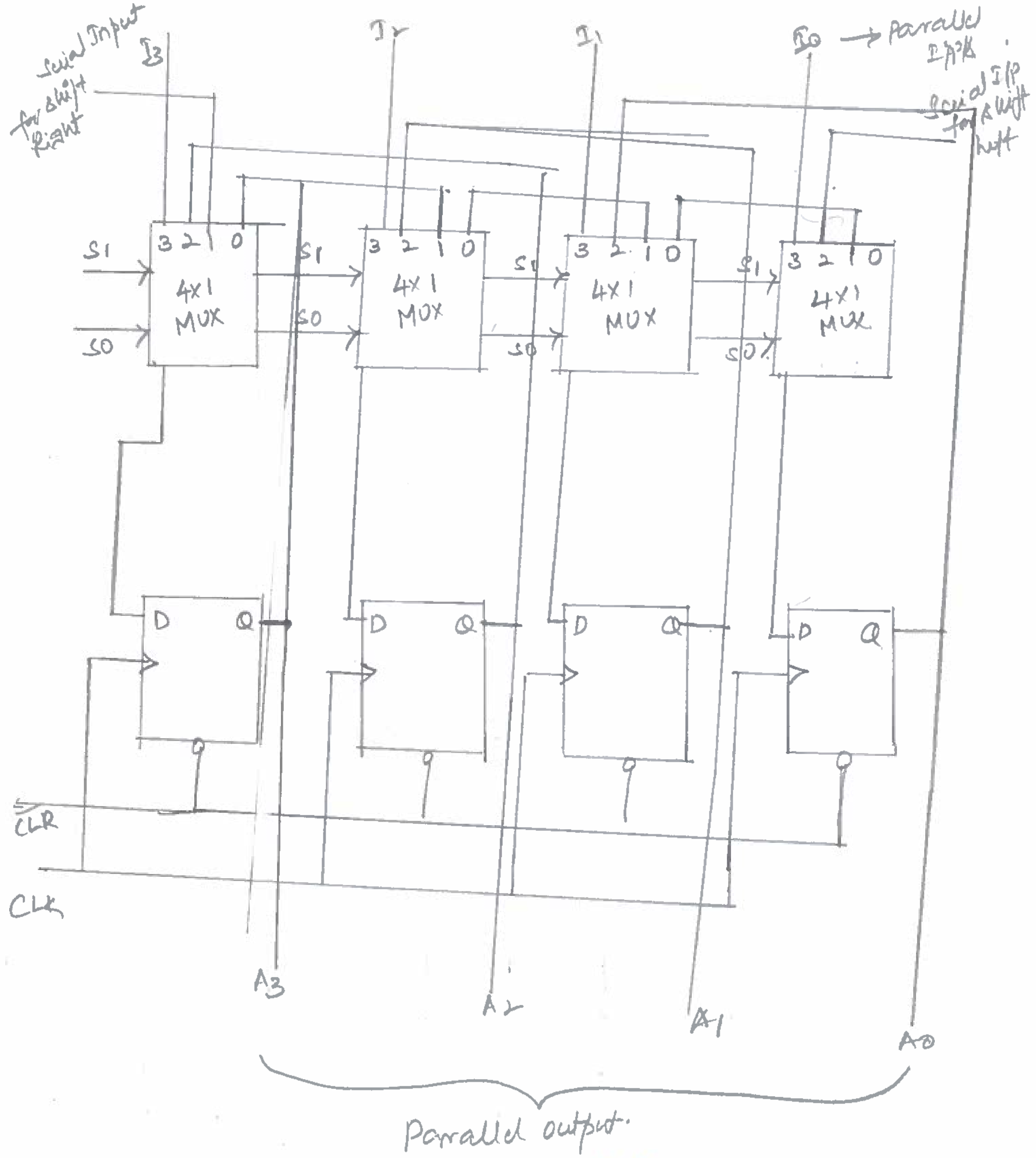


Fig:- 4 bit Universal Shift Register

VHDL code :-

```
Library IEEE;
use IEEE.std_logic_1164.all.
Entity univReg. is
    port (clk  : in std_logic;
          s    : in std_logic_vector (1 downto 0);
          il,ir : in std_logic;
          i    : in std_logic_vector (3 downto 0);
          q    : out std_logic_vector (3 downto 0));
end univreg;
Architecture behavioral of univreg is
signal x : std_logic_vector (3 downto 0);
begin
process (clk)
begin
  if (clk='1') then
case s is
when "00" => x <= "0000";
when "01" => x <= il & i(3 downto 1);
when "10" => x <= i (2 downto 1) & ir;
when "11" => x <= i;
when others => null;
end case;
end if;
end process;
q <= x;
end behavioral
```

Roja
01/12/2022

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE & CSE (AI & ML)	Academic Year	2022 – 2023
Course Code	20AI502	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Artificial Intelligence				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four applications of AI.	20AI502.1	L1
2	Distinguish between toy and real-world problems.	20AI502.2	L1
3	Define propositional calculus.	20AI502.3	L1
4	List the four kinds of knowledge approaches which need to be represented in the AI system.	20AI502.4	L1
5	Mention some applications of expert system.	20AI502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the four categories AI system.	6M	20AI502.1	L2
6 (b)	Summarize the applications of Artificial Intelligence.	6M	20AI502.1	L2
	OR			
7 (a)	How intelligence and artifact do helps artificial intelligence to succeed?	6M	20AI502.1	L2
7 (b)	Summarize the six foundations that compose most of AI concepts.	6M	20AI502.1	L2
8 (a)	Define state space search. Solve the water jug problem using production rules.	6M	20AI502.2	L2
8 (b)	Explain the various types hill climbing algorithm heuristic search techniques.	6M	20AI502.2	L2
	OR			
9 (a)	Discuss how well the standard approach to game playing would apply to games such which take place in a continuous physical state space.	6M	20AI502.2	L2
9 (b)	Compare Greedy best-first search and A* search strategies.	6M	20AI502.2	L2
10 (a)	Explain predicate logic with suitable example. .	6M	20AI502.3	L2
10 (b)	Write the propositional resolution refutation with an example.	6M	20AI502.3	L2
	OR			
11 (a)	Explain semantic tableau calculus for classical propositional logic.	6M	20AI502.3	L2
11 (b)	Discuss some examples of axiomatic systems.	6M	20AI502.3	L2
12 (a)	Explain the knowledge representation using semantic networks.	6M	20AI502.4	L2
12 (b)	Summarize the four approaches to knowledge representation in AI.	6M	20AI502.4	L2
	OR			
13 (a)	Explain how conceptual dependency is used to represent knowledge acquired from natural language input.	6M	20AI502.4	L2
13 (b)	Explain the components of script with an example of restaurant.	6M	20AI502.4	L2
14 (a)	Summarize the applications of the expert system.	6M	20AI502.5	L2
14(b)	Explain the phases in building an expert system.	6M	20AI502.5	L2
	OR			
15 (a)	Compare the expert system and traditional system with an example.	6M	20AI502.5	L2
15 (b)	Explain the truth maintenance system with an example.	6M	20AI502.5	L2

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SCHEME OF VALUATION
&
ANSWER KEY

Degree	B. Tech. (U. G.)	**Program**	CSE	**Test**	**END EXAM**	**Academic Year**	2022 - 2023
Course Code	20AI502	**Test Duration**	180 Min.	**Max. Marks**	70	**Semester**	I
Course	**ARTIFICIAL INTELLIGENCE**						
Assessment Pattern							

R (L1):		U (L2):		Apply (L3):		Analyze (L4):		E (L5):		C (L6)	

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	**List any four applications of AI.**	20AI502.1	**L1**
	For this we can give two marks for any four applications of AI AI in Business AI in Engineering AI in Manufacturing AI in Medical Field AI in Mining AI in Telecommunications AI in Banking AI in Agriculture or Farming AI in Education		
2	**Distinguish between toy and real-world problems.**	20AI502.2	**L1**
	A toy problem is intended to illustrate or exercise various problem-solving methods. These include sliding-block puzzles, N-Queens problem, missionaries and cannibals problem, tic-tac-toe,		

	chess, Tower of Hanoi and others. Real -World problems effectively solved by the AI are Online shopping made easy, consumer queries resolved, frauds preventing, in various medical diagnosis.		
3	**Define propositional calculus.**	20AI502.3	L1
	Propositional Calculus (PC) is a language of propositions basically refers to set of rules used to combine the propositions to form compound propositions using logical operators often called connectives such as Λ, V, ~, →, ↔		
4	**List the four kinds of knowledge approaches which need to be represented in the AI system.**	20AI502.4	L1
	The four kinds of knowledge approaches which need to be represented in the AI system are: 1. Knowledge representation using Semantic Networks 2. Knowledge representation using Frames 3. Knowledge representation using Extended Semantic Networks 4. Knowledge representation using Conceptual Dependency 5. Knowledge representation using Scripts		
5	**Mention some applications of expert system. 2M**	20AI502.5	L2
	Application of expert systems are: MYCIN DENDRAL CaDet Information management Hospitals and medical facilities Help desks management Employee performance evaluation Loan analysis Virus detection Useful for repair and maintenance projects Warehouse optimization Planning and scheduling		

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Learning Outcome (s)	DoK
6 (a)	**Explain the four categories AI system.** **For listing four categories – 1 M** **For the explanation of all categories – 5 M**	20AI502.1	L2
	Categories of AI System		

	1. Systems that think like humans 2. Systems that act like humans 3. Systems that think rationally 4. Systems that act rationally **Systems that think like humans** ▪ Most of the time it is a black box where we are not clear about our thought process. ▪ One has to know functioning of brain and its mechanism for prossessing information. ▪ It is an area of cognitive science. ❑ The stimuli are converted into mental representation. ❑ Cognitive processes manipulate representation to build new representations that are used to generate actions. ▪ Neural network is a computing model for processing information similar to brain. **Systems that act like humans** ▪ The overall behaviour of the system should be human like. ▪ It could be achieved by observation **Systems that think rationally** ▪ Such systems rely on logic rather than human to measure correctness. ▪ For thinking rationally or logically, logic formulas and theories are used for synthesizing outcomes. ▪ For example, ❑ given John is a human and all humans are mortal then one can conclude logically that John is mortal ▪ Not all intelligent behavior are mediated by logical deliberation. **Systems that act rationally** ▪ Rational behavior means doing right thing. ▪ Even if method is illogical, the observed behavior must be rational.		
6 (b)	Summarize the applications of Artificial Intelligence. For Summarizing or Explanation of the applications of AI – 6 M For writing any applications we can give 6 marks	20AI502.1	L2

	Some of the applications of AI are: AI in Medicine AI in Banking AI in telecommunications AI in Education AI in Gaming AI in Health Care AI in Astronomy AI in Finance AI in Data Security AI in Social Media AI in Travel& Transport AI in Automotive Industry AI in Agricuture AI in E-Commerce		
	OR		
7 (a)	**How do intelligence and artifact helps artificial intelligence to succeed?** **For writing, how the intelligence and artifact do helps in AI to succeed – 6 Marks**	20AI502.1	L2
	Artificial intelligence (AI) and cognitive science are two distinct disciplines, with overlapping methodologies but with rather different goals. AI is a branch of computer science and is concerned with construction and deployment of intelligent agents as computer programs, and also with understanding the behavior of these artifacts. The core scientific goal of AI is to understand the basic principles of intelligent behavior that apply equally to animal and artificial systems. Almost all of the work is mathematical or computational in character and much of the literature is technique oriented. Cognitive science is an explicitly interdisciplinary field that has participation not only from AI, but also from linguistics, philosophy, psychology, and subfields of other social and biological sciences. The unifying goal of cognitive science is to understand and model human intelligence, using the full range of findings and methodologies of the complementary disciplines. As one would expect, a wide range of techniques from the mathematical, behavioral, social, and biological sciences are employed. Cognitive science, in contrast with AI, is defined more by phenomena than by methodology. There are research groups that are active in both AI and cognitive science, but they tend to produce different types of reports for journals and conferences in the two areas.		
7 (b)	**Summarize the six foundations that compose most of AI concepts.** **For writing the six foundations of AI –2 Marks** **For Summarizing the foundations –4 Marks**	20AI502.1	L2
	Foundations of AI Foundation of AI is based on • Mathematics • Neuroscience • Control Theory • Linguistics • Psychology • Computer Science		

Foundations – Mathematics

- More formal logical methods
- Boolean logic
- Fuzzy logic
- Uncertainty

The basis for most modern approaches to handle uncertainty in AI applications can be handled by

- ✓ Probability theory
- ✓ Modal and Temporal logics

Foundations – Neuroscience

- How do the brain works?
- Early studies (1824) relied on injured and abnormal people to understand what parts of brain work
- More recent studies use accurate sensors to correlate brain activity to human thought

By monitoring individual neurons, monkeys can now control a computer mouse using thought alone

- Moore's law states that computers will have as many gates as humans have neurons in 2020
- How close are we to have a mechanical brain?
- Parallel computation, remapping, interconnections,....

Foundations – Control Theory

- Machines can modify their behavior in response to the environment (sense/action loop)
- Water-flow regulator, steam engine governor, thermostat
- The theory of stable feedback systems (1894)
- Build systems that transition from initial state to goal state with minimum energy
- In 1950, control theory could only describe linear systems and AI largely rose as a response to this shortcoming

Foundations – Linguistics

- ☐ Speech demonstrates so much of human intelligence
  - ❑ Analysis of human language reveals thought taking place in ways not understood in other settings
    - ■ Children can create sentences they have never heard before
    - ■ Language and thought are believed to be tightly intertwined

8 (a)	**Define state space search. Solve the water jug problem using production rules.** **For defining state space search --- 2 Marks** **For solving the water jug problem using production rules ------------ 4 Marks** **They can consider any problem statement related to water jug problem**	20AI502.2	**L2**

The state space is the configuration of the possible states and how they connect to each other i.e., legal moves between the states.
State Space Search is a process used in the field of computer science, including AI, in which successive configurations or states of an instance are considered with the intention of a goal state.

Example : Water Jug Problem
Problem statement:
- Given two jugs, a 4-gallon and 3-gallon having no measuring markers on them. There is a pump that can be used to fill the jugs with water. How can you get exactly 2 gallons of water into 4-gallon jug.

Solution:
- State for this problem can be described as the set of ordered pairs of integers (X, Y) such that
- X represents the number of gallons of water in 4-gallon jug and
- Y for 3-gallon jug.

- Start state is (0,0)
- Goal state is (2, N) for any value of N.

Production Rules
Following are the production rules for this problem.

R1: (X, Y | X < 4) → (4, Y)
{Fill 4-gallon jug}

R2: (X, Y | Y < 3) → (X, 3)
{Fill 3-gallon jug}

R3: (X, Y | X > 0) → (0, Y)
{Empty 4-gallon jug}

R4: (X, Y | Y > 0) → (X, 0)
{Empty 3-gallon jug}

R5: (X, Y | X+Y >= 4 Λ Y > 0) →(4, Y – (4 – X))
{Pour water from 3- gallon jug into 4-gallon jug until 4-gallon jug is full}

R6: (X, Y | X+Y >= 3 ☐ X > 0) ☐ (X – (3 – Y), 3)

{Pour water from 4-gallon jug into 3- gallon jug until 3-gallon jug is full}

R7: (X, Y | X+Y <= 4 ☐ Y > 0) ☐ (X+Y, 0)

{Pour all water from 3-gallon jug into 4-gallon jug }

R8: (X, Y | X+Y <= 3 ☐ X > 0) ☐ (0, X+Y)

{Pour all water from 4-gallon jug into 3-gallon jug }

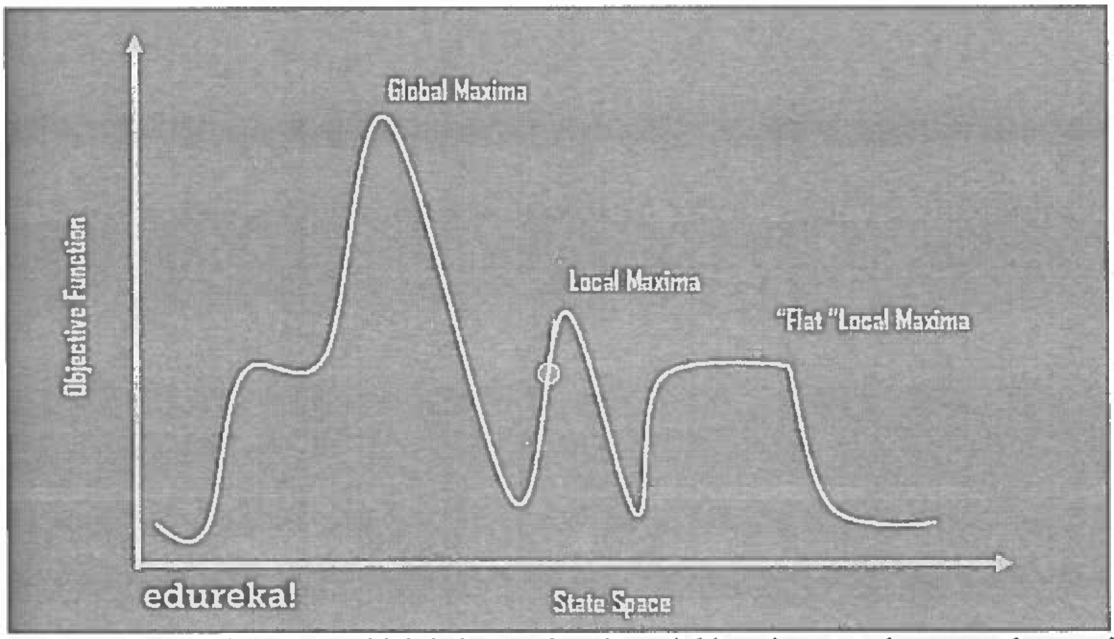
8 (b)	**Explain the various types hill climbing algorithm heuristic search techniques.** **For explaining Hill Climbing – 2 Marks** **For Describing types of Hill Climbing --- 4 Marks**	20AI502.2	L2

Hill Climbing is a heuristic search used for mathematical optimisation problems in the field of Artificial Intelligence.

So, given a large set of inputs and a good heuristic function, the algorithm tries to find the best possible solution to the problem in the most reasonable time period. This solution may not be the absolute best(global optimal maximum) but it is sufficiently good considering the time allotted.

The definition above implies that hill-climbing solves the problems where we need to maximise or minimise a given real function by selecting values from the given inputs.

State Space diagram for Hill Climbing



Local maxima: It is a state which is better than its neighbouring state however there exists a state which is better than it (global maximum). This state is better because here the value of the objective function is higher than its neighbours.

Global maxima: It is the best possible state in the state space diagram. This because at this state, objective function has the highest value.

Plateau/flat local maxima: It is a flat region of state space where neighbouring states have the same value.

Ridge: It is a region which is higher than its neighbour's but itself has a slope. It is a special kind of local maximum.

Types of Hill Climbing

1. **Simple Hill Climbing**
Simple hill climbing is the simplest way to implement a hill-climbing algorithm. It only evaluates the neighbour node state at a time and selects the first one which optimizes current cost and set it as a current state. It only checks it's one successor state, and if it finds better than the current state, then move else be in the same state. This algorithm has the following features:

Less time consuming

Less optimal solution

The solution is not guaranteed

	2. Steepest-Ascent hill climbing The steepest-Ascent algorithm is a variation of the simple hill-climbing algorithm. This algorithm examines all the neighbouring nodes of the current state and selects one neighbour node which is closest to the goal state. This algorithm consumes more time as it searches for multiple neighbours. **3. Stochastic hill climbing** Stochastic hill climbing does not examine for all its neighbours before moving. Rather, this search algorithm selects one neighbour node at random and evaluate it as a current state or examine another state.		
9 (a)	**Discuss how well the standard approach to game playing would apply to games such which take place in a continuous physical state space.** **----- 6 Marks**	20AI502.2	L2
	The space of actions of these new games is now continuous. For example, in pool, the cueing direction, angle of elevation, speed, and point of contact with the cue ball are all continuous quantities. The simplest solution for these types of games is to discretize the action space and then apply standard methods. This might work for tennis (modeled crudely as alternating shots with speed and direction), but for games such as pool and croquet it is likely to fail miserably because small changes in direction have large effects on action outcome. Instead, one must analyze the game to identify a discrete set of meaningful local goals, such as "potting the 4-ball" in pool or "laying up for the next hoop" in croquet. Then, in the current context, a local optimization routine can work out the best way to achieve each local goal, resulting in a discrete set of possible choices. Typically, these games are stochastic, so the backgammon model is appropriate provided that we use sampled outcomes instead of summing over all outcomes. Whereas pool and croquet are modeled correctly as turn-taking games, tennis is not. While one player is moving to the ball, the other player is moving to anticipate the opponent's return. In general, it may be reasonable to derive randomized strategies in tennis so that the opponent cannot anticipate where the ball will go.		
9 (b)	**Compare Greedy best-first search and A* search strategies.** **For writing the differences (at least 6) – 6 Marks**	20AI502.2	L2

S	Parameters	Best-First Search	A* Search
1	Evaluation Function	The evaluation function for best-first search is f(n) = h(n).	The evaluation function for A* search is f(n) = h(n) + g(n).
2	Past Knowledge	This search algorithm does not involve past knowledge.	This search algorithm involves past knowledge.
3	Completeness	Best-first search is not complete.	A* search is complete.

	4	Optimal	Best-first search is not optimal as the path found may not be optimal.	A* search is optimal as the path found is always optimal.
	5	Time and Space Complexity	Its time complexity is $O(b^m)$ and space complexity can be polynomial. where b is the branching and m is the maximum depth of the search tree	Its time complexity is $O(b^m)$ and space complexity is also $O(b^m)$. where b is the branching and m is the maximum depth of the search tree
	6	Memory	It requires less memory.	It requires more memory.
	7	Type of nodes kept	It keeps all the fringe or border nodes in the memory while searching.	It keeps all the nodes in the memory while searching.

10 (a)	**Explain predicate logic with suitable example.** **For Explaining Predicate Logic or First Order Logic – 4 Marks** **With an example – 2 Marks**	20AI502.3	**L2**

- In propositional logic, we have seen that how to represent statements using propositional logic. But unfortunately, in propositional logic, we can only represent the facts, which are either true or false. PL is not sufficient to represent the complex sentences or natural language statements. The propositional logic has very limited expressive power. Consider the following sentence, which we cannot represent using PL logic.

  "Some humans are intelligent", or
  "Sachin likes cricket."

- To represent the above statements, PL logic is not sufficient, so we required some more powerful logic, such as first-order logic.

First-Order logic or Predicate Logic

- First-order logic is another way of knowledge representation in artificial intelligence. It is an extension to propositional logic.
- FOL is sufficiently expressive to represent the natural language statements in a concise way.
- **First-order logic is also known as Predicate logic or First-order predicate logic**. First-order logic is a powerful language that develops information about the objects in an easier way and can also express the relationship between those objects.
- First-order logic (like natural language) does not only assume that the world contains facts like propositional logic but also assumes the following things in the world:
- Objects: A, B, people, numbers, colors, wars, theories, squares, pits, wumpus, ...
- Relations: It can be unary relation such as: red, round, is adjacent, or n-any relation such as: the sister of, brother of, has color, comes between

Function: Father of, best friend, third inning of, end of, ...

Quantifiers in First-order logic:

- A quantifier is a language element which generates quantification, and quantification specifies the quantity of specimen in the universe of discourse.
- These are the symbols that permit to determine or identify the range and scope of the variable in the logical expression. There are two types of quantifier:
    1. Universal Quantifier, (for all, everyone, everything)
    2. Existential quantifier, (for some, at least one).

1. Universal Quantifier:
    - Universal quantifier is a symbol of logical representation, which specifies that the statement within its range is true for everything or every instance of a particular thing.
    - The Universal quantifier is represented by a symbol ∀, which resembles an inverted A.

If x is a variable, then ∀x is read as:
For all x
For each x
For every x.
Example:
All man drink coffee.
Let a variable x which refers to a cat so all x can be represented
∀x man(x) → drink (x, coffee).
It will be read as: There are all x where x is a man who drink coffee.

2. **Existential Quantifier:**

Existential quantifiers are the type of quantifiers, which express that the statement within its scope is true for at least one instance of something.

It is denoted by the logical operator ∃, which resembles as inverted E. When it is used with a predicate variable then it is called as an existential quantifier.

If x is a variable, then existential quantifier will be ∃x or ∃(x). And it will be read as:

There exists a 'x.'
For some 'x.'
For at least one 'x.
Example:
Some boys are intelligent.
∃x: boys(x) ∧ intelligent(x)

It will be read as: There are some x where x is a boy who is intelligent.

10 (b)	Write the propositional resolution refutation with an example. For explaining the Propositional Resolution Refutation --- 5 Marks With an any one example – 2 Marks	20AI502.3	L2

Resolution Refutation in PL

Resolution refutation: Another simple method to prove a formula by contradiction.

- Here negation of goal is added to given set of clauses.

- If there is a refutation in new set using resolution principle then goal is proved

- During resolution we need to identify two clauses,
  - one with positive atom (P) and other with negative atom (~ P) for the application of resolution rule.

- Resolution is based on modus ponen inference rule.

Disjunctive & Conjunctive Normal Forms

- **Disjunctive Normal Form (DNF):** A formula in the form (L11 ∧ ..... ∧ L1n) V ..... V (Lm1 ∧ ..... ∧ Lmk), where all Lij are literals.
  - Disjunctive Normal Form is disjunction of conjunctions.

- **Conjunctive Normal Form (CNF):** A formula in the form (L11 V ..... V L1n) ...... ∧ (Lp1 V ..... V Lpm) , where all Lij are literals.
  - CNF is conjunction of disjunctions or
  - CNF is conjunction of clauses

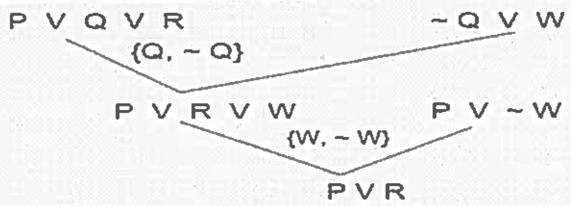
Clause: It is a formula of the form (L1V ... V Lm), where each Lk is a positive or negative atom

Resolvent of Clauses

- If two clauses C1 and C2 contain a complementary pair of literals {L, ~L},
  - then these clauses may be resolved together by deleting L from C1 and ~ L from C2 and constructing a new clause by the disjunction of the remaining literals in C1 and C2.
- The new clause thus generated is called resolvent of C1 and C2 .
  - Here C1 and C2 are called parents of resolved clause.

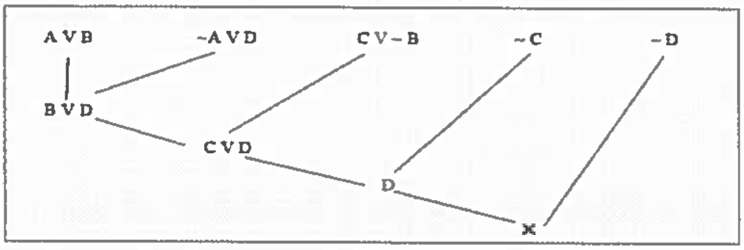
Inverted binary tree is generated with the last node (root) of the binary tree to be a resolvent. This is also called resolution tree.

- Find resolvent of the following clauses:
  - C_1 = P V Q V R; C_2 = ~ Q V W; C_3 = P V ~ W
- Inverted Resolution Tree



- Resolvent(C1,C2, C3) = P V R

- Show that C V D is a logical consequence of
  - S ={AVB, ~ AVD, C V~ B} using resolution refutation principle.
- First we will add negation of logical consequence
  - i.e., ~ (C V D) ≡ ~C Λ ~D to the set S.
  - Get S' = {A V B, ~ A V D, C V~ B, ~C, ~D}.
- Now show that S' is unsatisfiable by deriving contradiction using resolution principle.



11 (a)	**Explain semantic tableau calculus for classical propositional logic.** **For Explaining Sematic tableaux -- 3 Marks** **For writing Rules – 3 Marks**	20AI502.3	L2

Semantic Tableaux System in PL

- Earlier approaches require
  - construction of proof of a formula from given set of formulae and are called direct methods.
- In semantic tableaux,
  - the set of rules are applied systematically on a formula or set of formulae to establish its consistency or inconsistency.
- Semantic tableau
  - binary tree constructed by using semantic rules with a formula as a root
- Assume α and β be any two formulae.

Semantic Tableaux Rules

Rule 1: A tableau for a formula $(\alpha \wedge \beta)$ is constructed by adding both α and β to the same path (branch). This can be represented as follows:

$$\begin{array}{|l} \alpha \wedge \beta \\ \alpha \\ \beta \end{array}$$

Rule 2: A tableau for a formula $\sim(\alpha \wedge \beta)$ is constructed by adding two alternative paths one containing $\sim\alpha$ and other containing $\sim\beta$.

$$\begin{array}{ccc} & \sim(\alpha \wedge \beta) & \\ \sim\alpha & & \sim\beta \end{array}$$

Rule 3: A tableau for a formula $(\alpha \vee \beta)$ is constructed by adding two new paths one containing α and other containing β.

$$\begin{array}{ccc} & \alpha \vee \beta & \\ \alpha & & \beta \end{array}$$

Rule 4: A tableau for a formula $\sim(\alpha \vee \beta)$ is constructed by adding both $\sim\alpha$ and $\sim\beta$ to the same path. This can be expressed as follows:

$$\begin{array}{|l} \sim(\alpha \vee \beta) \\ \sim\alpha \\ \sim\beta \end{array}$$

Rule 5:

$$\begin{array}{|l} \sim\sim\alpha \\ \alpha \end{array}$$

Rule 6:

$$\begin{array}{ccc} & \alpha \rightarrow \beta & \\ \sim\alpha & & \beta \end{array}$$

Rule 7:

$$\begin{array}{|l} \sim(\alpha \rightarrow \beta) \\ \alpha \\ \sim\beta \end{array}$$

Rule 8: $\alpha \leftrightarrow \beta \cong (\alpha \wedge \beta) \vee (\sim\alpha \wedge \sim\beta)$

$$\begin{array}{ccc} & \alpha \leftrightarrow \beta & \\ \alpha \wedge \beta & & \sim\alpha \wedge \sim\beta \end{array}$$

Rule 9: $\sim(\alpha \leftrightarrow \beta) \cong (\alpha \wedge \sim\beta) \vee (\sim\alpha \wedge \beta)$

$$\begin{array}{ccc} & \sim(\alpha \leftrightarrow \beta) & \\ \alpha \wedge \sim\beta & & \sim\alpha \wedge \beta \end{array}$$

11 (b)	Discuss some examples of axiomatic systems. For Describing Axiomatic System and rules – 3 Marks For explaining with any one example – 3 Marks	20AI502.3	L2

Axiomatic System for Propositional Logic:

- It is based on the set of only three axioms and one rule of deduction.

– It is minimal in structure but as powerful as the truth table and natural deduction approaches.
– The proofs of the theorems are often difficult and require a guess in selection of appropriate axiom(s) and rules.

- These methods basically require forward chaining strategy where we start with the given hypotheses and prove the goal.

Axiom1 (A1): $\alpha \rightarrow (\beta \rightarrow \alpha)$

Axiom2 (A2): $(\alpha \rightarrow (\beta \rightarrow \gamma)) \rightarrow ((\alpha \rightarrow \beta) \rightarrow (\alpha \rightarrow \gamma))$

Axiom3 (A3): $(\sim \alpha \rightarrow \sim \beta) \rightarrow (\beta \rightarrow \alpha)$

Modus Ponen (MP) defined as follows:

Hypotheses: $\alpha \rightarrow \beta$ and α *Consequent:* β

Examples: Establish the following:

1. $\{Q\} \vdash (P \rightarrow Q)$ i.e., $P \rightarrow Q$ is a deductive consequence of $\{Q\}$.

{Hypothesis}	Q	(1)
{Axiom A1}	$Q \Rightarrow (P \rightarrow Q)$	(2)
{MP, (1,2)}	$P \rightarrow Q$	proved

2. $\{P \rightarrow Q, Q \rightarrow R\} \vdash (P \rightarrow R)$ i.e., $P \rightarrow R$ is a deductive consequence of $\{P \rightarrow Q, Q \rightarrow R\}$.

{Hypothesis}	$P \rightarrow Q$	(1)
{Hypothesis}	$Q \rightarrow R$	(2)
{Axiom A1}	$(Q \rightarrow R) \rightarrow (P \rightarrow (Q \rightarrow R))$	(3)
{MP, (2, 3)}	$P \rightarrow (Q \rightarrow R)$	(4)
{Axiom A2}	$(P \rightarrow (Q \rightarrow R)) \rightarrow ((P \rightarrow Q) \rightarrow (P \rightarrow R))$	(5)
{MP, (4, 5)}	$(P \rightarrow Q) \rightarrow (P \rightarrow R)$	(6)
{MP, (1, 6)}	$P \rightarrow R$	proved

12 a) Explain the knowledge representation using semantic networks. ---- 6 Marks

For Explaining Knowledge Representation using semantic networks – 4 Marks

Explaining with their own Example --- 2 Marks

Semantic Network:

- Formalism for representing information about objects, people, concepts and specific relationship between them.

- The syntax of semantic net is simple. It is a network of labeled nodes and links.
- It's a directed graph with nodes corresponding to concepts, facts, objects etc. and
- arcs showing relation or association between two concepts.

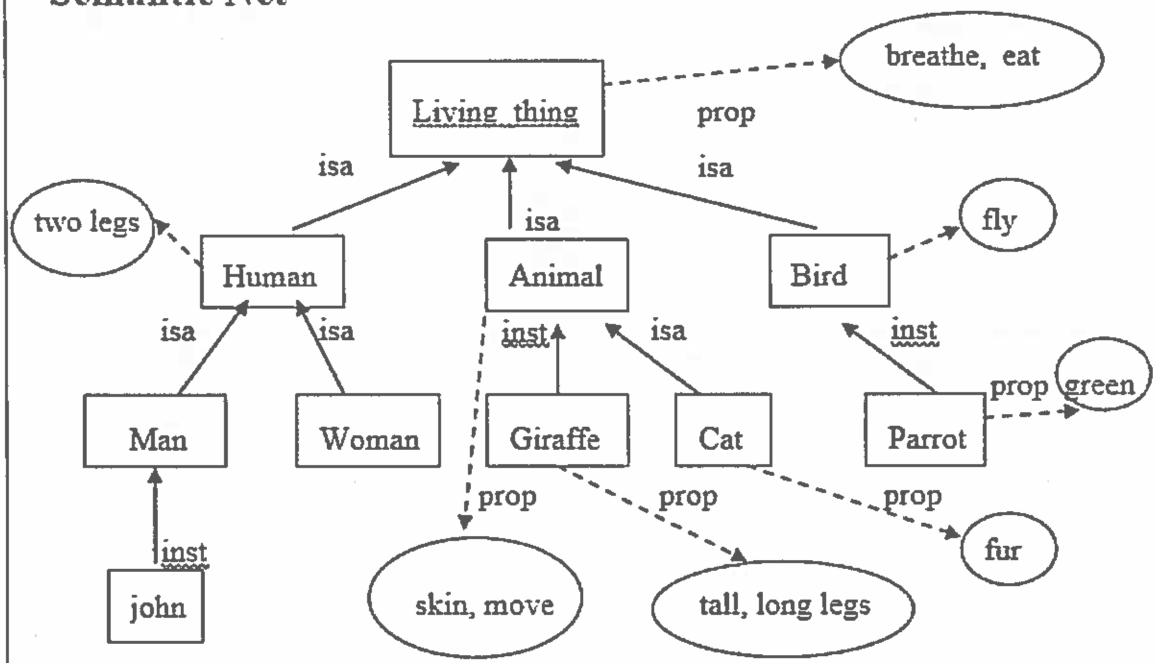
- The commonly used links in semantic net are of the following types.
  - isa → subclass of entity (e.g., child hospital is subclass of hospital)
  - inst → particular instance of a class (e.g., India is an instance of country)

prop → property link (e.g., property of dog is 'bark)

Representation of Knowledge in Sem Net

"Every human, animal and bird is living thing who breathe and eat. All birds can fly. All man and woman are humans who have two legs. Cat is an animal and has a fur. All animals have skin and can move. Giraffe is an animal who is tall and has long legs. Parrot is a bird and is green in color".

Semantic Net



Coding of Semantic Net in Prolog

Isa facts	Instance facts	Property facts
sa(living_thing, nil).	inst(john, man).	prop(breathe, living_thing).
sa(human, living_thing).	inst(giraffe, animal).	prop(eat, living_thing).
sa(animals, living_thing).	inst(parrot, bird)	prop(two_legs, human).
sa(birds, living_thing).		prop(skin, animal).
sa(man, human).		prop(move, animal).
sa(woman, human).		prop(fur, bird).
sa(cat, animal).		prop(tall, giraffe).
		prop(long_legs, giraffe).
		prop(tall, animal).
		prop(green, parrot).

12(b) **Summarize the four approaches to knowledge representation in AI.**

For explaining the any four approaches to represent knowledge – 6 Marks

The Four approaches to represent the knowledge are:

1. Knowledge representation using Semantic Networks **(Refer 12 (a))**
2. Knowledge representation using Frames
3. Knowledge representation using Extended Semantic Networks
4. Knowledge representation using Conceptual Dependency
5. Knowledge representation using Scripts

By Explaining any four approaches

13 a) **Explain how conceptual dependency is used to represent knowledge acquired from natural language input.**

For Explaining Conceptual dependency – 3 Marks

For Writing Primitive actions – 3 Marks

Conceptual Dependency (CD)

- CD theory was developed by Schank in 1973 to 1975 to represent the meaning of NL sentences.
  - It helps in drawing inferences
  - It is independent of the language
- CD representation of a sentence is not built using words in the sentence rather built using conceptual primitives which give the intended meanings of words.

	• CD provides structures and specific set of primitives from which representation can be built. **Primitive Acts of CD theory or Primitive Actions** • ATRANS Transfer of an abstract relationship (i.e. give) • PTRANS Transfer of the physical location of an object (e.g., go) • PROPEL Application of physical force to an object (e.g. push) • MOVE Movement of a body part by its owner (e.g. kick) • GRASP Grasping of an object by an action (e.g. throw) • INGEST Ingesting of an object by an animal (e.g. eat) • EXPEL Expulsion of something from the body of an animal (e.g. cry) • MTRANS Transfer of mental information (e.g. tell) • MBUILD Building new information out of old (e.g decide) • SPEAK Producing of sounds (e.g. say) • ATTEND Focusing of a sense organ toward a stimulus (e.g. listen)
13(b)	**Explain the components of script with an example of restaurant. ---- 6 Marks** **For writing the components –3 Marks** **For writing the script for restaurant – 3 Marks**
	Script Structure: • Scripts were introduced by Schank and Abelson introduced in 1977 that used CD framework. • The scripts are useful in describing certain stereotyped situations such as going to theater • It consists of set of slots containing default values along with some information about the type of values like frames. • It differs from FS as the values of the slots in scripts must be ordered and have more specialized roles. • In real world situations, we see that event tends to occur in known patterns because of clausal relationship to the occurrence of events **Script Components:** • **Each script contains the following main components.** – **Entry Conditions: Must be satisfied before events in the script can occur.** – **Results: Conditions that will be true after events in script occur.** – **Props: Slots representing objects involved in the events.** – **Roles: Persons involved in the events.**

	– **Track:Specific variation on more general pattern in the script. Different tracks may share many components of the same script but not all.** – **Scenes: The sequence of events that occur. Events are represented in conceptual dependency form.**
14 a)	**Summarize the applications of the expert system.** **About Expert System – 2 Marks** **For Writing Applications -- 4 Marks**

Expert Systems (ES):

- Expert systems are knowledge based programs which provide expert quality solutions to the problems in specific domain of applications.
- The core components of expert system are
  - knowledge base and
  - navigational capability (inference engine)
- Generally its knowledge is extracted from human experts in the domain of application by knowledge Engineer.
  - Often based on useful thumb rules and experience rather than absolute certainties.
- A process of gathering knowledge from domain expert and codifying it according to the formalism is called knowledge engineering.

Some of the popular examples of Expert Systems are:

DENDRAL

MYCIN

PXDES

CaDeT

Applications of Expert System

In designing and manufacturing domain

It can be broadly used for designing and manufacturing physical devices such as camera lenses and automobiles.

In the knowledge domain

These systems are primarily used for publishing the relevant knowledge to the users. The two popular ES used for this domain is an advisor and a tax advisor.

In the finance domain

In the finance industries, it is used to detect any type of possible fraud, suspicious activity, and advise bankers that if they should provide loans for business or not.

In the diagnosis and troubleshooting of devices

In medical diagnosis, the ES system is used, and it was the first area where these systems were used.

Planning and Scheduling

The expert systems can also be used for planning and scheduling some particular tasks for achieving the goal of that task.

14 b) **EXPLAIN THE PHASES IN BUILDING AN EXPERT SYSTEM.**

For Listing Phases – 1 Mark

For Explain Phases – 5 Marks

Phases in building Expert System

- There are different interdependent and overlapping phases in building an expert system as follows:
- **Identification Phase:**
  - Knowledge engineer finds out important features of the problem with the help of domain expert (human).
  - He tries to determine the type and scope of the problem, the kind of resources required, goal and objective of the ES.
- **Conceptualization Phase:**
  - In this phase, knowledge engineer and domain expert decide the concepts, relations and control mechanism needed to describe a problem solving.
- **Formalization Phase:**
  - It involves expressing the key concepts and relations in some framework supported by ES building tools.
  - Formalized knowledge consists of data structures, inference rules, control strategies and languages for implementation.
- **Implementation Phase:**
  - During this phase, formalized knowledge is converted to working computer program initially called prototype of the whole system.
- **Testing Phase:**
  - It involves evaluating the performance and utility of prototype systems and revising it if need be. Domain expert evaluates the prototype system and his feedback help knowledge engineer to revise it.

(OR)

15 a)	Compare the expert system and traditional system with an example. For Writing at least 6 differences ---- 6 Marks

Conventional System	Expert System
Knowledge and processing are combined in one unit.	Knowledge database and the processing mechanism are two separate components.
The programme does not make errors (Unless error in programming).	The Expert System may make a mistake.
The system is operational only when fully developed.	The expert system is optimized on an ongoing basis and can be launched with a small number of rules.
Step by step execution according to fixed algorithms is required.	Execution is done logically & heuristically.
It needs full information.	It can be functional with sufficient or insufficient information

(OR)

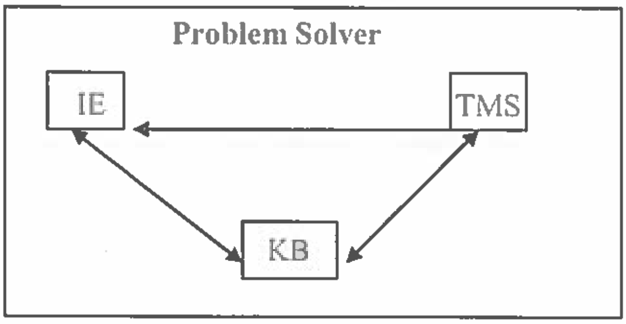
Conventional System	Expert System
Solves the generic numeric problems.	It solves the problem in very narrow domain.
It is sequential program where information and processing are combined.	The knowledge base is separated from the processing (inference engine). The program may not be sequential.
Tested program never makes mistakes	The well tested expert system may make mistakes and gives wrong answer.
No explanation is provided for output	An explanation is provided in most cases.
When incorrect information is provided, the system may not function.	The system can arrive at a conclusion, even when some information is missing or incomplete.

15 b)	**Explain the truth maintenance system with an example.** **About Truth Maintenance System – 4 Marks** **With suitable Example and diagram – 2 Marks**

Truth Maintenance System (TMS)

- Truth maintenance system (TMS) works with inference engines for solving problems within large search spaces.
- The TMS and inference engine both put together can solve problems where algorithmic solutions do not exist.

TMS maintains the beliefs for general problem solving systems.



- TMS can be used to implement monotonic or non-monotonic systems.
- In monotonic system, once a fact or piece of knowledge is stored in KB, it can not change.
  - In monotonic reasoning, the world of axioms continually increases in size and keeps on expending.
  - Predicate logic is an example of monotonic form of reasoning. It is a deductive reasoning system where new facts are derived from the known facts.
- Non-monotonic system allows retraction of truths that are present in the system whenever contradictions arise.
  - So number of axioms can both increase and decrease and depending upon the changes in KB, it can be updated.

Example – Monotonic TMS

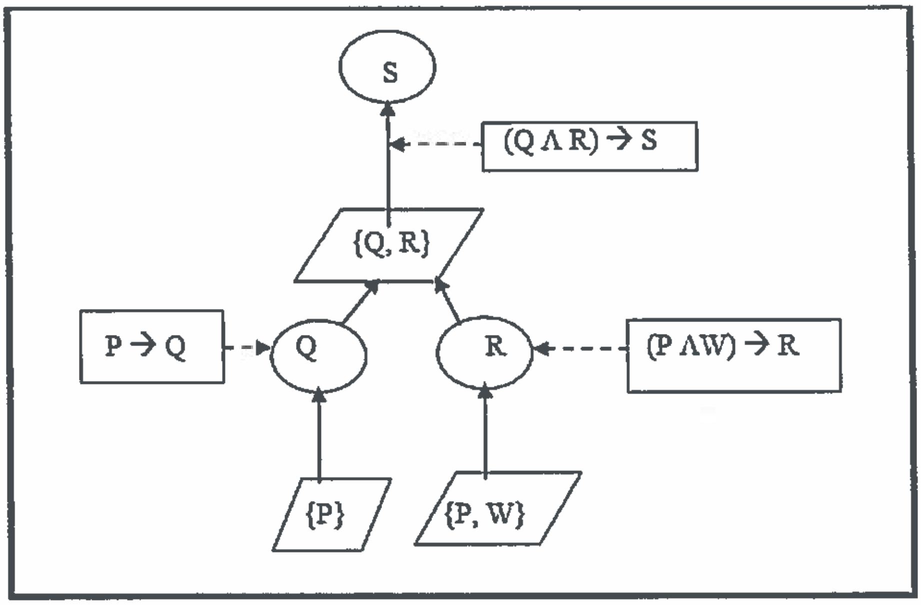
- Suppose we are given the premise set $\Sigma = \{P, W\}$ and the internal constraint set

$$\{P \rightarrow Q, (P \wedge W) \rightarrow R, (Q \wedge R) \rightarrow S\}.$$

- TMS are able to derive S from these constraints and the premise set Σ.

- TMS should provide the justifications of deriving S from constraints and premises.
- Therefore, for any given set of internal constraints and premise set Σ, if a formula S can be derived from these, then justification functions generate a justification tree for S.

Justification Tree



Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	CSE (Data Science)	**Academic Year**	2022 - 2023
Course Code	20DS502	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	V
Course	**Big Data**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List the five characteristics of big data.	20DS502.1	L1
2	List the three HDFS daemons.	20DS502.2	L1
3	Sketch the graphical representation of the spark dataframe.	20DS502.3	L1
4	Mention the two types of Apache Spark RDD operations.	20DS502.4	L1
5	Write the statement to drop database.	20DS502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

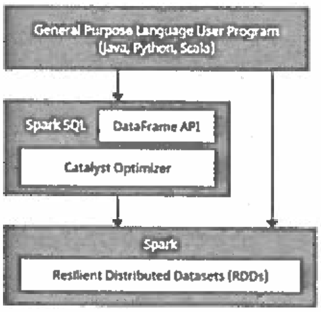
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Differentiate between structure and unstructured data.	5M	20DS502.1	L2
6 (b)	Illustrate the working of various phases of Map Reduce with appropriate example	7M	20DS502.1	L2
	OR			
7 (a)	Illustrate Hadoop architecture and its components with a neat diagram.	7M	20DS502.1	L2
7 (b)	Discuss a few benefits of Big Data.	5M	20DS502.1	L2
8 (a)	Summarize the steps to write data into HDFS.	8M	20DS502.2	L2
8 (b)	Compare pseudo-distributed and fully distributed modes.	4M	20DS502.2	L2
	OR			
9 (a)	Discuss the concept of Hadoop clusters.	5M	20DS502.2	L2
9 (b)	How will you configure XML files in Hadoop?	7M	20DS502.2	L2
10 (a)	Specify the Fundamental role of the dataframe.	5M	20DS502.3	L2
10 (b)	Illustrate the architecture of spark with a neat diagram.	7M	20DS502.3	L2
	OR			
11 (a)	Explicate how spark streaming can be used to stream live data and process the Stock Market data.	7M	20DS502.3	L2
11 (b)	Explain the process of reading a file from a local directory into DataFrame, and applying transformations before the DataFrame is written back to CSV file.	5M	20DS502.3	L2
12 (a)	Explain RDD Lineage with an example.	4M	20DS502.4	L2
12 (b)	List any ten spark pair RDD transformation functions along with its description.	8M	20DS502.4	L2
	OR			
13 (a)	Illustrate spark runtime architecture with a neat diagram.	8M	20DS502.4	L2
13 (b)	Discuss the behavior of spark streaming applications in the event of failures.	4M	20DS502.4	L2
14 (a)	Illustrate the Hive architecture with a neat diagram.	7M	20DS502.5	L2
14 (b)	Explain the hive partitioned table with an example.	5M	20DS502.5	L2
	OR			
15(a)	Illustrate the components used in Hive Query Processor with an example.	5M	20DS502.5	L2
15 (b)	Explain the table properties that can be altered with ALTER TABLE statement.	7M	20DS502.5	L2

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

1	**List the five characteristics of big data.** Five characteristics of big data are: - volume, value, variety, velocity, and veracity.
2	**List the three HDFS daemons.** Three daemons of HDFS are Namenode, Datanode, Secondary Namenode
3	**Sketch the graphical representation of the spark dataframe.** 
4	**Mention the two types of Apache Spark RDD operations.** Two types are: Transformations and Actions
5	**Write the statement to drop database.** Hive> drop database userdb;
6 (a)	**Differentiate between structure and unstructured data.** **Structured data:** Data which is in a particular format is known as structured data. It is most often categorized as quantitative data, and it's the type of data most of us are used to working with. Examples of structured data include spreadsheets, names, dates, addresses, credit card numbers, stock information, geolocation, and more. **Unstructured data:** Data which is not in a particular format is known as unstructured data. It is most often categorized as qualitative data, and it cannot be processed and analyzed using conventional data tools and methods. Examples of unstructured data include text, video files, audio files, mobile activity, social media posts, satellite imagery, surveillance imagery – the list goes on and on.

6 (b) **Illustrate the working of various phases of Map Reduce with appropriate example**

Phases of the MapReduce model: MapReduce model has 2 phases

<u>1. Map phase</u> It is the first phase of MapReduce programming and contains the coding logic of the mapper function. Mapper function accepts key-value pairs as input as (k, v), where the key represents the offset address of each record and the value represents the entire record content.

The output of the Mapper phase will also be in the key-value format as (k', v').

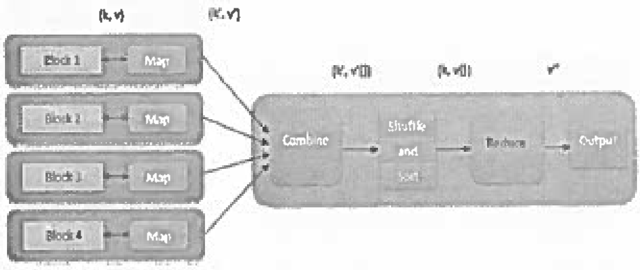
<u>Shuffle and Sort:</u> The output of various mappers (k', v'), then goes into Shuffle and Sort phase.

All the duplicate values are removed, and different values are grouped together based on similar keys.

The output of the Shuffle and Sort phase will be key-value pairs again as key and array of values (k, v[]).

<u>2. Reducer:</u> The output of the Shuffle and Sort phase (k, v[]) will be the input of the Reducer phase. In this phase reducer function's logic is executed and all the values are aggregated against their corresponding keys. Reducer consolidates outputs of various mappers and computes the final job output. The final output is then written into a single file in an output directory of HDFS.

<u>Combiner:</u> It is an optional phase in the MapReduce model. The combiner phase is used to optimize the performance of MapReduce jobs. In this phase, various outputs of the mappers are locally reduced at the node level.



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For example, MapReduce logic to find the word count on an array of words can be shown as below:

fruits_array = [apple, orange, apple, guava, grapes, orange, apple]

The mapper phase tokenizes the input array of words into the 'n' number of words to give the output as (k, v). For example, consider 'apple'. Mapper output will be (apple, 1), (apple, 1), (apple, 1).

Shuffle and Sort accept the mapper (k, v) output and group all values according to their keys as (k, v[]). i.e. (apple, [1, 1, 1]).

The Reducer phase accepts Shuffle and sort output and gives the aggregate of the values (apple, [1+1+1]), corresponding to their keys. i.e. (apple, 3).

7(a) **Illustrate Hadoop architecture and its components with a neat diagram.**

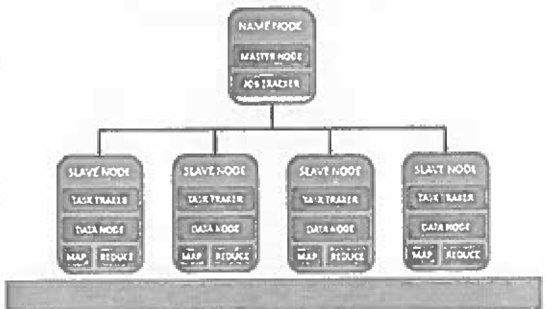
Hadoop has a Master-Slave Architecture for data storage and distributed data processing using MapReduce and HDFS methods.

The HDFS comprises the following components.

- **NameNode:** NameNode represented every files and directory which is used in the namespace
- **DataNode:** DataNode helps you to manage the state of an HDFS node and allows you to interacts with the blocks
- **Secondary Name Node:** As the name speaks, the Secondary Name Node is not a backup of the name node. It acts as a *Buffer* to the Name Node. It stores the intermediate updates the *FS-image* of the Name Node in the *Edit-log* and updates the information to the*FinalFS-image* when the name node is inactive.

Processing unit comprises the following components:

- **Job Tracker:** Assigns all the jobs to the task tracker for processing
- **Task Tracker:** Executes the jobs assigned by the job tracker.



(b) **Discuss a few benefits of Big Data.**

- Big data allows you to re-develop the products/services you are selling. Information on what others think about your products -such as through unstructured social networking site text- helps you in product development.
- Big data allows you to test different variations of CAD (computer-aided design) images to determine how minor changes

affect your process or product. This makes big data invaluable in the manufacturing process.

- Predictive analysis will keep you ahead of your competitors. Big data can facilitate this by, as an example, scanning and analyzing social media feeds and newspaper reports. Big data also helps you do health-tests on your customers, suppliers, and other stakeholders to help you reduce risks such as default.
- Big data allows you to diversify your revenue streams. Analyzing big data can give you trend-data that could help you come up with a completely new revenue stream.
- Big data is important in the healthcare industry, which is one of the last few industries still stuck with a generalized, conventional approach.

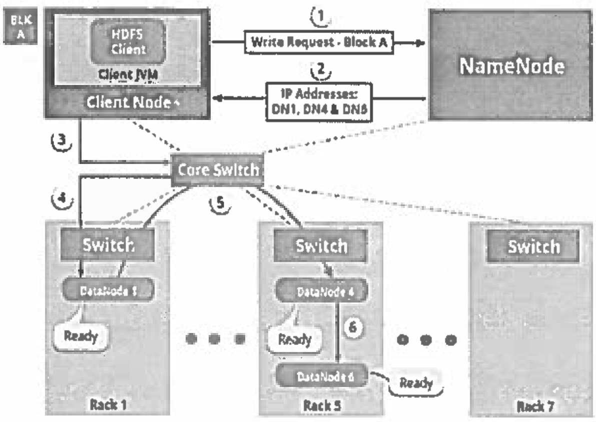
8 (a) Summarize the steps to write data into HDFS.

Write in HDFS is done in three stages:

- Set up of Pipeline
- Data streaming and replication/ HDFS write
- Shutdown of Pipeline (Acknowledgement stage)

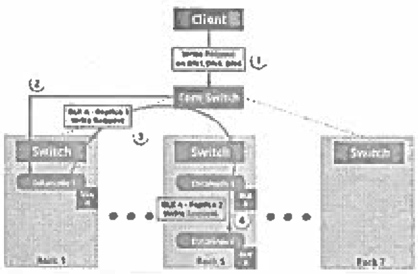
Pipeline setup:

Setting up HDFS - Write Pipeline



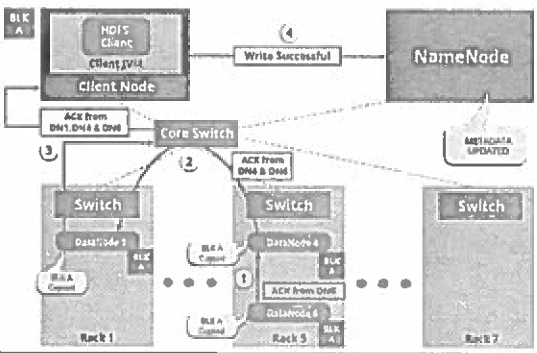
HDFS write:

HDFS - Write Pipeline



Acknowledgement:

Acknowledgement in HDFS - Write

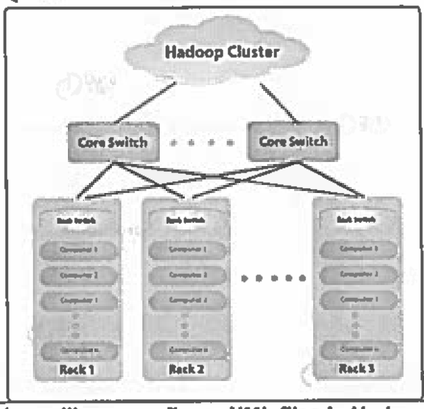


8 (b) **Compare pseudo-distributed and fully distributed modes.**

Pseudo-distributed Mode

The pseudo-distribute mode is also known as a single-node cluster where both NameNode and DataNode will reside on the same machine.

In pseudo-distributed mode, all the Hadoop daemons will be running on a single node. Such configuration is mainly used while testing

	when we don't need to think about the resources and other users sharing the resource. In this architecture, a separate JVM is spawned for every Hadoop components as they could communicate across network sockets, effectively producing a fully functioning and optimized mini-cluster on a single host. **Fully-Distributed Mode (Multi-Node Cluster)** This is the production mode of Hadoop where multiple nodes will be running. Here data will be distributed across several nodes and processing will be done on each node. Master and Slave services will be running on the separate nodes in fully-distributed Hadoop Mode. • Production phase of Hadoop • Separate nodes for master and slave daemons • Data are used and distributed across multiple nodes
9(a)	**Discuss the concept of Hadoop clusters.** Hadoop cluster is just a computer cluster which we use for Handling huge volume of data distributedly. Hadoop clusters have two types of machines, such as Master and Slave, where: • Master: HDFS NameNode, YARN ResourceManager. • Slaves: HDFS DataNodes, YARN NodeManagers. **Hadoop Cluster** 
9 (b)	**How will you configure XML files in Hadoop?** Configuration Files are the files which are located in the extracted tar.gz file in the etc/hadoop/ directory. All Configuration Files in Hadoop are listed below, **1) HADOOP-ENV.sh**->>It specifies the environment variables that affect the JDK used by Hadoop Daemon (bin/hadoop). We know that Hadoop framework is wriiten in Java and uses JRE so one of the environment variable in Hadoop Daemons is $Java_Home in Hadoop-env.sh. **2) CORE-SITE.XML**->>It is one of the important configuration files which is required for runtime environment settings of a Hadoop cluster.It informs Hadoop daemons where the NAMENODE runs in the cluster. It also informs the Name Node as to which IP and ports it should bind. **3) HDFS-SITE.XML**->>It is one of the important configuration files which is required for runtime environment settings of a Hadoop. It contains the configuration settings for NAMENODE, DATANODE, SECONDARYNODE. It is used to specify default block replication. The actual number of replications can also be specified when the file is created, **4) MAPRED-SITE.XML**->>It is one of the important configuration files which is required for runtime environment settings of a Hadoop. It contains the configuration settings for MapReduce. In this file, we specify a framework name for MapReduce, by setting the MapReduce.framework.name.
10(a)	**Specify the Fundamental role of the dataframe.** A Dataset is a distributed collection of data. Dataset is a new interface added in Spark 1.6 that provides the benefits of RDDs (strong typing, ability to use powerful lambda functions) with the benefits of Spark SQL's optimized execution engine. A Dataset can be constructed from JVM objects and then manipulated using functional transformations (map, flatMap, filter, etc.). The Dataset API is available in Scala and Java. Python does not have the support for the Dataset API. But due to Python's dynamic nature, many of the benefits of the Dataset API are already available (i.e. you can access the field of a row by name naturally row.columnName). The case for R is similar. A DataFrame is a *Dataset* organized into named columns. It is conceptually equivalent to a table in a relational database or a data frame in R/Python, but with richer optimizations under the hood. DataFrames can be constructed from a wide array of sources such as: structured data files, tables in Hive, external databases, or existing RDDs. The DataFrame API is available in Scala, Java, Python, and R. In Scala and Java, a DataFrame is represented by a Dataset of Rows. In the Scala API, DataFrame is simply a type alias of Dataset[Row]. While, in Java API, users need to use Dataset<Row> to represent a DataFrame.
10(b)	**Illustrate the architecture of spark with a neat diagram.** Spark Architecture, an open-source, framework-based component that processes a large amount of unstructured, semi-structured, and structured data for analytics, is utilised in Apache Spark.

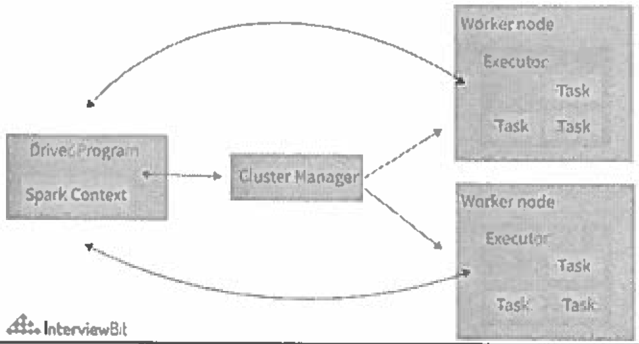
When the Driver Program in the Apache Spark architecture executes, it calls the real program of an application and creates a SparkContext. SparkContext contains all of the basic functions. The Spark Driver includes several other components, including a DAG Scheduler, Task Scheduler, Backend Scheduler, and Block Manager, all of which are responsible for translating user-written code into jobs that are actually executed on the cluster.

The Cluster Manager manages the execution of various jobs in the cluster. Spark Driver works in conjunction with the Cluster Manager to control the execution of various other jobs. The cluster Manager does the task of allocating resources for the job.

Once the job has been broken down into smaller jobs, which are then distributed to worker nodes, SparkDriver will control the execution. Many worker nodes can be used to process an RDD created in the SparkContext, and the results can also be cached.

The Spark Context receives task information from the Cluster Manager and enqueues it on worker nodes.

The executor is in charge of carrying out these duties. The lifespan of executors is the same as that of the Spark Application. We can increase the number of workers if we want to improve the performance of the system. In this way, we can divide jobs into more coherent parts.



11(a) Explicate how spark streaming can be used to stream live data and process the Stock Market data.

Apache Spark Streaming is a scalable fault-tolerant streaming processing system that natively supports both batch and streaming workloads. Spark Streaming is an extension of the core Spark API that allows data engineers and data scientists to process real-time data from various sources including (but not limited to) Kafka, Flume, and Amazon Kinesis. This processed data can be pushed out to file systems, databases, and live dashboards. Its key abstraction is a Discretized Stream or, in short, a DStream, which represents a stream of data divided into small batches. DStreams are built on RDDs, Spark's core data abstraction. This allows Spark Streaming to seamlessly integrate with any other Spark components like MLlib and Spark SQL. Spark Streaming is different from other systems that either have a processing engine designed only for streaming, or have similar batch and streaming APIs but compile internally to different engines. Spark's single execution engine and unified programming model for batch and streaming lead to some unique benefits over other traditional streaming systems.

Four Major Aspects of Spark Streaming

Fast recovery from failures and stragglers

Better load balancing and resource usage

Combining of streaming data with static datasets and interactive queries

Native integration with advanced processing libraries (SQL, machine learning, graph processing)

11(b) Explain the process of reading a file from a local directory into DataFrame, and applying transformations before the DataFrame is written back to CSV file.

PySpark provides csv("path") on DataFrameReader to read a CSV file into PySpark DataFrame and dataframeObj.write.csv("path") to save or write to the CSV file.

PySpark supports reading a CSV file with a pipe, comma, tab, space, or any other delimiter/separator files.

```
spark = SparkSession.builder().master("local[1]")
        .appName("SparkByExamples.com")
        .getOrCreate()
df = spark.read.csv("/tmp/resources/zipcodes.csv")
df.printSchema()
```

Read Multiple CSV Files

Using the read.csv() method you can also read multiple csv files, just pass all file names by separating comma as a path, for example :

```
df = spark.read.csv("path1,path2,path3")
```

Read all CSV Files in a Directory

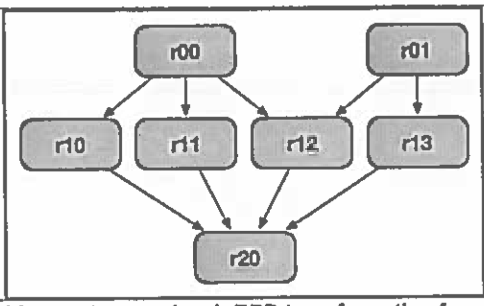
We can read all CSV files from a directory into DataFrame just by passing directory as a path to the csv() method.

```
df = spark.read.csv("Folder path")
```

12(a) Explain RDD Lineage with an example.

DD Lineage (aka RDD operator graph or RDD dependency graph) actually is a graph of all the parent RDDs of an RDD. It is built as a consequence of applying transformations to the RDD and creates a logical execution plan.

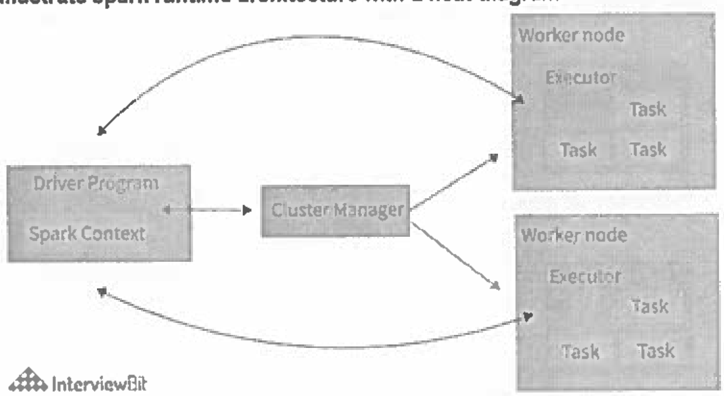
The execution DAG or physical execution plan is that the DAG of stages.



12(b) **List any ten spark pair RDD transformation functions along with its description.**

PAIR RDD FUNCTIONS	FUNCTION DESCRIPTION
aggregateByKey	Aggregate the values of each key in a data set. This function can return a different result input RDD.
combineByKey	Combines the elements for each key.
combineByKeyWithClassTag	Combines the elements for each key.
flatMapValues	It's flatten the values of each key with out changing key values and keeps the original RD
foldByKey	Merges the values of each key.
groupByKey	Returns the grouped RDD by grouping the values of each key.
mapValues	It applied a map function for each value in a pair RDD with out changing keys.
reduceByKey	Returns a merged RDD by merging the values of each key.
reduceByKeyLocally	Returns a merged RDD by merging the values of each key and final result will be sent to
sampleByKey	Returns the subset of the RDD.
subtractByKey	Return an RDD with the pairs from this whose keys are not in other.
keys	Returns all keys of this RDD as a RDD[T].

13(a) **Illustrate spark runtime architecture with a neat diagram**



- Using spark-submit, the user submits an application.
- In spark-submit, we invoke the main() method that the user specifies. It also launches the driver program.
- The driver program asks for the resources to the cluster manager that we need to launch executors.
- The cluster manager launches executors on behalf of the driver program.
- The driver process runs with the help of user application. Based on the actions and transformation on RDDs, the driver sends work to executors in the form of tasks.
- The executors process the task and the result sends back to the driver through the cluster manager. So, this was all in how Apache Spark works.

13(b) **Discuss the behavior of spark streaming applications in the event of failures.**

The recovery of failures of machines is already inbuilt in Apache spark streaming. This feature is what we call spark streaming fault tolerance property.

Fault tolerance depends on two points. The failure scenario and the type of receiver, for input sources on the basis of receivers. Basically, receivers are of two types:

Reliable receiver

As soon as the received data has been replicated, reliable sources acknowledge it. Although, as receiver fails, the source will not acknowledge it for the buffered data. Hence, as the receiver restarts next time, data will be automatically, send to the source again. Thus, there will be no data loss due to failure.

Unreliable Receiver

Unreliable receivers do not send any acknowledgment. Hence, if any failure occurs in the worker or driver, we can easily lose the data. In other words, as the receiver is reliable there will be no data loss, even if the worker node fails. While there is a definite data loss condition, in case of an unreliable receiver.

Spark Streaming Fault Tolerance – Write – ahead logs

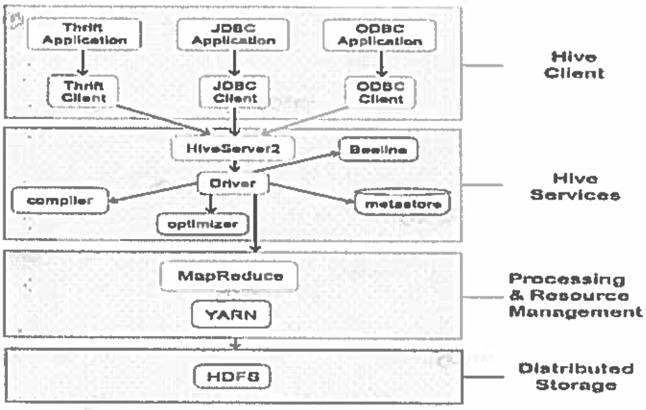
To ensure the durability of any data operations in database and file systems we use write-ahead logs, this is what we call journaling in streaming. Further, there are following steps of the process at first, an operation is written down into a durable log. Afterwards, it applies the operation to the data. Although, if in the middle of applying the operation, the system fails. We can easily recover by again applying the operations it had intended to do. Also, possible to recover it through reading the log. To receive data, sources like Kafka and Flume use receivers. In the executors, they run as long-running tasks. Also, responsible for receiving the data from the source. It acknowledges the received data if source supports it. It runs tasks on the executors to process the tasks. Also, store the received data in the memory of the executors and the driver. As soon as write ahead logs are enabled, received data is saved to log files in a fault-tolerant file system. Hence, across any failure in spark streaming, it allows the received data to be durable.

There are two conditions, ensures zero data loss, such as either recover all the data logs or all data again sent by the sources.

14(a) **Illustrate the Hive architecture with a neat diagram.**

The major components of HIVE are:

- Hive Client
- Hive Services
- Processing and Resource Management
- Distributed Storage



Hive Architecture & Its Components

1. Hive Client: Hive supports applications written in any language like Python, Java, C++, Ruby, etc. using JDBC, ODBC, and Thrift drivers, for performing queries on the Hive. Hence, one can easily write a hive client application in any language of its own choice.

Hive clients are categorized into three types:

1. Thrift Clients
2. JDBC client
3. ODBC client

2. Hive Service

To perform all queries, Hive provides various services like the Hive server2, Beeline, etc. The various services offered by Hive are:

- Beeline
- Hive Server 2
- Hive Compiler
- Optimizer
- 6. Execution Engine
- 7. Metastore

3. Processing Framework and Resource Management

Hive internally uses a MapReduce framework as a defacto engine for executing the queries.

MapReduce is a software framework for writing those applications that process a massive amount of data in parallel on the large clusters of commodity hardware. MapReduce job works by splitting data into chunks, which are processed by map-reduce tasks.

4. Ditributed Storage

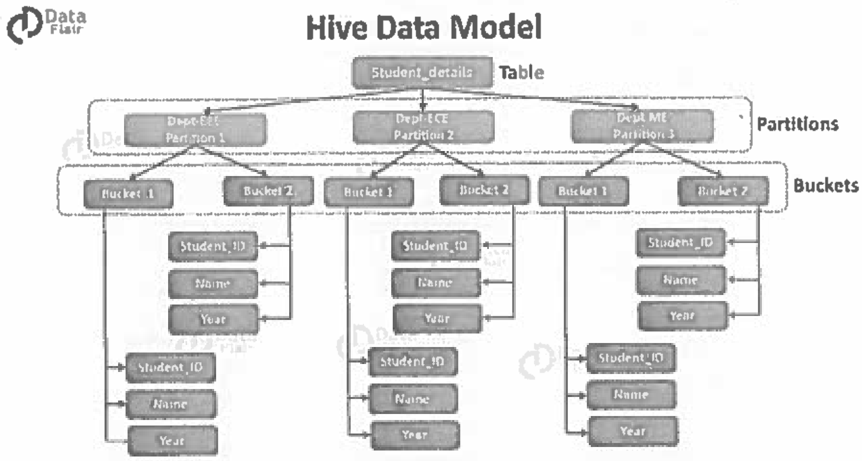
Hive is built on top of Hadoop, so it uses the underlying Hadoop Distributed File System for the distributed storage.

14(b) **Explain the hive partitioned table with an example.**

Partitioning – Apache Hive organizes tables into partitions for grouping same type of data together based on a column or partition key. Each table in the hive can have one or more partition keys to identify a particular partition. Using partition we can make it faster to do queries on slices of the data.

The Hive command for Partitioning is:

CREATE TABLE table_name (column1 data_type, column2 data_type) PARTITIONED BY (partition1 data_type, partition2 data_type,....);



Hive Partitioning Example

For example, we have a table Student_details containing the student information from different departments like ***student_id, name, year***, etc. Now, if we want to perform partitioning on the basis of department column.
Then the information of all the students belonging to a particular department will be stored together in that very partition. Physically, a partition in Hive is nothing but just a sub-directory in the table directory.

For example, we have data for three departments in our student_details table –Dept-EEE, ECE, ME. Thus we will have three partitions in total for each of the departments as we can see clearly in diagram above.
For each department we will have all the data regarding that very department residing in a separate sub – directory under the table directory.
So for example, all the student data regarding EEE department will be stored in user/hive/warehouse/student_details/dept.=EEE. So, the queries regarding EEE employee would only have to look through the data present in the EEE partition. Similarly for all the other departments.
Therefore from above example, we can conclude that partitioning is very useful. It reduces the query latency by scanning only relevant partitioned data instead of the whole data set.

15(a) **Illustrate the components used in Hive Query Processor with an example.**

The main components of Apache Hive Query Processor are as follows:

- Parse and SemanticAnalysis (ql/parse)
- Optimizer (ql/optimizer)
- Plan Components (ql/plan)
- MetaData Layer (ql/metadata)
- Map/Reduce Execution Engine (ql/exec)
- Sessions (ql/session)
- Type interfaces (ql/typeinfo)
- Hive Function Framework (ql/udf)
- Tools (ql/tools)

15(b) **Explain the table properties that can be altered with ALTER TABLE statement.**

In Hive, we can perform modifications in the existing table like changing the table name, column name, comments, and table properties. It provides SQL like commands to alter the table.
Rename a Table
If we want to change the name of an existing table, we can rename that table by using the following signature: -
Hive>Alter table old_table_name rename to new_table_name;
Now, change the name of the table by using the following command: -
Hive>Alter table emp rename to employee_data;
Adding column
In Hive, we can add one or more columns in an existing table by using the following signature:
Hive>Alter table table_name add columns(column_name datatype);

SCHEME OF EVALUATION

1	List the five characteristics of big data.	2M
2	List the three HDFS daemons.	2M
3	Sketch the graphical representation of the spark dataframe.	2M
4	Mention the two types of Apache Spark RDD operations.	2M
5	Write the statement to drop database.	2M
6(a)	Differentiate between structure and unstructured data. Structured data description- 2 M Examples- ½ M Unstructured data description- 2 M Examples- ½ M	5M
6(b)	Illustrate the working of various phases of Map Reduce with appropriate example Two phases- Map Phase description- 2M Reduce phase description – 2M Diagram – 1M Example – 2M	7M
7(a)	Illustrate Hadoop architecture and its components with a neat diagram. Hadoop description- 1M 5 components description -5M Diagram- 1M	7M
7(b)	Discuss a few benefits of Big Data. 5 benefits- 5M	5M
8(a)	Summarize the steps to write data into HDFS. Write mechanism- 3 steps – 6M 3 diagrams- 2M	8M
8(b)	Compare pseudo-distributed and fully distributed modes. Pseudo-distributed mode- 2M Fully-distributed mode- 2M	4M
9(a)	Discuss the concept of Hadoop clusters. Hadoop clusters description – 4M Diagram- 1M	5M
9(b)	How will you configure XML files in Hadoop? Configuration files description- 7M	7M
10(a)	Specify the Fundamental role of the dataframe. Role of dataframe in Spark- 5M	5M
10(b)	Illustrate the architecture of spark with a neat diagram. Spark components description- 5M Diagram- 2M	7M
11(a)	Explicate how spark streaming can be used to stream live data and process the Stock Market data. Spark streaming description- 2M Live data processing – 5M	7M
11(b)	Explain the process of reading a file from a local directory into DataFrame, and applying transformations before the DataFrame is written back to CSV file. Read file – 3M Apply transformations- 2M	5M
12(a)	Explain RDD Lineage with an example. RDD lineage description- 3M Example- 1M	4M
12(b)	List any ten spark pair RDD transformation functions along with its description. Pair RDD transformation description- 3M	8M

	10 Functions- 5M	
13(a)	Illustrate spark runtime architecture with a neat diagram. Spark description – 2M Spark architecture- 4M Diagram- 2M	8M
13(b)	Discuss the behavior of spark streaming applications in the event of failures. Spark streaming description – 2M Fault tolerant – 2M	4M
14(a)	Illustrate the Hive architecture with a neat diagram. Hive components description – 6M Diagram – 1M	7M
14(b)	Explain the hive partitioned table with an example. Hive partitions description- 2M Example- 1M Diagram- 2M	5M
15(a)	Illustrate the components used in Hive Query Processor with an example. Hive query processor description – 2M Components – 2M Example – 1M	5M
15(b)	Explain the table properties that can be altered with ALTER TABLE statement. Alter table description – 2M Syntaxes – 5M	7M

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Civil Engineering	Academic Year	2022 – 2023
Course Code	20CE503	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Foundation Engineering				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Give the purpose of subsurface exploration	20CE503.1	L1
2	What is shallow foundation?	20CE503.2	L1
3	List the types of slope failures	20CE503.3	L1
4	What do you mean by sinking of wells?	20CE503.4	L1
5	List the types of lateral earth pressures	20CE503.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Describe the procedure of conducting a plate load test with the help of a neat sketch	12 M	20CE503.1	L2
	OR			
7	Explain in detail about preparation of soil investigation report	12M	20CE503.1	L2
8	Explain Terzaghi's analysis of bearing capacify of soil in general shear failure with assumptions	12M	20CE503.2	L2
	OR			
9	Describe the procedure of conducting a standard penetration test & explain about the correction factors to be applied	12M	20CE503.2	L2
10	Explain in detail about Swedish circle method of Stability analysis of finite slopes with the help of neat sketch	12M	20CE503.3	L2
	OR			
11	Explain in detail about standard method of slices for Stability analysis of finite slopes with the help of neat sketch	12M	20CE503.3	L2
12	What are the precautions to be taken during sinking of wells? How are tilts and shifts of wells rectified?	12M	20CE503.4	L2
	OR			
13 (a)	What are different forces to be considered in analysis of well foundation in different shapes of well	6M	20CE503.4	L2
13 (b)	Classify various components of well foundation with the help sketch	6M	20CE503.4	L2
14	A retaining wall is 4 metres high. Its back is vertical and it has got sandy backfill up to its top. The top of the fill is horizontal and carries a uniform surcharge of 85 kN/m^2. Dry density of soil = 18.5 KN/m^3. Moisture content of soil above water table = 12%. Angle of internal friction of soil = 30°, specific gravity of soil particles = 2.65. Porosity of backfill = 30%. The wall friction may be neglected. Determine (i) Passive pressure acting on the wall (ii) Active pressure acting on the wall	12M	20CE503.5	L3
	OR			
15	Explain Rankine's Active earth pressure theory for cohesion less soil and cohesive soil	12M	20CE503.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

PART - A

1) Give the Purpose of subsurface Exploration?

A:- Soil Exploration is generally done to obtain the information that is useful for one or more of the following Purposes.

* To Select the type of soil and depth of Exploration of foundation for a given structure.
* To Determine Load - Bearing capacity of foundation of soil
* To Establish the Ground water level and to determine the Properties of water
* To Predict and to solve Potential foundation Problems.

2) What is shallow foundation?

A:- * A shallow foundation is one whose width is greater than the depth i.e., $[D_f/B] \leq 1$

* shallow foundation are located just below the lower Point of the wall or a column which they support.

3) List the types of slope failures?

A:- The slope may have any one of the following types of Failures.

1) Rotational Failure

2) Translational Failure

3) compound Failure

4) wedge Failure

5) Miscellaneous Failure

4) what do you mean by sinking of wells?

A:- well sinking operation involves lowering of well or dredging in the drege hole with the help of plate grab in the case of soft strata like sandy silt, soft clay, sandy clay, dense sand etc...,

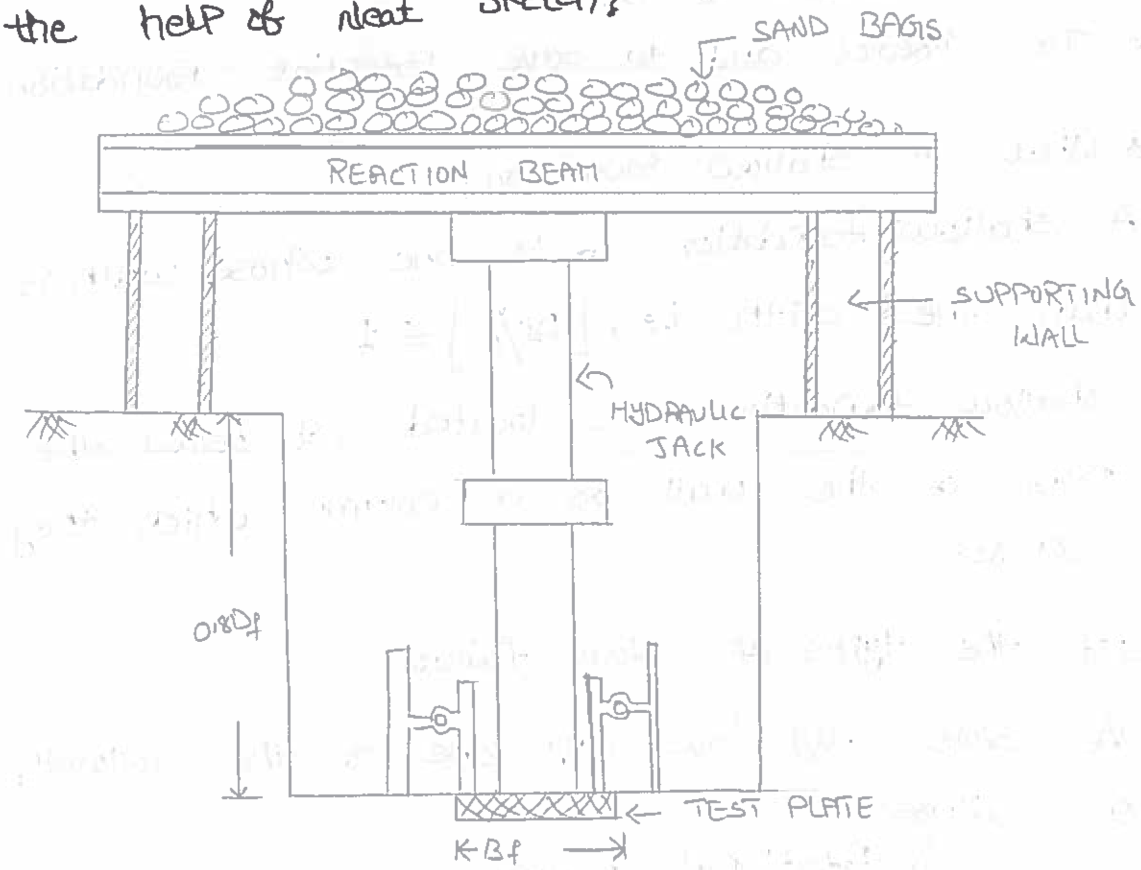
5) List the types of lateral Earth Pressure?

A:- The following are the types of lateral Earth Pressure

1) At rest Earth Pressure
2) Active Earth Pressure
3) Passive Earth Pressure

PART-B

6) Describe the Procedure of conducting a Plate load test with the help of neat sketch?



⇒ It is the most direct method to get information on bearing capacity & settlement characteristics of foundation soil

⇒ Since load Test on Actual foundation are not Practical due to required heavy loading and higher cost and time

⇒ Plate load test is done using square (or) circular steel Plates of size range from 30 cm to 70 cm, the thickness of Plate shall be 25 mm (minimum).

⇒ The Test done in a Pit which is digged as shown in figure.

⇒ The cross-section of Pit is $5B_P \times 5B_P$ where B_P is width of Test Plate.

⇒ The Test Plate is loaded with the help of hydraulic Jack by Taking reaction from gravity Plat form

⇒ The seating load of 7 KPa is applied for establish proper contact between test Plate underlying soil.

⇒ The Load on test Plate is applied in increment of 1 kg/cm^2 (or) 1/10th of anticipated UBC of soil which ever is less.

⇒ Loading is continued untill failure (or) untill have a Settlement of 25 mm occurs.

⇒ The maximum load is generally Restricted to 3 times the anticipated safe load

⇒ In saturated clay

$$q_{u(F)} = q_{u(P)}$$

⇒ In Granular soil.

$$q_{u(F)} = \frac{q_{u(P)}}{B_P} \times B_F$$

where;

$q_{u(F)}$ = U.B.C of footing

$q_{u(P)}$ = U.B.C of Plate

⇒ settlement of Actual Footing (S_F)

At same intensity of soil

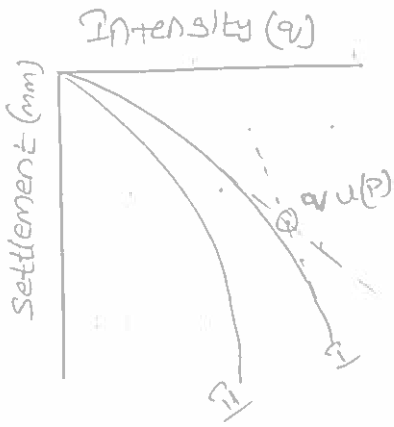
a) saturated clay,

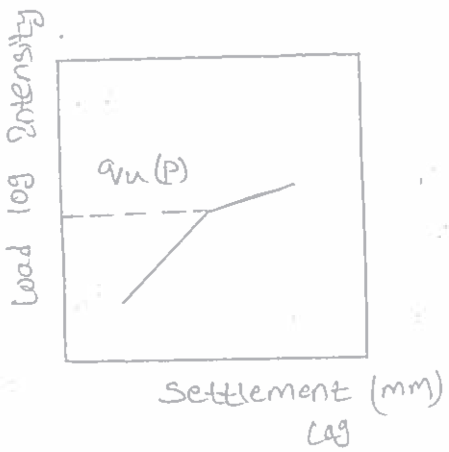
$$S_F = \frac{S_P}{B_P} \times B_F$$

b) Granular soils,

$$S_F = S_P\left[\frac{B_F(B_P+0.3)}{B_P(B_F+0.3)}\right]^2$$

where;
S_P = settlement of Plate
S_F = settlement of footing
B_F = width of footing
B_P = width of Plate.





7) EXPLAIN in Detail about Preparation of soil investigation Report ?

Ar SOIL INVESTIGATION REPORT :-

A consolidated report consisting of the results of soil Exploration. Field investigations, Laboratory and Field testing and their analysis along with suitable recommenda-tions is known as geotechnical soil Investigation Report.

ESSENTIAL DATA :-

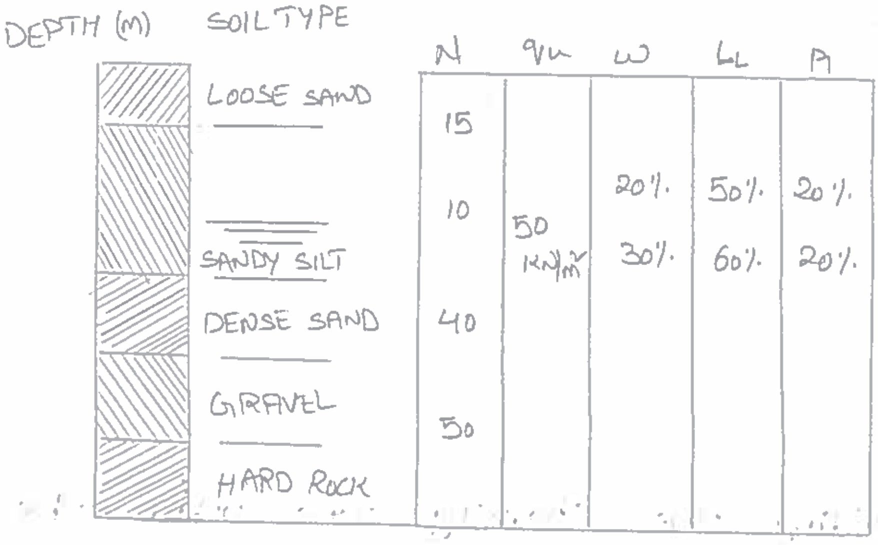
For preparation of a clear and precise report are as follows,

1) Details of proposed structure
2) Site conditions
3) Geological and Topographical Features
4) Details of adjacent structures
5) Laboratory test Results

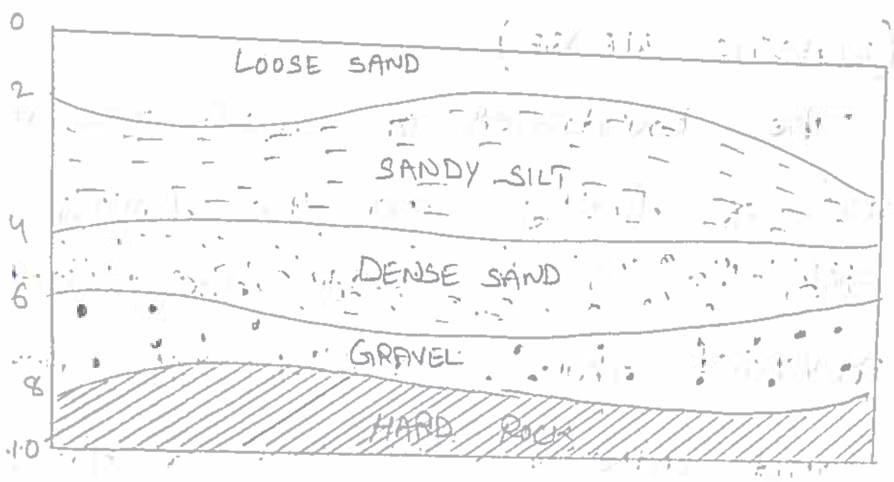
6) Details of Field investigations, Field testing and Sampling

* A good report is believed to be developed in the following format.

1) Introduction
2) objectives of the geotechnical investigations
3) Details of Proposed structures
4) site conditions
5) Field Investigations
6) Laboratory Test results
7) sub surface profile and Bore log report



BORE LOG REPORT



SUB SURFACE PROFILE

8) Allowable bearing capacity

9) Analysis and interpretation of results

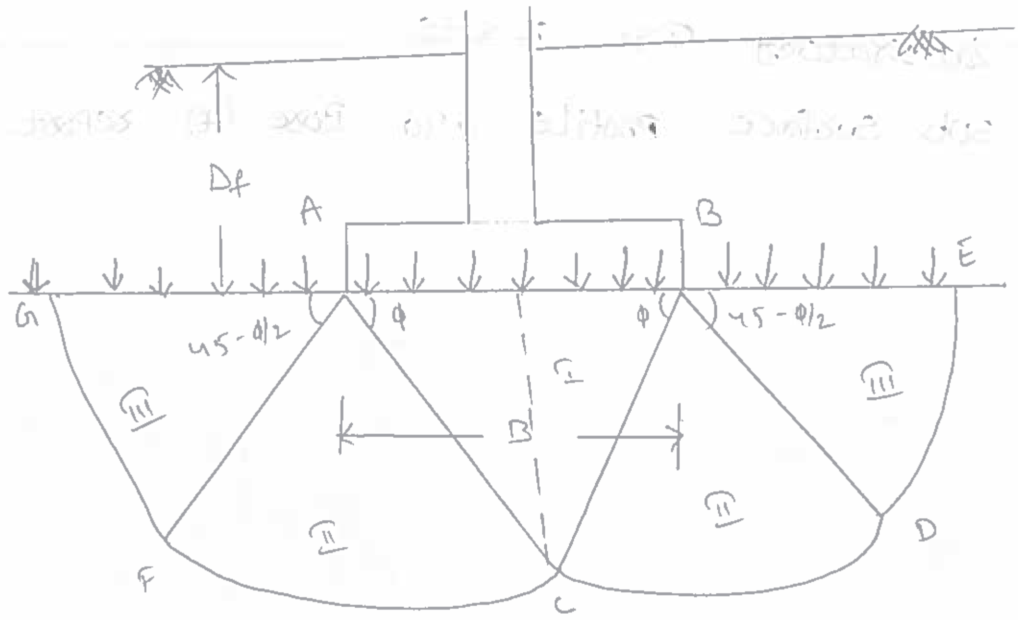
10) Foundation alternatives

11) Recommendations

12) Limitations and uncertains of soil Exploration

13) Annexures.

8) EXPLAIN TERZAGHI'S analysis of bearing capacity of soil in general shear failure with Assumptions.



Terzaghi's dividing the shearing zone under the footing as described below.

ZONE - I :- (ELASTIC WEDGE)

The boundaries of zone I rise at an angle φ w.r.t base of footing. when the footing starts sinking into soil, the soil in wedge ABC beneath the footing is prevented from undergoing any lateral yeild by friction and adhesion b/w base of footing.

The soil in wedge portion remains in a state of Elastic equillibrium it acts as if it is part of footing

ZONE - II :- [RADIAL SHEAR]

The lines that constitute the zones radiate from outer edges of base of footing. The radial lines are straight while the other set are logarathemic spirals with their centres located at outer edges of Base footing.

ZONE - III :- [PASSIVE ZONE]

The zone is also called as Rankine's Passive Zone. The boundaries of zone are inclined at an angle $(45 - \phi/2)$ with horizontal. when a load q_u is applied, the footing tends to push wedge of soil ABC into ground with lateral Displacement of zones II & III

The various forces acting on elastic wedge of limiting equilibrium are

1. Load from footing q_u
2. Self wt of wedge, $\frac{1}{2} B (\gamma) \frac{B}{2} \tan\phi \gamma$
3. cohesion resisting force along the planes CA and CB

$$CA = CB = \frac{B}{2\cos\phi}$$

ASSUMPTIONS:-

1) Soil is homogeneous and isotropic
2) Footing is longer (i.e., L/B ratio is large)
3) Base of footing is rough
4) Foundation soil fails in general shear failure mode
5) shear strength of soil is governed by Mohr's columb equation

$$S = c + \sigma \tan\phi$$

6) Loading on footing is verticle and is uniformly Distributed

7) Footing is laid at shallow depth i.e $[D_f/B \leq 1]$.

9) DESCRIbe the Procedure of conducting a standard Penetration test & Explain about the correction factors to be applied.

Ai:- STANDARD PENETRATION TEST (SPT) :-

[IS 2131 - 1981 split spoon sampler]

⇒ The Test is conducted in bore holes of 150 mm diameter

⇒ The test uses split spoon sampler (or) split tube sampler whose inside diameter is 35 mm & outside diameter is 51mm. The sampler can be split into two halves. The length of the sampler is 60 cm.

⇒ The test is performed by lowering split spoon sampler attached to drill rods upto the desired depths, in already made bore holes.

⇒ The split spoon sampler is made to penetrate in to soil at bottom of bore hole under the blows of 63.5 kg hammer filling through a height of 75cm.

⇒ The No. of blows taken for next 30 cm penetration of sampler is recorded as standard penetration Resistance (N)

⇒ The No. of blows for first 15 cm penetration of sampler is recorded as seating drive

⇒ Refusal is set to have reached if the sampler penetrates less than 2.5 cm under 50 blows.

The observed values of N are corrected for

1) over burden effects
2) Dialatancy effect.

OVER BURDEN CORRECTION :-

These are of two types.

i) Gibbs and Holtz correction

$$N_c = N_R \times \frac{350}{\bar{\sigma}_0 + 70} \qquad \bar{\sigma}_0 \leq 280 \ kN/m^2$$

where,

N_c is value of N correction for over burden

$\bar{\sigma}_0$ = effective over burden pressure at SPT level (kN/m^2)

N = measured (or) observed (SPT)

ii) PEAK Hansen & Thorn burn :- [IS 2131]

$$N_0 = \left[0.77 \log \frac{2000}{\bar{\sigma}_0'}\right] N$$

$N_0 = 2N$ at Ground surface

$\bar{\sigma}_0'$ = effective over burden pressure at the level of SPT (kN/m^2)

DIALATANCY CORRECTION :-

This correction is applied to N values in saturated silt sand and saturated fine sand.

$$N_c = 15 + \frac{1}{2}(N_0 - 15) \qquad \text{if } N > 15$$

$$= N_0 \qquad \text{if } N \leq 15$$

NOTE :- The values of N need not be corrected for over burden.

10) EXPLAIN in detail about swedish circle method of stability analysis of finite slopes with the help of Neat sketch.

A:- STABILITY ANALYSIS OF FINITE SLOPES BY

SWEDISH ARC (OR) SLIP CIRCLE METHOD:-

The actual shape of slip ortrail surface in case of finite slopes is curvilinear for convinience it is taken as circale approx.

* the assumptions of circular slip surface and its application in stability analysis was developed in sweden.

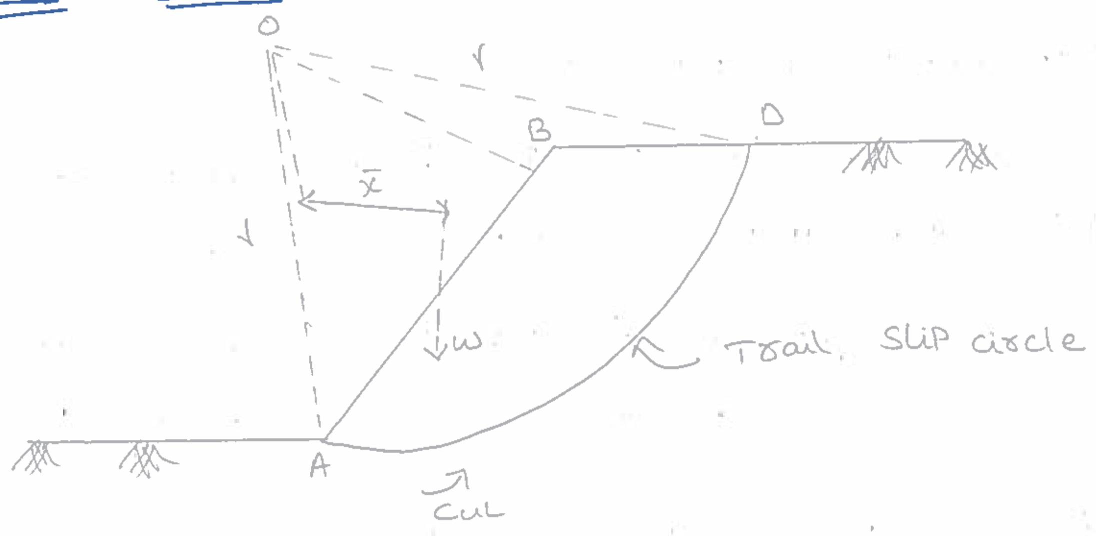
* This method is known as swedish circle method.

Two cases of soil conditions will be Analyzed.

i] Analysis of purely cohesive soil ($\phi = 0$) analysis

ii) Analysis of a soil possessing both cohesion and friction ($c - \phi$) Analysis

i) $\phi = 0$ ANALYSIS :-



↳ For the slope AB the stability has to be determined

↳ This method Assuming a number of Trail slip circles and finding the factor of safety of each

→ Let AD be the trail slip circle

W → wt of soil to the wedge ABDA

M_D → Driving moment $= W\bar{x}$

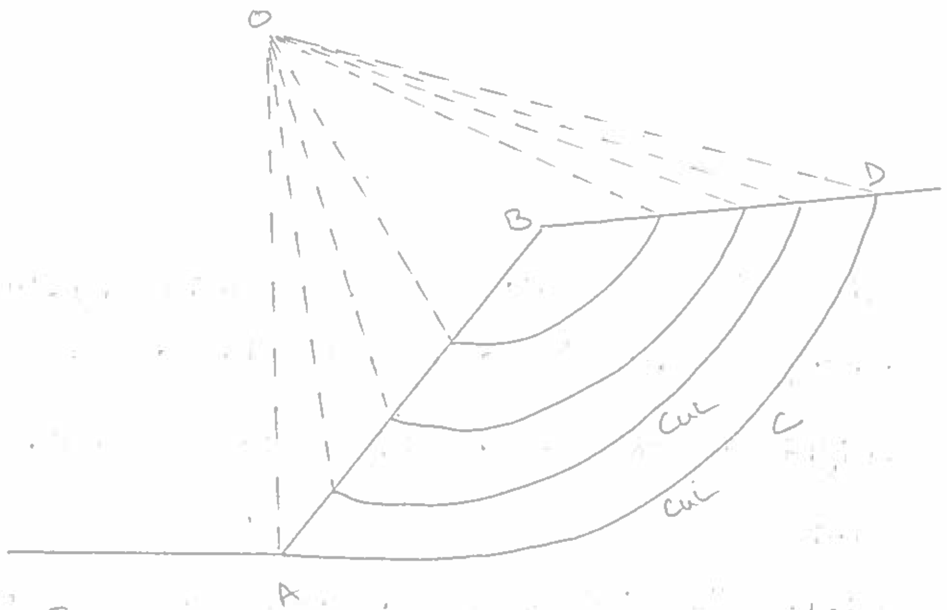
C_u → unit cohesion

L → length of slip arc AD $= \frac{2\pi r \delta}{360}$

Factor of safety $(F) = \frac{M_R}{M_D}$

$$\boxed{F = \frac{C_u L r}{W\bar{x}}}$$

~~c-φ analysis~~:-



* Find F.o.s for all slopes & the minimum fos value is "critical slip circle".

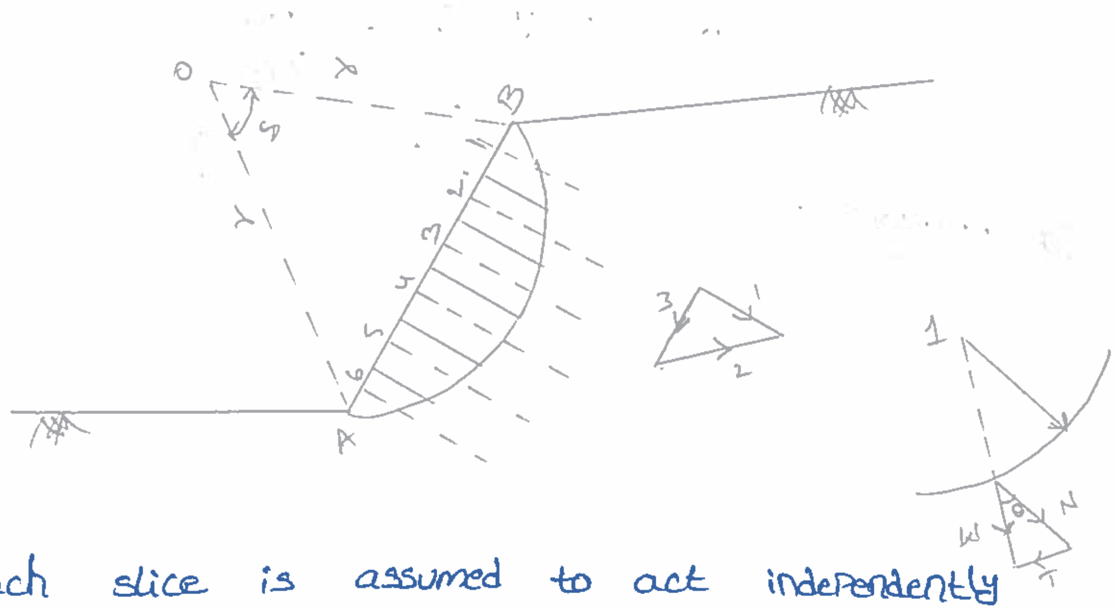
11) EXPLAIN IN DETAIL About standard method of slices for stability analysis of finite slopes with the help of Neat Sketch.

Ar <u>c-φ</u> <u>Analysis</u> [Standard method of slices].

The ordinary method of slices

⇒ The first & simples method of slices was developed by fellenius in 1936 and is known as the ordinary or normal method of slices.

⇒ In order to test the stability of the slopes of a $c-\phi$, trail slip circle is drawn and the material above assumed slip surface is divided into a convienient number of verticle slices.



⇒ Each slice is assumed to act independently and having width (b) and unit thickness.

⇒ The weight (w) of each slice is assumed to act at its centre

⇒ The weight (w) is resolved into two components are N (normal) T (Tangential).

⇒ The Tangential moment (T) causes a driving moment

$$\boxed{M_D = T \times r}$$

⇒ c is the unit cohesion and ΔL is the curved length of each slice. then the resisting force from Colomb's Equation is equal to

$$S = c + \sigma \tan\phi$$

$$= c\Delta L + N \tan\phi$$

For entire slip surface AB;

Driving moment $\Rightarrow M_D = \gamma \Sigma T$

Resisting moment $M_R = \gamma [c \Delta L + \Sigma N \tan\phi]$

ΣT = Algebraic sum of all tangential components

ΣN = sum of all Normal components

$\Sigma \Delta L = \bar{L} = \frac{2\pi r \delta}{360^\circ}$ = length AB of slip circle

$$\text{Factor of Safety} = \frac{M_R}{M_D} = \frac{c\bar{L} + \tan\phi \, \Sigma N}{\Sigma T \times \gamma}$$

$\Rightarrow$ The circle giving the minimum Factor of Safety is the "critical slip circle".

2) Q) What are the Precautions to be taken during sinking of wells? How are tilts and shifts of wells Rectified?

Ans PRECAUTIONS DURING SINKING OF WELLS :-

The following are the Precautions must be during sinking of wells.

↳ when the wells to be sunk close to each other and the distance b/w them is not greater than the diameter of the wells.

↳ similarly when two Parallel rows of wells have to be sunk with centre of each at about 1 meter apart. One row should be sunk before the other or they can be started on different ends of from the centre towards two ends.

↳ The Purpose of this is to disturb least Possible area of soil in the vicinity of well at one time.

It is also advisable to sink the alternate wells in a row in preferance to sinking them one after the others.

↳ In sinking of wells joined together, for example dumb bell shaped wells, the excavation in both the dredge. holes should be carried out simultaneously and equality to facilitate even sinking.

↳ The sinking operations should be carried on with great caution whenever cutting edge approaches the junction of different types of strata. To control this boring chart should be consulted regular.

CORRECTION OF TILTS AND SHIFTS:-

The following are the method of correction of tilts and shifts.

1) Eccentric loading
2) Packing low side of well

↳ Inspite of all possible precautions, During the course of sinking the well do get tilted due to reasons which are beyond the control, Any of the following methods may be adopted either separately or in combination.

13a) what are different forces to be considered in analysis of well foundation in Different ~~~~ shapes of wells.

Ar DEAD LOADS:- The dead load carried by the well include the wt of superstructure and the self weight

LIVE LOADS:- The Design of live loads for railways bridges are taken according to Indian Railway Bridges

Rules. For road bridges, the live loads are specified by the Indian Road congress standard specifications and code of practice for Road Bridges. (15)

3) IMPACT LOADS:- Impact effect due to live load is considered only in the Design of pier cap.

4) WIND LOADS:- wind loads on the live load, super structure, and the part of sub structure which lies above the water level calculated.

5) WATER PRESSURE:- water pressure due to water current acts on the part of substructure which lies b/w the water level.

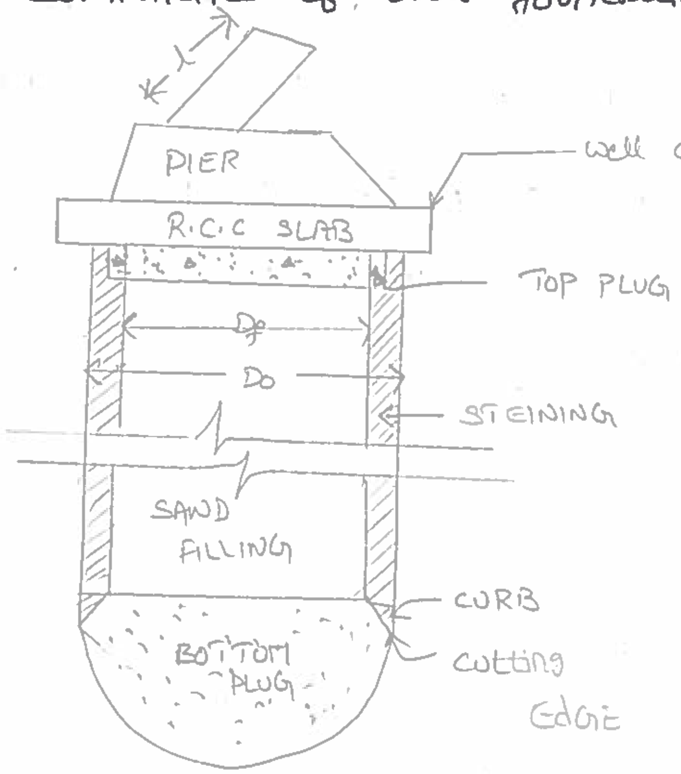
$$P = KV^2$$

where P = intensity of pressure (kN/m²)

K = a constant.

6) LONGITUDINAL FORCES:- longitudinal forces ~~acti~~ occur due to tractive and braking forces. These forces depend upon the type of vehicles & bearings

3b) CLASSIFY VARIOUS components of well foundation with the help of sketch.



WELL FOUNDATION

COMPONENTS :-

1. WELL CAP :- It consists of RCC slab laid at the top of well steining and is usually cast monolithically with the steining. It transmits load of superstructure to the steining.

2. STEINING :- It is constructed of RCC. It is main body of well which transfer load to the subsoil. It also acts as a cofferdam during sinking and provides weights for sinking.

3. CURB :- It is the lower wedge shaped portion of well steinning. It facilitates the process of sinking.

4. cutting Edge :- It is the lower most portion of the well curb. It cuts into soil during sinking

5. BOTTOM PLUG :- It is formed of 1:2:4 concrete with 10% extra cement. It transmits load to sub soil.

6. DREDGE HOLE :- The hole formed inside the well due to excavation of soil is called dredge soil. It is later filled with sand. sand filling helps in distribution of load of super structure to bottom plug.

15) Explain Rankine's Active Earth Pressure theory for cohesion less soil and cohesive soil.

A :- RANKINE'S Active Earth Pressure COHESION LESS SOILS

$\Rightarrow$ For Horizontal Backfill

Active Earth Pressure $P_0 = K_a \gamma z$

where,

K_a = coefficient of horizontal backfill

$$K_a = \frac{1-\sin\phi}{1+\sin\phi} = \tan^2(45^\circ - \phi/2)$$

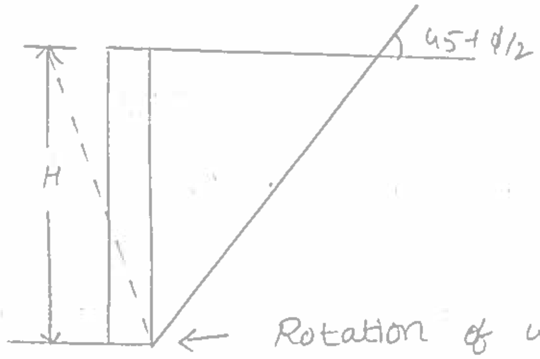
$\Rightarrow$ For inclined Back fill

Active pressure $\quad P_a = K_a \gamma z$

where $\quad K_a = \cos i \times \dfrac{\cos i - \sqrt{\cos^2 i - \cos^2\phi}}{\cos i + \sqrt{\cos^2 i - \cos^2\phi}}$

$\Rightarrow$ For ~~ipctt~~ frictionless back fill soil $(c = 0, \phi)$

$\Rightarrow$ For inclination of back fill rising at an angle α



RANKINE'S EARTH PRESSURE (COHESIVE SOILS):-

$\rightarrow$ Rankines Active earth Pressure

active Pressure $\quad P_a = K_a \gamma z - 2c\sqrt{K_a}$

where

$$K_a = \frac{1-\sin\phi}{1+\sin\phi} = \tan^2(45^\circ - \phi/2)$$

$\rightarrow$ Passive earth Pressure

Passive Pressure $\quad P_a = K_a \gamma z + 2c\sqrt{K_p}$

where

$$K_p = \frac{1+\sin\phi}{1-\sin\phi} = \tan^2(45^\circ + \phi/2)$$

14) A Retaining wall is 4 meters high. Its back is verticle and it has got sandy backfill up its top. The top of the fill is horizontal and carries a uniform surcharge of 85 kN/m². Dry density of soil = 18.5 kN/m³. Mosture content of soil above water table = 12%. Angle of internal friction of soil = 30°. Specific gravity of soil particles 2.65 Porosity of backfill = 30%. The wall friction may be neglected. Determine

i) Passive pressure acting on the wall
ii) Active pressure Acting on the wall

Sol:- ACTIVE PRESSURE ACTING ON THE WALL

$$P_a = K_a \gamma z$$

$$K_a = \frac{1-\sin\phi}{1+\sin\phi} = \tan^2(45° - \phi/2)$$

$$K_a = \frac{1-\sin 30°}{1+\sin 30°} = \tan^2(45° - 30/2)$$

$$= 0.666$$

$$P_a = K_a \gamma z \Rightarrow 0.666 \times 18.5 \times 85 \Rightarrow 1047.2$$

PASSIVE

$$P_p = K_p \gamma z$$

$$K_p = \frac{1+\sin\phi}{1-\sin\phi} = \tan^2(45° - \phi/2)$$

$$K_p = \frac{1+\sin 30°}{1-\sin 30°} = \tan^2(45 + 30/2)$$

$$= 2 \Longrightarrow P_p = K_p \gamma z = 2 \times 18.5 \times 85$$

$$= \underline{3145}$$

3/12/22.

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Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	Mechanical Engineering	**Academic Year**	2022 - 2023
Course Code	20ME503	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	V
Course	**Metal Cutting and Machine Tools**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four essential characteristics of cutting fluids	20ME503.1	L1
2	State the working principle of Automatic lathe	20ME503.2	L1
3	List any three operations performed on planner	20ME503.3	L1
4	What are the types of milling machines?	20ME503.4	L1
5	List any two types of abrasives used in grinding wheel	20ME503.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain single point cutting tool geometry with a neat diagram	6M	20ME503.1	L2
6 (b)	Draw Merchants circle and derive the expression to show the relation among the cutting forces involved in metal cutting	6M	20ME503.1	L2
	OR			
7 (a)	Explain the various types of chips	4M	20ME503.1	L2
7 (b)	Give two examples of orthogonal cutting and explain with neat sketch	8M	20ME503.1	L2
8	Explain working principle of Lathe machine with neat sketch	12M	20ME503.2	L2
	OR			
9 (a)	Explain any three types of operations performed on lathe machine	6M	20ME503.2	L2
9 (b)	Outline the specifications and working principle of Lathe	6M	20ME503.2	L2
10	Differentiate Planner, Shaper and Slotter	12M	20ME503.3	L2
	OR			
11 (a)	Describe any three operations performed on shaper	6M	20ME503.3	L2
11 (b)	Explain the principal parts and specifications of shaping machine	6M	20ME503.3	L2
12	Explain the working principle of vertical milling machines with neat sketch	12M	20ME503.4	L2
	OR			
13 (a)	Describe any three operations performed on milling machine with a neat sketch	6M	20ME503.4	L2
13 (b)	Label the specifications and working principle of milling machine	6M	20ME503.4	L1
14	Illustrate briefly about Centre less Grinding Machine and its methods with neat sketch	12M	20ME503.5	L2
	OR			
15 (a)	Classify NC machine tools	6M	20ME503.5	L1
15 (b)	Differentiate between Honing and Lapping	6M	20ME503.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

Part A (Short Answer Questions 5 x 2 = 10 Marks)

1. List any four essential characteristics of cutting fluids

- High Heat Absorption Capacity
- Good lubricating quality
- Very High Flash Point
- High Stability
- Neutral Properties
- Must be odorless
- It must be harmless
- Transparent
- Low Viscosity and
- Low Price.

2. State the working principle of Automatic lathe

Automatic lathes are high speed, heavy duty lathes and are adapted in mass production. It is provided with automatic controls for movement of work and tools at the proper rates and sequences. Once the Cutting tools are set up and the machine starts, it performs automatically all the operations to complete the work. The cycle is repeated automatically (i.e., without attention of operators) to complete the next job. The machine in which loading and unloading of work is done by operator, and other operations are completed automatically is called as semi-automatic machines.

3. List any three operations performed on planner

- Planing flat horizontal, vertical and curved surfaces.
- Planing at an angle and machining dovetails.
- Planing slots and grooves.

4. What are the types of milling machines?

Milling machines are broadly classified as follows:

- Column and knee type:
  Hand milling machine.
  Plain or horizontal milling machine.
  Universal milling machine.
  Omniversal milling machine.
  Vertical milling machine.
- Manufacturing or bed type:
  Simplex milling machine.
  Duplex milling machine.
  Triplex milling machine.
- Planer type :
- Special type:
  Drum milling machine.
  Rotary table milling machine.
  Profile milling machine.
  Pantograph milling machine.

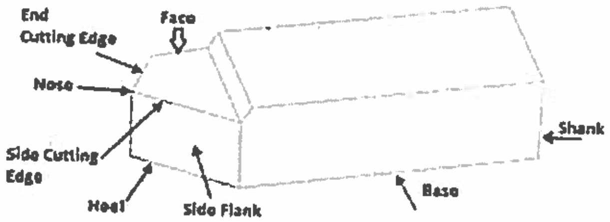
5. List any two types of abrasives used in grinding wheel

- Sand stone or solid quartz
- Emery (50-60% crystalline al2o3+ironoxide)
- Diamond
- Garnet
- Silicon carbide
- Aluminum oxide
- Boron nitrite

Part B (Long Answer Questions 5 x 12 = 60 Marks)

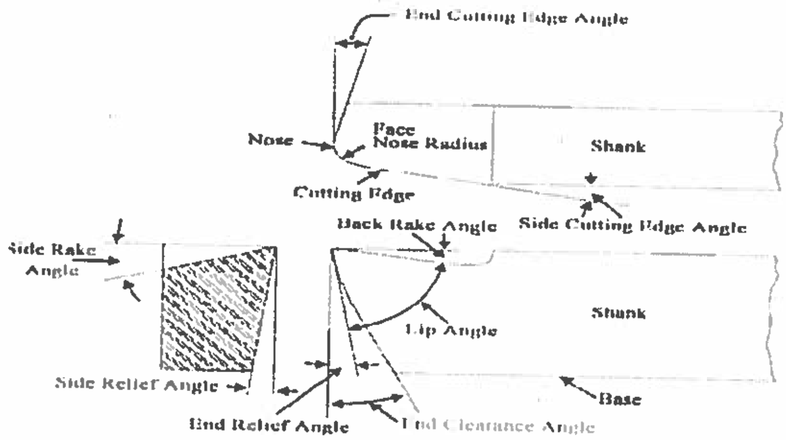
6.A.Explain single point cutting tool geometry with a neat diagram

Geometry of The Single Point Tool Angles:



1. Shank: It forms the main body of a solid tool and it is this part of the tool. Which is gripped in the tool holder.
2. Face: it is the top surface of the tool between the shank and point of the tool.
3. Point: It is the wedge-shaped portion, where the face and flank of the tool meet. It is the cutting part of the tool. It is also called nose.
4. Flank: portion of the tool which faces the work is termed as flank. It is the surface adjacent to and below the cutting edge when the tool lies in a horizontal position.
5. Heel: it is the curved portion at the bottom of the tool where the base and flank of the tool meet.

Angles of Single Point Tools :

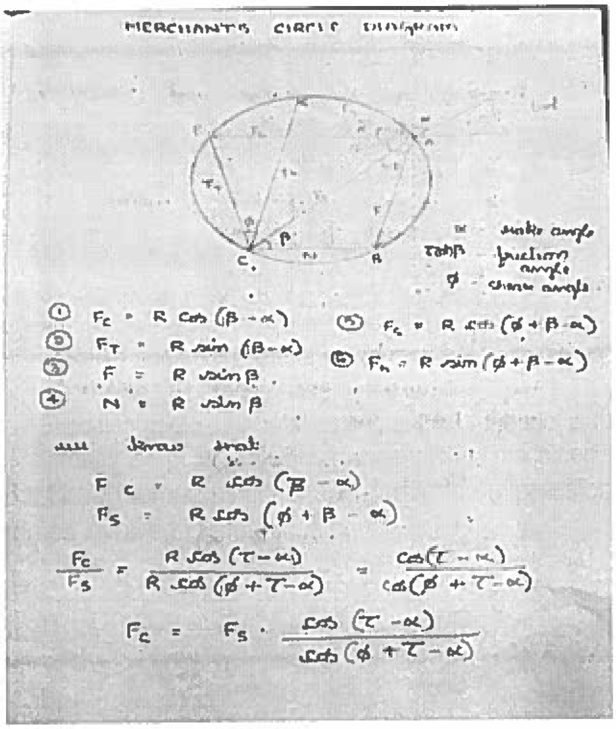


Elements of tool signature or nomenclature of single point cutting tool

1. Rake angle: It is the angle formed between the face of the tool and a plane parallel to its base. If this inclination is towards the shank, it is known as back rake or top rake. When it is measured towards the side of the tool, it is called the side rake.
2. Lip angle : The angle between the face and the flank of the tool is known as lip angle. Strength of the cutting edge or point of the tool is directly affected by this angle larger lip angle stronger will be the cutting edge and vice versa.

3. Clearance angle: It is the angle formed by the front or side surfaces of the tool which are adjacent and below the cutting edge. when the tool is held in a horizontal position.
4. Relief angle : it is the angle formed between the flank of the tool and a perpendicular line draw from the cutting point to the base of the tool.
5. cutting angle : the total cutting angle of the tool is the angle formed between the fool face and a line through the point, which is a tangent to the machined surface of the work at that point.

6.b.Draw Merchants circle and derive the expression to show the relation among the cutting forces involved in metal cutting


MERCHANT'S CIRCLE DIAGRAM

α - rake angle
$\tan\beta$ - friction angle
ϕ - shear angle

① $F_c = R\cos(\beta-\alpha)$
② $F_T = R\sin(\beta-\alpha)$
③ $F = R\sin\beta$
④ $N = R\sin\beta$
⑤ $F_s = R\cos(\phi+\beta-\alpha)$
⑥ $F_n = R\sin(\phi+\beta-\alpha)$

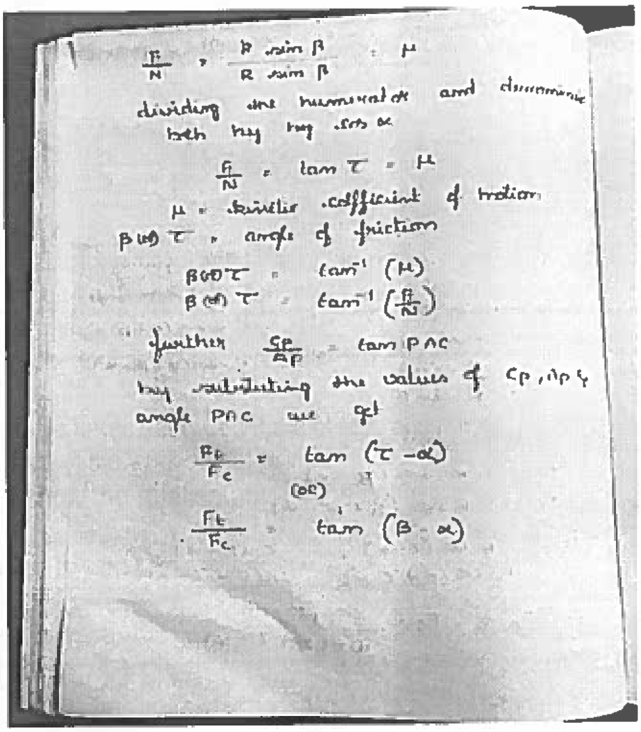
we know that

$F_c = R\cos(\beta-\alpha)$

$F_s = R\cos(\phi+\beta-\alpha)$

$$\frac{F_c}{F_s} = \frac{R\cos(\tau-\alpha)}{R\cos(\phi+\tau-\alpha)} = \frac{\cos(\tau-\alpha)}{\cos(\phi+\tau-\alpha)}$$

$$F_c = F_s \cdot \frac{\cos(\tau-\alpha)}{\cos(\phi+\tau-\alpha)}$$


$\frac{F}{N} = \frac{R\sin\beta}{R\sin\beta} = \mu$

dividing the numerator and denominator both by by $\cos\alpha$

$\frac{F}{N} = \tan\tau = \mu$

μ = kinetic coefficient of friction

β (or) τ = angle of friction

β (or) $\tau = \tan^{-1}(\mu)$

β (or) $\tau = \tan^{-1}\left(\frac{F}{N}\right)$

further $\frac{CP}{AP} = \tan PAC$

by substituting the values of CP, AP & angle PAC we get

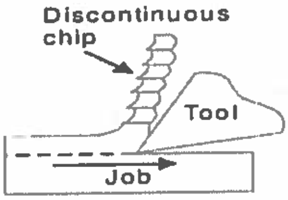
$\frac{F_t}{F_c} = \tan(\tau-\alpha)$

(or)

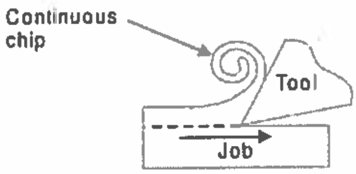
$\frac{F_t}{F_c} = \tan(\beta-\alpha)$

7.a.Explain the various types of chips

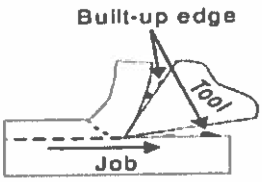
The chips produced during machining of various metals can be broadly classified into the following three types:
1. Discontinuous or segmental chips
2. Continuous chip
3. Continuous chip with built up edge.



Discontinuous chips



Continuous chips



Chips with built-up edge

1. **Discontinuous chip:**This type of chips are produced during machining of brittle materials like cast iron and bronze. These chips are produced in the form of small segments, as shown in fig. in

machining of such materials, as the tool advances forward, the shear plane angle gradually reduces until the value of compressive stress acting on the shear plane becomes too low to prevent rupture. Such chips are also some times produced in the machining of ductile materials. When low cutting speeds are used and lubrication is not provided.

Effects:

1. Friction between chip and tool is high
2. Chip into small segments
3. Wear of the tool is high
4. Poor surface finish
5. Smaller rake angle on the tool and too much depth of cut.

2. continuous chip: This type of chip is produced during the machining of ductile materials like mild steel, under favorable cutting conditions, such as high cutting speed and minimum friction between the chip and the tool face. If other wise, it will break and chip tool interface can be minimized by polishing the tool face and used coolant. In this process the chip moving smoothly up the tool face.

Advantages:

1. rake angle is high and cutting speed is high.
2. good surface finish
3. less friction between chip and tool.

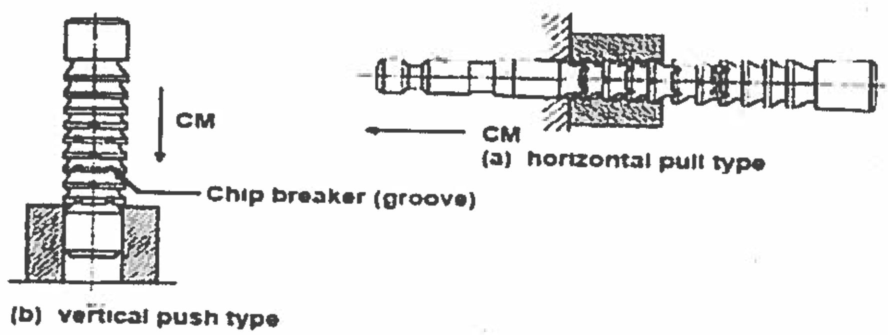
3. continuous chip with built up edge: These chips are usually formed while machining of ductile materials, when high friction exists at the chip tool interface. The upward flowing chip exerts pressure on the tool face. The normal reaction of the chip on the tool face is quite high and is maximum at the cutting edge and nose of the tool. This gives rise to an high temperature and the compressed metal adjacent to the tool nose gets welded to it. The extra metal welded to the nose or point of the tool is called built up edge. This metal is highly strain hardened and brittle with the result, as the chip flow up the tool, the built up edge is broken and carried away with the chip.

Effects:

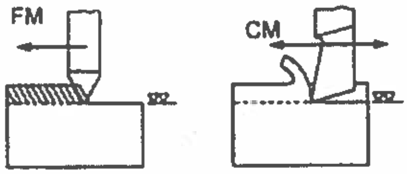
1. rough surface finish on the work piece
2. vibrations in cutting tool
3. low cutting speeds
4. small rake angle.

7.b.Give two examples of orthogonal cutting and explain with neat sketch

Broaching on Lathe Machine : Broaching is a machining process that uses a toothed tool, called a broach, to remove material. There are two main types of broaching: linear and rotary. In linear broaching, which is the more common process, the broach is run linearly against a surface of the workpiece to produce the cut. Linear broaches are used in a broaching machine, which is also sometimes shortened to broach. In rotary broaching, the broach is rotated and pressed into the workpiece to cut an Axisymmetric shape. A rotary broach is used in a lathe or screw machine. In both processes the cut is performed in one pass of the broach, which makes it very efficient.



Machining flat surfaces in slotting machine:-For machining a flat surface the work is mounted on a parallel strip such that the tool will not touch the work table at the end of the cut. Metal is removed as the tool travels past the work. The cross feed is given at the beginning of each cutting stroke. For machining cylindrical surfaces the tool is set radially on the work. The feed is given by rotary table at the beginning of the cut.

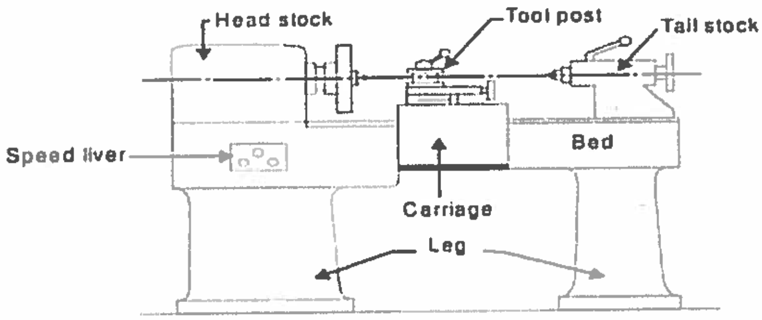


(a) horizontal surface

8. Explain working principle of Lathe machine with neat sketch

Principle: A lathe is a machine tool which use to removes unwanted materials from a work piece in the form of chips with the help of a tool that travels across the work piece and can be fed deep in work

Parts of the lathe :-



The lathe carries the following main parts,

1. Bed 2. Headstock 3. Spindle 4. Tailstock 5. Carriage 6. Feed mechanism 7. Legs.

1. Bed:
This is the base of the lathe, which will supports the all other parts. This is a heavy rigid structure made by casting. This is a single piece structure which is having high damping capacity for the vibration generated by machine during machining. The rigid structure will helps to avoid deflections.

2. Headstock:
The head stock will gives the rotation motion to the job at different speeds. It is a fixed part which will present on the left side of the lathe bed. Head stock will consists of a hallow spindle and drive unit like main spindle, feed reverse level, live centre, cone pulley etc.

3. Spindle:
The spindle rotates on two large bearings housed on the headstock casting. A hole extends through the spindle so that a long bar stock may be passed through the hole. The front end of the spindle is threaded on which chucks, faceplate, driving plate and catch plate are screwed. The front end of the hole is tapered to receive live center which supports the work.
On the other side of the spindle, a gear known as a spindle gear is fitted. Through this gear, tumbler gears and a main gear train, the power is transmitted to the gear on the lead screw.

4. Tailstock:
The tail stock is located at the right hand side of the lathe bed, it is called loose head stock because it is a movable member for alignment of work with head stock. The moving action of tailstock can be done by hand or by a wheel depending upon the design parameters and requirements.
The use of the tailstock:
i. It supports the other end of the long workpiece when it is machined between centres.
ii. It is useful in holding tools like drills, reamers and taps when performing drilling, reaming and tapping.

5. Carriage:
The carriage will present between headstock and tailstock which will slides on the bed ways of the lathe bed. The carriage will give feed to the tool and it holds the tool.
The carriage consists of the following parts,
(a) Saddle, (b) Cross slide, (c) Compound rest, (d) Tool post, (e) Apron.

(a) Saddle:

The base of the carriage is the saddle and slides along the bed ways. It contains the cross feed mechanism for moving the cutting tool at right angles to the ways. It supports the cross slide on which compounds rest and tool post are mounted.

(b) Cross slide:

The cross slide function to provide cutting action to the tool and the action of cutting tool will be perpendicular to centre line lathe.

(c) Compound rest:

It is supported on cross slide and is equipped with hand feed only. It can be swiveled horizontally on its base to any angle through 360°. It is used for turning and boring short angles and tapers.

(d) Tool post:

The tool post holds the tool in position and it is mounted on the top of the carriage.

(e) Apron:

The apron is secured underneath the saddle of the carriage unit. The controls and mechanism for all movements of carriage are housed in the apron. Its feed mechanism converts rotary motion of the lead screw or feed rod into linear motion of the carriage on which cutting tool is clamped.

6. Feed mechanism:

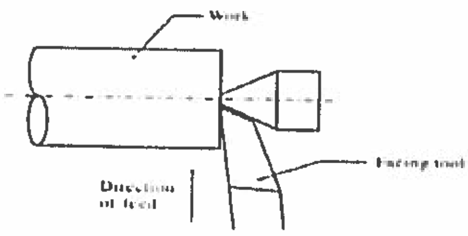
Most standard engine lathes are equipped with fed rod and a lead screw. The feed rod is used to provide automatic feed for all turning operations except for thread cutting. The lead screw transmits feed motion for screw cutting. In the absence of feed rod, the lead screw may be used for carriage feeds as well as thread cutting. The feed rod and lead screw obtain motion from the lathe spindle (Via gears) and transmit it to the carriage through gears and feed clutches.

7. Legs:

The legs are the main support for the lathe to withstand the load, generally materials made by casting method will be used for legs.

9.a. Explain any three types of operations performed on lathe machine

1. Facing:-Facing is the operation of machining the ends of a piece of work to produce flat surface square with the axis. The operation involves feeding the tool perpendicular to the axis of rotation of the work. Facing operation is shown in Fig.

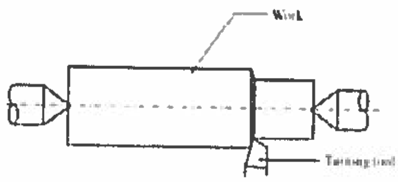


Facing

2. Turning:- Turning in a lathe is to remove excess material from the workpiece to produce a cylindrical surface of required shape and size. Straight turning operation is illustrated in Fig.

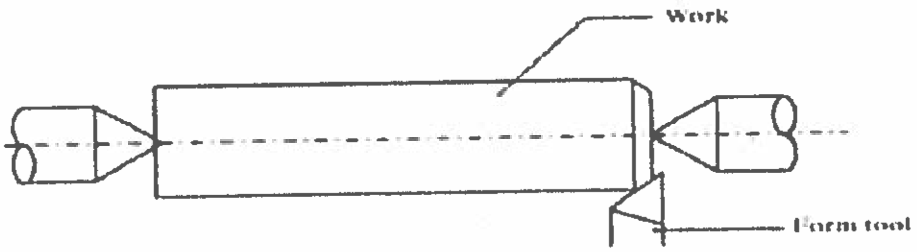
a. Straight turning:- The work is turned straight when it is made to rotate about the lathe axis and the tool is fed parallel to the lathe axis. The straight turning produces a cylindrical surface by removing excess metal from the work pieces.

b. Stepped turning:-Step turning is the process of turning different surfaces having different diameters. The work is held between centres and the tool is moved parallel to the axis of the lathe. It is also called shoulder turning.



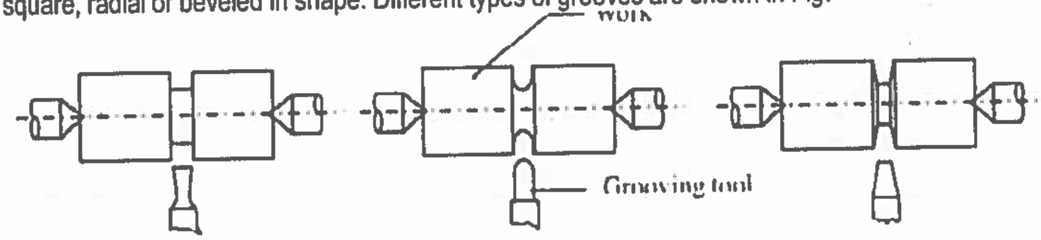
Straight turning

3. Chamfering:- Chamfering is the operation of beveling the extreme end of the workpiece. The form tool used for taper turning may be used for this purpose. Chamfering is an essential operation after thread cutting so that the nut may pass freely on the threaded workpiece. Chamfering is shown in Fig.



Chamfering

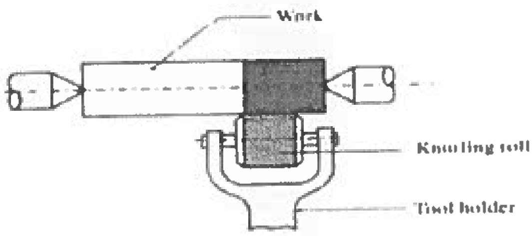
4. Grooving:-Grooving is the process of cutting a narrow groove on the cylindrical surface of the workpiece. It is often done at end of a thread or adjacent to a shoulder to leave a small margin. The groove may be
square, radial or beveled in shape. Different types of grooves are shown in Fig.



Grooving

5. Forming:-Forming is a process of turning a convex, concave or any irregular shape. For turning a small length formed surface, a forming tool having cutting edges conforming to the shape required is fed straight
into the work.

6. Knurling:- Knurling is the process of embossing a diamond shaped pattern on the surface of the workpiece. The knurling tool holder has one or two hardened steel rollers with edges of required pattern. The tool holder is pressed against the rotating work. The rollers emboss the required pattern. The tool holder is fed automatically to the required length.



Knurling

7. Undercutting:-It is a process of enlarging the diameter if done internally and reducing the diameter if done externally over a short length. It is useful mainly to make fits perfect. Boring tools and parting tools are used for this operation.

9. b. Outline the specifications and working principle of Lathe

Specifications of engine lathe:-

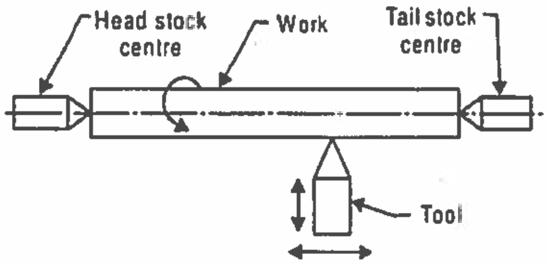
The size of lathe is specified or designated by the following items.

1. Height of the centers above the lathe bed.
2. The largest diameter of work that can be revolved over the ways of lathe bed i.e., swing diameter over the bed.
3. The largest diameter that can be accommodated over the carriage i.e., the swing diameter over the carriage.
4. Maximum diameter that can be turned over the gap of bed (in case of gap bed) i.e., the swing diameter over the gap of bed.
5. The maximum length of work that can be mounted between the centres.

Working principle of lathe:-The working principle of lathe is to remove the excess material in the form of chips from rotating work piece held between two centers with the help of a cutting tool fed against the work piece.

The centers between which the work piece is rotating are head stock Centre (live Centre) and tail stock centre (dead Centre). The tool can be fed parallel to the work piece (horizontal) or perpendicular to the work piece (vertical).To cut material properly,

i. The tool should be harder than the work piece.
ii. Work piece should be rigidly held on the machine.
iii. The cutting tool should be fed in different way relative to the work.



10. Differentiate Planner, Shaper and Slotter

	Shaper	Planer	Slotter
1	The work is stationary and the tool on the ram is moved back and forth across the work.	The tool is stationary and the workpiece on the table travels back and forth under the tool	The work is held stationary and the tool on the ram is moved up and down across the work.
2	Used for shaping much smaller jobs	Meant for much larger jobs. Jobs as large as 6 metre wide and twice as long can be machined.	It is used for making slots in smaller jobs.
3	Is a light machine	It is a heavy duty machine.	Slotting is light machine
4	Can employ light cuts and finer feed.	Can employ heavier cuts and coarse feed.	Can employ light cuts and finer feed.
5	Uses one cutting tool at a time	Several tools can cut simultaneously.	Shaper uses one cutting tool at a time
6	Driven using quick- return link mechanism	The drive on the planer table is either by gears or by hydraulic means	The rams are either crank-driven or hydraulically driven.
7	It is less rigid and less robust	Better rigidity that give more accuracy on machined surfaces.	It is less rigid and less robust
8	**applications** This machine is used for machining flat surfaces like horiontal, vertical and inclined.	**applications** This machine is used for machining large flat surfaces.	**applications** This machine used for machining various slots,grooves,keyways etc.

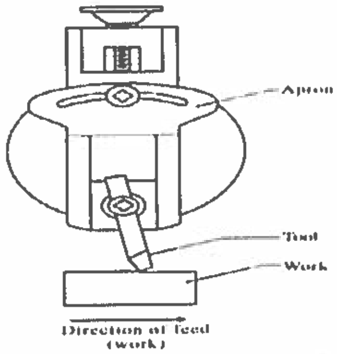
11.a. Describe any three operations performed on shaper

1. Horizontal cutting:-It is most common operation performed on a shaping machine. In this, the work is feding a horizontal direction under the reciprocating tool and the surface produced is horizontal and flat. For this, the work is either held in a vice or clamped on the machine table, depending upon it size. Before clamping the work, the vice jaws, work seat or table top are tested for accuracy. Parallels are used for clamping the work, if it is held in the vice. The tool is held in a proper tool holder. The depth of cut is adjusted and the machine started cross feed to the table is given initially by hand, till the cut starts. After that power feed can be employed. After the cut is finished, the machine is stopped and the work inspected. If more material is to be removed, the procedure is repeated till the desired surface is obtained.

A special precautions is required in setting the tool for horizontal cutting.

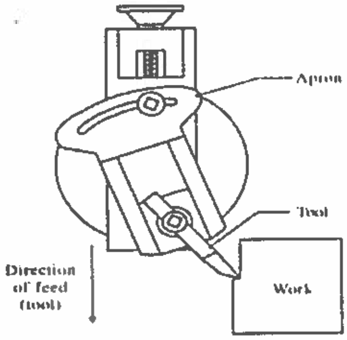
i) The tool should be held vertically in such a way that it's cutting edge point in a direction slightly away from the work as show in fig.

ii) Tool setting is that its cutting edge should not be projected much below the tool holder and vertical slide of the tool head should not be made to over hand too far below the ram. If otherwise the tool will be weakened.



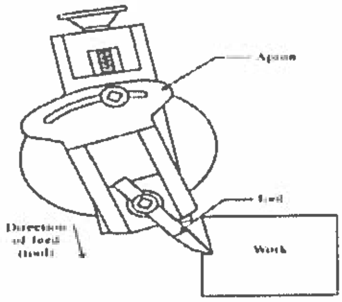
Machining a horizontal surface

2. Vertical cutting:- The tool is fed downward in vertical cutting. This sort of tool feed is commonly employed in cutting grooves, keyways, tongues, parting off and squaring ends and shoulders. For machining the vertical surface on the work piece, the job is held in a vice or is directly fitted on the table. The surface to be machined is carefully aligned with the axis of the ram. The type of tool is used for vertical machining is a side cutting tool. This side cutting tool is set in a tool post and the position and length of the stroke is adjusted. The vertical slide is set exactly at zero position and the apron is swiveled in a direction away from the surface being cut. This enables the tool to move upwards and away from the work during return stroke.



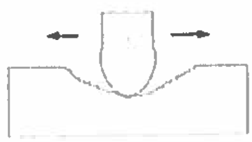
Machining a vertical surface

3. Angular cutting:- The operation of Angular cutting is employed for machining inclined surfaces, beveled surfaces and dovetails, etc. Here again, the down feed of tool is used. The work piece is set on the table and the vertical slide of the tool head is swiveled to the required angle towards right or left from the vertical position. In this type of machining, down feed of the tool is used and the cutting tool set properly. The down feed is provided by rotating the down feed screw.



Machining a angular surface

4. Irregular cutting:- For machining irregular surface, the required shape is scribed on the surface of the work. The required shape is obtained by manipulating vertical and horizontal feeds. Beside the above operations, the shaper can be conveniently used for cutting splines and gears by using index Centre.



Irregular machining

11.b. Explain the principal parts and specifications of shaping machine

Specifications of a shaper:-

Shaping machine is specified with the following specifications,

(i) Length of stroke, which indicates the general size of the machine and the size of a cube that can be held and planed in the shaper.

(ii) Types of drive belt drive or individual motor drive.

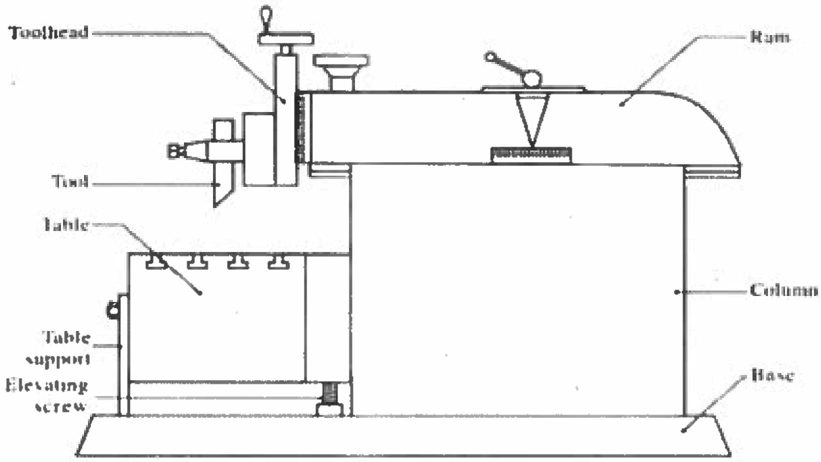
(iii) Power input.

(iv) Floor space required.

(v) Weight of the machine.

(vi) Cutting to return stroke ratio.

(vii) Number and amount of feed.



Shaping machine

Construction:- The main parts of the Shaper machine is Base, Body (Pillar, Frame, Column), Cross rail, Ram and tool head (Tool Post, Tool Slide, Clamper Box Block).

Base:- The base is a heavy cast iron casting which is fixed to the shop floor. It supports the body frame and the entire load of the machine. The base absorbs and withstands vibrations and other forces which are likely to be induced during the shaping operations.

Column:- It is box type cast iron body, mounted on the base and acts as a housing for the operating mechanism of the machine. It can also act as a support for other parts of the machine such as cross rail and ram..etc. In case of hydraulic shaper, it carries the hydraulic drive mechanism inside it.

Cross rail:- It is a heavy cast iron construction attached to the column at its front on the vertical guide ways. It carries two mechanisms one for elevating the table and other for cross travel of the table.

Table:- It is made of cast iron and has a box type construction. It holes and supports the work during the operation and slides along the cross rail.

Ram:- It is also a cast iron semi-circular in shape and provided with a ribbed constructions for inside rigidity and strength. It carries the tool head. It carries the mechanism for adjustment for ram position inside it. The tool head is attached to the front end of the ram and can be swiveled at any angle on either for making angular cuts. It is the device in which is held the tool it can slide up and down.

Vice:- It is a job holding device and is mounted on the table. It holds and supports the work during the operation. Alternatively, the job can be directly clamped to the machine table.

12. Explain the working principle of vertical milling machines with neat sketch

Vertical milling machine:

This machine is very similar to a horizontal milling machine. The only difference is the spindle is vertical. The work table may or may not have swiveling features. The spindle head may be swiveled at an angle, permitting the milling cutter to work on angular surfaces. In some machines, the spindle can also be adjusted up or down relative to the work piece. This machine works using end milling and face milling cutters. This machine is adapted for machining grooves, slots and flat surfaces. *Fig. 3.56 schematically shows the basic configuration of a vertical milling machine.*

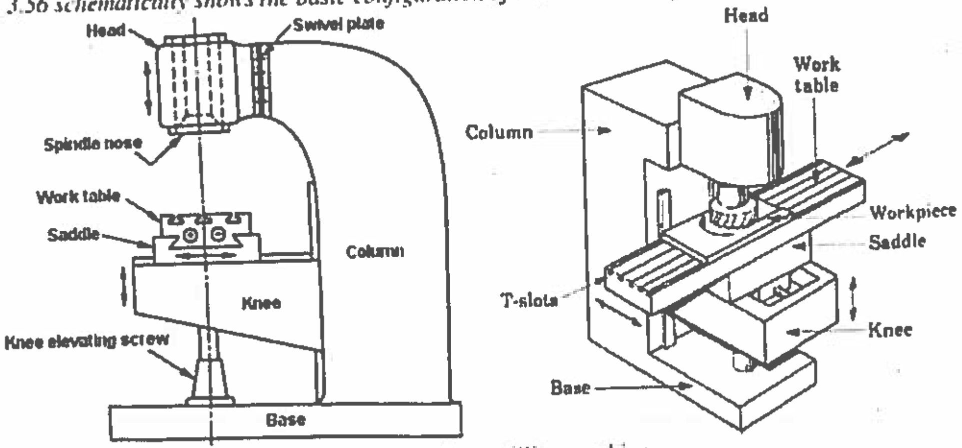
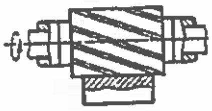


Fig. 3.56 Vertical milling machine

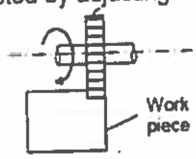
13.a. Describe any three operations performed on milling machine with a neat sketch

Plain Milling Operation: Producing plain, flat horizontal surface. This is called slab milling if performed with a peripheral cutter and called face Milling if a face milling cutter is used.



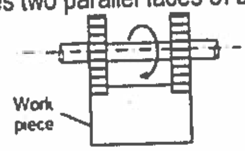
Face Milling Operation: This operation produces flat surface at the face o the work piece. This surface is perpendicular to the surface prepared in plain milling operation. This operation is performed by face milling cutter mounted on stub arbor of milling machine. Depth of cutis set according to the need and cross feed is given to the work table.

Side Milling Operation: This operation produces flat and vertical surfaces at the sides of the work piece. In this operation depth of cut is adjusted by adjusting vertical feed screw of the work piece.



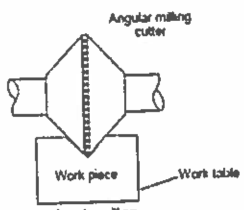
(3) Side Milling

Straddle Milling Operation: This is similar to the side milling operation. Two side milling cutters are mounted on the same arbor. Distance between them is so adjusted that both sides of the work piece can be milled simultaneously. Hexagonal bolt can be produced by this operation by rotating the work piece only two times as this operation produces two parallel faces of bolt simultaneously.



(4) Straddle Milling

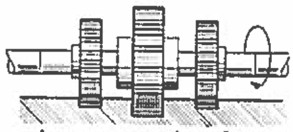
Angular Milling Operation



Angular milling.

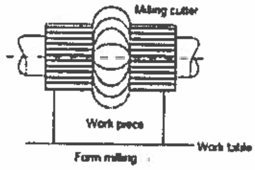
Angular milling operation is used to produce angular surface on the work piece. The produced surface makes an angle with the axis of spindle which is not right angle.Production of V shaped groove is the example of angular milling operation. Angular milling is shown.

Gang Milling Operation



As the name indicates, this operation produces several surfaces of a work piece simultaneously using a gang of milling cutters. During this operation, the work piece mounted on the table is fed against the revolving milling cutters. This operation is illustrated.

Form Milling Operation



Form milling

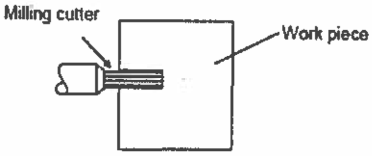
Form milling operation is illustrated in below figure. This operation produces irregular contours on the work surface. These irregular contours may be convex, concave, or of any other shape. This operation is done comparatively at very low cutter speed than plain milling operation.

Profile Milling Operation

Gang milling

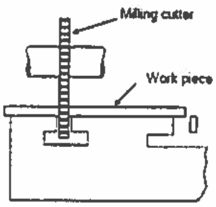
In this operation a template of complex shape or master die is used. A tracer and milling cutter are synchronized together with respect to their movements. Tracer reads the template or master die and milling cutter generates the same shape on the work piece. Profile milling is an operation used to generate shape of a template or die. This operation is demonstrated in the below figure.

End Milling Operation

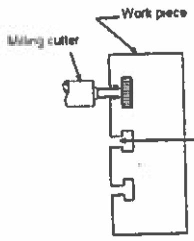


End milling operation produces flat vertical surfaces, flat horizontal surfaces and other flat surfaces making an angle from table surface using milling cutter named as end mill. This operation is preferably carried out on vertical milling machine. This operation is illustrated in the below figure.

Saw Milling Operation: Saw milling operation produces narrow slots or grooves into the work piece using saw milling cutter. This operation is also used to cut the work piece into two equal or unequal pieces which cut is also known as "parting off". Incase of parting off operation cutter and work piece are set in a manner so that the cutter is directly placed over one of the slot of the worktable as illustrated in Figure .



Slot Milling Operation: The operation of producing keyways, grooves, slots of varying shapes and sizes is called slot milling operation. Slot milling operation can use any type of milling cutter like plain milling cutter, metal slitting saw or side milling cutter. Selection of a cutter depends upon type and size of slot or groove to be produced. Right placement of milling cutter is very important in this operation as axis of cutter should be at the middle of geometry of slot or groove to be produced. The operation is illustrated in the below figure.



13. b. Label the specifications and working principle of milling machine

Working Principle of Milling Machine

Working of a milling machine is based on the fact that milling cutter is fed against work piece. This is achieved by developing relative motion with precise control between work piece and rotating milling cutter. Feed motion is generally given to the work piece through its holding device. Cutting mechanism of the work piece in milling operations is

(1) Conventional
(2) Partial Face Milling
(3) End Milling
(4) Profile Milling
(5) Pocket Milling
(6) Surface Contouring

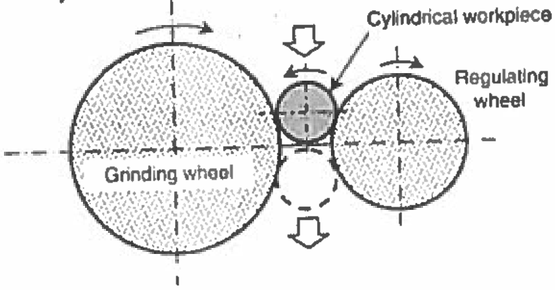
same as that in turning operation on lathe. This cutting takes place due to plastic deformation of metal by the cutting tool. Milling machine can also hold more than one cutter at a time. The holding device is supported by mechanism that can offer a selective portion of the work piece to milling cutter for its processing. Indexing is one of the examples of this type of processing.

Specifications of a Milling Machine

1. The max. length of longitudinal, cross and vertical travel of the table.
2. No. of spindle speeds,
3. No. of table speeds and feeds
4. Floor space required
5. Net weight required
6. Spindle nose taper (for vertical milling machine spindle and arbors)
7. Total power available

14. Illustrate briefly about Centre less Grinding Machine and its methods with neat sketch

Centerless Grinding Machines: In center less grinding, the workpiece is held between two grinding wheels, rotating in the same direction at different speeds. One grinding wheel is on a fixed axis and rotates so that the force applied to the workpiece is directed downward. This wheel usually performs the grinding action by having a higher linear speed than the workpiece at the point of contact. The other movable wheel is positioned to apply lateral pressure to the workpiece and usually has either a rough or rubber-bonded abrasive to trap the workpiece. The relative speed of the two wheels provides the grinding action and determines the rate at which material is removed from the workpiece surface as shown in Figure 4.79. In the first of three types of center less grinding, the through-feed type, the workpiece is fed through the grinding wheels completely, entering on one side and exiting on the opposite. The regulating wheel in through-feed grinding is canted away from the plane of the grinding wheel in such a way as to provide a lateral force component, feeding the workpiece through between the two wheels. Through-feed grinding can be highly efficient because it does not require a separate feed mechanism; however, it can only be used for a simple cylindrical shape.



Different types of feeding methods in centre less grinding

- Through-feed
- In feed
- End feed

15. a. Classify NC machine tools

Types of NC systems

Machine controls are divided into three groups,

1. Traditional numerical control (NC):

The original numerical control machines were referred to as NC machine tool. They have "hardwired" control, whereby control is accomplished through the use of punched paper (or plastic) tapes or cards. Tapes tend to wear, and become dirty, thus causing misreadings. Many other problems arise from the use of NC tapes, for example the need to manual reload the NC tapes for each new part and the lack of program editing abilities, which increases the lead time. The end of NC tapes was the result of two competing developments, CNC and DNC.

2. Computer numerical control (CNC):

CNC refers to a system that has a local computer to store all required numerical data. While CNC was used to enhance tapes for a while, they eventually allowed the use of other storage media, magnetic tapes and hard disks. The advantages of CNC systems include but are not limited to the possibility to store and execute a number of large programs (especially if a three or more dimensional machining of complex shapes is considered), to allow editing of programs, to execute cycles of machining commands, etc.

3. Distributed numerical control (DNC):

The development of CNC over many years, along with the development of local area networking, has evolved in the modern concept of DNC. Distributed numerical control is similar to CNC, except a remote computer is used to control a number of machines. An off-site mainframe host computer holds programs for all parts to be produced in the DNC facility. Programs are downloaded from the mainframe computer, and then the local controller feeds instructions to the hardwired NC machine.

The recent developments use a central computer which communicates with local CNC computers (also called Direct Numerical Control)

15.b. Differentiate between Honing and Lapping

Lapping	Honing
In the lapping process, flats and outer cylindrical surfaces are preferred for a smooth finish.	In honing process internal cylindrical surfaces are preferred for a smooth finish.
Give fine finish like glass and It is finer than honing.	It is less fine than lapping.
IT 01 tolerance grade can be obtained in this process.	Up to IT 4 tolerance grade can be obtained in this process.
Sliding is not suitable for the surface because it has more adhesive.	Sliding is suitable for surfaces.
In lapping, process metal can be cut from 0.003 mm to 0.03 mm.	In honing process metal can be cut from 0.010 mm to 0.15 mm.
The lapping speed is slightly less than honing.	The honing speed is slightly higher than lapping.
Lapping slurry or fluid suspension of very small abrasive particles is used between the workpiece and the lapping tool.	Bonded abrasive sticks/stones are used.
The lap is pressed against the work and moved back and forth over the surface.	Honing stones are pressed against the work and given oscillatory and rotary motion.

Lapping	Honing
Surface finishes of around 0.08 to 0.025 um or slightly better can be achieved.	Surface finishes of around 0.13 to 1.25um
Commonly used grit sizes are 300- 600.	Commonly used grit sizes are 100- 150.
The cutting mechanism is different than honing.	Bonded abrasives remove the material. Similar to as in grinding.
Used for external as well as internal cylindrical surfaces and flat surfaces.	Mainly used for internal cylindrical surfaces. However, its variant short-stroke honing or superfinishing is used for all types of surfaces.
Used in finishing of optical lenses, metallic bearing surfaces, gauges, etc.	Used in finishing of bores of internal combustion engines, bearings, hydraulic cylinders, gun barrels, etc.
It is done by using a lap or lapping compound.	It is done by using honing stone.
This operation is done with the help of a machine and manually.	This operation is done with the help of a machine.
Coolant is not important.	Coolant is most important.
A small amount of material is removed.	A large amount of material is removed.
During the lapping process, there is no burr produced.	During honing process burr can be produced.
Heat generation in the lapping process is negligible.	The heat is generated during the honing process.

3/12/20
Signature of Faculty

Head of the Department
3/12/20

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech. (U. G.)	Program	EEE	Academic Year	2022-2023
Course Code	20EC305	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Digital System Design				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the significance of BCD code?	20EC305.1	L1
2	Give the truth table for half adder.	20EC305.2	L1
3	What is ring counter?	20EC305.3	L1
4	What is race around condition?	20EC305.4	L1
5	Write the structure flow of VHDL.	20EC305.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Represent the decimal number 3452 in i)BCD ii)Excess-3	8M	20EC305.1	L2
6 (b)	Convert the following numbers i) $(615)_{10} = (\quad)_{16}$ ii) $(658.825)_{10} = (\quad)_{8}$	4M	20EC305.1	L2
	OR			
7 (a)	Perform the Excess- 3 Subtraction for 87 & 32.	6M	20EC305.1	L2
7 (b)	Perform the BCD Subtraction for 798 & 389.	6M	20EC305.1	L2
8 (a)	Given the Boolean function Y(A,B,C,D) =A + B C +ABD' +ABCD. Convert to standard SOP.	6M	20EE405.2	L2
8 (b)	Find the reduced POS form using K-map F(A,B,C,D) = πM(0,6,7,8,12,13,14,15). Implement using NOR gates.	6M	20EE405.2	L2
	OR			
9 (a)	Convert to Canonical forms (i) F(X,Y,Z)= X Y + Z (ii) F 2 (X, Y, Z) = (X + Y')(X'+Z) .	6M	20EC305.2	L3
9 (b)	Using K map, simplify the following expression and implement using NAN D gates $F_1(A, B, C, D) = \varepsilon_m (1, 5, 6, 7, 11, 12, 13) + \varepsilon_d (10, 15)$.	6M	20EC305.2	L3
10(a)	Implement the function using only one 4: 1 mux and gates. F (A, B, C, D) = εm (0, 2, 3, 6, 8, 9, 11. 13).	6M	20EC305.3	L2
10 (b)	Explain the working of the priority encoder and draw the diagram.	6M	20EC305.3	L2
	OR			
11 (a)	Implement the following functions using PLA having 3 i/ps, 4 product terms and 2 outputs. $F1(A, B, C) = \varepsilon_m (3,5,6,7)$ $F2(A, B, C) = \varepsilon_m (0,2,4,7)$.	6M	20EE405.3	L3
11 (b)	Draw the basic structures of PAL and PLA.	6M	20EE405.3	L2
12 (a)	State the excitation table of JK Flip Flop.	4M	20EE405.4	L2
12 (b)	Design a binary counter using T flip flops to count in the following sequence: (i) 000, 001, 010, 011, 100, 101, 111, 000	8M	20EE405.4	L2
	OR			
13 (a)	Implement T flip flop using D flip flop.	6M	20EE405.4	L2
13 (b)	Design a synchronous counter that counts the sequence 000,001,010,011,100,101,110,111,000 Using D flip flop.	6M	20EE405.4	L2

14 (a)	Design 2-bit magnitude comparator and write a verilog HDL code.	8M	20EE405.5	L2
14 (b)	Define logic synthesis and simulation.	4M	20EE405.5	L2
	OR			
15 (a)	Explain VHDL programming using structural modeling.	6M	20EE405.5	L2
15 (b)	Write the IC design flow of VHDL.	6M	20EE405.5	L2

DSD Scheme

1) BCD Code def → 1M
analysis → 1M

2) Half adder def → 1M
Truth table → 1M

3) Ring Counter – def → 1M
diagram → 1M

4) Condition → 1M
remedy → 1M

5) Modeling styles → 1M
Structure flow & example → 1M.

6) a) BCD conversion – 4M
Ex-3 Conversion – 4M

b) Hexadecimal Conversion – 2M
octal Conversion – 2M.

7) a) for process — 8M
for correct result – 3M

7) b) for process – 3M
for correct result – 3M.

8) a) for process – 3M
for Correct Conversion – 3M

8) b) for K-Map – 3M
for implementation – 3M.

9) a) (1) for 1st Cannonical form – 3 M
(2) for F2 conversion – 3 M

9) b) for k-maps – 3 M
Implementation using NAND gates – 3 M

10) a) for Process – 3 M
for correct Implementation – 3 M

10) b) For diagram — 3M
for working – 3 M

11) a) for Process – 3M
for PLA Implementation – 3M

11) b) for PLA structure – 3M
for PAL Structure – 3 M

12) a) for Truth Table – ~~1~~3M
for excitation table – 3M

12) b) for table — 4M
for logic diagram – 4M.

13) a) for Implementation logic diagram — 3M
for conversion table – 3M

13) b) for Table – 3M
for diagram – 3M

14) a) for logic diagram of Comparator — 4M
for Coding – 4M

14) b) Logic Synthesis def – 2M
Simulation def – 2M

15) a) for process – 3M, for example – 3M.

15) b) for diagram – 3M, for explanation – 3M.

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ANSWER KEY AND SCHEME OF EVALUATION

1) a) BCD Code: It is also known as Numeric code. Numeric codes are codes which represent numeric information. i.e. only numbers as a Series of 0's and 1's. Numeric codes used to represent the decimal digits is called Binary Coded Decimal (BCD) Codes.

Examples of BCD codes: 8421, 2421, 5211 .....

A BCD code is one which represents all the digits of a decimal number in a coded form – one at a time – into groups of four binary digits.

BCD Codes combine the features of decimal and binary numbers.

The 8421 BCD code is known as Natural BCD code.

2) a) Half Adder:- A half-adder is a combinational circuit with two binary inputs (augend and addend bits) and two binary outputs. (Sum and carry bits.)

It adds two I/PS A and B and produces the sum (S) and the carry (C) bits.

Block diagram

$S = A\bar{B} + \bar{A}B = A \oplus B$

$C = AB$

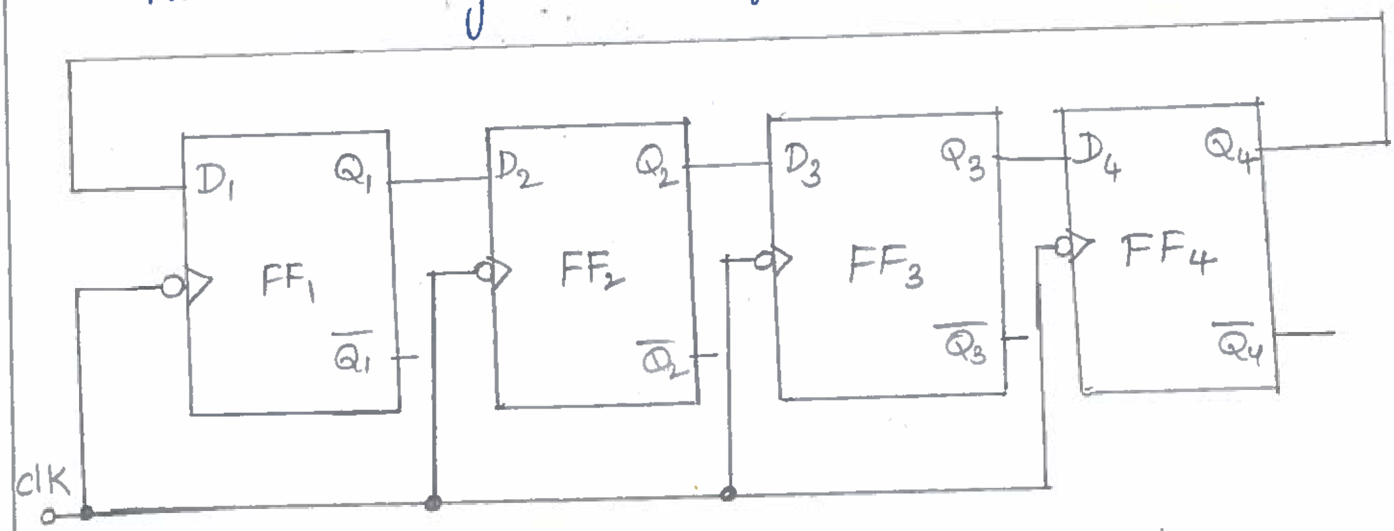
Truth table

Inputs		Outputs	
A	B	S	C
0	0	0	0
0	1	1	0
1	0	1	0
1	1	0	1

3) What is ring counter?

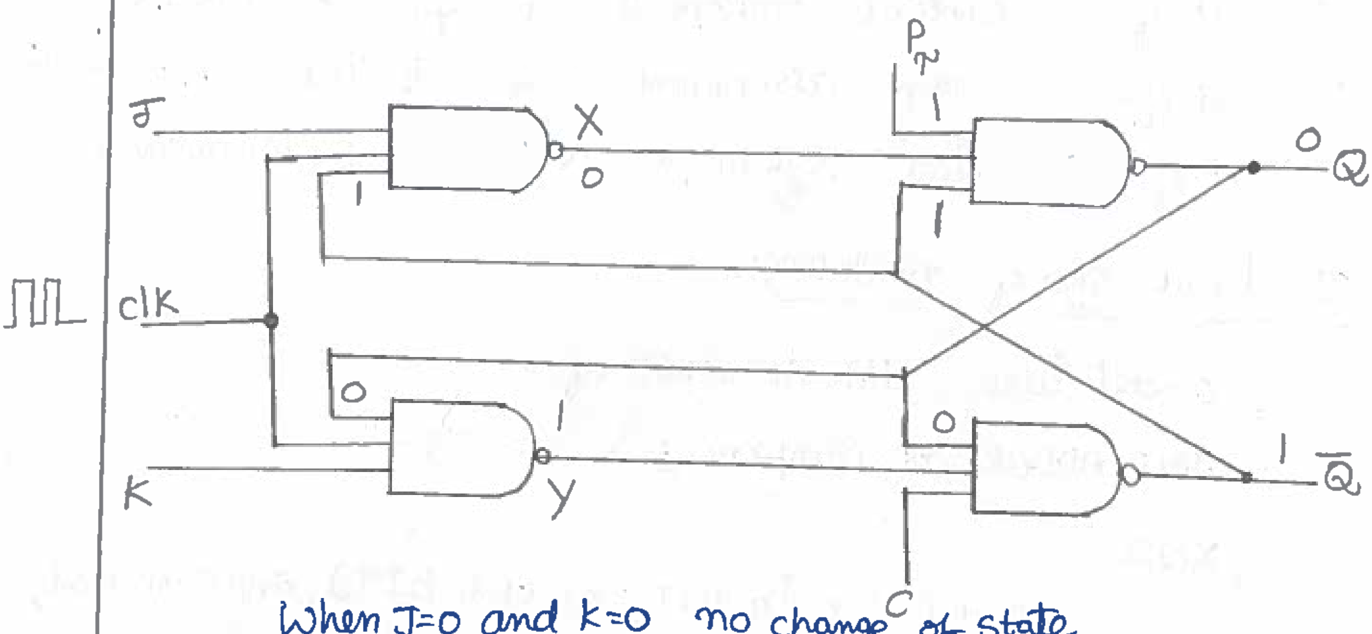
a) It is the simplest shift register counter. Shift register counters are obtained from serial-in, serial-out shift registers by providing feedback from the output of the last FF to the input of the first FF. These devices are called counters because they exhibit a specified sequence of states.

The basic ring counter using D FFS is shown below.



Logic diagram of a 4-bit ring counter
(using D flip flops)

4) what is race around condition?



When J=0 and k=0 no change of state
when J=0 and K=1 the flip-flop resets
when J=1 and k=0 the flip-flop sets
when J=1 and k=1 the flip-flop toggles i.e goes to the opposite state i.e. at the positive going edge of the clock pulse. (Positive edge triggered J-K FliP FloP.)

Consider the excitations J=K=1. If the width of the clock pulse t_p is too long the state of the flip-flop will keep on changing from 0 to 1, 1 to 0, 0 to 1 and so on and at the end of the clock pulse it's state will remain uncertain. This phenomenon is called the race around condition. The o/ps Q and $\bar{Q}$ will change on their own if t_p is too long compared with the propagation delay τ of each NAND gate.

5) VHDL is an acronym for VHSIC Hardware Description Language

VHSIC is an acronym for Very High Speed Integrated circuits. It is a hardware description Language.

Architecture body:- The internal details of an entity are specified by

an architecture body using any of the following modeling styles.
① A set of interconnected Components (to represent structure)
② A set of Concurrent assignment statements (to represent dataflow)
③ A set of Sequential Statements (to represent behaviour)

structural style of modeling:-

```
architecture HA_STRUCTURE of
HALF_ADDER is component
XOR2
    port (X, Y: in BIT; Z: out BIT); end component;
    Component AND2
    Port( L, M: in BIT; N: out BIT); end component
begin
    X1: XOR2 Port map (A, B, SUM);
    A: AND2 Port map (A, B, CARRY);
end HA_STRUCTURE;
```

The name of the architecture body is HA_STRUCTURE.

6) a) Given decimal number is 3452

BCD Code:- 3 4 5 2
0011 0100 0101 0010

Excess-3 Code:- 3 4 5 2
3+3 4+3 5+3 2+3
6 7 8 5
0110 0111 1000 0101

6) b)

① $615_{(10)} = (\ )_{16}$

$$\begin{array}{r|l} 16 & 615 \\ 16 & 38-7 \\ 16 & 2-6 \\ & 0-2 \end{array}$$

$267_{(16)}$

② $658.825_{(10)} = (\ )_8$

Conversiong 658 to octal:

$$\begin{array}{r|l} 8 & 658 \\ 8 & 82-2 \\ 8 & 10-2 \\ 8 & 1-2 \\ & 0-1 \end{array}$$

$658_{(10)} = 1222_{(8)}$

Conversion of 0.825 to octal

$0.825 \times 8 = 6.6$

$0.6 \times 8 = 4.8$

$0.8 \times 8 = 6.4$

$0.4 \times 8 = 3.2$

$0.825_{(10)} = 0.6463_{(8)}$

Therefore $658.825_{(10)} = 1222.6463_{(8)}$

7) a) Perform the Excess-3 Subtraction for 87 & 32.

$$\begin{array}{rr} & 8\ 7 \\ - & 3\ 2 \\ \hline \end{array}$$

$8+3=11$ $7+3=10$

$3+3=6$ $2+3=5$

Now 1011 1010 (87 in XS-3) — ①

0110 0101 (32 in XS-3)

Complement the subtrahend

Complement of 32 in XS-3 1001 1010 — ②

Now add ① and ②

$$
\begin{array}{rrl}
 & 1011 & 1010 \\
+ & 1001 & 1010 \\
\hline
1 & 0101 & 0100 \\
\hline
+ & 0010 & 0010 \\
\hline
+ & 1000 & 0010 \\
 & & 1 \\
\hline
 & \underbrace{1000}_{8} & \underbrace{1000}_{8}
\end{array}
$$

end around carry

add 0001

add end around carry

(55 in XS-3)

$$
\begin{array}{r}
87 \\
-32 \\
\hline
55
\end{array}
$$

7) b) Perform the BCD subtraction for 798 & 389

7 9 8
3 8 9

9's complement of 389 = 610

$$
\begin{array}{rrrr}
798 \rightarrow & 0111 & 1001 & 1000 \\
610 \rightarrow & 0110 & 0001 & 0000 \\
\oplus & & & \\
\hline
 & 1101 & 1010 & 1000 \\
\boxed{1} & 0110 & 0110 & \\
\hline
 & 0110 & 0000 & 1000 \\
\hline
 & & & +1 \\
\hline
 & \underbrace{0101}_{4} & \underbrace{0000}_{0} & \underbrace{1001}_{9}
\end{array}
$$

1101 and 1010 are illegal codes.

add 0110

end around carry

Correct difference

$$
\begin{array}{r}
798 \\
-389 \\
\hline
409
\end{array}
$$

8 a) Given the Boolean function

$$Y(A,B,C,D) = A + BC + AB\bar{D} + ABCD$$

Convert to Standard SOP.

$A + BC + AB\bar{D} + ABCD$

$= A(B+\bar{B})(C+\bar{C})(D+\bar{D}) + BC(A+\bar{A})(D+\bar{D}) + AB(C+\bar{C})\bar{D} + ABCD$

$= (AB + A\bar{B})(CD + C\bar{D} + \bar{C}D + \bar{C}\bar{D}) + BC[AD + A\bar{D} + \bar{A}D + \bar{A}\bar{D}] + AB[C\bar{D} + \bar{C}\bar{D}] + ABCD$

$= ABCD + ABC\bar{D} + AB\bar{C}D + AB\bar{C}\bar{D} + A\bar{B}CD + A\bar{B}C\bar{D} + A\bar{B}\bar{C}D + A\bar{B}\bar{C}\bar{D} + ABCD + ABC\bar{D} + \bar{A}BCD + \bar{A}BC\bar{D} + ABC\bar{D} + AB\bar{C}\bar{D} + ABCD$

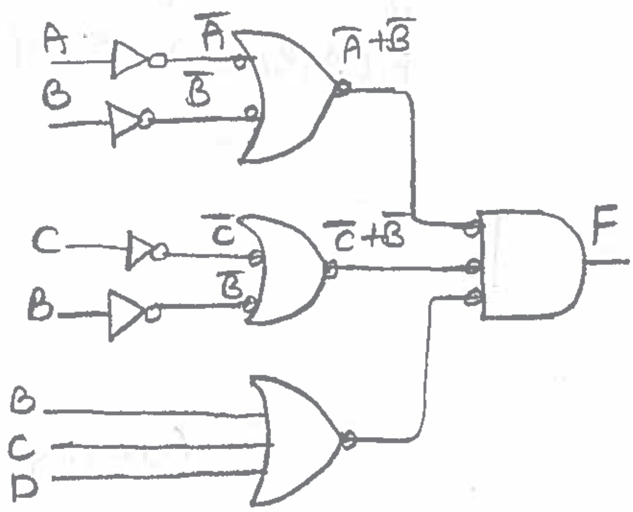
$= ABCD + ABC\bar{D} + AB\bar{C}D + AB\bar{C}\bar{D} + A\bar{B}CD + A\bar{B}C\bar{D} + A\bar{B}\bar{C}D + A\bar{B}\bar{C}\bar{D} + \bar{A}BCD + \bar{A}BC\bar{D}$

---

8 b) Find the reduced POS form using K-Map $F(A,B,C,D) = \pi M(0,6,7,8,12,13,14,15)$ implement using NOR gates.

AB \ CD	$C+D$	$C+\bar{D}$	$\bar{C}+\bar{D}$	$\bar{C}+D$
$A+B$	0 (0)	(1)	(3)	(2)
$A+\bar{B}$	(4)	(5)	0 (7)	0 (6)
$\bar{A}+\bar{B}$	0 (12)	0 (13)	0 (15)	0 (14)
$\bar{A}+B$	0 (8)	(9)	(11)	(10)

Logic diagram



$F_1 = (\bar{A}+\bar{B})$ $\quad F_2 = (\bar{C}+\bar{B})$

$F_3 = B+C+D$

$F = (\bar{A}+\bar{B})(\bar{C}+\bar{B})(B+C+D)$

—0—

9) a) Convert to Canonical forms.

(i) $F(X,Y,Z) = XY + Z = \overline{XY+Z} = \overline{XY} \cdot \bar{Z}$

$= (\bar{X}+\bar{Y}) \cdot \bar{Z}$

— o —

(ii) $F2(X,Y,Z) = (X+\bar{Y})\,(\bar{X}+Z) = \overline{(X+\bar{Y})\,(\bar{X}+Z)}$

$= \overline{(X+\bar{Y})} + \overline{(\bar{X}+Z)} = (\bar{X} \cdot \bar{\bar{Y}}) + (\bar{\bar{X}} \cdot \bar{Z})$

$= (\bar{X}Y) + (X\bar{Z})$

— o —

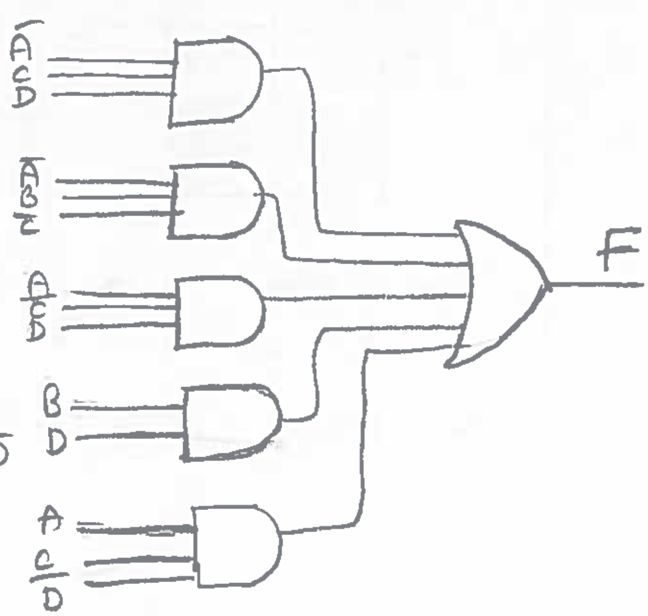
9) b) using K-map simplify the following expression and implement using NAND gates.

$F(A,B,C,D) = \Sigma m(1,5,6,7,11,12,13) + \Sigma d(10,15)$

CD \ AB	$\bar{A}\bar{B}$	$\bar{A}B$	AB	$A\bar{B}$
$\bar{C}\bar{D}$	^{0}	^{1} 1	^{3}	^{2}
$\bar{C}D$	^{4}	^{5} 1	^{7} 1	^{6} 1
CD	^{12} 1	^{13} 1	^{15} X	^{14}
$C\bar{D}$	^{8}	^{9}	^{11} 1	^{10} X

$F = \bar{A}CD + \bar{A}B\bar{C} + BD + A\bar{C}D + AC\bar{D}$

Logic diagram:



— o —

10) a) Implement the function using only one 8:1 mux and gates.

$F(A,B,C,D) = \Sigma m(0,2,3,6,8,9,11,13)$

Consider D as data I/P and A,B,C are the select lines

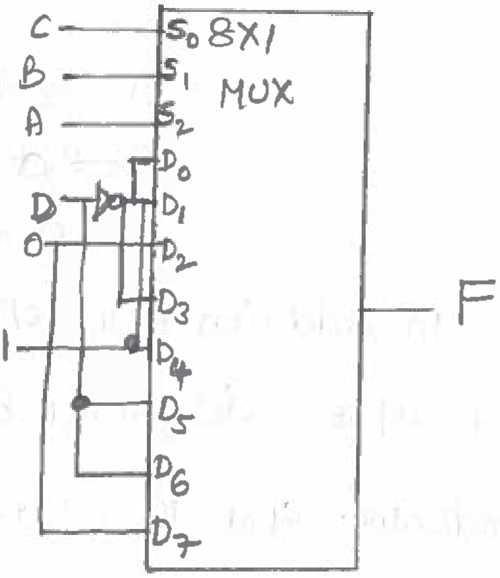
$CD \backslash AB$	$\bar A\bar B$	$\bar A B$	AB	$A\bar B$
$\bar C\bar D$	1 (0)	(1)	1 (3)	1 (2)
$\bar C D$	(4)	(5)	(7)	1 (6)
CD	(12)	1 (13)	(15)	(14)
$C\bar D$	1 (8)	1 (9)	1 (11)	(10)

$F = \bar C A \bar B + AB\bar D + \bar A\bar B\bar D + \bar A C\bar D + A\bar C\bar D + AB\bar D + \bar A BCD$

Truth Table

Logic diagram

		S_2 A	S_1 B	S_0 C	D	F	
D_0	0	0	0	0	0	1	$F=\bar D$
	1	0	0	0	1	0	
D_1	2	0	0	1	0	1	$F=1$
	3	0	0	1	1	1	
D_2	4	0	1	0	0	0	$F=0$
	5	0	1	0	1	0	
D_3	6	0	1	1	0	1	$F=\bar D$
	7	0	1	1	1	0	
D_4	8	1	0	0	0	1	$F=1$
	9	1	0	0	1	1	
D_5	10	1	0	1	0	0	$F=D$
	11	1	0	1	1	1	
D_6	12	1	1	0	0	0	$F=D$
	13	1	1	0	1	1	
D_7	14	1	1	1	0	0	$F=D$
	15	1	1	1	1	0	



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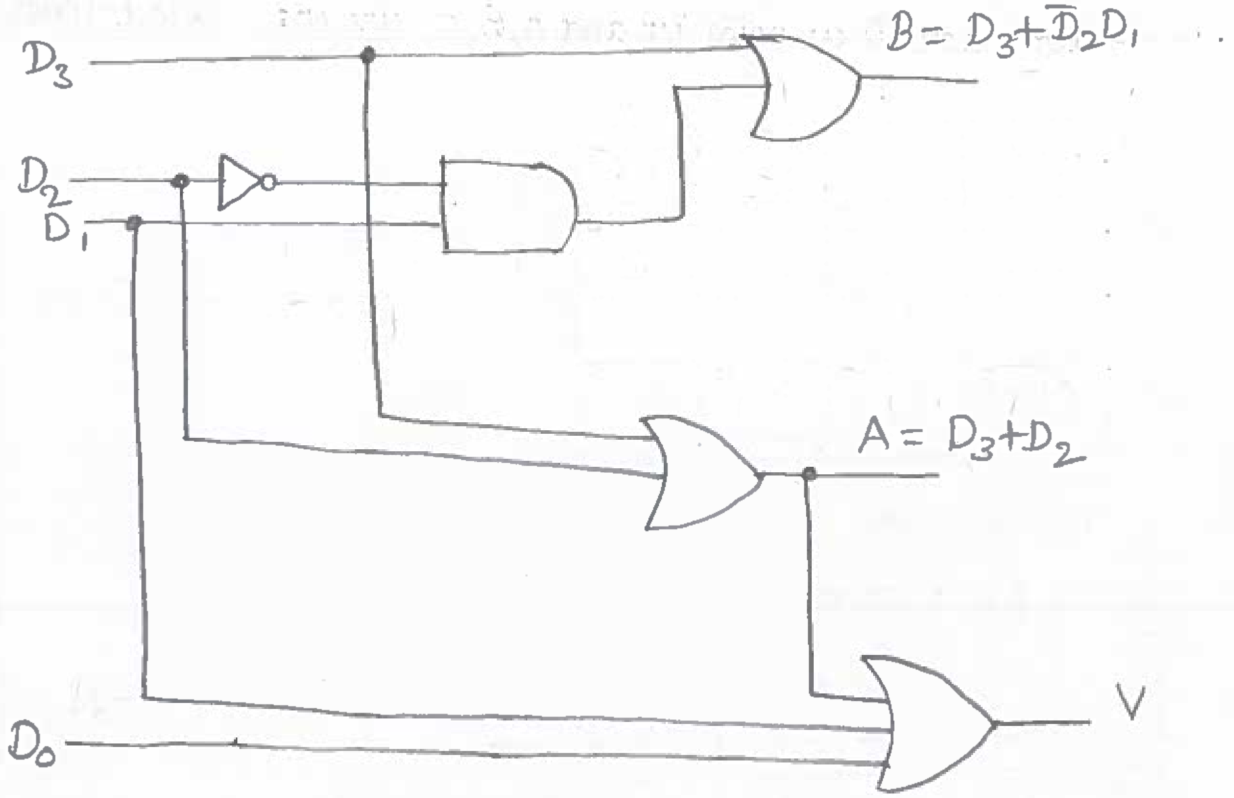
10 b) Working of a Priority encoder:–

A Priority encoder is a logic circuit that responds to just one input in accordance with some priority system, among all those that may be simultaneously HIGH.

4-bit Priority encoder Logic diagram



$$A = D_3 + D_2$$
$$B = D_3 + \bar{D}_2 D_1$$
$$V = D_3 + D_2 + D_1 + D_0.$$

In addition to the o/ps A and B the circuit has a 3rd output designated by V. This is a valid bit indicator that is set to 1 when one or more inputs are equal to 1.

If all i/ps are 0 there is no valid i/p. and V is equal to 0.

The other two outputs are not inspected when V=0 and are specified as don't care conditions.

Truth table

inputs				outputs		
D_0	D_1	D_2	D_3	A	B	V
0	0	0	0	X	X	0
1	0	0	0	0	0	1
X	1	0	0	0	1	1
X	X	1	0	1	0	1
X	X	X	1	1	1	1

According to the Truth table, the higher the subscript number, the higher the Priority of the i/p.

input D_3 has the highest Priority.

So regardless of the values of the i/ps when this i/p is 1 the o/p for AB is 11 (Binary 3) D_2 has the next Priority level. The o/p is 10. if $D_2=1$ Provided that $D_3=0$ regardless of the values of the other two lower Priority i/ps.

The o/p for D_1 is generated only if higher Priority i/ps are 0 and so on down the Priority levels.

11) a) Implement the following functions using PLA having 3 i/ps, 4 Product terms and 2 o/ps.

$$F1(A,B,C) = \Sigma m(3,5,6,7)$$

$$F2(A,B,C) = \Sigma m(0,2,4,7)$$

(P.T.O)

K-Map for F_1

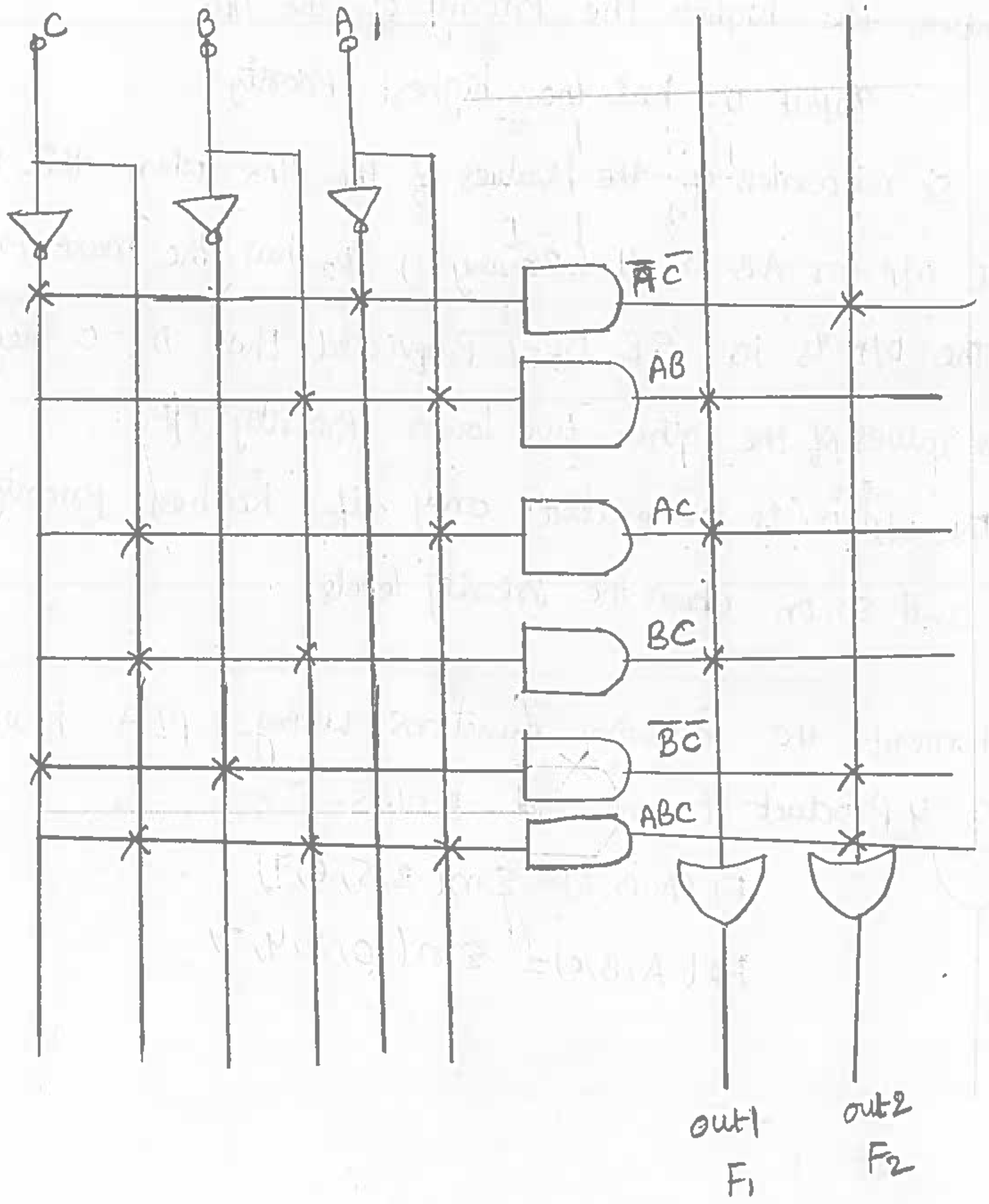
$A \backslash BC$	$\bar{B}\bar{C}$	$\bar{B}C$	BC	$B\bar{C}$
$\bar{A}$			1	
A		1	1	1

$$F_1 = AB + AC + BC$$

K-Map for F_2

$A \backslash BC$	$\bar{B}\bar{C}$	$\bar{B}C$	BC	$B\bar{C}$
$\bar{A}$	1			1
A	1		1	

$$F_2 = \bar{B}\bar{C} + \bar{A}\bar{C} + ABC$$



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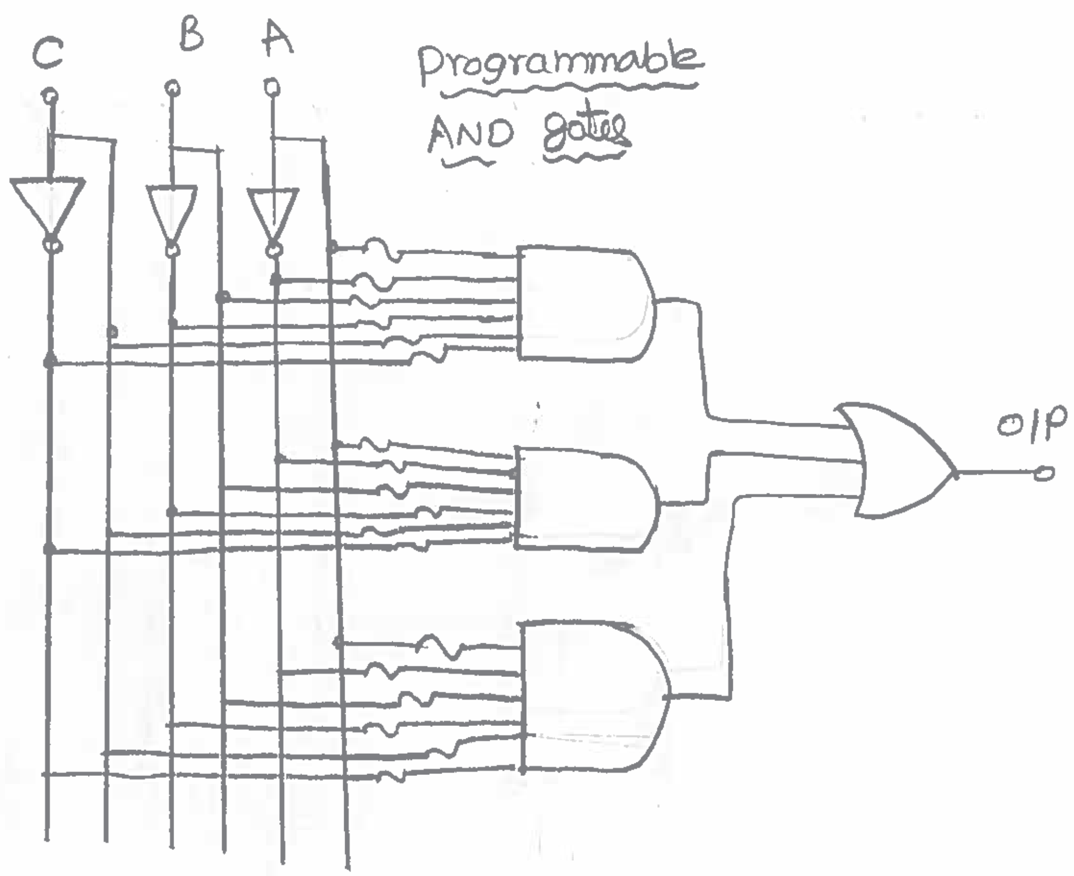
11) b) Draw the basic structures of PAL and PLA.

Basic structure of a PAL circuit



Programmable Array logic (PAL) is a particular family of PLDS.

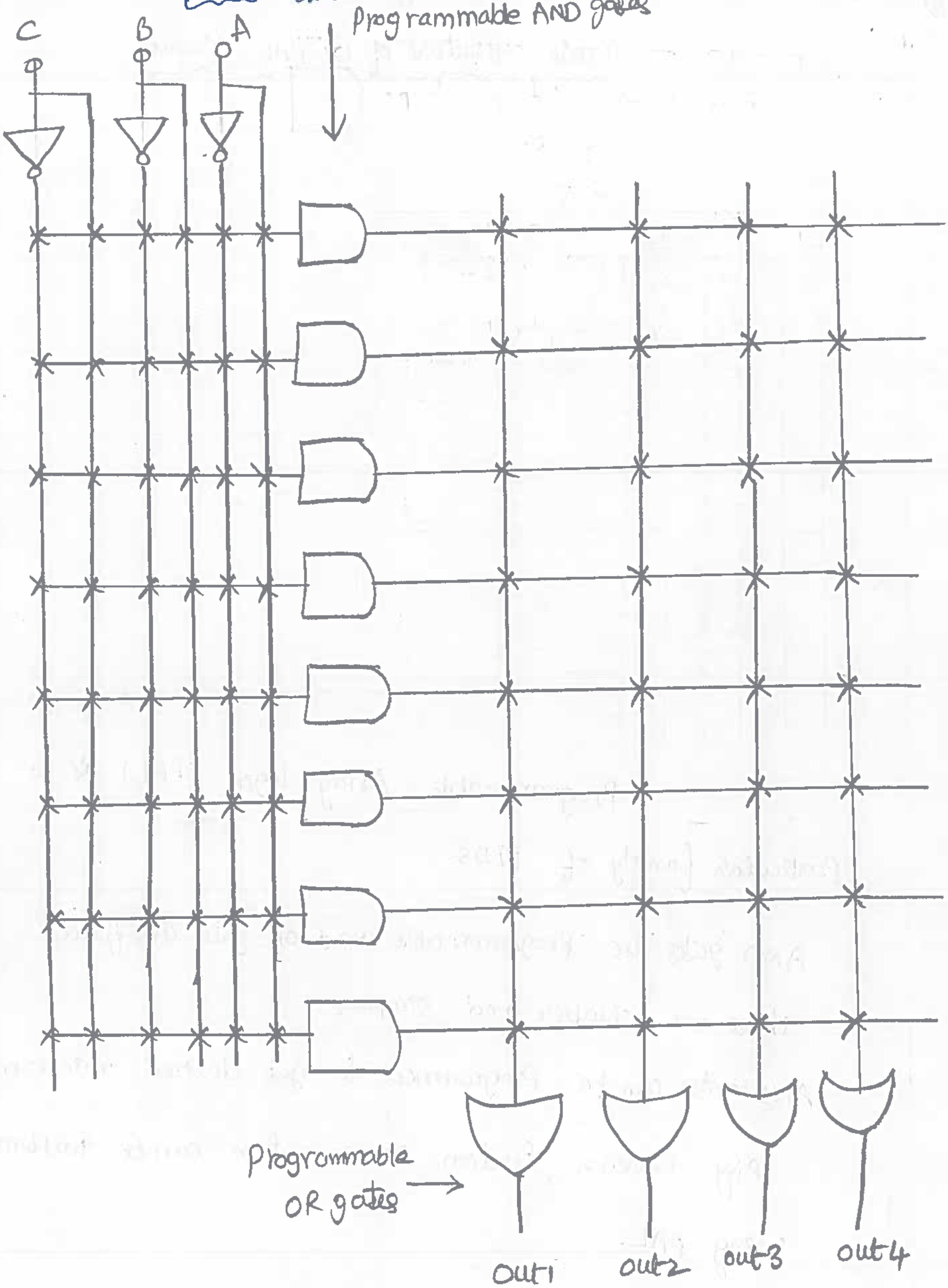
AND gates are programmable and OR gates are fixed.

These are cheaper and simpler.

AND gates can be programmed to get desired minterms

Any Boolean function in SOP form can be implemented using PAL.

Structure of a PLA :-



Both AND gates and OR gates are Programmable.

— o —

12) a) State the excitation table of J-K FF:-

J-K Truth table

J	K	Q_{n+1}
0	0	Q_n
0	1	0
1	0	1
1	1	$\overline{Q_n}$

J-K excitation Table

PS	NS	required J	flips K
0	0	0	X
0	1	1	X
1	0	X	1
1	1	X	0

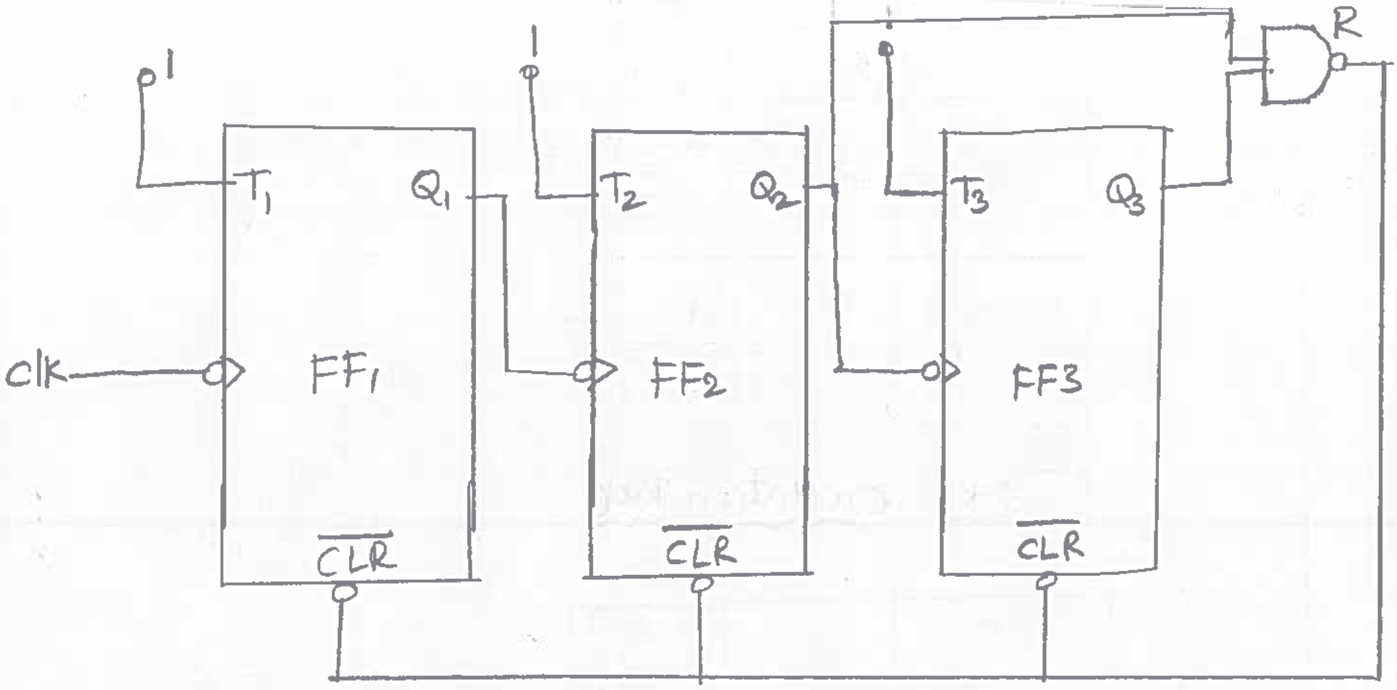
12) b) Design a binary counter using T- flip flops to count in the following sequence.

000, 001, 010, 011, 100, 101, 110, 000

After pulses	States			R
	Q_3	Q_2	Q_1	
0	0	0	0	0
1	0	0	1	0
2	0	1	0	0
3	0	1	1	0
4	1	0	0	0
5	1	0	1	0
6	1	1	0	1
⋮	↓ 0	↓ 0	↓ 0	0
7	0	0	1	0

R=0 for 000 to 101. R=1 for 110 hence R=X for 111

$$R = Q_3 Q_2 \bar{Q}_1 + Q_3 Q_2 Q_1 = Q_3 Q_2$$



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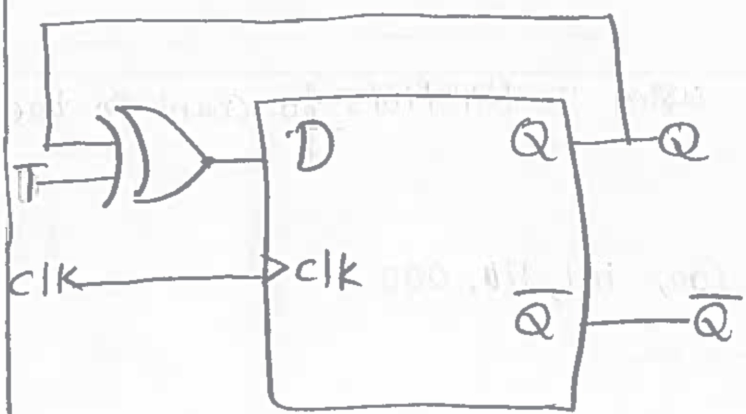
13) a) Implement T- FF using D FF.



D to T Conversion table

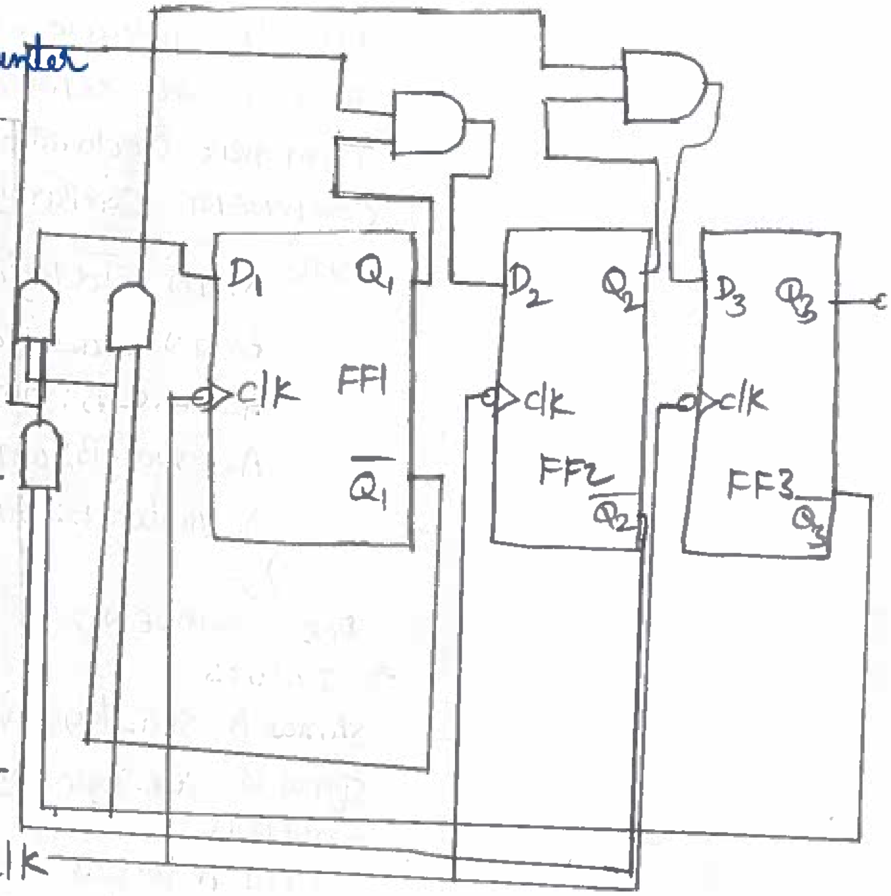
I/P	Intermediate I/PS			O/PS	
T	Q	$\bar{Q}$	$D = T \oplus Q$	Q	$\bar{Q}$
0	0	1	0	0	1
0	1	0	1	1	0
1	0	1	1	1	0
1	1	0	0	0	1

— 0 —

B) b) Design a synchronous Counter that counts the Sequence 000, 001, 010, 011, 100, 101, 110, 111, 000 using D- flip flop.

Excitation table for the counter

PS				NS			
Q_D	Q_C	Q_B	Q_A	Q_{D+1}	Q_{C+1}	Q_{B+1}	Q_{A+1}
0	0	0	0	0	0	0	1
0	0	0	1	0	0	1	0
0	0	1	0	0	0	1	1
0	0	1	1	0	1	0	0
0	1	0	0	0	1	0	1
0	1	0	1	0	1	1	0
0	1	1	0	0	1	1	1
0	1	1	1	1	0	0	0
1	0	0	0	1	0	0	1
1	0	0	1	0	0	0	0



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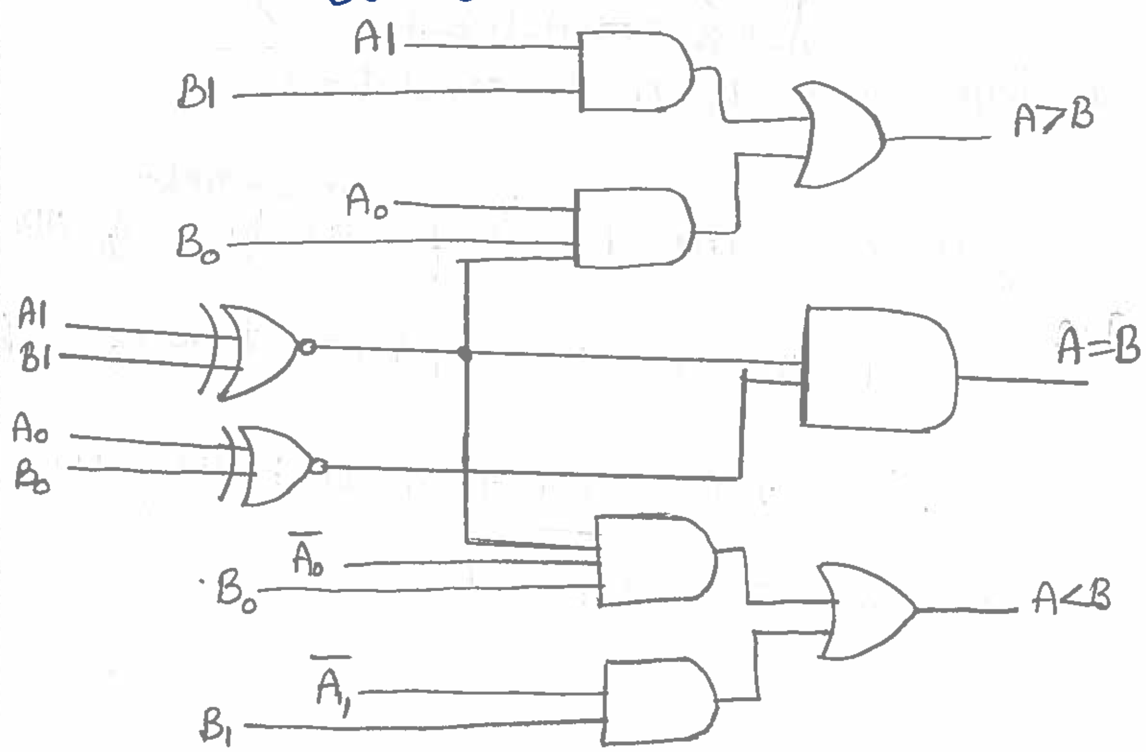
14) a) Design a 2-bit Magnitude Comparator and write a VHDL Code.

2-bit Magnitude Comparator logic diagram



VHDL Code:-

```
library ieee;
USE ieee.std_logic_1164.ALL;
USE ieee.numeric_std.all;
ENTITY tb_Comparator_VHDL IS
END tb_Comparator_VHDL;
ARCHITECTURE behaviour OF tb_comparator_VHDL
Component declaration for the comparator
COMPONENT Comparator_VHDL
PORT(  A: IN  Std_logic_vector(1 downto 0);
       B: IN  Std_logic_vector(1 down to 0);
       A_Less_B: OUT Std_logic;
       A_equal_B: OUT std_logic;
       A_greater_B: OUT Std_logic;
      );
END COMPONENT;
-- Inputs
signal A: std_logic_Vector(1 down to 0); =(others =>'0');
signal B: Std_logic_Vector(1 down to 0); =(others =>'0');
--- outputs
signal A_less_B : std_logic;
signal A_equal_B; std_logic;
signal A_greater_B: Std_logic;
BEGIN
  Comparator_VHDL PORT MAP(
      A => A;
      B => B,
      A_less_B => A_Less_B
      A_equal_B => A_equal_B
      A_greater_B => A_greater
  );
```

(4) b) Define logic synthesis and simulation

Logic synthesis:- The Process of ~~analysis~~ transformation of HDL code to primitive gates and FFs ~~for the~~ implementation of data registers, buses, logic units and their Controlling H/W is known as logic synthesis.

Logical Synthesis transforms HDL code into a net list describing the H/W.

Simulation:- The model of a set of problems or events that can be used to teach someone how to do something or the process of making such a model is known as simulation. It describes the behaviour of the ckt in terms of i/p signals, o/p signals and knowledge of delays.

15) a) Explain VHDL Programming using structural modeling.

Structural style of modeling:-

In this style of modeling, an entity is described as a set of interconnected components. Such as a model for the HALF ADDER is shown below.

```
architecture HA-STRUCTURE of
HALFADDER is component
XOR2
      Port(X,Y: in BIT; Z: out BIT);
      end component;
      Component AND2
         Port (L, M: in BIT; N: out BIT);
      end component;
begin
      X1: XOR2 Port map(A,B, SUM);
      A1: AND2 Port map( A,B, CARRY);
end  HA-STRUCTURE;
```

The name of the architecture body is HA- STRUCTURE.

The entity declaration for HA-ADDER specifies the interface ports for this architecture body.

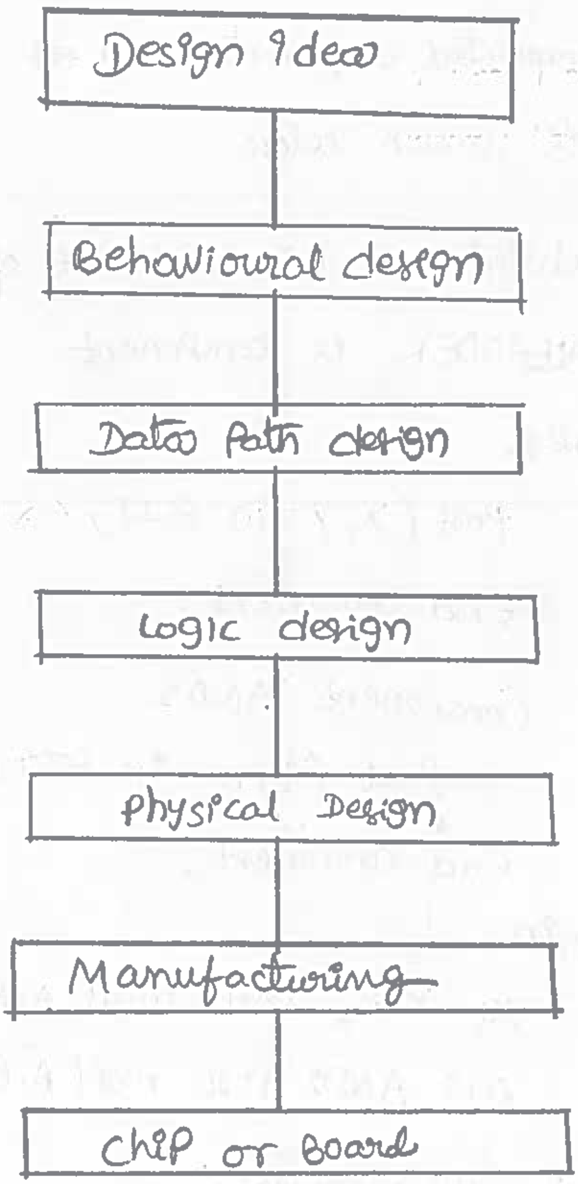
X1 and A1 are the component labels

A and B i/p ports

Z is the output port. SUM.

15) b)

Generic IC Design flow



The design team forms a general idea about the solution to the problem.

At the end of this stage a general Block diagram solution of the design is agreed upon.

CAD tools are generally not needed at this stage.

Behavioural Design— These are used to model the design idea. ckt details and electrical Components are not specified.

The behaviour of each block is modeled.

Data Path design:-

In this phase, designer specifies the registers and logic units necessary for implementation of the system.

Logic design:- It involves the use of primitive gates and flip flops for the implementation of data registers, buses, logic units and their controlling H/w.

The result is a net list of gates and FFs.

physical design:- This stage transforms the net list into Transistor list or lay out. It involves replacement of gates and FFs with Transistors.

Manufacturing:- The final step is manufacturing. It uses the Transistor list or layout specification to burn fuses of FPGA or to generate masks for IC.

— o —

Y.H.D. Aparna

8/12/22
HOD—ECE

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	ECE			**Academic Year**	2022 - 2023
Course Code	20EC503	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	V
Course	**Antennas & Wave Propagation**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Radiation Pattern	20EC503.1	L1
2	Define Radiation Resistance of An Antenna	20EC503.2	L1
3	What are the applications of spiral antenna?	20EC503.3	L1
4	What is Concept of adaptive beam forming?	20EC503.4	L1
5	Define Virtual Height and Fading	20EC503.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain the following terms with proper expressions i)Directivity ii)Field pattern ii)Half power beam width iv)Beam efficiency v) Radiation Intensity vi) Polarization	12M	20EC503.1	L2
	OR			
7 (a)	Derive Friss transmission equation	6M	20EC503.1	L3
7 (b)	Derive the relationship between directivity and effective area	6M	20EC503.1	L3
8	Derive the field components and radiation resistance of a half wave dipole antenna.	12M	20EC503.2	L3
	OR			
9 (a)	Explain in detail about Broadside and End-fire arrays	6M	20EC503.2	L2
9 (b)	Explain the Operating Principle of Yagi -Uda Antanna	6M	20EC503.2	L2
10 (a)	Explain the Design and construction of Helical Antenna	6M	20EC503.3	L2
10 (b)	Discuss the design considerations of Pyramidal Horn Antenna	6M	20EC503.3	L2
	OR			
11 (a)	Explain about construction of Spiral Antenna	6M	20EC503.3	L2
11 (b)	What are different types of feed mechanism used in parabolic reflector antenna?	6M	20EC503.3	L2
12 (a)	Explain the principle of Lens antenna	7M	20EC503.4	L2
12 (b)	Explain about different types of smart antennas.	5M	20EC503.4	L2
	OR			
13 (a)	Explain the measurement of gain of an antenna	7M	20EC503.4	L2
13 (b)	Explain about microstrip antenna	5M	20EC503.4	L2
14	Derive the Refractive Index of an ionosphere layer in Sky Wave Propagation	12M	20EC503.5	L3
	OR			
15 (a)	Explain in detail about Ground Wave Propagation	6M	20EC503.5	L2
15 (b)	Explain the following terms: (i) Critical Frequency (ii) Skip Distance (iii) Virtual Height	6M	20EC503.5	L2

Sub:- Antennas & wave propagation
Code : 20EC503
Prepared by:
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ANSWER KEY AND SCHEME OF EVALUATION

PART–A

① Radiation Pattern:-

The graphical representation of radiation of an antenna as a function of direction is given the name radiation pattern of Antenna → (2) Marks

② Radiation Resistance:-

Input impedance of an antenna is given by

$$Z_A = R_A + jX_A$$

R_A = resistive part of input imedence

$$R_A = R_r + R_L$$

where $R_r \rightarrow$ Radiation resistance

Radiation resistance is a effective resistance, due to the power carried away from the antenna as radiowaves. Radiation Resistance is unlike conventional resistance, it is the loss not due to the opposition to current, resistance due to radiation of radiowaves. → (2) M

③ Applications of Spiral Antenna:-

→ used in defence industry for sensing applications
→ used in military aircraft as it does not occupy more space.
→ used in GPS, Microwave direction finding applications
→ (2) M

(4) Adaptive beam foaming :-

By using Adaptive Algorithms, the array produce maximum radiation of the antenna pattern towards the SOI and places Nulls in the pattern towards SNOIs.

The Adaptive beam foaming technique doesnot need DOA information but instead uses the reference signal (or) Training Sequence the magnitude and phases.

$\longrightarrow$ 2M

(5) Virtual Height :-

Virtual height of an ionosphere layer may be defined as the height to which a short pulse of energy sent vertically upwards and travelling with a speed of light would reach taking the same two ways travel time as does the actual pulse reflected from the layer.

Fading :-

Fading is the fluctuation in the received signal strength at the receiver or a random variation in the received signal is known as fading

$\longrightarrow$ 2M

Part B (Long Answer Questions)

6) $\longrightarrow$ total 12 M

i) Directivity :- (2M)

The Directive gain (or) Directivity in a given direction is defined as the ratio of the radiation-Intensity in that direction to the Avg radiation Intensity.

$$G_d = \frac{\text{Radiation Intensity in a particular direction}}{\text{Average radiation Intensity}}$$

$$G_d(\theta,\phi) = \frac{\phi(\theta,\phi)}{\phi_{avg}} = \frac{\phi(\theta,\phi)}{P_r/4\pi}$$

$$G_d(\theta,\phi) = \frac{4\pi\,\phi(\theta,\phi)}{P_r} = \frac{4\pi\,\phi(\theta,\phi)}{\int \phi\, d\Omega}$$

* The max. directive gain is called directivity.

ii) Field pattern :- (2M)

If the radiation from the antenna is expressed in terms of Field strength E(V/m), then radiation pattern is called as the "Field strength pattern". Radiation Field strength may have components E_θ & E_ϕ.

$$E = \sqrt{E_\theta^2 + E_\phi^2}$$

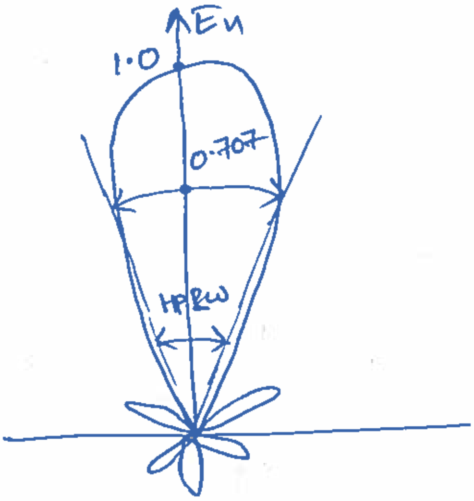
$E \rightarrow$ Total electric field strength

$E_\theta \rightarrow$ Amplitude of θ component

$E_\phi \rightarrow$ Amplitude of ϕ component

iii) HPBW :- (2M)

HPBW is an angular width (in degrees), measured on the major lobe of an antenna radiation-pattern at half power points.



(iv) Beam efficiency :- (2M)

Antenna Beam efficiency is a parameter that is frequently used to judge the quality of transmitting and receiving antennas.

$$BE = \frac{\text{Power transmitted (or received) with in cone angle } \theta}{\text{Power transmitted (or received) by the antenna}}$$

$$BE = \frac{\int_0^{2\pi}\int_0^{\theta_1} \phi(\theta,\phi)\sin\theta \, d\theta \, d\phi}{\int_0^{2\pi}\int_0^{\pi} \phi(\theta,\phi)\sin\theta \, d\theta \, d\phi}$$

In terms of Beam Area

$$BE = \frac{\text{Main beam Area}}{\text{Total beam area}} = \frac{\Omega_M}{\Omega_A}$$

v) Radiation Intensity :- (2M)

Radiation Intensity is a quantity which doesnot depend upon the distance from the radiator and it is denoted by the capital letter U (or) ϕ.

Radiation Intensity defined as "power per unit solid angle". units are watts/steradian

Radiation Intensity in a given direction is defined as "The power radiated from an antenna per unit solid angle".

Radiation intensity of Isotropic Source is $\boxed{U = \frac{P_{rad}}{4\pi}}$

$$\boxed{U = r^2 P_{rad}}$$

(vi) Polarization :- (2M)

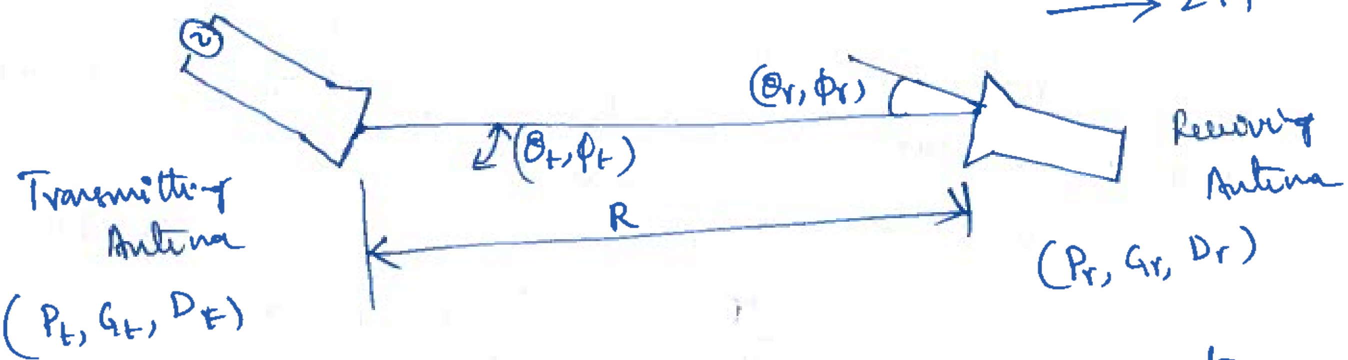
The polarization of Antenna defined as, the direction of the EM fields produced by the antenna as energy radiates away from it.

These directional Fields determine the direction in which the energy moves away from or is received by an antenna.

7(a) Friis Transmission equation :- Total (6 M)

The Friis transmission eq relates to the power received to the power transmitted b/w 2-antennas separated by a distance $R > \frac{2D^2}{\lambda}$, where D is the largest dimension of either Antenna.

$\longrightarrow$ 2 M



Let us assume that initially transmitting antenna is isotropic

If the i/p power at the terminals of the transmitting antenna is P_t, then its isotropic power density 'w_0' at distance R from antenna

$$w_0 = e_t \frac{P_t}{4\pi R^2}$$

$e_t \rightarrow$ radiation efficiency of transmitting antenna

For a nonisotropic transmitting antenna, the power density in the direction (θ_t, ϕ_t) can be written as

$$w_t = \frac{P_t G_t(\theta_t, \phi_t)}{4\pi R^2} = \frac{e_t P_t D_t(\theta_t, \phi_t)}{4\pi R^2}$$

$G(\theta_t, \phi_t) \rightarrow$ is the gain $\longrightarrow$ 2 M

$D_t(\theta_t, \phi_t) \rightarrow$ Directivity

Effective area A_r of an antenna is related to its efficiency e_r as

$$A_r = e_r D_r(\theta_r, \phi_r)\left(\frac{\lambda^2}{4\pi}\right)$$

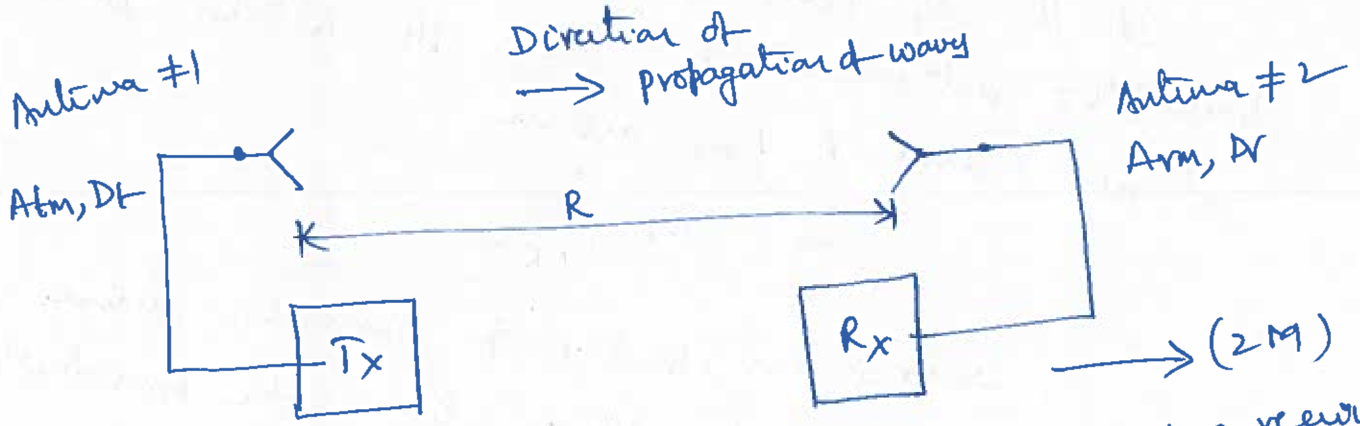
Power Collected $$P_r = e_r D_r(\theta_r, \phi_r) \frac{\lambda^2}{4\pi} \omega_t = e_t e_r \frac{\lambda^2 D_t(\theta_t, \phi_t) D_r(\theta_r, \phi_r)}{4\pi R^2} P_t$$

$$\frac{P_r}{P_t} = e_t e_r \frac{\lambda^2 D_t(\theta_t, \phi_t) D_r(\theta_r, \phi_r)}{4\pi R^2}$$

For reflection and polarisation-matched antennas aligned for maximum directional radiation and reception the above equation reduces to

$$\frac{P_r}{P_t} = \left(\frac{\lambda}{4\pi R}\right)^2 G_{0t} G_{0r}. \rightarrow 2M$$

7 (b) Relationship b/w Directivity and effective area :- Total (6M)



$\rightarrow (2M)$

Antenna 1 used as transmitter and 2 as a receiver. The effective areas and directivities are designated as A_t, A_r and D_t, D_r.

If antenna-1 is isotropic, then radiated power density at a distance R

$$w_0 = \frac{P_t}{4\pi R^2}$$

Because of directive properties, the actual density is

$$w_t = w_0 D_t = \frac{P_t D_t}{4\pi R^2}$$

Power Received $\quad P_r = w_t \cdot A_r = \frac{P_t D_t A_r}{4\pi R^2}$

(or)

$$D_t A_r = \frac{P_r}{P_t}(4\pi R^2) \quad \text{— (1)}$$

If antenna-2 used as transmitter, similarly

$$D_r A_t = \frac{P_r}{P_t}(4\pi R^2) \quad \text{— (2)}$$

$\longrightarrow$ (2M)

(1) & (2) reduces to

$$\frac{D_t}{A_t} = \frac{D_r}{A_r}$$

$$\frac{D_{0t}}{A_{tm}} = \frac{D_{0r}}{A_{rm}}$$

A_{tm}, A_{rm}, (D_{0t}, D_{0r}) are max effective areas (directivities) of antennas 1, 2 respectively

If Antenna-1 is isotropic then $D_{0t} = 1$

$$A_{tm} = \frac{A_{rm}}{D_{0r}}$$

For short Dipole $\quad A_{tm} = \frac{A_{rm}}{D_{0r}} = \frac{0.119\lambda^2}{1.5} = \frac{\lambda^2}{4\pi}$

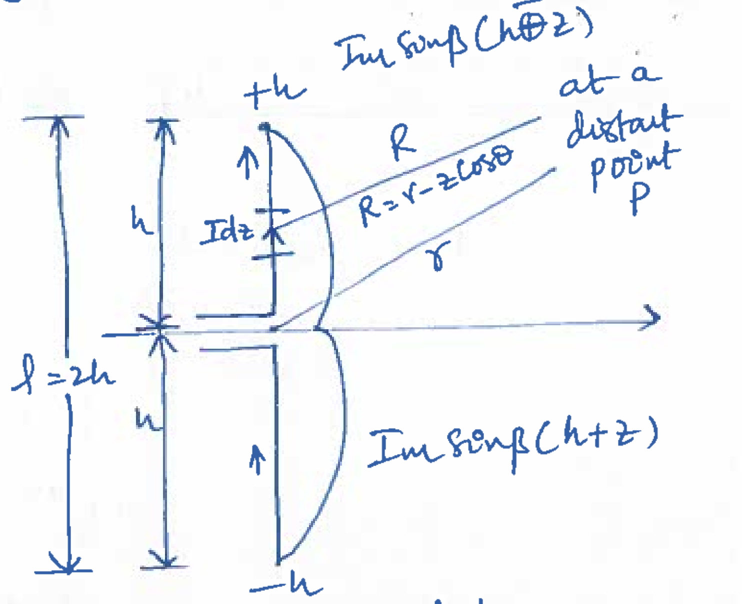
$$A_{rm} = D_{0r} A_{tm}$$

$$A_{rm} = D_{0r}\left(\frac{\lambda^2}{4\pi}\right)$$

In general $\quad \boxed{A_e = \frac{\lambda^2}{4\pi} D} \quad \longrightarrow$ (2M)

(8). $\longrightarrow$ Total 12 Marks

A Dipole antenna may be defined as a symmetrical antenna in which the two ends are at equal potential relative to the midpoint.



Since current is assumed to be sinusoidal

$I = I_m \sin\beta(h-z)$ for $z>0$

$I = I_m \sin\beta(h+z)$ for $z<0$

Vector potential at a distant point P due to current element Idz is given by $\longrightarrow$ (6 Marks)

$$dA_z = \frac{\mu I e^{-j\beta R}}{4\pi R} dz$$

$R \rightarrow$ Distance b/w Idz to distant point

The total vector potential due to all such elements is given by

$$\int dA_z = \int_{-h}^{0} \frac{\mu I dz\, e^{-j\beta R}}{4\pi R} + \int_{0}^{+h} \frac{\mu I dz}{4\pi R} e^{-j\beta R}$$

$$A_z = \frac{\mu}{4\pi}\int_{-h}^{0} \frac{I_m \sin\beta(h+z) e^{-j\beta R}}{R} dz + \frac{\mu}{4\pi}\int_{0}^{+h} \frac{I_m \sin\beta(h-z) e^{-j\beta R}}{R} dz$$

Since P is at large distance $R \simeq r$

$$A_z = \frac{\mu}{4\pi r}\left[\int_{-h}^{0} I_m \sin\beta(h+z) e^{-j\beta(r - z\cos\theta)} dz + \int_{0}^{h} I_m \sin\beta(h-z) e^{-j\beta(r-z\cos\theta)} dz\right.$$

$$A_z = \frac{\mu I_m e^{-j\beta r}}{4\pi r}\left[\int_{-h}^{0} \sin\beta(h+z)\, e^{j\beta z\cos\theta} + \int_{0}^{h} \sin\beta(h-z)\, e^{j\beta z\cos\theta}\, dz\right]$$

For halfwave dipole $l = 2h = \lambda/2$

$h = \lambda/4 = \pi/2$

$$\sin\beta(h+z) = \sin\beta(h-z) = \cos\beta z$$

$$A_z = \frac{\mu I_m e^{-j\beta r}}{4\pi r}\left[\int_{h}^{0} \cos\beta z\, e^{+j\beta z\cos\theta}\, dz + \int_{0}^{h} \cos\beta z\, e^{+j\beta z\cos\theta}\, dz\right]$$

$$A_z = \frac{\mu I_m e^{-j\beta r}}{2\pi\beta r}\left[\frac{\cos\left(\frac{\pi}{2}\cos\theta\right)}{\sin^2\theta}\right]$$

$$\mu H = \nabla\times A$$

$$\mu H = \begin{vmatrix} r & \theta & \phi \\ \frac{\partial}{\partial r} & \frac{1}{r}\frac{\partial}{\partial\theta} & \frac{1}{r\sin\theta}\frac{\partial}{\partial\phi} \\ A_r & A_\theta & A_\phi \end{vmatrix}$$

$A_\phi = 0$

$A_\theta = A_z\cos\theta$

$A_r = -A_z\sin\theta$

$H_r = H_\theta = 0$

$$H_\phi = \frac{1}{\mu}\left[\frac{\partial}{\partial r}(A_\theta) - \cancel{\frac{1}{r}\frac{\partial}{\partial\theta}(A_r)}\right] \quad 0$$

$$H_\phi = \frac{1}{\mu}\left[\frac{\partial}{\partial r}(A_\theta)\right] = \frac{1}{\mu}\left[\frac{\partial}{\partial r}\left[A_z\sin\theta\right]\right]$$

$$H_\phi = \frac{j I_m e^{-j\beta r}}{2\pi r}\,\frac{\cos\left(\frac{\pi}{2}\cos\theta\right)}{\sin\theta} \quad A/m^2$$

$$\boxed{|H_\phi| = \frac{I_m}{2\pi r}\,\frac{\cos\left(\frac{\pi}{2}\cos\theta\right)}{\sin\theta}} \quad A/m^2$$

$$\frac{|E_\theta|}{|H_\phi|} = \eta = 120\pi$$

$$E_\theta = 120\pi\, |H_\phi|$$

$$\boxed{|E_\theta| = \frac{60 I_m}{r} \frac{\cos\left(\frac{\pi}{2}\cos\theta\right)}{\sin\theta} \quad V/m} \longrightarrow (4M)$$

$$P_{avg} = \frac{E_\theta}{\sqrt{2}} \cdot \frac{H_\phi}{\sqrt{2}} = \frac{1}{2} E_\theta \cdot H_\phi$$

$$P_{avg} = \frac{30 I_{rms}^2}{\pi r^2} \left[\frac{\cos^2\left(\frac{\pi}{2}\cos\theta\right)}{\sin^2\theta}\right] \frac{W}{m^2}$$

$$W = \oint_s P_{av}\, ds$$

$$W = \int_0^\pi \frac{30 I_{rms}^2}{\pi r^2} \left[\frac{\cos^2\left(\frac{\pi}{2}\cos\theta\right)}{\sin^2\theta}\right] 2\pi r^2 \sin\theta\, d\theta$$

$$W = 60 \cdot I_{rms}^2 \cdot I$$

where $I = \frac{1}{2}\int_0^\pi \left[\frac{1 + \cos(\pi\cos\theta)}{\sin\theta}\right] d\theta = 1.219$

$$W = 60\, I_{rms}^2 \cdot 1.219$$

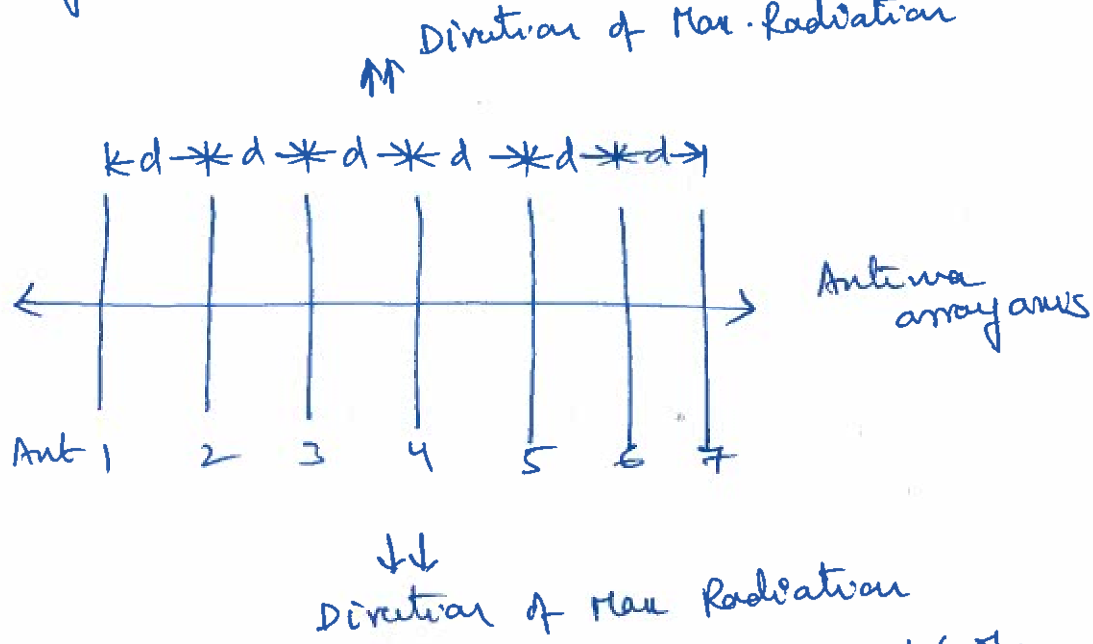
$$W = 73.14 \times I_{rms}^2$$

$$W = R_r \cdot I_{rms}^2$$

$$\boxed{R = 73.14 \simeq 73} \longrightarrow (2M)$$

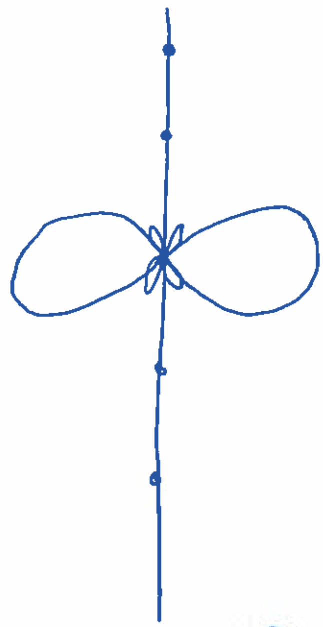
9) a) → total 6 M

Broadside Array :-



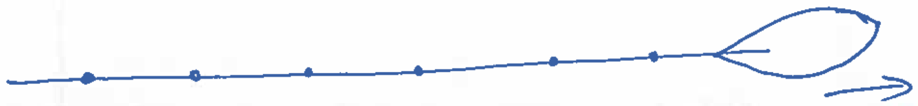
Broadside array is one in which a number of identical parallel antennas are setup along a line drawn perpendicular to their respective axis. In broad side array, the individual elements are equally spaced along a line and each element is fed with current of equal magnitude and phase.

By doing so the arrangement fires in broadside directions (i.e $\perp$ar to array axis). The radiation pattern is bidirectional. → (3 M)



9) (a) End fire array:-

The endfire array is nothing but broadside array except that individual elements are fed in out of phase (usually 180°)



In endfire array, a number of identical antenna are spaced equally along a line and individual elements are fed with currents of equal magnitude. Radiation pattern of end fire array is unidirectional. The individual elements are excited in a such a manner that a progressive phase difference b/w adjacent elements becomes equal to the spacing b/w the elements. → (3 M)

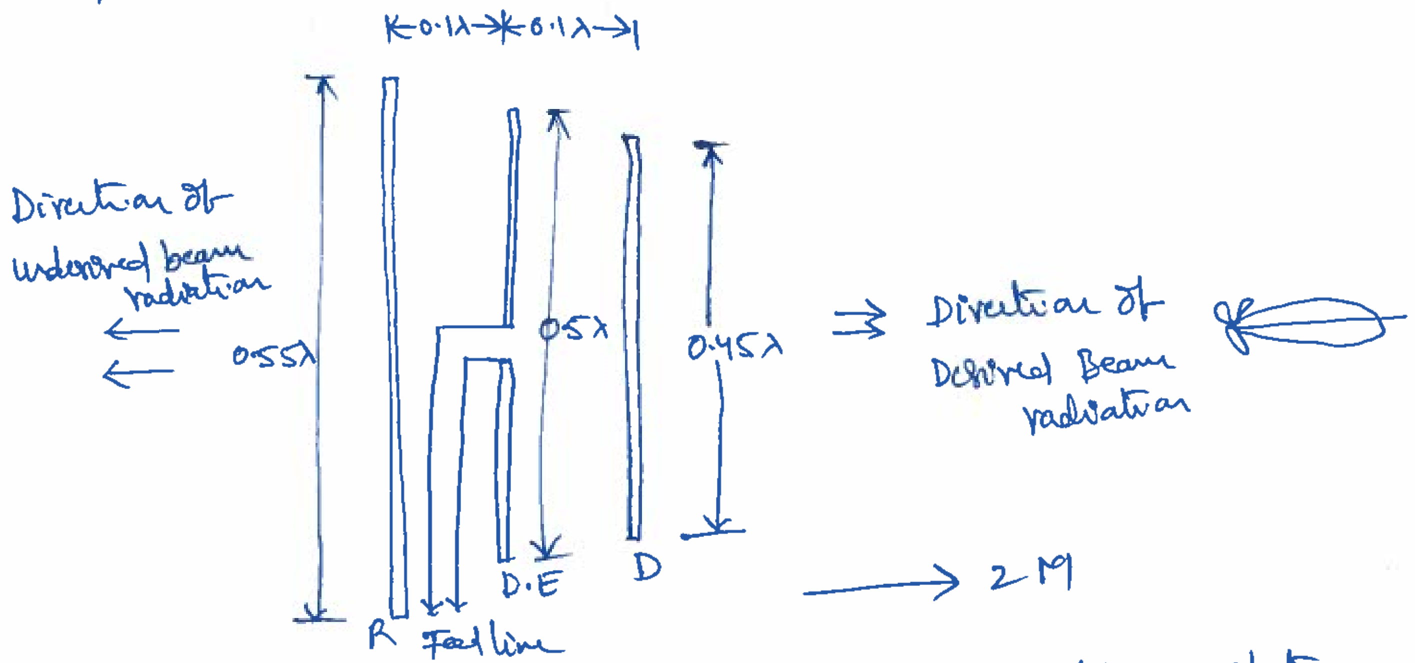
9(b) Yagi-uda Antenna : →Total 6 M

Yagi-uda Antenna is the most high gain antennas and are known after the names of Prof. S. uda and H. Yagi.

It consists of a driven element, a reflector and one (or) more Directors, i.e Yagi-uda antenna is an array of a driven element (Active element) and one or more parasitic elements (Passive elements)

The driven element is a resonant halfwave dipole usually of metalic rod at frequency of operation. The parasitic elements of continuous metalic rods arranged parallel to driven element. They arranged collinearly.

→ 2 M



→ 2 M

The parasitic elements receive their excitation from the voltages induced in them by the current flow in driven element.

The phase and currents flowing due to the induced voltage depend on the spacing b/w the elements and the length of the elements.

The spacing b/w driven element and parasitic elements are in order of 0.1λ to 0.15λ.

The parasitic element in front of driven element is known as Director and its number may be more than one, where as element back of it is known as reflector.

The Reflector is 5% more and director is 5% less length than the driven element, which is $\lambda/2$ at resonant frequency.

In Practice for 3-element array

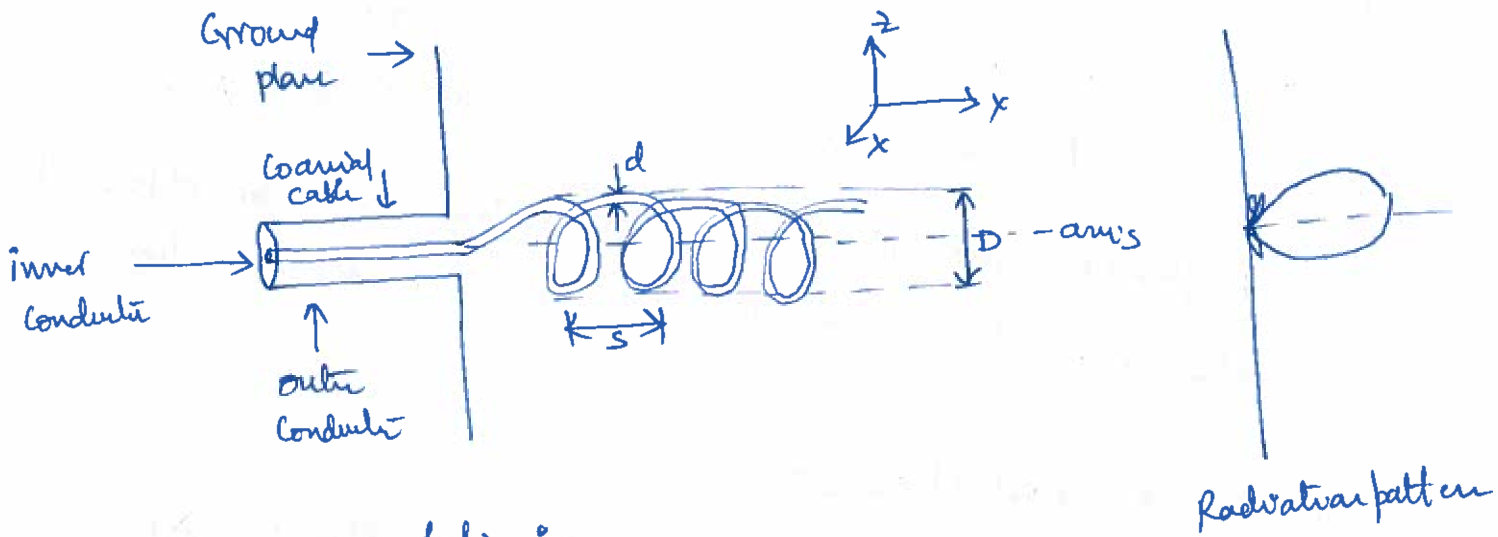
$$\text{Reflector length} = \frac{500}{f\,MHz}\ \text{feet}$$

$$\text{Driven element length} = \frac{475}{f\,MHz}\ \text{feet}$$

$$\text{Director length} = \frac{455}{f\,MHz}\ \text{feet}$$

$\longrightarrow$ 2M

10) a) Hellical Antenna : (6 Marks)



The dimensions of helix :

$C \rightarrow$ Circumfrance of helix $= \pi D$

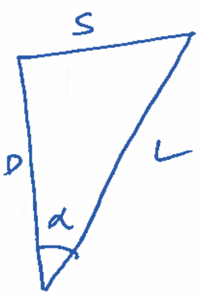
$\alpha \rightarrow$ Pitch angle $= \tan^{-1}\left(\frac{S}{\pi D}\right)$

$d \rightarrow$ Dia of helix conductor

$A \rightarrow$ Axial length $= N \cdot S$

$L \rightarrow$ length of one turn

$l \rightarrow$ spacing from helix to ground plane $\longrightarrow$ (3 M)



Hellical antenna is the simplest antenna to provide circularly polarized waves.

Hellical antenna is broad band UHF & VHF antenna to provide circular polarization characteristics.

It consists of helix of thick copper wire in a shape of screw thread and used as an antenna in conjunction with a flat metal plate called ground plane.

The ground plane is simply made of sheet or screen or of radial conducting wires.

The helix fed by a coaxial cable. The one end of helix is connected to the center conductor of the cable and other conductor is connected to the ground plane.

The parameter on which mode of radiation depend are the diameter of helix 'D' & turn spacing 'S'

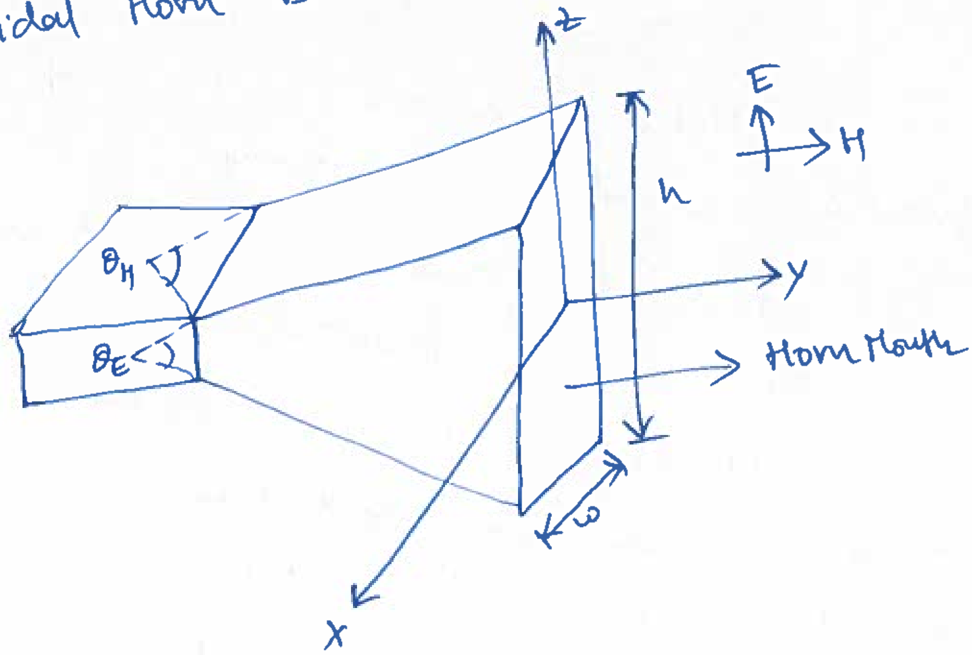
Different radiation characteristics can be obtained by changing the parameters in relation to wavelength

Modes of radiation :-

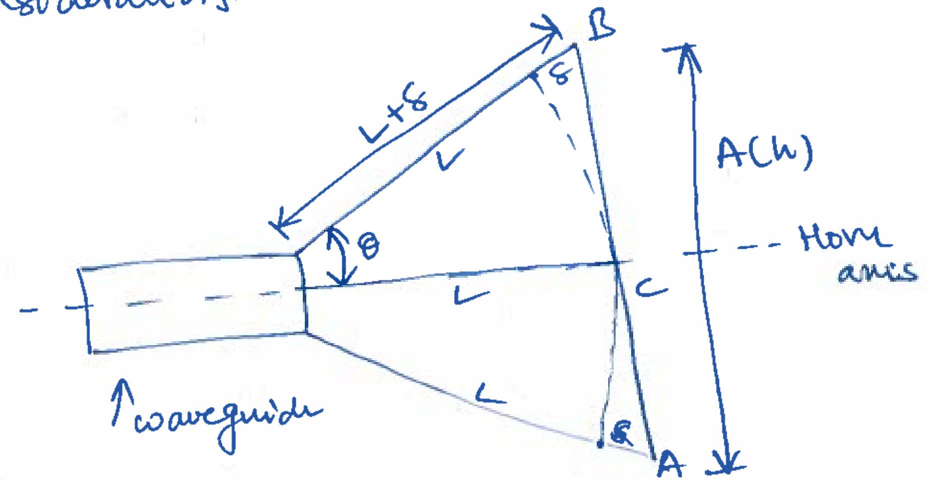
1) Normal mode of radiation $(NL << \lambda)$
2) Axial (or) Beam Mode of radiation $(NL \cong \lambda)$ $\longrightarrow$ (3 M)

10) b) Pyramidal Horn Antenna :- (6 Marks)

If the flaring is done in both walls of rectangular waveguide (E & H Directions) then the pyramidal Horn is obtained.



Design considerations.



L → axial length

A → Aperture

θ → Flare Angle

δ → Path difference ⟶ 4 Marks

From geometry $\cos\theta = \dfrac{L}{L+\delta}$; $\tan\theta = \left(\dfrac{h/2}{L}\right)$

$$\theta = \cos^{-1}\left(\frac{L}{L+\delta}\right) = \tan^{-1}\left(\frac{h}{2L}\right)$$

~~Fo~~ ~~oat~~ From Δ^{le} OBC

$$(L+\delta)^2 = L^2 + \left(\frac{h}{2}\right)^2$$

$$L^2+\delta^2+2L\delta = L^2 + \frac{h^2}{4}$$

If δ is small, δ^2 can be neglected

$$2L\delta = h^2/4$$

$$L^2 = \frac{h^2}{8\delta}$$

⟶ 2 M

11. a) Explain about construction of Spiral Antenna. (6 Marks)

→ It is a frequency independent antenna, i.e the antenna characteristics such as polarization, impedance radiation pattern etc., are constant over a frequency range.

→ It has very large Bandwidth (UWB – ultra wide band)

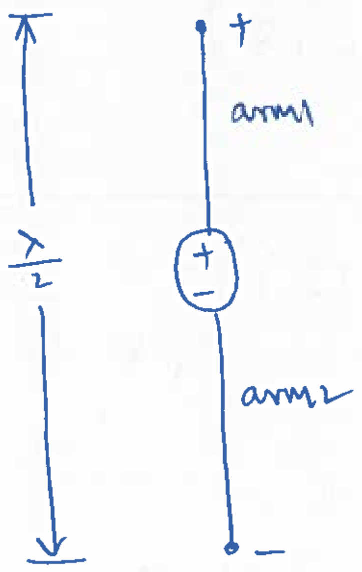
→ Its fractional bandwidth $\frac{f_H}{f_L} = 30$; which means lower cutoff frequency is 1 GHz and upper cutoff freq is 30 GHz

→ It produces circularly polarized waves.

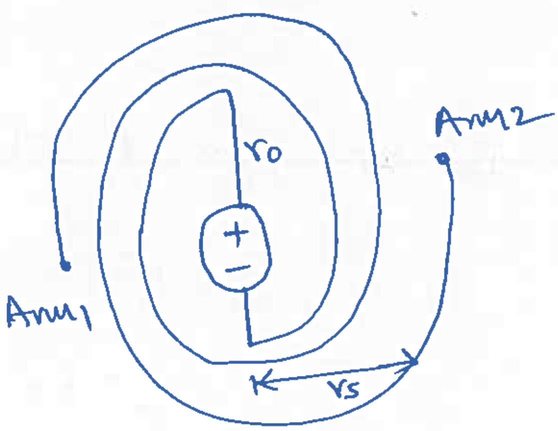
→ The maximum radiation pattern is in the direction perpendicular to the plane of spiral structure (Broadside Direction)

→ HPBW is in the range of 70° to 90°. → 2M

SPIRAL Antenna



$\lambda/2$ – Dipole



r_0 → inner radius

r_s → outer radius → 2M

Each arm of spiral Antenna is defined as

$$\boxed{r = r_0 e^{a\phi}}$$

The above equation states that spiral Antenna radius grows exponentially as it turns.

$r_0 \rightarrow$ Inner radius, Is a constant radius which controls the initial radius of the spiral.

$r_s \rightarrow$ outer radius.

$r_s \rightarrow$ Determines the lowest frequency of operation for spiral Antenna

$$f_{low} = \frac{c}{\lambda_{low}} = \frac{c}{2\pi R_s}$$

$r_0 \rightarrow$ The heighest frequency of operation depends on "r_0".

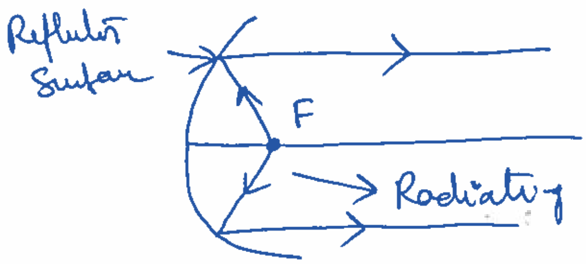
$$f_{upper} = \frac{c}{\lambda_{upper}} = \frac{c}{4 R_0} \longrightarrow 2M$$

11) b) Feed mechanism in parabolic reflector Antenna :- (6 Marks)

1) Focal feed system
2) Cassegrain feed system
3) Gregorian feed system
4) offset feed system.

1) Focal feed system :-

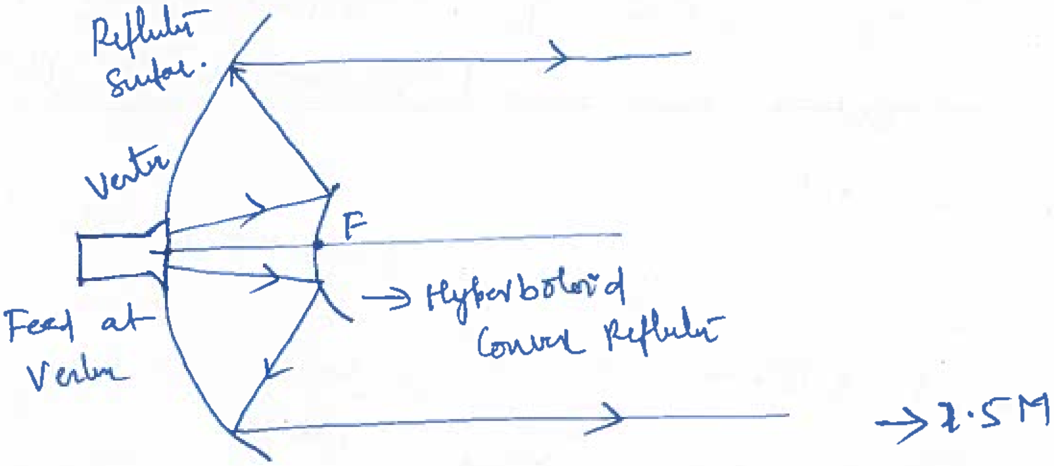
The primary antenna (radiating element) is placed at the focal point of parabolic Reflecting Surface



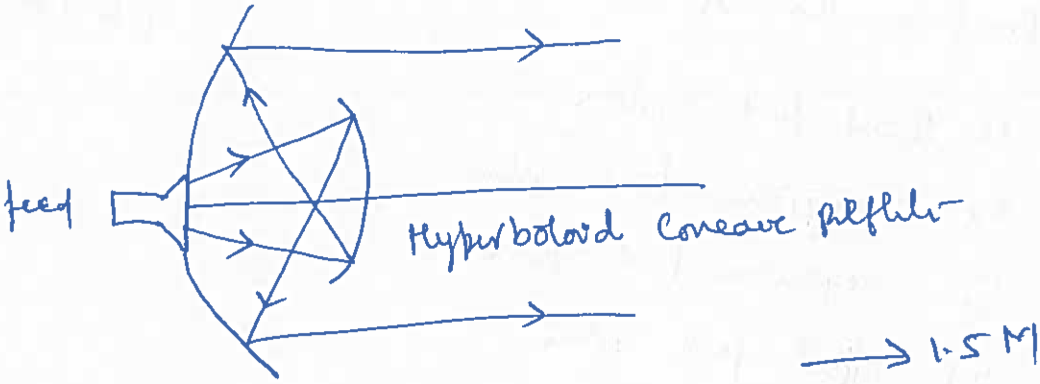
$\longrightarrow$ 1.5 M

2) Cassegrain feed system :-

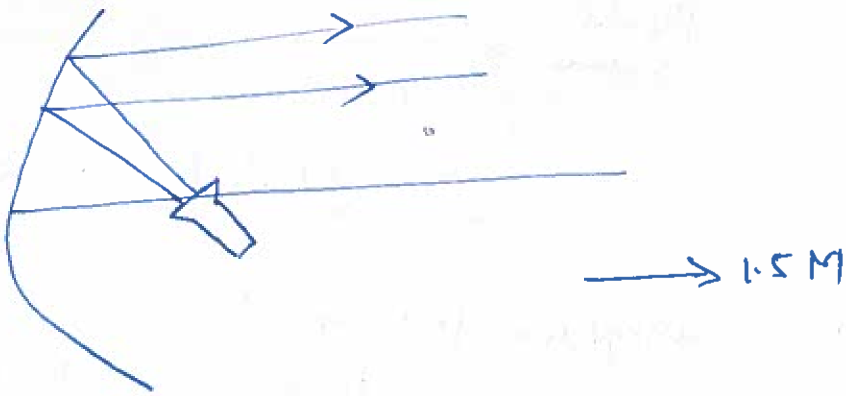
Radiation Source is placed at vertex instead of at focus (F). Required Second Reflecting Surface



Gregorian Feed :-
It is similar to cassegrain system, the major difference is the secondary reflector is concave or mostly ellipsoidal in shape.



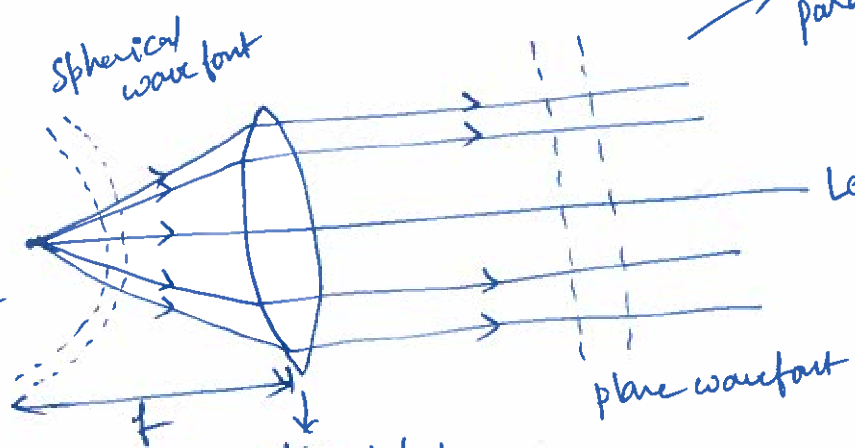
Offset Feed :-
The feed is placed at offset from the center of actual antenna.



12) a) Lens Antenna (6 Marks)

Lens antenna is also a microwave antenna which is used at a bit higher frequencies. At lower frequencies lens antennas become bulky and heavy. They are just as a glass lens as used in optics.

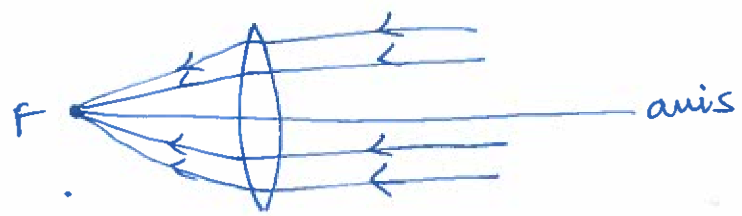
Principle :-



→ 3M

Assuming the source at focal point, at a distance of focal length, along the lens axis, it is seen that on the other side collimated (or) parallel rays are obtained due to refraction.

From optical point of view a divergent beam is collimated because refraction takes place as a result of which ray at centre are refracted less than at the edge



lens antenna as receiver

→ 3M

The lens antenna also obeys " The principle of equality of path-length".

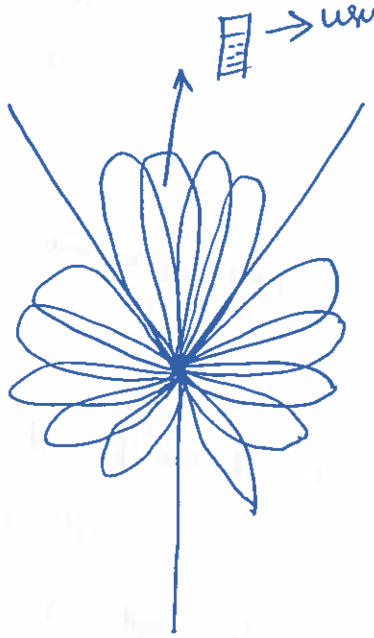
12) b) Smart Antennas :- (Total 6M)

These systems are classified as

A) switched Beam systems

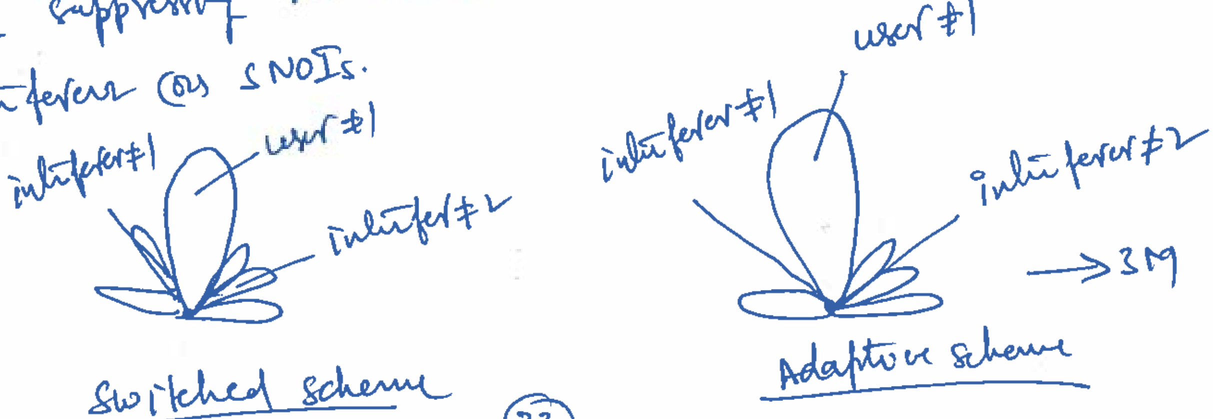
B) Adaptive array systems

A switched Beam system is a system, that can be chosen from one of many predefined patterns in order to enhance the received signal, and it is an extension of cell sectoring as each sector is sub-divided into smaller sectors.



→ 3M

Adaptive array systems provide more degree of freedom since they have the ability to adapt in real time, the radiation pattern to the RF signal environment. They can direct the main beam towards the SOI while suppressing the antenna pattern in the direction of interferers (or) SNOIs.



→ 3M

13) a) Measurement of gain :- (Total 6M)

Absolute Gain Measurement :-

These methods are based on Friis transmission formula. Antennas are separated by distance 'R' and it must satisfy the far field criteria of each Antenna.

Polarizations must be matched to maximum directional radiation. $\longrightarrow$ (2M)

Two Antenna Method :-

$$(G_{ot})_{dB} + (G_{or})_{dB} = 20 \log_{10}\left(\frac{4\pi R}{\lambda}\right) + 10 \log_{10}\left(\frac{P_r}{P_t}\right)$$

where

$(G_{ot})_{dB} \rightarrow$ Gain of transmitting antenna (dB)

$(G_{or})_{dB} \rightarrow$ Gain of receiving antenna (dB)

$P_r \rightarrow$ Power received, W

$P_t \rightarrow$ Power transmitted, W

$\lambda \longrightarrow$ operating wavelength, m

$R \longrightarrow$ Antenna separation, m

If the transmitting antenna & receiving antennas are identical $\Rightarrow G_{or} = G_{ot}$

$$(G_{or})_{dB} = (G_{ot})_{dB} = \frac{1}{2}\left[20 \log_{10}\left(\frac{4\pi R}{\lambda}\right) + 10 \log_{10}\left(\frac{P_r}{P_t}\right)\right]$$

By measuring R, λ and the ratio of P_r/P_t, the gain of antenna can be found. —— (4M)

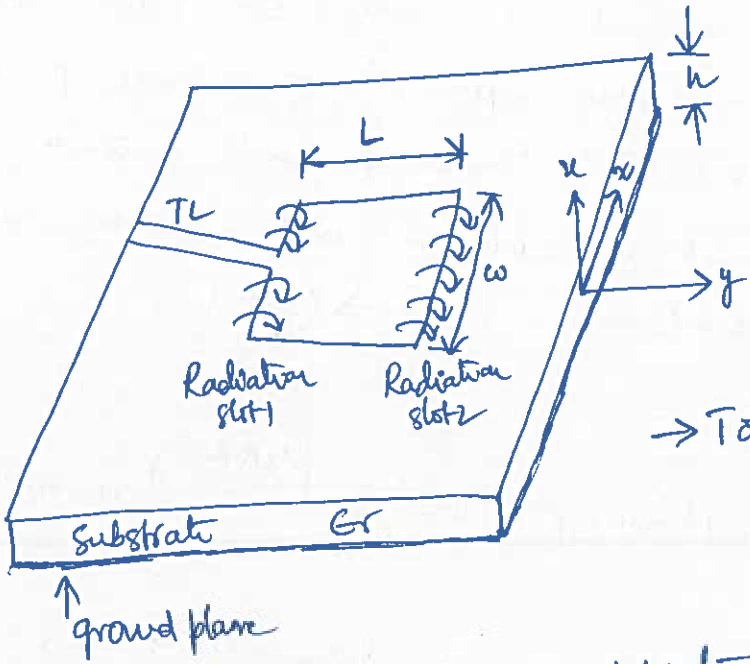
(or)

Three Antenna Method

(or)

Gain Comparison Method

13) b) Microstrip Antenna :- (6 Marks)



→ Top view of Patch Antenna

h → height of the substrate

L → length of the Patch

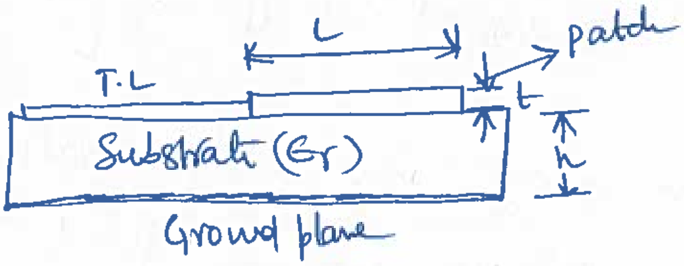
ω → width of the Patch

Єr → Dielectric const of substrate. → (2 M)

Typical Dimensions of Patch Antenna :-

1. $t << \lambda_0$ (λ_0 → free space wavelength)
2. $0.0003\lambda_0 \leq h \leq 0.05\lambda_0$.
3. $\lambda_0/3 \leq L \leq \lambda_0/2$
4. $2.2 \leq \epsilon_r \leq 12.0$

Side view :-



t → thickness of Patch

→ (2 M)

→ In high performance aircraft, spacecraft, satellite and missile applications, where size, weight, cost, performance, ease of installation and aerodynamic profile are constraints low-profile antennas like Patch antennas are required.

→ It consists of a very thin metal strip (Patch – copper) placed above a ground plane.

→ The microstrip patch is designed such that its pattern maximum is normal to the patch. (Broadside Radiator)

→ For rectangular patch of length 'L', the strip and ground plane are seperated by a dielectric sheet (substrate, ϵ_r).

→ The radiating elements and the feed lines are usually photoetched on the dielectric substrate.

→ The radiating patch may be square, rectangular, circular, triangular, elliptical etc. → (2 M)

14. Refractive index of Ionosphere layer :- (Total 12 M)

Let an Electric field of value 'E' is acting across a cubic meter of space in the ionosphere.

$$E = E_m \sin \omega t \text{ volts/m}$$

where $\omega \rightarrow$ Angular velocity

$E_M \rightarrow$ Maximum amplitude.

Force exerted by electric field on each electron

$$F = -eE \text{ Newton}$$

$e \rightarrow$ charge of electron in Coulombs.

Let us assume that there is no collision, then the electron will have a instantaneous velocity v m/sec in the direction opposite to the field. —— (4 M)

$$\text{Force} = \text{mass} \times \text{Acceleration}$$

$$-eE = m \frac{dv}{dt}$$

$m \rightarrow$ mass of e^-

$\frac{dv}{dt} \rightarrow$ Acceleration

$$\frac{dv}{dt} = -\frac{Ee}{m} \quad (or) \quad dv = -\frac{Ee}{m}dt$$

$$\int dv = -\int \frac{eE}{m}\, dt$$

$$v = \frac{-e}{m}\int E_m \sin\omega t\, dt$$

$$v = +\frac{e E_m \cos\omega t}{m\cdot\omega}$$

$$v = \left(\frac{e}{m\omega}\right) E_m \cos\omega t$$

if N be the number of electrons / cubic metre, then instantaneous electric current

$$i_e = -Nev \text{ amp}/m^2$$

$$i_e = -Ne\left(\frac{e}{m\omega}\right) E_m \cos\omega t$$

$$i_e = -\left(\frac{Ne^2}{m\omega}\right) E_m \cos\omega t$$

which shows i_e is lags behind 'E' by $90°$

Beside the inductive current, there is also capacitive current

capacitive current (or) displacement current

$$i_c = \frac{dD}{dt} = \frac{d}{dt}(K_0 E) = K_0 \frac{d}{dt} E_m \sin\omega t$$

$$K_0 = 8.854 \times 10^{-12} \text{ F/m}$$

Since $D = \epsilon_0 E = K_0 E$; K_0 = constant

$$i_c = K_0 E_m \cos\omega t\, \omega \xrightarrow{(1)} 4M$$

The total current i flow through a cubic metre of ionized medium is

$$i = i_c + i_e = K_0 E_m \omega\cos\omega t - \frac{Ne^2}{m\omega} E_m \cos\omega t$$

$$i = E_m \cos\omega t \, \omega \left[K_0 - \frac{Ne^2}{m\omega^2} \right] \quad \text{——(2)}$$

Comparing ① & ②

The effective dielectric const K of ionosphere given by

$$K = K_0 - \frac{Ne^2}{m\omega^2} = K_0 \left[1 - \frac{Ne^2}{m\omega^2 K_0} \right]$$

The relative dielectric const w.r.t vacuum (or) air

$$K_r = \frac{K}{K_0} = \left[1 - \frac{Ne^2}{m\omega^2 K_0} \right]$$

Refractive index (μ) of ionosphere

$$\mu = \sqrt{K_r}$$

$$\mu = \sqrt{1 - \frac{Ne^2}{m\omega^2 K_0}}$$

$m = 9.1 \times 10^{-31}$ kg , $e = 1.6 \times 10^{-18}$ coulombs

$K_0 = 8.854 \times 10^{-12}$ F/m & $\omega = 2\pi f$

$$\boxed{\mu = \sqrt{1 - \frac{81N}{f^2}}} \longrightarrow (4M)$$

$N \rightarrow$ Number of $\bar{e}$ / cubic meter (or) Ionic density

$f \rightarrow$ Frequency in Hz.

15) a) Ground wave propagation :- (up to 2 MHz) (6 Marks)

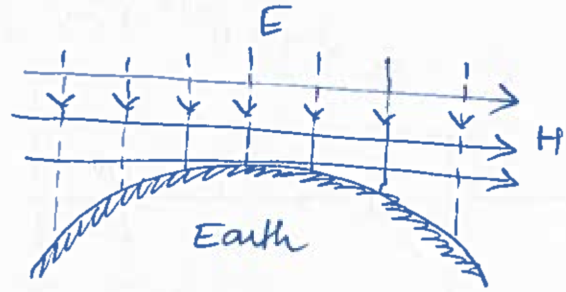
— The groundwaves is of practical importance at broadcast and lower frequencies i.e for mediumwaves, longwaves and very long waves.

— The groundwaves is a wave that is guided along the surface of the earth just as an EM waves are guided by a waveguide (or) transmission line.

— Surface wave permits the propagation around the curvature of the earth.

— This mode of propagation exist when the transmitting and receiving antennas are close to the surface of the earth and is supported at its lower edge by the presence of ground

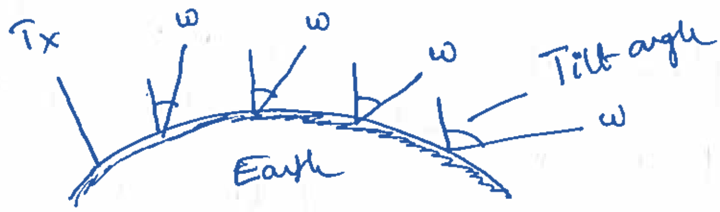
— The groundwaves are usually produced by vertical antennas, are vertically polarized i.e Electric field vectors of EM waves are vertical w.r.t to ground.



Any components (horizontal) of Electric field in contact with the earth is short circuited by the earth

→ when surface waves glides over the surface of the earth energy is absorbed from the surface wave to supply the losses in the earth. Thus while passing over the surface of the earth, the surface wave losses some of its energy by absorption.

- The groundwave suffers varying amount of attenuation while propagating along the curvature of the earth, depending on the frequency, surface irregularities, permittivity and conductivity.
- Earth's attenuation increases as the frequency increases and hence the mode of propagation is suitable for low and medium frequency up to 2MHz only.



→ 2M

- As the wave progress over the curvature of the earth, the wavefronts start gradually tilting more and more.
- This increase in the tilt angle causes more short circuit of the electric field component and hence the fieldstrength goes on reducing.
- Maximum range of surfacewave propagation depends not only on the frequency but power as well.
- The Field strength at a distance 'd' from the transmitting antenna due to groundwave has been calculated from the Maxwell Eqns as

$$\boxed{E = \frac{120\pi \; h_t \cdot h_r \; I_s}{\lambda \cdot d}} \quad \text{volt/mtr}$$

→ 2M

$h_t, h_r \rightarrow$ Effective heights of transmitting and receiving antennas

$I_s \rightarrow$ Antenna current

$\lambda \rightarrow$ wavelength

$d \rightarrow$ Distance b/w transmitting and receiving Antennas

If the distance 'd' is very large, then the reduction in Field strength due to ground attenuation and atmosphere absorption increases thus the actual voltage received at the receiving point decreases. So it is suitable for the short distance communication only.

15) b) (6 Marks)

(i) Critical frequency :- (fc)

The critical frequency of an ionized layer of the ionosphere is defined as the highest frequency which can be reflected by a particular layer at vertical incidence

Critical frequency for a particular layer is proportional to the square root of the maximum electron density in the layer.

$$\mu = \frac{\sin i}{\sin r} = \sqrt{1 - \frac{81 N}{f^2}}$$

By definition of vertical incidence

$$\angle i = 0 ; \quad N = N_{max} \text{ and } f = f_c$$

$$\Rightarrow \sqrt{1 - \frac{81 N_m}{f_c^2}} = 0$$

$$1 = \frac{81 N_m}{f_c^2} \quad \text{(or)} \quad f_c = \sqrt{81 \cdot N_m}$$

$$\boxed{f_c = 9\sqrt{N_m}} \longrightarrow 2M$$

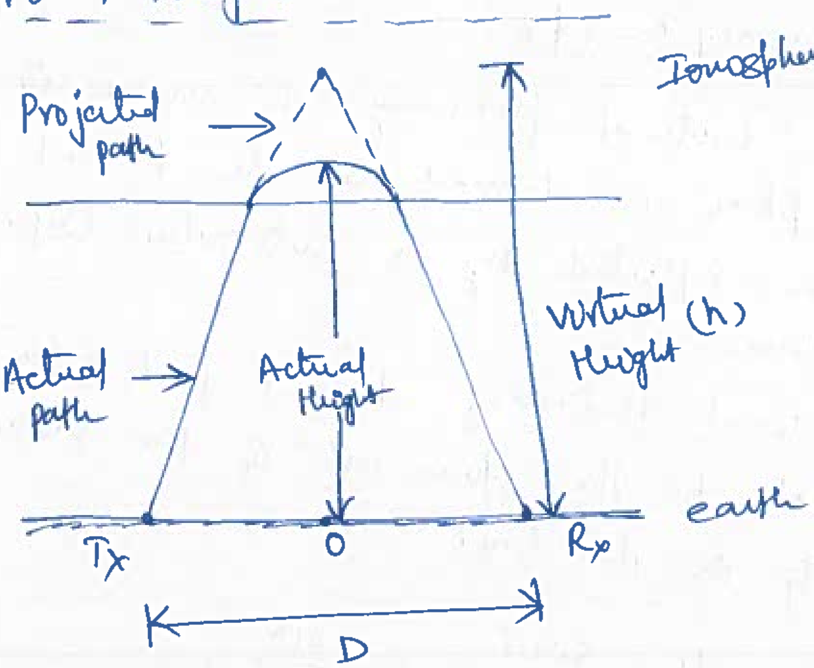
ii) Skip Distance :- (D)

The minimum distance from the transmitter at which a skywave of given frequency is returned to earth by the ionosphere.

$$\frac{f_{muf}}{f_c} = \sqrt{1 + \left(\frac{D}{2h}\right)^2} \quad \text{(or)} \quad \frac{f_{muf}^2}{f_c^2} - 1 = \left(\frac{D}{2h}\right)^2$$

$$\Rightarrow \boxed{D_{skip} = 2h\sqrt{\left(\frac{f_{muf}}{f_c}\right)^2 - 1}} \longrightarrow 2M$$

iii) Virtual Height :-



Virtual height of an ionosphere layer is defined as the height to which a short pulse of energy sent vertically upward and travelling with the speed of light would reach taking the same two ways travel time as does the actual pulse reflected from the layer.

For flat earth condition

$$TR = \frac{2h}{\tan\beta} = D$$

$\longrightarrow 2M.$

B

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE			Academic Year	2022 – 2023
Course Code	20CS502	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Computer Networks						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the layers in TCP/IP Model?	20CS502.1	L1
2	What is a Tropology?	20CS502.2	L1
3	What is Transmission Media?	20CS502.3	L1
4	What is Error Control?	20CS502.4	L1
5	What is a WWW?	20CS502.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain OSI reference model	12M	20CS502.1	L2
	OR			
7	Explain about different types of transmission media	12M	20CS502.1	L2
8	Explain Cyclic Redundancy Check (CRC) with an example	12M	20CS502.2	L2
	OR			
9	What is flow control? Explain different types of flow control methods	12M	20CS502.2	L2
10	What is Routing? Explain different types of routing	12M	20CS502.3	L2
	OR			
11	Explain Leakey bucket and Token bucket	12M	20CS502.3	L2
12 (a)	Define network	3M	20CS502.4	L2
12 (b)	Explain TCP header format	9M	20CS502.4	L2
	OR			
13 (a)	What is IP?	2M	20CS502.4	L1
13 (b)	Explain IPV4 header format	10M	20CS502.4	L2
14	Explain below protocols a) FTP b) DNS c) SMTP d) HTTP	12M	20CS502.5	L2
	OR			
15	What is Email? Explain any two scenarios of Email	12M	20CS502.5	L2

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SCHEME OF VALUATION
&
ANSWER KEY

Degree	B. Tech. (U. G.)	Program	CSE	Test	END EXAM	Academic Year	2022 - 2023
Course Code	20CS502	Test Duration	180 Min.	Max. Marks	70	Semester	I
Course	COMPUTER NETWORKS						
Assessment Pattern							

R (L1):		U (L2):		Apply (L3):		Analyze (L4):		E (L5):		C (L6)	

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	**What are the layers in TCP/IP Model?**	20CS502.1	L1
	For this we can give two marks for 4 layers of TCP/IP TCP/IP reference model has only 4 layers. They are, 1. Host-to-Network Layer 2. Internet Layer 3. Transport Layer 4. Application Layer		
2	**What is a Topology?**	20CS502.2	L1
	Topology defines the structure of the network of how all the components are interconnected to each other. There are two types of topology:		

	1. Physical topology 2. logical topology		
3	**What is Transmission Media?**	20CS502.3	L1
	Transmission media is a communication channel that transmits information from the source/transmitter to the receiver. There are 2 types of transmission media.1.guided transmission media 2.unguided transmission media		
4	**What is Error Control?**	20CS502.4	L1
	Error control is basically process in data link layer of detecting or identifying and re-transmitting data frames that might be lost or corrupted during transmission.		
5	**What is a WWW?**	20CS502.5	L1
	The World Wide Web -- also known as the web, WWW or W3 -- refers to all the public websites or pages that users can access on their local computers and other devices through the internet. These pages and documents are interconnected by means of hyperlinks that users click on for information.		

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Learning Outcome (s)	DoK
6 (a)	**Explain OSI reference model** For daigram -3M For the explanation of all layers – 9 M	20CS502.1	L2

OSI reference model

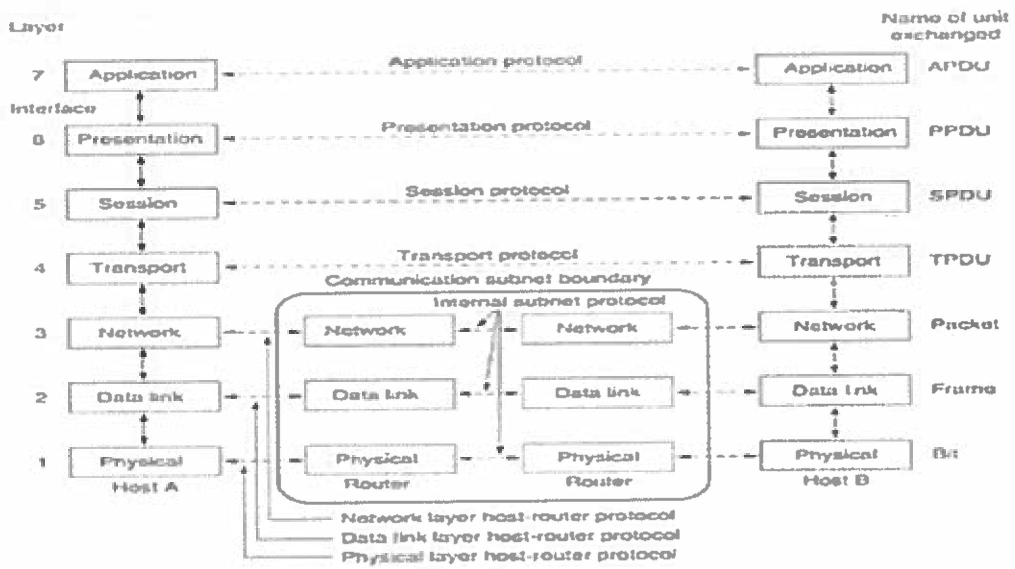


Fig.4: The OSI reference model

Physical layer:

- It is the lowest layer of the OSI model.
- It establishes, maintains and deactivates the physical connection.
- It specifies the mechanical, electrical and procedural network interface specifications.

Data link layer:

- This layer is responsible for the error-free transfer of data frames.
- It defines the format of the data on the network.
- It provides a reliable and efficient communication between two or more devices.
- It is mainly responsible for the unique identification of each device that resides on a local network.

Network layer:

- It is a layer 3 that manages device addressing, tracks the location of devices on the network.
- It determines the best path to move data from source to the destination based on the network conditions, the priority of service, and other factors.
- The Data link layer is responsible for routing and forwarding the packets.
- Routers are the layer 3 devices, they are specified in this layer and used to provide the routing services within an inter network.
- The protocols used to route the network traffic are known as Network layer protocols. Examples of protocols are IP and Ipv6.

Transport layer:

- The Transport layer is a Layer 4 ensures that messages are transmitted in the order in which they are sent and there is no duplication of data.
- The main responsibility of the transport layer is to transfer the data completely.
- It receives the data from the upper layer and converts them into smaller units known as segments.
- This layer can be termed as an end-to-end layer as it provides a point-to-point connection between source and destination to deliver the data reliably.

Session layer:

- It is a layer 3 in the OSI model.
- The Session layer is used to establish, maintain and synchronizes the interaction between communicating devices.

Presentation layer:

- A Presentation layer is mainly concerned with the syntax and semantics of the information exchanged between the two systems.
- It acts as a data translator for a network.
- This layer is a part of the operating system that converts the data from one presentation format to another format.
- The Presentation layer is also known as the syntax layer

Application layer

- An application layer serves as a window for users and application processes to access network service.
- It handles issues such as network transparency, resource allocation, etc.
- An application layer is not an application, but it performs the application layer functions.
- This layer provides the network services to the end-users.

OR

7	**Explain about different types of transmission media – 12Marks** **For writing guided media transmission three types with diagrams - 9Marks** **For writing unguided media transmission and types -3Marks**	**20CS502.1**	**L2**

Guided Transmission Media:

Guided Media. It is defined as the physical **medium** through which the signals are transmitted. It is also known as Bounded **media.**

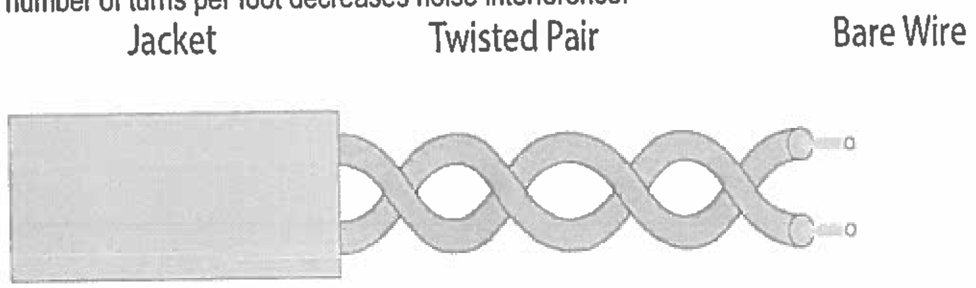
Types Of Guided media:

Twisted pair:

Twisted pair is a physical media made up of a pair of cables twisted with each other. A twisted pair cable is cheap as compared to other transmission media. Installation of the twisted pair cable is easy, and it is a lightweight cable. The frequency range for twisted pair cable is from 0 to 3.5KHz.

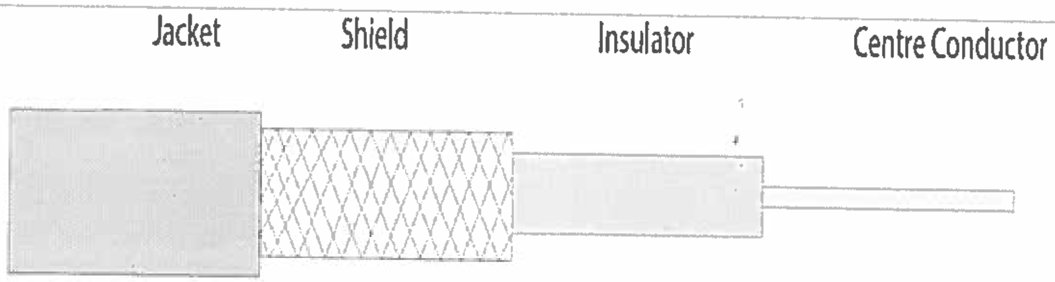
A twisted pair consists of two insulated copper wires arranged in a regular spiral pattern.

The degree of reduction in noise interference is determined by the number of turns per foot. Increasing the number of turns per foot decreases noise interference.



Coaxial Cable

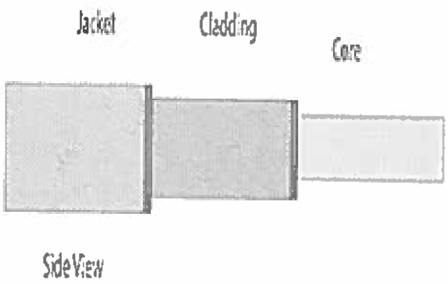
- Coaxial cable is very commonly used transmission media, for example, TV wire is usually a coaxial cable.
- The name of the cable is coaxial as it contains two conductors parallel to each other.
- It has a higher frequency as compared to Twisted pair cable.
- The inner conductor of the coaxial cable is made up of copper, and the outer conductor is made up of copper mesh. The middle core is made up of non-conductive cover that separates the inner conductor from the outer conductor.
- The middle core is responsible for the data transferring whereas the copper mesh prevents from the EMI(Electromagnetic interference).

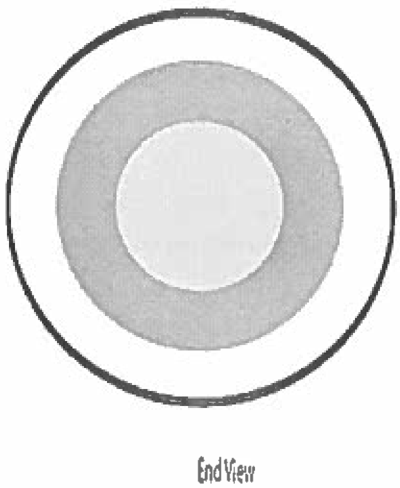


Fibre Optic:

- Fibre optic cable is a cable that uses electrical signals for communication.
- Fibre optic is a cable that holds the optical fibers coated in plastic that are used to send the data by pulses of light.
- The plastic coating protects the optical fibers from heat, cold, electromagnetic interference from other types of wiring.
- Fibre optics provide faster data transmission than copper wires.

Diagrammatic representation of fibre optic cable:





Unguided media:

- An unguided transmission transmits the electromagnetic waves without using any physical medium. Therefore it is also known as **wireless transmission**.
- In unguided media, air is the media through which the electromagnetic energy can flow easily.

Unguided transmission is broadly classified into three categories:

1.radio waves
2.Microwaves

3.Satellite Microwave Communication

RADIO WAVES :

- Radio waves are the electromagnetic waves that are transmitted in all the directions of free space.
- Radio waves are omnidirectional, i.e., the signals are propagated in all the directions.
- The range in frequencies of radio waves is from 3Khz to 1 khz.

Micro waves:

Microwave is a line-of-sight wireless communication technology that uses high frequency beams of radio waves to provide high speed wireless connections that can send and receive voice, video, and data information.

Microwaves are of two types:

1. Terrestrial microwave
2. Satellite microwave communication.

Satellite microwave communication:

It is used for broadcasting and receiving signals. The signals are transmitted to space where these satellites are positioned and it re transmits the signal to the appropriate location.

It acts as a repeater as it only receives the signal and re transmits it. The satellites should be aligned properly with the earth for this system to work. It is a physical object which revolves around the earth at a known height.

Satellite communication is more flexible and reliable nowadays than cable and fibre optic systems.

8	Explain Cyclic Redundancy Check (CRC) with an example For this answer about CRC -4Marks For example 8 marks	20CS502.2	L2

Cyclic Redundancy Check (CRC)

The Cyclic Redundancy Checks (CRC) is the most powerful method for Error-Detection and Correction.

It is given as a k bit message and the transmitter creates an (n – k) bit sequence called frame check sequence. The out coming frame, including n bits, is precisely divisible by some fixed number. Modulo 2 Arithmetic is used in this binary addition with no carries, just like the XOR operation.

Redundancy means duplicate. The redundancy bits used by CRC are changed by splitting the data unit by a fixed divisor. The remainder is CRC.

Qualities of CRC

- It should have accurately one less bit than the divisor.
- Joining it to the end of the data unit should create the resulting bit sequence precisely divisible by the divisor.

CRC generator and checker

Process

- A string of n 0s is added to the data unit. The number n is one smaller than the number of bits in the fixed divisor.
- The new data unit is divided by a divisor utilizing a procedure known as binary division; the remainder appearing from the division is CRC.
- The CRC of n bits interpreted in phase 2 restores the added 0s at the end of the data unit.

Example

Message D = 1010001101 (10 bits)

Predetermined P = 110101 (6 bits)

FCS R = to be calculated 5 bits

Hence, n = 15 K = 10 and (n – k) = 5

The message is generated through 25:accommodating 1010001101000

The product is divided by P

```
              1101010110 <--- Q
        ___________________
110101 ) 101000110100000
          110101
         -------
           111011
           110101
          -------
             111010
             110101
            -------
              111110
              110101
             -------
                101100
                110101
               -------
                  110010
                  110101
              ----------
                   01110 <--- R
```

The remainder is inserted to 25D to provide T = 101000110101110 that is sent.

Suppose that there are no errors, and the receiver gets T perfect. The received frame is divided by P.

```
              1101010110
110101 ) 101000110101110
         110101
         -------
          1110111
          1101101
         -------
           111010
           110101
         -------
            111110
            110101
         -------
              101100
              110101
         ---------
                110101
                110101
         ---------
                     0  <--- R
```

Because of no remainder, there are no errors

	OR		
9	What is flow control? Explain different types of flow control methods ----- 12 Marks For writing about flow control -2 marks For writing 2 types of flow control methods with one example-10 marks	20CS502.2	L2

Flow control:

It is a set of procedures that tells the sender how much data can be transmitted before it overwhelms the receiver's system. A normal user's device usually has limited speed and memory and it can't handle high volumes of data. Therefore, the receiving device must be able to inform the sending device to stop the transmission temporarily before the limits are reached. It requires a buffer- a block of memory for storing the information until they are processed

Whereas pool and croquet are modeled correctly as turn-taking games, tennis is not. While one player is moving to the ball, the other player is moving to anticipate the opponent's return. In general, it may be reasonable to derive randomized strategies in tennis so that the opponent cannot anticipate where the ball will go.

Types of flow control:

There are two types of flow control protocols

1.stop and wait protocol

2.Sliding window protocol

Stop and wait protocol:

It is the simplest flow control method. In this, the sender will transmit one frame at a time to the receiver. The sender will **stop and wait** for the acknowledgement from the receiver.

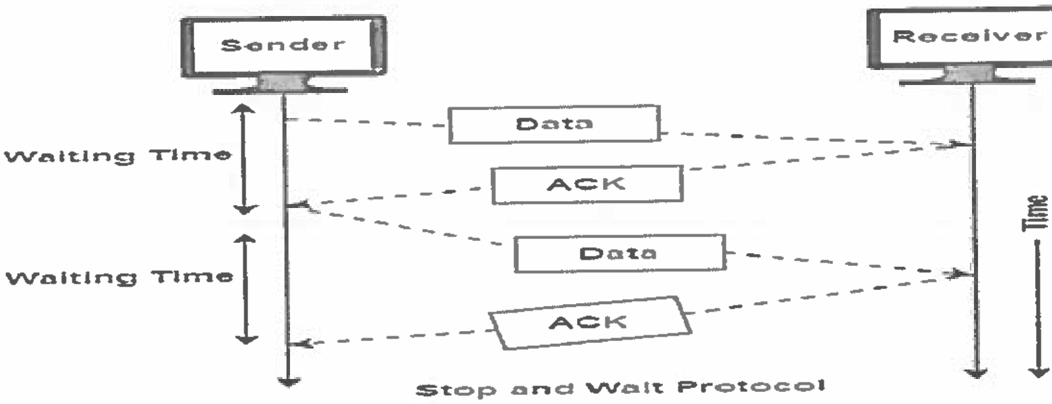
This time (i.e. the time joining message transmitting and acknowledgement receiving) is the sender's waiting time, and the sender is idle during this time.

When the sender gets the acknowledgement (ACK), it will send the next data packet to the receiver and wait for the disclosure again, and this process will continue as long as the sender has the data to send.

While sending the data from the sender to the receiver, the data flow needs to be controlled. If the sender is transmitting the data at a rate higher than the receiver can receive and process it, the data will get lost.

The Flow-control methods will help in ensuring that the data doesn't get lost. The flow control method will check that the senders send the data only at a rate that the receiver can receive and process.

The working of Stop and Wait Protocol is shown in the figure below –



Stop and Wait Protocol

The main advantage of stop & wait protocols is their accuracy. The next frame is transmitted only when the first frame is acknowledged. So there is no chance of the frame being lost.

The drawback of this approach is that it is inefficient. It makes the transmission process slow. An individual frame travels from source to destination in this method, and a single acknowledgement travels from destination to source.

The features of Stop and Wait Protocol are as follows

It is used in Connection-oriented communication.
It offers error and flows control.
It can be used in data Link and transport Layers.
Stop and Wait ARQ executes Sliding Window Protocol with Window Size 1.

Sliding Window Protocol

The sliding window is a technique for sending multiple frames at a time. It controls the data packets between the two devices where reliable and gradual delivery of data frames is needed. It is also used in TCP (Transmission Control Protocol).

In this technique, each frame has sent from the sequence number. The sequence numbers are used to find the missing data in the receiver end. The purpose of the sliding window technique is to avoid duplicate data, so it uses the sequence number.

Types of Sliding Window Protocol

Sliding window protocol has two types:

1. Go-Back-N ARQ

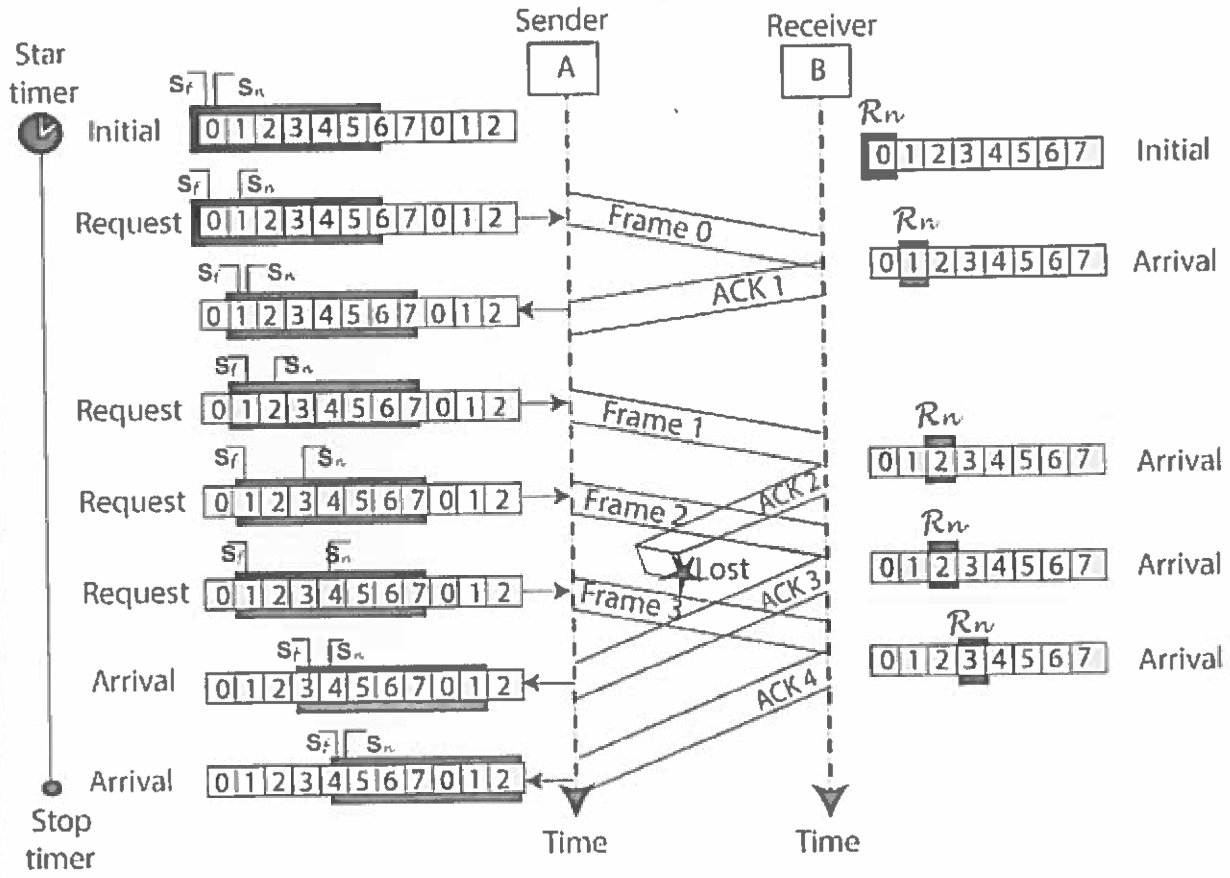
2. Selective Repeat ARQ

Go-Back-N ARQ

Go-Back-N ARQ protocol is also known as Go-Back-N Automatic Repeat Request. It is a data link layer protocol that uses a sliding window method. In this, if any frame is corrupted or lost, all subsequent frames have to be sent again.

The size of the sender window is N in this protocol. For example, Go-Back-8, the size of the sender window, will be 8. The receiver window size is always 1.Page 23

If the receiver receives a corrupted frame, it cancels it. The receiver does not accept a corrupted frame. When the timer expires, the sender sends the correct frame again. The design of the Go-Back-N ARQ protocol is shown below



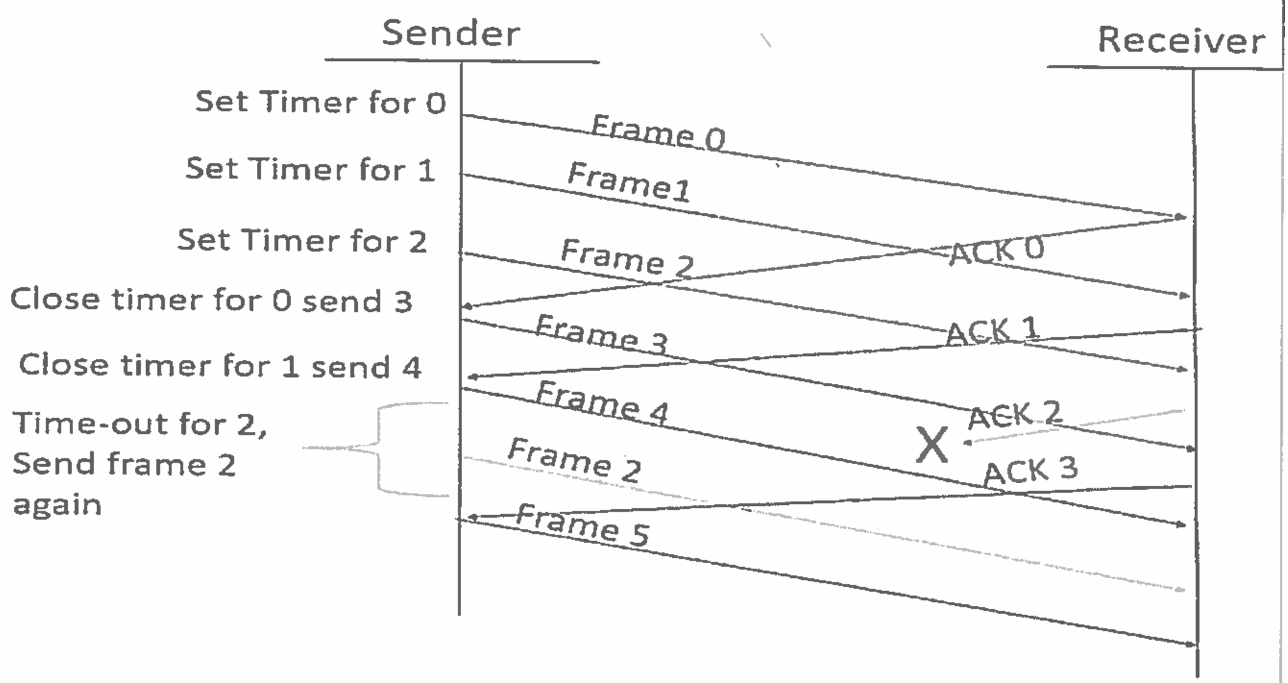
Selective Repeat ARQ

It is also known as Sliding Window Protocol and used for error detection and control in the data link layer.

In the selective repeat, the sender sends several frames specified by a window size even without the need to wait for individual acknowledgement from the receiver as in Go-Back-N ARQ. In selective repeat protocol, the re transmitted frame is received out of sequence.

In Selective Repeat ARQ only the lost or error frames are re transmitted, whereas correct frames are received and buffered.

The receiver while keeping track of sequence numbers buffers the frames in memory and sends NACK for only frames which are missing or damaged. The sender will send/re transmit a packet for which NACK is received.

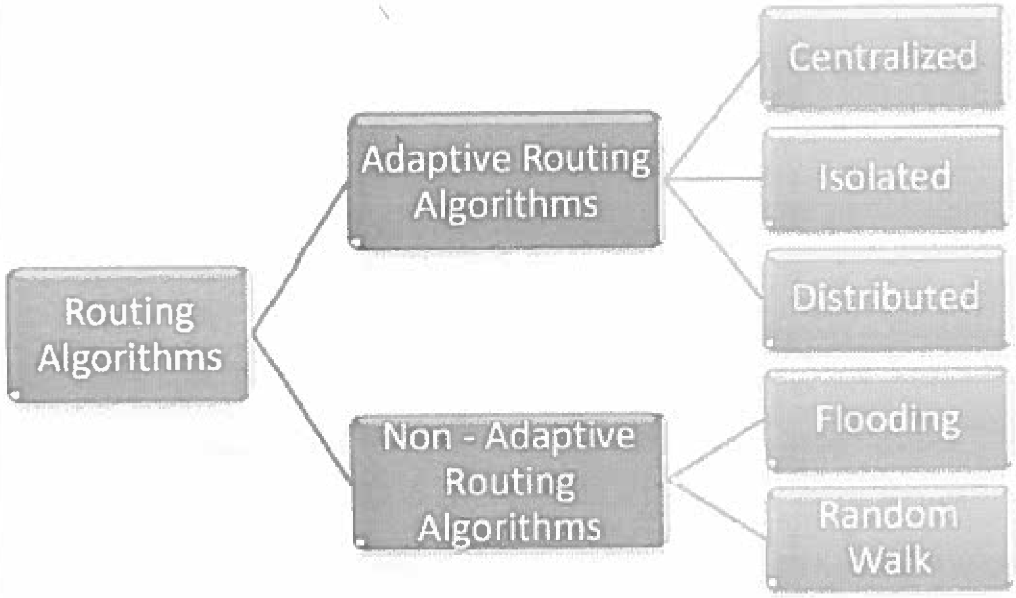


10	**What is Routing? Explain different types of routing -12 Marks** **For writing about routing-2Marks** **For writing about types of routing -10 marks**	**20CS502.3**	**L2**

Routing: Routing is the process of path selection in any network. A computer network is made of many machines, called nodes, and paths or links that connect those nodes. Communication between two nodes in an interconnected network can take place through many different paths. When a packet reaches the router's input link, the router will move the packets to the router's output link. For example, a packet from S1 to R1 must be forwarded to the next router on the path to S2.

TYPES OF ROUTING ALGORITHMS

Routing algorithms can be broadly categorized into two types, adaptive and non adaptive routing algorithms. They can be further categorized as shown in the following diagram –



ADAPTIVE ROUTING ALGORITHMS

Adaptive routing algorithms, also known as dynamic routing algorithms, makes routing decisions dynamically depending on the network conditions. It constructs the routing table depending upon the network traffic and topology. They try to compute the optimized route depending upon the hop count, transit time and distance.

The three popular types of adaptive routing algorithms are –

- **Centralized algorithm** – It finds the least-cost path between source and destination nodes by using global knowledge about the network. So, it is also known as global routing algorithm.
- **Isolated algorithm** – This algorithm procures the routing information by using local information instead of gathering information from other nodes.
- **Distributed algorithm** – This is a decentralized algorithm that computes the least-cost path between source and destination iteratively in a distributed manner.

Adaptive routing have 2 algorithms:

1. link state routing algorithm
2. Distance vector routing algorithm

link state routing algorithm

Link state routing is a method in which each router shares its neighborhood's knowledge with every other router in the inter network. In this algorithm, each router in the network understands the network topology then makes a routing table depend on this topology

Distance vector routing algorithm:

Distributed: It is distributed in that each node receives information from one or more of its directly attached neighbors, performs calculation and then distributes the result back to its neighbors.

Iterative: It is iterative in that its process continues until no more information is available to be exchanged between neighbors.

Asynchronous: It does not require that all of its nodes operate in the lock step with each other.

- The Distance vector algorithm is a dynamic algorithm.
- It is mainly used in ARPANET, and RIP.
- Each router maintains a distance table known as **Vector.**

NON – ADAPTIVE ROUTING ALGORITHMS

Non-adaptive Routing algorithms, also known as static routing algorithms, construct a static routing table to determine the path through which packets are to be sent. The static routing table is constructed based upon the routing information stored in the routers when the network is booted up.

The two types of non – adaptive routing algorithms are –

- **Flooding** – In flooding, when a data packet arrives at a router, it is sent to all the outgoing links except the one it has arrived on. Flooding may be uncontrolled, controlled or selective flooding.
- **Random walks** – This is a probabilistic algorithm where a data packet is sent by the router to any one of its neighbors randomly.

Non adaptive have 3 algorithms

1. **shortest path routing algorithm**
2. **Flooding algorithm**
3. **Flow based algorithm**

shortest path routing algorithm

The goal of shortest path routing is to find a path between two nodes that has the lowest total cost, where the total cost of a path is the sum of arc costs in that path.In shortest path routing, the topology communication network is defined using a directed weighted graph. The nodes in the graph define switching components and the directed arcs in the graph define communication connection between switching components. Each arc has a weight that defines the cost of sharing a packet between two nodes in a specific direction

Flooding algorithm

Flooding is a non-adaptive routing technique following this simple method: when a data packet arrives at a router, it is sent to all the outgoing links except the one it has arrived on.

Flow based algorithm

Flow routing uses *adaptive routing algorithms that base their decisions on the traffic conditions between a computer and all the other computers it is connected to on the sub network.* Based on the traffic within this sub network, flow routing makes the decision about which computer to send a packet to.

OR

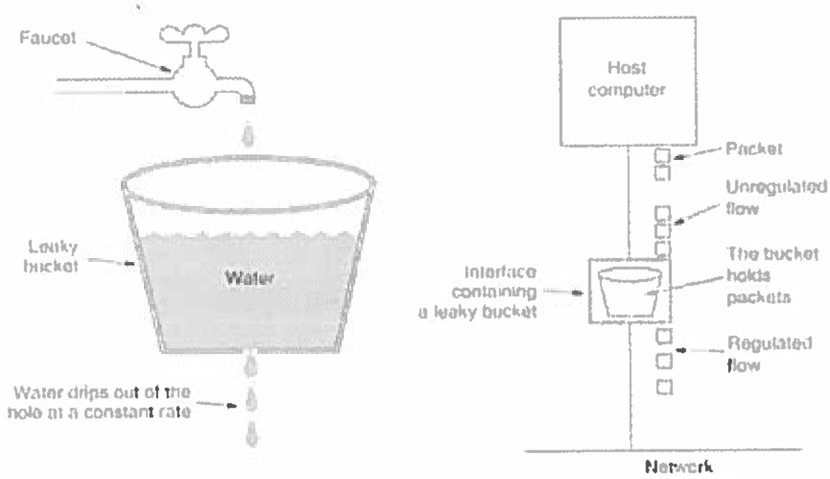
11	**Explain Leakey bucket and Token bucket-12 Marks** **For writing leaky bucket with diagram -6 Marks** **For writing token bucket with diagram-6Marks**	20CS502.3	L2

LEAKY BUCKET

The leaky bucket algorithm discovers its use in the context of network traffic shaping or rate-limiting. The algorithm allows controlling the rate at which a record is injected into a network and managing burstiness in the data rate.

A leaky bucket execution and a token bucket execution are predominantly used for traffic shaping algorithms. This algorithm is used to control the rate at which traffic is sent to the network and shape the burst traffic to a steady traffic stream.

The figure shows the leaky bucket algorithm.



(a) (b)

(a) A leaky bucket with water. (b) a leaky bucket with packets.

In this algorithm, a bucket with a volume of, say, b bytes and a hole in the Notes bottom is considered. If the bucket is null, it means b bytes are available as storage. A packet with a size smaller than b bytes arrives at the bucket and will forward it. If the packet's size increases by more than b bytes, it will either be discarded or queued. It is also considered that the bucket leaks through the hole in its bottom at a constant rate of r bytes per second.

The outflow is considered constant when there is any packet in the bucket and zero when it is empty. This defines that if data flows into the bucket faster than data flows out through the hole, the bucket overflows.

The disadvantages compared with the leaky-bucket algorithm are the inefficient use of available network resources. The leak rate is a fixed parameter. In the case of the traffic, volume is deficient, the large area of network resources such as bandwidth is not being used effectively. The leaky-bucket algorithm does not allow individual flows to burst up to port speed to effectively consume network resources when there would not be resource contention in the network.

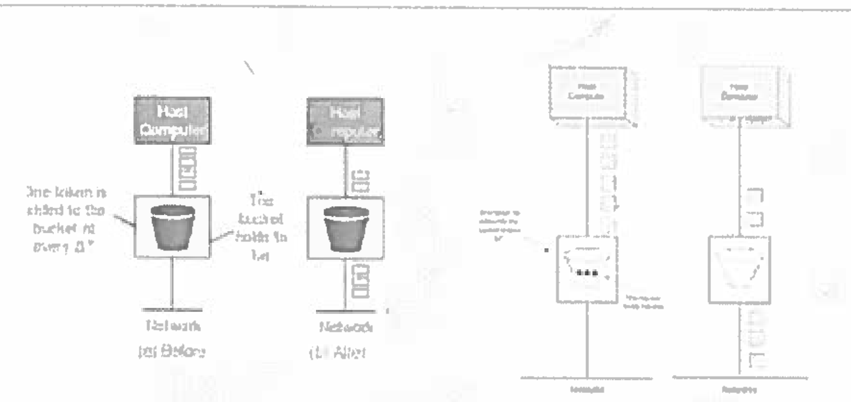
TOKEN BUCKET ALGORITHM

The leaky bucket algorithm has a rigid output design at the average rate independent of the bursty traffic. In some applications, when large bursts arrive, the output is allowed to speed up. This calls for a more flexible algorithm, preferably one that never loses information. Therefore, a token bucket algorithm finds its uses in network traffic shaping or rate-limiting.

It is a control algorithm that indicates when traffic should be sent. This order comes based on the display of tokens in the bucket. The bucket contains tokens. Each of the tokens defines a packet of predetermined size. Tokens in the bucket are deleted for the ability to share a packet.

When tokens are shown, a flow to transmit traffic appears in the display of tokens. No token means no flow sends its packets. Hence, a flow transfers traffic up to its peak burst rate in good tokens in the bucket.

Thus, the token bucket algorithm adds a token to the bucket each 1 / r seconds. The volume of the bucket is b tokens. When a token appears, and the bucket is complete, the token is discarded. If a packet of n bytes appears and n tokens



network. **Leaky Bucket** **Token Bucket**

When a packet of n bytes appears but fewer than n tokens are available. No tokens are removed from the bucket in such a case, and the packet is considered non-conformant. The non-conformant packets can either be dropped or queued for subsequent transmission when sufficient tokens have accumulated in the bucket.

They can also be transmitted but marked as being non-conformant. The possibility is that they may be dropped subsequently if the network is overloaded.

12 a)	**Define network -3marks**	**20CS502.4**	**L2**

Network:

A network is defined as the connection of atleast two computer systems either by a cable or a wireless connection. The simplest network is a combination of two computers connected by a cable. This type of network is called a **peer-to-peer network**. There is no hierarchy in this network; both participants have equal privileges. Each computer has access to the data of the other device and can share resources such as disk space, applications or peripheral devices (printers, etc.).

Today's networks tend to be a bit more complex and don't just consist of two computers. Systems with more than ten participants usually use client server network. In these networks, a central computer (server) provides resources to the other participants in the network (clients).

12(b)	**Explain TCP header format-9Marks** **For writing Tcp header format theory -7 Marks** **For diagram -2Marks**	**20CS502.4**	**L2**

TCP

- TCP stands for Transmission Control Protocol.
- It provides full transport layer services to applications.
- It is a connection-oriented protocol means the connection established between both the ends of the transmission. For creating the connection, TCP generates a virtual circuit between sender and receiver for the duration of a transmission.

TCP SEGMENT FORMAT

<table>
<tr><td colspan="8">Source port address 16 bits</td><td>Destination port address 16 bits</td></tr>
<tr><td colspan="9">Sequence number 32 bits</td></tr>
<tr><td colspan="9">Acknowledgement number 32 bits</td></tr>
<tr><td>HLEN 4 bits</td><td>Reserved 6 bits</td><td>URG</td><td>ACK</td><td>PSH</td><td>RST</td><td>SYN</td><td>FIN</td><td>Window size 16 bits</td></tr>
<tr><td colspan="8">Checksum 16 bits</td><td>Urgent pointer 16 bits</td></tr>
<tr><td colspan="9">Options & padding</td></tr>
</table>

- **Source port address:** It is used to define the address of the application program in a source computer. It is a 16-bit field.
- **Destination port address:** It is used to define the address of the application program in a destination computer. It is a 16-bit field.
- **Sequence number:** A stream of data is divided into two or more TCP segments. The 32-bit sequence number field represents the position of the data in an original data stream.
- **Acknowledgement number:** A 32-field acknowledgement number acknowledge the data from other communicating devices. If ACK field is set to 1, then it specifies the sequence number that the receiver is expecting to receive.
- **Header Length (HLEN):** It specifies the size of the TCP header in 32-bit words. The minimum size of the header is 5 words, and the maximum size of the header is 15 words. Therefore, the maximum size of the TCP header is 60 bytes, and the minimum size of the TCP header is 20 bytes.
- **Reserved:** It is a six-bit field which is reserved for future use.
- **Control bits:** Each bit of a control field functions individually and independently. A control bit defines the use of a segment or serves as a validity check for other fields.

There are total six types of flags in control field:

- **URG:** The URG field indicates that the data in a segment is urgent.
- **ACK:** When ACK field is set, then it validates the acknowledgement number.
- **PSH:** The PSH field is used to inform the sender that higher throughput is needed so if possible, data must be pushed with higher throughput.
- **RST:** The reset bit is used to reset the TCP connection when there is any confusion occurs in the sequence numbers.
- **SYN:** The SYN field is used to synchronize the sequence numbers in three types of segments: connection request, connection confirmation (with the ACK bit set), and confirmation acknowledgement.
- **FIN:** The FIN field is used to inform the receiving TCP module that the sender has finished sending data. It is used in connection termination in three types of segments: termination request, termination confirmation, and acknowledgement of termination confirmation.
    - **Window Size:** The window is a 16-bit field that defines the size of the window.
    - **Checksum:** The checksum is a 16-bit field used in error detection.

- **Urgent pointer:** If URG flag is set to 1, then this 16-bit field is an offset from the sequence number indicating that it is a last urgent data byte.
- **Options and padding:** It defines the optional fields that convey the additional information to the receiver.

OR

13 a) **What is IP -2Marks** **20CS502.4** **L1**

IP:

The Internet Protocol (IP) is protocol, or set of rules, for routing and addressing packets of data so that they can travel across networks and arrive at the correct destination. Data traversing the Internet is divided into smaller pieces, called packets. IP information is attached to each packet, and this information helps routers to send packets to the right place. Every device or domain that connects to the Internet is assigned an IP address, and as packets are directed to the IP address attached to them, data arrives where it is needed.

13(b) **Explain IPV4 header format -10Marks** **20CS502.4** **L2**
For writing about ipv4 header format theory -7Marks
For diagram -3Marks

IPV4

IPv4 short for Internet Protocol Version 4 is the fourth version of the internet protocol (IP).

IP is responsible to deliver data packets from the source host to the destination host.

This delivery is solely based on the IP address in the packet headers.

IPv4 is the first major version of IP.

IPv4 is a connectionless protocol for use of packet switched network

IPv4 Header

Version:

- Version is a 4 bit field that indicates the IP version used.
- The most popularly used IP versions are version-4 (IPv4) and version-6 (IPv6).
- Only IPv4 uses the above header.
- So, this field always contains the decimal value 4.

Header length:

- Header length is a 4 bit field that contains the length of the IP header.
- It helps in knowing from where the actual data begins.

Types of service:

- Type of service is a 8 bit field that is used for Quality of Service (QoS).
- The data gram is marked for giving a certain treatment using this field.

Total length :

- Total length is a 16 bit field that contains the total length of the data gram (in bytes).

Total length = Header length + Payload length

Identification:

- Identification is a 16 bit field.
- It is used for the identification of the fragments of an original IP data gram.

When an IP data gram is fragmented,

- Each fragmented data gram is assigned the same identification number.
- This number is useful during the re assembly of fragmented data grams.
- It helps to identify to which IP data gram, the fragmented data gram belongs to.

DF-Bit:

- DF bit stands for Do Not Fragment bit.
- Its value may be 0 or 1.

When DF bit is set to 0,

- It grants the permission to the intermediate devices to fragment the data gram if required.

When DF bit is set to 1,

- It indicates the intermediate devices not to fragment the IP data gram at any cost.
- An error message is sent to the sender saying that the data gram has been discarded due to its settings.

MF-Bit :

- MF bit stands for More Fragments bit.
- Its value may be 0 or 1.

When MF bit is set to 0,

- It indicates to the receiver that the current data gram is either the last fragment in the set or that it is the only fragment.

When MF bit is set to 1,

- It indicates to the receiver that the current data gram is a fragment of some larger data gram.
- More fragments are following.
- MF bit is set to 1 on all the fragments except the last one.

Fragment offer:

- Fragment Offset is a 13 bit field.
- It indicates the position of a fragmented data gram in the original unfragmented IP data gram.
- The first fragmented data gram has a fragment offset of zero.

Time to Live:

- Time to live (TTL) is a 8 bit field.
- It indicates the maximum number of hops a data gram can take to reach the destination.
- The main purpose of TTL is to prevent the IP data grams from looping around forever in a routing loop.

Protocol:

- Protocol is a 8 bit field.
- It tells the network layer at the destination host to which protocol the IP data gram belongs to.
- In other words, it tells the next level protocol to the network layer at the destination side.

Header checksum:

- Header checksum is a 16 bit field.
- It contains the checksum value of the entire header.

- The checksum value is used for error checking of the header.

Source IP-Address:

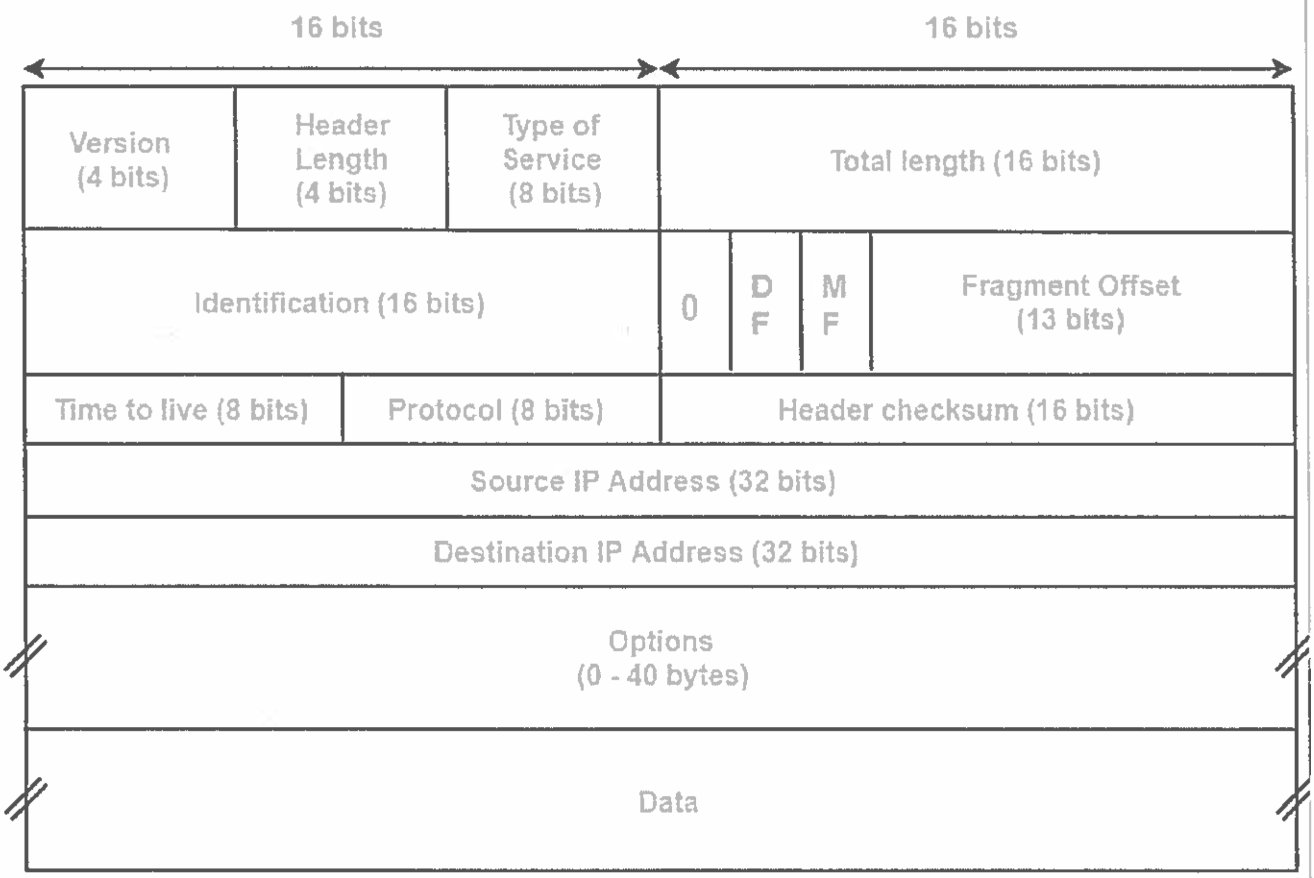
- Source IP Address is a 32 bit field.
- It contains the logical address of the sender of the data gram.

Destination IP address:

- Destination IP Address is a 32 bit field.
- It contains the logical address of the receiver of the data gram.

Options:

- Options is a field whose size vary from 0 bytes to 40 bytes.



IPv4 Header

14	**Explain below protocols a) FTP b) DNS c) SMTP d) HTTP-12Marks 20CS502.5 L2** **For writing FTP -3 Marks** **For writing DNS-3Marks** **For writing SMTP-3Marks** **For writing HTTP-3Marks**

FTP

File Transfer Protocol(FTP) is an application layer protocol which moves files between local and remote file systems. It runs on the top of TCP, like HTTP.

To transfer a file, 2 TCP connections are used by FTP in parallel: control connection and data connection.Although transferring files from one system to another is very simple and straightforward, but sometimes it can cause problems. For example, two systems may have different file conventions. Two systems may have different ways to represent text and data. Two systems may have different directory structures. FTP protocol overcomes these problems by establishing two connections between hosts. One connection is used for data transfer, and another connection is used for the control connection.

Simple Mail Transfer Protocol (SMTP)

Email is emerging as one of the most valuable services on the internet today. Most internet systems use SMTP as a method to transfer mail from one user to another. SMTP is a push protocol and is used to send the mail whereas POP (post office protocol) or IMAP (internet message access protocol) are used to retrieve those emails at the receiver's side.

SMTPFundamentals: SMTP is an application layer protocol. The client who wants to send the mail opens a TCP connection to the SMTP server and then sends the mail across the connection. The SMTP server is an always-on listening mode. As soon as it listens for a TCP connection from any client, the SMTP process initiates a connection through port 25. After successfully establishing a TCP connection the client process sends the mail instantly.

DNS:The **domain name system** (i.e., "DNS") is responsible for translating domain names into a specific IP address so that the initiating client can load the requested Internet resources. The domain name system works much like a phone book where users can search for a requested person and retrieve their phone number.DNS is a host name to IP address translation service. DNS is a distributed database implemented in a hierarchy of name servers. It is an application layer protocol for message exchange between clients and servers.

Requirement: Every host is identified by the IP address but remembering numbers is very difficult for the people and also the IP addresses are not static therefore a mapping is required to change the domain name to IP address. So DNS is used to convert the domain name of the websites to their numerical IP address.

HTTP:

- HTTP stands for HyperText Transfer Protocol. .
- The protocols that are used to transfer hypertext between two computers is known as Hyper Text Transfer Protocol.
- HTTP provides standard between a web browser and web server to establish communication. It is set of rules for transferring data from one computer to another. Data such as text, images, and other multimedia files are shared on the World Wide Web. Whenever a web user opens their web browser, user indirectly uses HTTP. It is an application protocol which is used for distributed, collaborative, hypermedia information systems.

OR

15	**What is Email? Explain any two scenarios of Email-12Marks 20CS502.5 L2**

For writing Email-2Marks

For Email architecture -2Marks

For 2 scenarios each one -3marks -6marks

ELECTRONIC MAIL:

Electronic mail, or more commonly email, has been around for over three decades. Faster and cheaper than paper mail, email has been a popular application since the early days of the Internet. It is one of the most widely used internet services. This service allows an Internet user to send a message in formatted manner (mail) to the other Internet user in any part of world. Message in mail not only contain text, but it also contains images, audio and videos data.

Architecture and Services

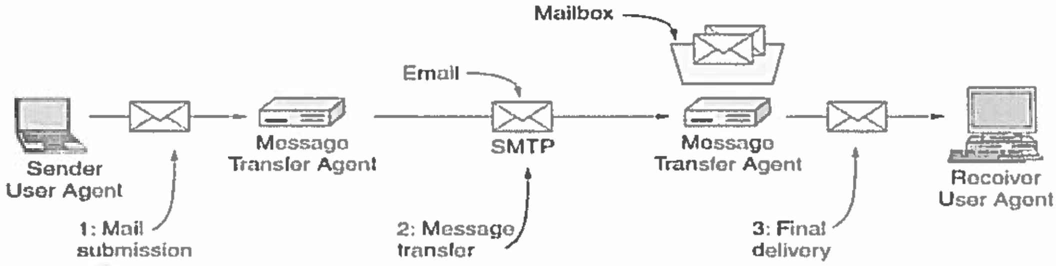


Figure 7-7. Architecture of the email system.

The user agent is a program that provides a graphical interface, or sometimes a text- and command-based interface that lets users interact with the email system. It includes a means to compose messages and replies to messages, display incoming messages, and organize messages by filing, searching, and discarding them.

The message transfer agents are typically system processes. They run in the background on mail server machines and are intended to be always available.

Their job is to automatically move email through the system from the originator to the recipient with SMTP (Simple Mail Transfer Protocol)

Message transfer agents also implement mailing lists, in which an identical copy of a message delivered to everyone on a list of email addresses.

Mailboxes store the email that is received for a user. They are maintained by mail servers. User agents simply present users with a view of the contents of their mailboxes

The message transport agents use the envelope for routing, just as the post office does.

Scenarios of EMAIL:

The User Agent

- A user agent is a program (sometimes called an email reader) that accepts a variety of commands for composing, receiving, and replying to messages, as well as for manipulating mailboxes.

- There are many popular user agents, including Google gmail, Microsoft Outlook, Mozilla Thunderbird, and Apple Mail. They can vary greatly in their appearance.

- Most user agents have a menu- or icon driven graphical interface that requires a mouse, or a touch interface on smaller mobile devices.

Message Formats:

Now we turn from the user interface to the format of the email messages themselves.

- Messages sent by the user agent must be placed in a standard format to be handled by the message transfer agents.

- First we will look at basic ASCII email using RFC 5322, which is the latest revision of the original Internet message format as described in RFC 822. After that, we will look at multimedia extensions to the basic format.

Header	Meaning
To:	Email address(es) of primary recipient(s)
Cc:	Email address(es) of secondary recipient(s)
Bcc:	Email address(es) for blind carbon copies
From:	Person or people who created the message
Sender:	Email address of the actual sender
Received:	Line added by each transfer agent along the route
Return-Path:	Can be used to identify a path back to the sender

Figure 7-10. RFC 5322 header fields related to message transport.

Message Transfer

- Now that we have described user agents and mail messages, we are ready to look at how the message transfer agents relay messages from the originator to the recipient.

- The mail transfer is done with the SMTP protocol.

- The simplest way to move messages is to establish a transport connection from the source machine to the destination machine and then just transfer the message. This is how SMTP originally worked. Over the years, however, two different uses of SMTP have been differentiated. The first use is mail submission.

Final Delivery

IMAP is an improvement over an earlier final delivery protocol, POP3 (Post Office Protocol, version 3), which is specified in RFC 1939. POP3 is a simpler protocol but supports fewer features and is less secure in typical usage. Mail is usually downloaded to the user agent computer, instead of remaining on the mail server. This makes life easier on the server, but harder on the user. It is not easy to read mail on multiple computers, plus if the user agent computer breaks, all email may be lost permanently.

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE (AI &ML)	Academic Year	2022 - 2023
Course Code	20AI503	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	High Performance Computing				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any two advantages and limitations of CUDA.	20AI503.1	L1
2	When multiple GPUs are preferred?	20AI503.2	L1
3	List any four benefits of Open CL.	20AI503.3	L1
4	List any four applications of Parallel Computing.	20AI503.4	L1
5	What is heterogeneous cluster computing?	20AI503.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain CUDA architecture with a neat diagram.	6M	20AI503.1	L2
6 (b)	Explain three parallel architecture schemes.	6M	20AI503.1	L2
	OR			
7	Define GPU. Explain architecture of GPU.	12M	20AI503.1	L2
8 (a)	What are the libraries and software components in CUDA 8.0?	8M	20AI503.2	L1
8 (b)	What are the advantages of CUDA over traditional general-purpose computation on GPUs?	4M	20AI503.2	L2
	OR			
9	List any five applications of CUDA.	12M	20AI503.2	L2
10 (a)	Why synchronization is important? Describe about implicit and explicit synchronization.	8M	20AI503.3	L2
10 (b)	Explain parallel programming techniques.	4M	20AI503.3	L2
	OR			
11 (a)	What is Wait protocol for synchronization?Compare Busy-wait and Sleep-wait protocols.	8M	20AI503.3	L2
11 (b)	Describe Error Handling in CUDA.	4M	20AI503.3	L2
12	Explain the concept of Host Device Interaction.	12M	20AI503.4	L2
	OR			
13 (a)	Describe memory consistency models in detail.	6M	20AI503.4	L2
13 (b)	What are the Pit falls of OpenCL applications?	6M	20AI503.4	L2
14 (a)	Write a "Hello World "program in open MPI.	8M	20AI503.5	L2
14 (b)	Differentiate MPI and Open MPI.	4M	20AI503.5	L2
	OR			
15	Explain prefix sum matrix multiplication algorithm with example.	12M	20AI503.5	L2

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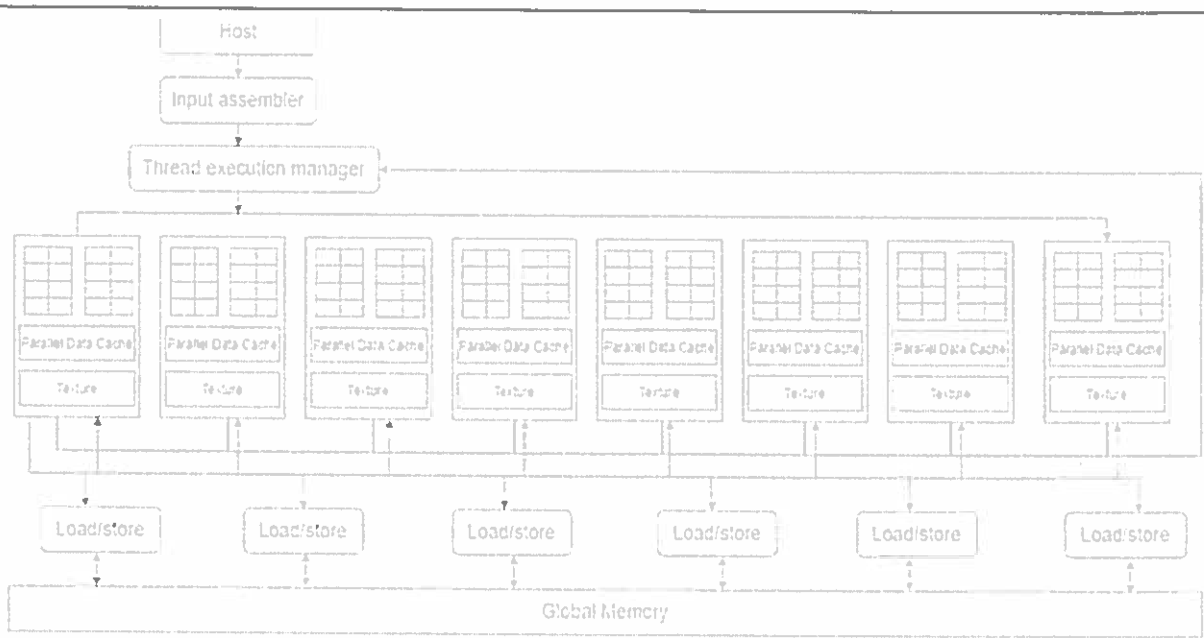
ANSWER KEY AND SCHEME OF EVALUATION

HPC ANSWER KEY

1	**List any two advantages and limitations of CUDA.** **advantages of CUDA.** • Scattered reads – code can read from arbitrary addresses in memory. • Shared memory – CUDA exposes a fast shared memory region that can be shared among threads. • Faster downloads and read backs to and from the GPU. **limitations of CUDA.** • CUDA source code is given on the host machine or GPU, as defined by the C++ syntax rules. Longstanding versions of CUDA use C syntax rules, which mean that up-to-date CUDA source code may or may not work as required. • CUDA has unilateral interoperability (the ability of computer systems or software to exchange and make use of information) with transferor languages like OpenGL. OpenGL can access CUDA registered memory, but CUDA cannot access OpenGL memory.
2	**When multiple GPUs are preferred?** • Enhanced performance. • Multiple applications. • Multi-tasking inside an application. • High throughput and responsiveness. • Hardware sharing among CPUs.
3	**List any four benefits of Open CL.** A primary benefit of OpenCL is substantial acceleration in parallel processing. OpenCL takes all computational resources, such as multi-core CPUs and GPUs, as peer computational units and correspondingly allocates different levels of memory, taking advantage of the resources available in the system. OpenCL also complements the existing OpenGL® visualization API by sharing data structures and memory locations without any copy or conversion overhead. A second benefit of OpenCL is cross-vendor software portability. This low-level layer draws an explicit line between hardware and the upper software layer. All the hardware implementation specifics, such as drivers and runtime, are invisible to the upper-level software programmers through the use of high-level abstractions,

	allowing the developer to take advantage of the best hardware without having to reshuffle the upper software infrastructure. The change from proprietary programming to open standard also contributes to the acceleration of general computation in a cross-vendor fashion.
4	**List any four applications of Parallel Computing.** There are various applications of Parallel Computing, which are as follows: o One of the primary applications of parallel computing is Databases and Data mining. o The real-time simulation of systems is another use of parallel computing. o The technologies, such as Networked videos and Multimedia. o Science and Engineering. o Collaborative work environments. o The concept of parallel computing is used by augmented reality, advanced graphics, and virtual reality.
5	**What is heterogeneous cluster computing?** Heterogeneous computing involves the use of a various types of computational units. A computation unit can be a general-purpose processing unit (such as a CPU), a graphics processing unit (such as a GPU), or a special-purpose processing unit (such as digital signal processor, or DSP). In the past, most computer applications were able to scale with advances in CPU technologies. With modem computer applications requiring interactions with various systems (such as audio/video systems, networked applications, etc.) even the advances in CPU technology proved insufficient to cope with this need. To achieve greater performance gains, specialized hardware was required, making the system heterogeneous. The addition of various types of computation units in these heterogeneous systems allows application designers to select the most suitable one on which to perform tasks
6 (a)	**Explain CUDA architecture with a neat diagram.** CUDA stands for Compute Unified Device Architecture. It is an extension of C/C++ programming. CUDA is a programming language that uses the Graphical Processing Unit (GPU). It is a parallel computing platform and an API (Application Programming Interface) model, Compute Unified Device Architecture was developed by Nvidia. This allows computations to be performed in parallel while providing well-formed speed. Using CUDA, one can harness the power of the Nvidia GPU to perform common computing tasks, such as processing matrices and other linear algebra operations, rather than simply performing graphical calculations. **Architecture of CUDA**



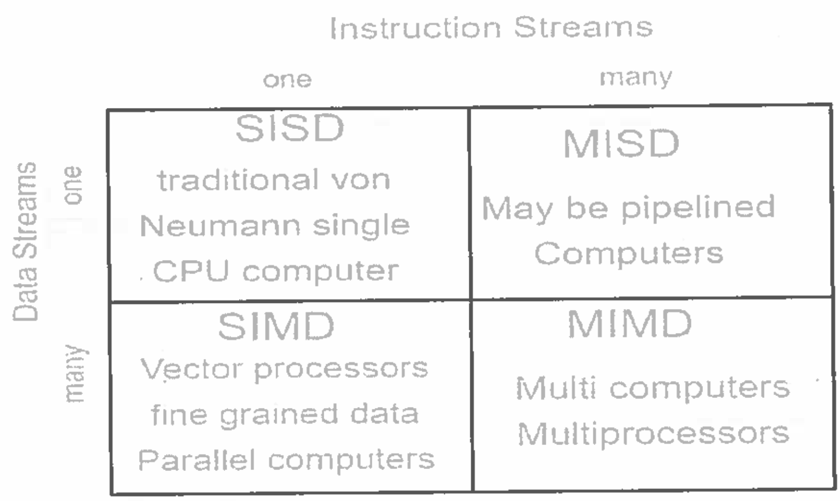
- 16 Streaming Multiprocessor (SM) diagrams are shown in the above diagram.
  Each Streaming Multiprocessor has 8 Streaming Processors (SP) ie, we get a total of 128 Streaming Processors (SPs).
- Now, each Streaming processor has a MAD unit (Multiplication and Addition Unit) and an additional MU (multiplication unit).
- The GT200 has 240 Streaming Processors (SPs), and more than 1 TFLOP processing power.
- Each Streaming Processor is gracefully threaded and can run thousands of threads per application.
- The G80 card supports 768 threads per Streaming Multiprocessor (note: not per SP).
- Eventually, after each Streaming Multiprocessor has 8 SPs, each SP supports a maximal of 96 threads. Total threads that can run – **128 * 96 = 12,228 times.**
- Therefore these processors are called **massively parallel.**
- The G80 chips have a memory bandwidth of 86.4GB/s.
- It also has an 8GB/s communication channel with the CPU (4GB/s for uploading to the CPU RAM, and 4GB/s for downloading from the CPU RAM).

6 (b) **Explain three parallel architecture schemes.**

Parallel computing is a computing where the jobs are broken into discrete parts that can be executed concurrently. Each part is further broken down to a series of instructions. Instructions from each part execute simultaneously on different CPUs. Parallel systems deal with the simultaneous use of multiple computer resources that can include a single computer with multiple processors, a number of computers connected by a network to form a parallel processing cluster or a combination of both.

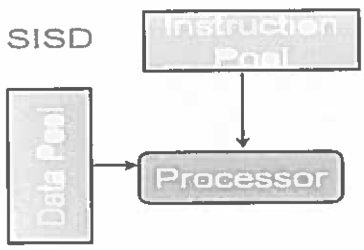
Parallel systems are more difficult to program than computers with a single processor because the architecture of parallel computers varies accordingly and the processes of multiple CPUs must be coordinated and synchronized.

The crux of parallel processing are CPUs. Based on the number of instruction and data streams that can be processed simultaneously, computing systems are classified into four major categories:



1. Single-instruction, single-data (SISD) systems –

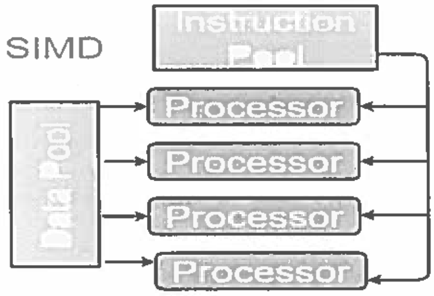
An SISD computing system is a uni processor machine which is capable of executing a single instruction, operating on a single data stream. In SISD, machine instructions are processed in a sequential manner and computers adopting this model are popularly called sequential computers. Most conventional computers have SISD architecture. All the instructions and data to be processed have to be stored in primary memory.



The speed of the processing element in the SISD model is limited (dependent) by the rate at which the computer can transfer information internally. Dominant representative SISD systems are IBM PC, workstations.

2. Single-instruction, multiple-data (SIMD) systems –

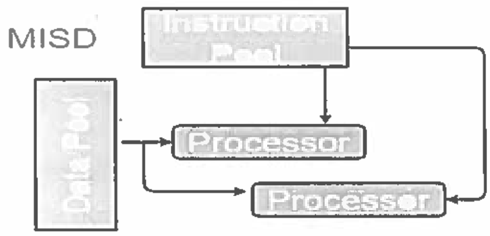
A SIMD system is a multiprocessor machine capable of executing the same instruction on all the CPUs but operating on different data streams. Machines based on a SIMD model are well suited to scientific computing since they involve lots of vector and matrix operations. So that the information can be passed to all the processing elements (PEs) organized data elements of vectors can be divided into multiple sets(N-sets for N PE systems) and each PE can process one data set.



Dominant representative SIMD systems are Cray's vector processing machine.

3. Multiple-instruction, single-data (MISD) systems –

An MISD computing system is a multiprocessor machine capable of executing different instructions on different PEs but all of them operating on the same dataset .

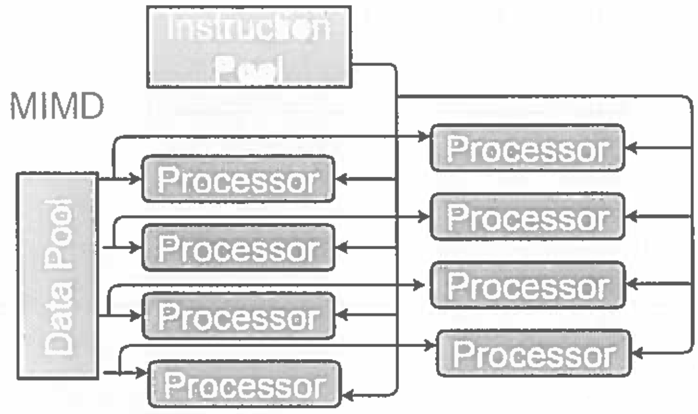


Example Z = sin(x)+cos(x)+tan(x)

The system performs different operations on the same data set. Machines built using the MISD model are not useful in most of the application, a few machines are built, but none of them are available commercially.

4. Multiple-instruction, multiple-data (MIMD) systems –

An MIMD system is a multiprocessor machine which is capable of executing multiple instructions on multiple data sets. Each PE in the MIMD model has separate instruction and data streams; therefore machines built using this model are capable to any kind of application. Unlike SIMD and MISD machines, PEs in MIMD machines work asynchronously.



MIMD machines are broadly categorized into shared-memory MIMD and distributed-memory MIMD based on the way PEs are coupled to the main memory.

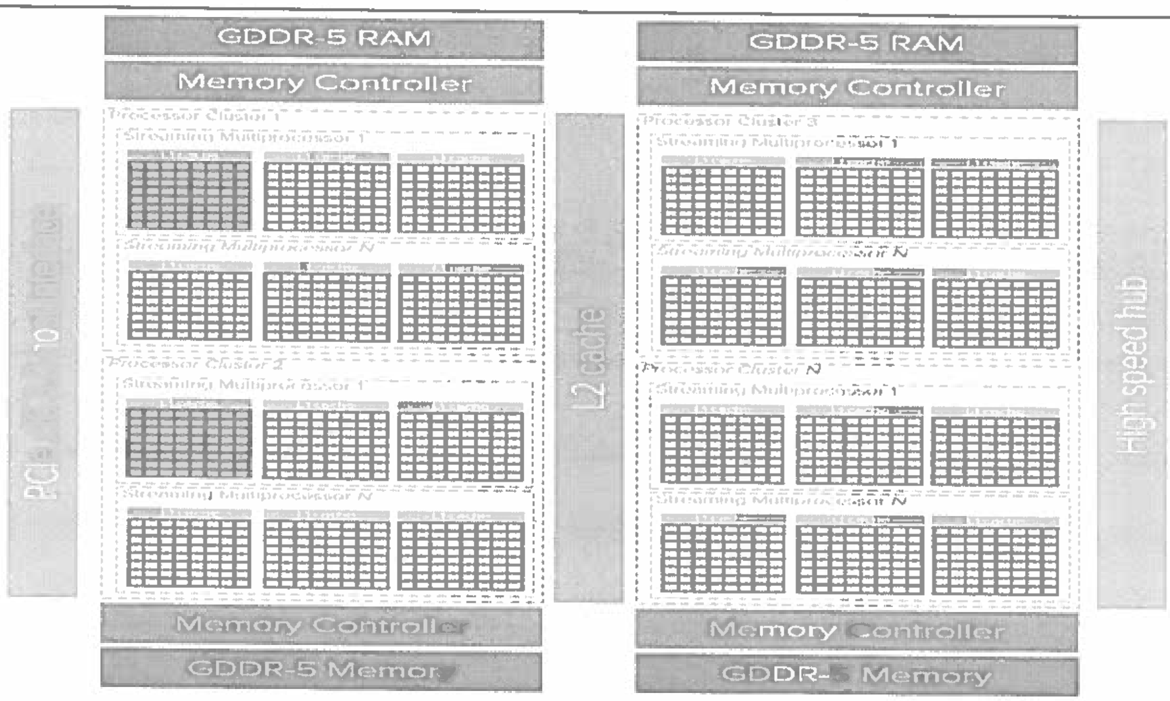
In the shared memory MIMD model (tightly coupled multiprocessor systems), all the PEs are connected to a single global memory and they all have access to it. The communication between PEs in this model takes place through the shared memory; modification of the data stored in the global memory by one PE is visible to all other PEs. Dominant representative shared memory MIMD systems are Silicon Graphics machines and Sun/IBM's SMP (Symmetric Multi-Processing).

7 **Define GPU. Explain architecture of GPU.**

GPU stands for GPU stands for **Graphics Processing Unit**. GPUs are also known as video cards or graphics cards. In order to display pictures, videos, and 2D or 3D animations, each device uses a GPU. A GPU performs fast calculations of arithmetic and frees up the CPU

To do different things. A GPU has lots of smaller cores made for multi-tasking, while a CPU makes use of some cores primarily based on sequential serial processing. In the world of computing, graphics processing technology has advanced to offer specific benefits. The modern GPUs enable new possibilities in **content creation, machine learning, gaming**, etc.**Graphics Processing Unit**. GPUs are also known as video cards or graphics cards. In order to display pictures, videos, and 2D or 3D animations, each device uses a GPU. A GPU performs fast calculations of arithmetic and frees up the CPU
to do different things. A GPU has lots of smaller cores made for multi-tasking, while a CPU makes use of some cores primarily based on sequential serial processing. In the world of computing, graphics processing technology has advanced to offer specific benefits. The modern GPUs enables new possibilities in **content creation, machine learning, gaming**, etc.

If we inspect the high-level architecture overview of a GPU (again, strongly depended on make/model), it looks like the nature of a GPU is all about putting available cores to work and it's less focused on low latency cache memory access.



A single GPU device consists of multiple Processor Clusters (PC) that contain multiple Streaming Multiprocessors (SM). Each SM accommodates a layer-1 instruction cache layer with its associated cores. Typically, one SM uses a dedicated layer-1 cache and a shared layer-2 cache before pulling data from global GDDR-5 (or GDDR-6 in newer GPU models) memory. Its architecture is tolerant of memory latency.

8 (a) **what are the libraries and software components in CUDA 8.0?**

The CUDA platform is accessible to software developers through CUDA-accelerated libraries, compiler directives such as Opencast and extensions to industry-standard programming languages including C, C++ and FORTRAN. C/C++ programmers can use 'CUDA C/C++', compiled to PTX with nvcc, Nvidia's LLVM-based C/C++ compiler, or by clang itself. Fortran programmers can use 'CUDA Fortran', compiled with the PGI CUDA Fortran compiler from The Portland Group.

In addition to libraries, compiler directives, CUDA C/C++ and CUDA Fortran, the CUDA platform supports other computational interfaces, including the Khronos Group's OpenCL, Microsoft's Direct Compute, OpenGL Compute Shadier and C++ AMP Third party wrappers are also available for Python, Perl, Fortran, Java, Ruby, Lua, Common Lisp, Haskell, R, MATLAB, IDL, Julia, and native support in Mathematical

CUDA 8.0 comes with the following libraries (for compilation & runtime, in alphabetical order):

- cuBLAS – CUDA Basic Linear Algebra Subroutines library
- CUDART – CUDA Runtime library
- cuFFT – CUDA Fast Fourier Transform library
- cuRAND – CUDA Random Number Generation library
- cuSOLVER – CUDA based collection of dense and sparse direct solvers
- cuSPARSE – CUDA Sparse Matrix library
- NPP – NVIDIA Performance Primitives library
- nvGRAPH – NVIDIA Graph Analytics library
- NVML – NVIDIA Management Library
- NVRTC – NVIDIA Runtime Compilation library for CUDA C++

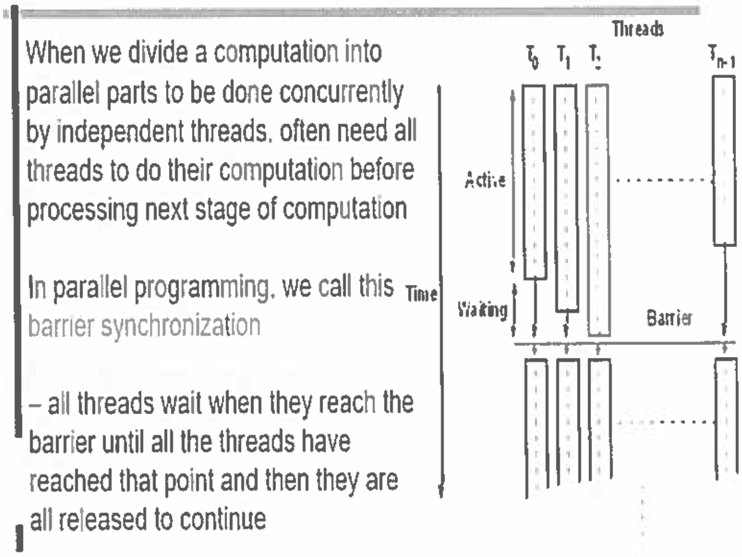
	CUDA 8.0 comes with these other software components: • nView – NVIDIA nView Desktop Management Software • NVWMI – NVIDIA Enterprise Management Toolkit • GameWorks PhysX – is a multi-platform game physics engine
8 (b)	**what are the advantages of CUDA over traditional general-purpose computation on GPUs?** A special approach is required to move computations to a GPU according to this model. Even element wise addition of two vectors will require drawing a figure on screen or to an off-screen buffer. A figure is rasterized; the color of each pixel is calculated by a given program (pixel shader). The program reads input data from textures for each pixel, adds them, and records them to an output buffer. And all these multiple operations do what a single operator can do in a usual programming language! So GPGPU usage for general-purpose computations is limited due to its steep learning curve. There are also other limitations – a pixel shader is just a formula describing how a resulting pixel color depends on its coordinates. And the language of pixel shaders just records these formulas with C-like syntax. Early GPGPU methods are a cunning trick to use the power of GPU, but they are not convenient. Data are represented by images (textures), and an algorithm is a raster process. We should also note a specific model of memory and execution. **Advantages of CUDA over the traditional approach to GPGPU computing:** • Programming interface of CUDA applications is based on the standard C language with extensions, which facilitates the learning curve of CUDA • CUDA provides access to 16 KB of memory (per multiprocessor) shared between threads, which can be used to setup cache with higher bandwidth than texture lookups • More efficient data transfers between system and video memory • No need in graphics APIs with their redundancy and overheads • Linear memory addressing, gather and scatter, writing to arbitrary addresses • Hardware support for integer and bit operations
9.	**List any five applications of CUDA.** **CUDA applications must run parallel operations on a lot of data, and be processing-intensive.** **Fast Video Transcoding** Transcoding is a very common, and highly complex procedure which easily involves trillions of parallel computations, many of which are floating point operations. Applications such as Bad boom have been created which harness the raw computing power of GPUs in order to transpose video much faster than ever before. For example, if you want to transcode a DVD so it will play on your iPod, it may take several hours to fully transcode. However, with Bad boom, it is possible to transcode the movie or any video file faster than real time. **Video Enhancement** Complicated video enhancement techniques often require an enormous amount of computations. For example, there are algorithms that can upscale a movie by using information from frames surrounding the current frame. This involves too many computations for a CPU to handle in real time. ArcSoft was able to create a plugin for it's

	movie player which uses CUDA in order to perform DVD upscaling in real time! This is an amazing feat, and greatly enhances any movie watching experience if you have a high definition monitor. This is fine example of how mainstream programs are harnessing the computational power of CUDA in order to delight their customers. Another fine example would be vReveal, which is able to perform a variety of enhancements to motion video, and then save the resulting video. **Oil and Natural Resource Exploration** The first two topics I talked about had to do with video, which is naturally suited for the video card. Now it's time to talk about more serious technologies involving oil, gas, and other natural resource exploration. Using a variety of techniques, it is overwhelmingly difficult to construct a 3d view of what lies underground, especially when the ground is deeply submerged in a sea. Scientists used to work with very small sample sets, and low resolutions in order to find possible sources of oil. Because the ground reconstruction algorithms are highly parallel, CUDA is perfectly suited to this type of challenge. Now CUDA is being used to find oil sources quicker. **Medical Imaging** CUDA is a significant advancement for the field of medical imaging. Using CUDA, MRI machines can now compute images faster than ever possible before, and for a lower price. Before CUDA, it used to take an entire day to make a diagnosis of breast cancer. Now with CUDA, this can take 30 minutes. In fact, patients no longer need to wait 24 hours for the results, which will benefit many people. **Computational Sciences** In the raw field of computational sciences, CUDA is very advantageous. For example, it is now possible to use CUDA with MATLAB, which can increase computations by a great amount. Other common tasks such as computing Eigen values, or SVD decompositions, or other matrix mathematics can use CUDA in order to speed up calculations. **Neural Networks** I personally worked on a program which required the training of several thousand neural networks to a large set of training data. Using the Core 2 Duo CPU that was available to me, it would have taken over a month to get a solution. However, with CUDA, I was able to reduce my time to solution to under 12 hours. **Fluid Dynamics** Fluid dynamics simulations have also been created. These simulations require a huge number of calculations, and are useful for wing design, and other engineering tasks.
10(a)	**why synchronization is important? Describe about implicit and explicit synchronization.** Synchronization in CUDA is the term used for sharing of information between threads within a block, or between blocks within a grid. A thread can access register space or local memory space, both of which are private to the thread. in CUDA is the term used for sharing of information between threads within a block, or between blocks within a grid. A thread can access register space or local memory space, both of which are private to the thread. The CUDA API has a method,

__syncthreads(),CudaThreadSynchronize() to synchronize threads. When the method is encountered in the kernel, all threads in a block will be blocked at the calling location until each of them reaches the location.

It ensure phase synchronization. That is, all the threads of a block will now start executing their next phase only after they have finished the previous one. There are certain nuances to this method.

Explicit synchronization with child kernels from a parent block (i.e. using cudaDeviceSynchronize() in device code) is deprecated in CUDA 11.6, removed for compute_90+ compilation, and is slated for full removal in a future CUDA release.



Implicit synchronization Implicit Wait tells the Web Driver to Wait until the stated time before throwing the No Such Element /Element Not Visible exception. Waiting time across the test script between each consecutive steps are taken by default. Hence, next test Step will execute only when the specified time is elapsed post executing the previous test Step.

Explicit synchronization: Explicit waits are very good to use when page loads dynamically. Explicit Wait tells the Web Driver to Wait until the specified condition is met or maximum time elapses before throwing No Such Element (or) Element Not Visible Exceptions. Explicit waits are applied for the specified test Step in test script.

CUDA runtime operations from any thread, including kernel launches, are visible across a thread block. This means that an invoking thread in the parent grid may perform synchronization on the grids launched by that thread, by other threads in the thread block, or on streams created within the same thread block. Execution of a thread block is not considered complete until all launches by all threads in the block have completed. If all threads in a block exit before all child launches have completed, a synchronization operation will automatically be triggered.

10(b) **Explain parallel programming techniques.**

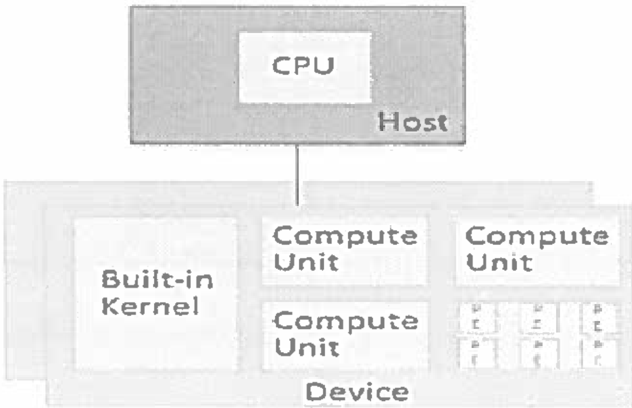
OpenMP

OpenMP is an API that supports multi-platform shared memory multiprocessing programming in C, C++, and Fortran. It is prevalent only on a multi-core computer platform with a shared memory subsystem.

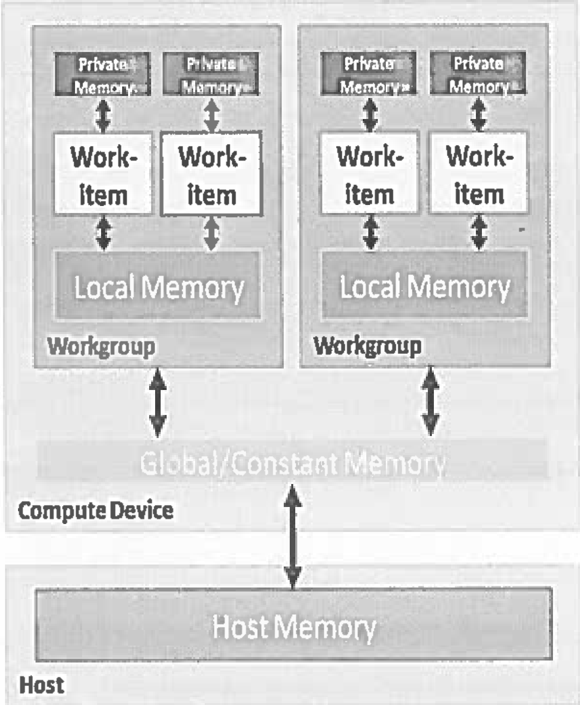
	MPI **Message Passing Interface (MPI)** has an advantage over OpenMP, that it can run on either the shared or distributed memory architecture. Distributed memory computers are less expensive than large shared memory computers. But it has its own drawback with inherent programming and debugging challenges. One major disadvantage of MPI parallel framework is that the performance is limited by the communication network between the nodes.Supercomputers have a massive number of processors which are interconnected using a high speed network connection or are in computer clusters, where computer processors are in close proximity to each other. In clusters, there is an expensive and dedicated data bus for data transfers across the computers. MPI is extensively used in most of these compute monsters called supercomputers. **OpenACC** The **OpenACC Application Program Interface (API)** describes a collection of compiler directives to specify loops and regions of code in standard C, C++, and Fortran to be offloaded from a host CPU to an attached accelerator, providing portability across operating systems, host CPUs, and accelerators. OpenACC is similar to OpenMP in terms of program annotation, but unlike OpenMP which can only be accelerated on CPUs, OpenACC programs can be accelerated on a GPU or on other accelerators also. OpenACC aims to overcome the drawbacks of OpenMP by making parallel programming possible across heterogeneous devices. OpenACC standard describes directives and APIs to accelerate the applications. The ease of programming and the ability to scale the existing codes to use the heterogeneous processor warrantees a great future for OpenACC programming.
11(a)	**What is Wait protocol for synchronization?Compare Busy-wait and Sleep-wait protocols.** The wait protocol is used for resolving the conflicts, which arise because of a number of multiprocessors demanding the same resource. There are two types of wait protocols: busy-wait and sleep-wait. In busy-wait protocol, process stays in the process context register, which continuously tries for processor availability. In sleep-wait protocol, wait protocol process is removed from the processor and is kept in the wait queue. The hardware complexity of this protocol is more than busy-wait in multiprocessor system; if locks are used for synchronization then busy-wait is used more than sleep-wait **Sleep():** This Method is used to pause the execution of current thread for a specified time in Milliseconds. Here, Thread does not lose its ownership of the monitor and resume's it's execution **Wait():** This method is defined in object class. It tells the calling thread (a.k.a Current Thread) to wait until another thread invoke's the notify() or notifyAll() method for this object, The thread waits until it reobtains the ownership of the monitor and Resume's Execution.

Busy Wait()	Sleep wait ()
Wait() method belongs to Object class.	Sleep() method belongs to Thread class.

	Wait() method releases lock during Synchronization.	Sleep() method does not release the lock on object during Synchronization.
	Wait() should be called only from Synchronized context.	There is no need to call sleep() from Synchronized context.
	Wait() is not a static method.	Sleep() is a static method.
	Wait() Has Three Overloaded Methods: • wait() • wait(long timeout) • wait(long timeout, int nanos)	Sleep() Has Two Overloaded Methods:
11(b)	**Describe Error Handling in CUDA.** **CUDA error handling** we introduced the CUDA_CALL macro. All of the CUDA API functions return an error code. Anything other than cudaSuccess generally indicates you did something wrong in calling the API. There are, however, a few exceptions, such as cudaEventQuery, which returns the event status as opposed to an error status. The CUDA API is by nature asynchronous, meaning the error code returned at the point of the query, may have happened at some distant point in the past. In practice, it will usually be as a result of the call immediately prior to the error being detected. You can, of course, force this by synchronizing (i.e., calling the cudaDeviceSynchronize function) after every API call The CUDA error handling can be somewhat rudimentary. Most of the time, you'll get a useful error message. However, often you will get a not-so-useful message such as unknown error, usually after a kernel invocation. **Kernel launching and bounds checking** One of the most common failings in CUDA is an array overrun. You should ensure all your kernel invocations start with a check to ensure the data they will access, both for read and write purposes, is guarded by a conditional. **Invalid device handles** The other types of errors you typically see are incorrect mixing of handles, most often pointers. When you allocate memory on the device or on the host, you receive a pointer to that memory **Volatile qualifiers** The C "volatile" keyword specifies to the compiler that all references to this variable, read or write, must result in a memory reference, and those references must be in the order specified in the program.	
12.	**Explain the concept of Host Device Interaction.** **The Platform Model** The OpenCL platform model defines the logical representation of all hardware capable of executing an OpenCL program. OpenCL platforms are defined by the grouping of a host processor and one or more OpenCL compute devices. The host processor, which runs the OS for the system, is also responsible for the general bookkeeping	

	and task launch duties associated with the execution of OpenCL applications. The device is the hardware element in the system on which the compute kernels of an OpenCL application are executed. Each device is further divided into a set of compute units. The number of compute units depends on the target hardware. A compute unit is further subdivided into processing elements. A processing element is the fundamental computation engine in the compute unit, which is responsible for executing the operations of one work item.  The preceding figure shows a conceptual view of the OpenCL platform model. An OpenCL platform always starts with a host processor. For platforms created with Xilinx® devices, the host processor is an x86 based processor communicating to the devices using PCIe®. The host processor has the following responsibilities: Manage the operating system and enable drivers for all devices. Execute the application host program. Set up all global memory buffers and manage data transfer between the host and the device. Monitor the status of all compute units in the system. In all OpenCL platforms, the host processor tasks are executed using a common set of OpenCL API. The implementation of the OpenCL API functions is provided by the hardware vendor and is referred to as the OpenCL runtime library. The OpenCL runtime library is responsible for translating user commands described by the OpenCL API into hardware specific commands for a given device. For example, when the application programmer allocates a memory buffer using the clCreateBuffer API, it is the responsibility of the runtime library to keep track of where the allocated buffer physically resides in the system, and of the mechanism required for buffer access. It is important for the application programmer to keep in mind that the OpenCL API is portable across vendors, but the runtime library provided by a vendor is not. Therefore, OpenCL applications have to be linked at compile time with the runtime library that is paired with the target execution device. The other component of a platform is the device. A device in the context of OpenCL is the physical collection of hardware resources onto which the application kernels are executed. A platform must have at least one device available for the execution of kernels. Also, per the OpenCL platform model, all devices in a platform do not have to be of identical type. A host is any computer with a CPU running a standard operating system. OpenCL devices can be a GPU, DSP, or a multi-core CPU. An OpenCL device consists of a collection of one or more compute units (cores). A compute unit is further composed of one or more processing elements. Processing elements execute instructions as SIMD (Single Instruction, Multiple Data) or SPMD (Single Program, Multiple Data). SPMD instructions are typically executed on general purpose devices such as CPUs, while SIMD instructions require a vector processor such as a GPU or vector units in a CPU.
13(a)	**Explain memory hierarchy question** **The Memory Model**

Since common memory address space is unavailable on the host and the OpenCL devices, the OpenCL



memory model defines four regions of memory accessible to work-items when executing a kernel. The following figure shows the regions of memory accessible by the host and the compute device:

Host Memory
The host memory is defined as the region of system memory that is directly (and only) accessible from the host processor. Any data needed by compute kernels must be transferred to and from OpenCL device global memory using the OpenCL API.

Global Memory
The global memory is defined as the region of device memory that is accessible to both the OpenCL host and device. Global memory permits read/write access to the host processor as well to all compute units in the device. The host is responsible for the allocation and de-allocation of buffers in this memory space. There is a handshake between host and device over control of the data stored in this memory. The host processor transfers data from the host memory space into the global memory space. Then, once a kernel is launched to process the data, the host loses access rights to the buffer in global memory. The device takes over and is capable of reading and writing from the global memory until the kernel execution is complete. Upon completion of the operations associated with a kernel, the device turns control of the global memory buffer back to the host processor. Once it has regained control of a buffer, the host processor can read and write data to the buffer, transfer data back to the host memory, and de-allocate the buffer.

Constant Global Memory
Constant global memory is defined as the region of system memory that is accessible with read and write access for the OpenCL host and with read only access for the OpenCL device. As the name implies, the typical use for this memory is to transfer constant data needed by kernel computation from the host to the device.

Local Memory
Local memory is a region of memory that is local to a single compute unit. The host processor has no visibility and no control on the operations that occur in this memory space. This memory space allows read and write operations by all the processing elements with a compute units. This level of memory is typically used to store data that must be shared by multiple work-items. Operations on local memory are un-ordered between work-items but synchronization and consistency can be achieved using barrier and fence operations. In SDAccel, the structure of local memory can be customized to meet the requirements of an algorithm or application.

Private Memory

Private memory is the region of memory that is private to an individual work-item executing within an OpenCL processing element. As with local memory, the host processor has no visibility into this memory region. This memory space can be read from and written to by all work-items, but variables defined in one work-item's private memory are not visible to another work-item. In SDAccel, the structure of private memory can be customized to meet the requirements of an algorithm or application.

For devices using an FPGA device, the physical mapping of the OpenCL memory model is the following:

- Host memory is any memory connected to the host processor only.
- Global and constant memories are any memory that is connected to the FPGA device. These are usually memory chips (e.g. SDRAM) that are physically connected to the FPGA device, but might also include distributed memories (e.g. Block RAM) within the FPGA fabric. The host processor has access to these memory banks through infrastructure provided by the FPGA platform.
- Local memory is memory inside of the FPGA device. This memory is typically implemented using registers or Block Rams in the FPGA fabric.
- Private memory is memory inside of the FPGA device. This memory is typically implemented using registers or Block Rams in the FPGA fabric.

13(b) **What are the Pit falls of OpenCL applications?**

OpenCL offers many applications in computational physics in comparison to traditional MPI/OpenMP parallelization. We present an MPI/OpenCL based plasma simulation code, as an example of how computational physics can benefit from OpenCL. The code utilizes a hybrid modeling approach which combines elements from both fluid and particle-in-cell (PIC) methods. Most applications in computational physics and engineering are based on either of these models. Typical application range from fluid simulations for the characterization of flow behavior to PIC simulations of different plasma applications like nuclear fusion. Hence the hybrid model includes many of the common problems encountered in computational science. This includes for example solving differential equations, systems of linear equations and parallel reduction. Therefore it is well suited to show the common advantages and problems that arise with an OpenCL based approach. For this purpose we compare performance results between CPUs and GPUs for the above mentioned problems as they arise in the simulation code. Overall the runtime per FLOP using GPUs is significantly lower compared to CPUs. One reason for this advantage in performance is that numerical solvers can be implemented more efficiently on GPUs compared to CPUs due to the inherent advantage of GPU architecture.

14(a) **Write a "Hello World "program in open MPI.**

```
#include <mpi.h>
#include <stdio.h>

int main(int argc, char** argv) {
  // Initialize the MPI environment
  MPI_Init(NULL, NULL);

  // Get the number of processes
  int world_size;
  MPI_Comm_size(MPI_COMM_WORLD, &world_size);

  // Get the rank of the process
  int world_rank;
  MPI_Comm_rank(MPI_COMM_WORLD, &world_rank);
```

```
// Get the name of the processor
char processor_name[MPI_MAX_PROCESSOR_NAME];
int name_len;
MPI_Get_processor_name(processor_name, &name_len);

// Print off a hello world message
printf("Hello world from processor %s, rank %d out of %d processors\n",
    processor_name, world_rank, world_size);

// Finalize the MPI environment.
MPI_Finalize();
```

14(b) **Differentiate MPI and Open MPI.**

OpenMP	OpenMPI
High-level API allowing shared-memory parallel computing	High-level implementation of Message Passing Interace (MPI) for distributed-memory systems.
Allows parallel code to run on a single multi-core system	Allows parallel code to run on multiple systems connected by a network
Automatically creates multiple threads and deals with synchronization	Provides API that allows programmer to control communication between distributed nodes
Automatically reduces/compiles the final results	Programmer has to manually receive and compile the results
Can run offline in isolation	Needs a network to function
Can execute directly on any multi-core system	Needs a setup before execution where we identify each node in the network
Owned by OpenMP Architecture Review Board	Open-source BSD license
Available in C, C++, and Fortran	Available in many languages through bindings. Comes with wrapper compilers for C, C++, and Fortran

15. **Explain prefix sum matrix multiplication algorithm with example.**

Prefix sum

Given an array arr[] of size n, its prefix sum array is another array prefixSum[] of the same size, such that the value of prefixSum[i] is arr[0] + arr[1] + arr[2] ... arr[i].

The operation can be implemented using the all-to-all broadcastkernel.

We must account for the fact that in prefix sums the node with label *k*uses information from only the *k*-node subset whose labels are less than or equal to *k*.

This is implemented using an additional result buffer. The content of anincoming message is added to the result buffer only if the message comes from a node with a smaller label than the recipient node.

The contents of the outgoing message (denoted by parentheses in thefigure) are updated with every incoming message.

Examples :

Input : arr[] = {10, 20, 10, 5, 15}

Output : prefixSum[] = {10, 30, 40, 45, 60}

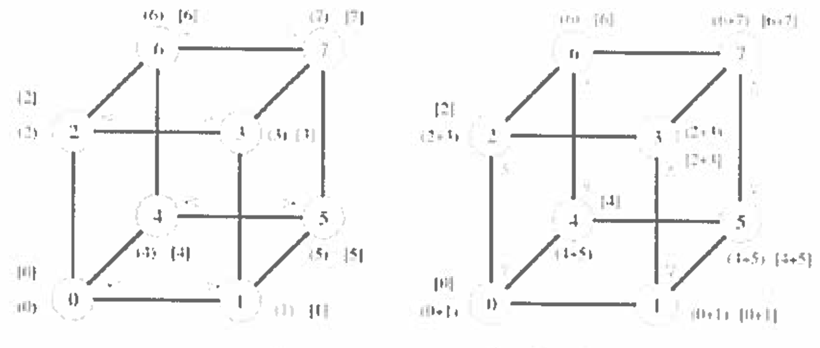
Explanation : While traversing the array, update the element by adding it with its previous element.

prefixSum[0] = 10,
prefixSum[1] = prefixSum[0] + arr[1] = 30,
prefixSum[2] = prefixSum[1] + arr[2] = 40 and so on.

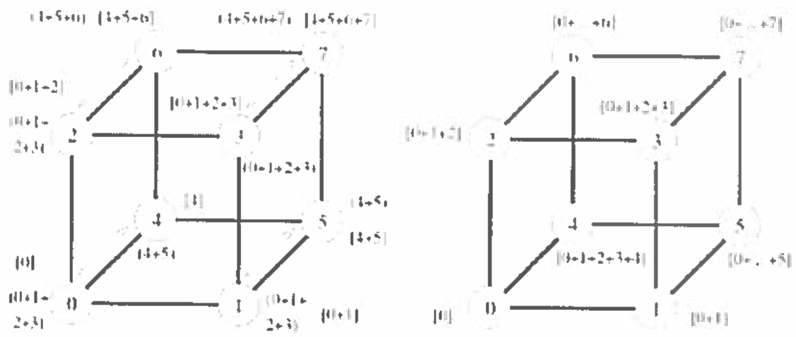
Prefix sum algorithm

- **Declare an array prefixSum of size n with all its entries initialized to 0.**
- **Assign a[0] to prefixSum[0].**
- **Iterate from i = 0 to i = n - 1 and do the following -**
  - ➢ **Store the value of prefixSum[i-1] in a varaible (say prev).**
  - ➢ **Assign the value prev + a[i] to prefixSum[i].**

Return prefixSum



(a) Initial distribution of values (b) Distribution of sums before second step



(c) Distribution of sums before third step (d) Final distribution of prefix sums

Application of Prefix Sum:

- **Equilibrium index of an array -** The equilibrium index can be defined as the index in the array such that the sum of elements of lower indices is equal to the sum of elements of higher indices. This can easily be found by traversing the prefix Sum array once and for each index i checking if the sum of range [0, i] is equal to the sum of range [i+1, n - 1].

- **Find if there exists a sub array with sum 0 -** Given an array consisting of integers (possibly negative integers). Check if there exists a non-empty sub array such that the sum of elements in it is 0. This can be checked using the prefixSum and some simple hashing concepts.

Example2

Let d_M be –

2	4	1
8	7	4
7	4	9

The above matrix will be stored as –

2	4	1	8	7	4	7	4	9

And let d_N be –

4	8	9
1	7	0
2	5	4

The above matrix will be stored as –

4	8	9	1	7	0	2	5	4

Since d_P will be a 3x3 matrix, we will be launching 9 threads, each of which will compute one element of d_P.

d_P matrix

(0,0)	(0,1)	(0,2)
(1,0)	(1,1)	(1,2)
(2,0)	(2,1)	(2,2)

Let us compute the (2,1) element of d_P by doing a dry-run of the kernel –

row=2;
col=1;

When (k=0)

```
product_val = 0 + d_M[2*3+0] * d_N[0*3+1]
product_val = 0 + d_M[6]*d_N[1] = 0+7*8=56
```

1st Iteration

```
product_val = 56 + d_M[2*3+1]*d_N[1*3+1]
product_val = 56 + d_M[7]*d_N[4] = 84
```

Final Iteration

```
product_val = 84+d_M[2*3+2]*d_N[2*3+1]
product_val = 84+d_M[8]*d_N[7] = 129
```

	Now, we have d_P[7] = 129 •

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE (DS)	Academic Year	2022 – 2023
Course Code	20AI603	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Machine Learning				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List out any four applications of machine learning.	20AI603.1	L1
2	What is a Decision Tree?	20AI603.2	L1
3	Define a Dendrogram.	20AI603.3	L1
4	Explain Posteriori Probability.	20AI603.4	L2
5	Define Reinforcement Learning.	20AI603.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Differentiate in between Classification and Regression.	6M	20AI603.1	L2
6 (b)	Explain about binary classification.	6M	20AI603.1	L2
	OR			
7 (a)	How to classify the data with multiple class labels? Explain in detail	6M	20AI603.1	L2
7 (b)	With example explain how the Concept Learning task determines the Hypothesis for given target concept.	6M	20AI603.1	L2
8 (a)	Explain about decision tree based learning? How it is represented. Give some problems for which decision tree learning is appropriate.	6M	20AI603.2	L2
8 (b)	Discuss in detail about Learning Ordered Rule Lists.	6M	20AI603.2	L2
	OR			
9 (a)	Explain Tree Learning as variance reduction.	6M	20AI603.2	L2
9 (b)	Describe in detail about descriptive rule learning.	6M	20AI603.2	L2
10 (a)	Describe the procedure used by Least Square Methods for predicting the target classes.	6M	20AI603.3	L2
10 (b)	Explain about nearest neighbor classification.	6M	20AI603.3	L3
	OR			
11 (a)	Explain SVM Algorithm and its Kernel methods.	6M	20AI603.3	L2
11 (b)	Explain hierarchical clustering with an example.	6M	20AI603.3	L2
12 (a)	Demonstrate Normal or Gaussian distribution with an example.	6M	20AI603.4	L2
12 (b)	Explain how Discriminative Learning is used to perform pattern classification.	6M	20AI603.4	L2
	OR			
13 (a)	Explain Naïve Bayes Classifier with an example.	6M	20AI603.4	L2
13 (b)	Demonstrate in detail about Compression based models.	6M	20AI603.4	L2
14 (a)	Define Q learning? Explain with an example about Q-learning.	6M	20AI603.5	L2
14(b)	Describe the random forest algorithm to improve classifier accuracy.	6M	20AI603.5	L2
	OR			
15 (a)	Compare and Contrast Bagging and Boosting ensemble techniques.	6M	20AI603.5	L2
15 (b)	Explain in detail the main components of reinforcement learning. Is RNN reinforcement learning algorithm	6M	20AI603.5	L2

16
16/5/24

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SCHEME OF VALUATION
&
ANSWER KEY

Degree	B. Tech. (U. G.)	Program	CSE	Test	END EXAM	Academic Year	2022 - 2023
Course Code	20AI603	Test Duration	180 Min.	Max. Marks	70	Semester	I
Course	MACHINE LEARNING						
Assessment Pattern							
R (L1):	U (L2):	Apply (L3):	Analyze (L4):		E (L5):		C (L6)

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	**List any four applications of ML.** For this we can give two marks for any four applications of AI ML in Business ML in Engineering ML in Manufacturing ML in Medical Field ML in Telecommunications ML in Banking ML in Agriculture or Farming ML in Education	20AI603.1	L1
2	**What is a Decision Tree?** A decision tree is a type of supervised machine learning used to categorize or make predictions based on how a previous set of questions were answered. The model is a form of supervised learning, meaning that the model is trained and tested on a set of data that contains the desired categorization.	20AI603.2	L1

3	**Define a Dendrogram.**	20AI603.3	L1
	A *dendrogram* is a diagram that shows the hierarchical relationship between objects. It is most commonly created as an output from *hierarchical clustering*. The main use of a dendrogram is to work out the best way to allocate objects to clusters. The dendrogram below shows the hierarchical clustering of six *observations* shown on the *scatterplot* to the left.		

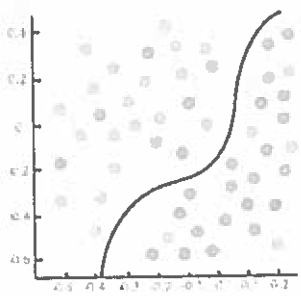
4	**Explain Posteriori Probability.**	**20AI603.4**	**L2**
	A posterior probability, in Bayesian statistics, is the revised or updated probability of an event occurring after taking into consideration new information. The posterior probability is calculated by updating the prior probability using Bayes' theorem. In statistical terms, the posterior probability is the probability of event A occurring given that event B has occurred.		

5	**Define Reinforcement Learning.**	**20AI603.5**	**L1**
	Reinforcement learning is an area of Machine Learning. It is about taking suitable action to maximize reward in a particular situation. It is employed by various software and machines to find the best possible behavior or path it should take in a specific situation. Reinforcement learning differs from supervised learning in a way that in supervised learning the training data has the answer key with it so the model is trained with the correct answer itself whereas in reinforcement learning, there is no answer but the reinforcement agent decides what to do to perform the given task. In the absence of a training dataset, it is bound to learn from its experience.		

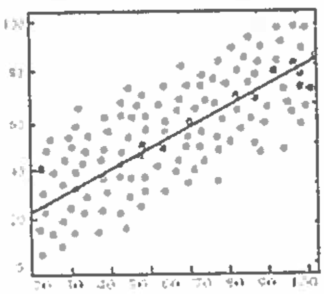
Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Learning Outcome (s)	DoK
6 (a)	**Differentiate in between Classification and Regression.** **For writing 6differences – 6 marks**	**6M**	**20AI603.1**

The Difference Between Regression vs. Classification



Classification



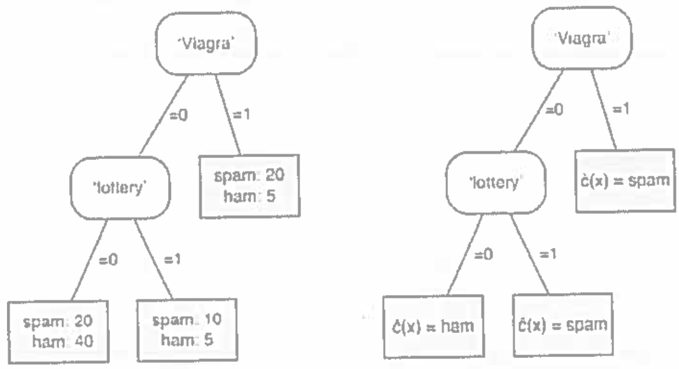
Regression

Regression Algorithms	Classification Algorithms
The output variable must be either continuous nature or real value.	The output variable has to be a discrete value.
The regression algorithm's task is mapping input value (x) with continuous output variable (y).	The classification algorithm's task mapping the input value of x with the discrete output variable of y.
They are used with continuous data.	They are used with discrete data.
It attempt to find the best fit line, which predicts the output more accurately.	Classification tries to find the decision boundary, which divides the dataset into different classes.
Regression algorithms solve regression problems such as house price predictions and weather predictions.	Classification algorithms solve classification problems like identifying spam e-mails, spotting cancer cells, and speech recognition.
We can further divide Regression algorithms into Linear and Non-linear Regression.	We can further divide Classification algorithms into Binary Classifiers and Multi-class Classifiers.

6 (b)	**Explain about binary classification.** **For explaning binary classification – 4 marks** **For explaining accuracy – 2marks**	**6M**	**20AI603.1**

Classification is the most common task in machine learning. A classifier is a mapping $\hat{c} : X \rightarrow C$, where C = {C1,C2,...,Ck } is a finite and usually small set of class labels. We will sometimes also use Ci to indicate the set of examples of that class. We use the 'hat' to indicate that $\hat{c}(x)$ is an estimate of the true but unknown function c(x). Examples for a classifier take the form (x,c(x)), where $x \in X$ is an instance and c(x) is the true class of the instance.

In the simplest case we have only two classes which are usually referred to as pos itive and negative, ⊕and ⊖ , or +1 and −1. Two-class classification is often called bi nary classification.



Spam e-mail filtering is a good example of binary classification, in which spam is conventionally taken as the positive class, and ham as the negative class (clearly, positive here doesn't mean 'good'!). Other examples of binary classification include medical diagnosis (the positive class here is having a particular disease) and credit card fraud detection.

Assessing classification performance :

The performance of such classifiers can be summarised by means of a table known as a contingency table or confusion matrix (Table 2.2 (left)). In this table, each row refers to actual classes as recorded in the test set, and each column to classes as predicted by the classifier. So, for instance, the first row states that the test set contains 50 positives, 30 of which were correctly predicted and 20 incorrectly.

The last column and the last row give the marginals (i.e., column and row sums). Marginals are important because they allow us to assess statistical significance. For instance, the contingency table in Table 2.2 (right) has the same marginals, but the classifier clearly makes a random choice as to which predictions are positive and which are negative – as a result the distribution of actual positives and negatives in either predicted class is the same as the overall distribution (uniform in this case).

	Predicted ⊕	Predicted ⊖	
Actual ⊕	30	20	50
Actual ⊖	10	40	50
	40	60	100

⊕	20	30	50
	20	30	50
	40	60	100

OR

7 (a)	**How to classify the data with multiple class labels? Explain in detail** **One vs one -- 3marks** **One vs rest – 3marks**	**6M**	**20AI603.1**

Multi-class classification Classification tasks with more than two classes are very common. For instance, once a patient has been diagnosed as suffering from a rheumatic disease, the doctor will want to classify him or her further into one of several variants. If we have k classes, performance of a classifier can be assessed using a k-by-k contingency table. Assessing performance is easy if we are interested in the classifier's accuracy, which is still the sum of the descending diagonal of the contingency table, divided by the number of test instances.

Imagine now that we want to construct a multi-class classifier, but we only have the ability to train two-class models – say linear classifiers. There are various ways to combine several of them into a single k-class classifier. The one-versus-rest scheme is to train k binary classifiers, the first of which separates class C1 from C2,...,Cn, the second of which separates C2 from all other classes, and so on. When training the i th classifier we treat all instances of class Ci as positive examples, and the remaining instances as negative examples. Sometimes the classes are learned in a fixed order, in which case we learn k −1 models, the i-th one separating Ci from Ci+1,...,Cn with $1 \le i < n$. An alternative to one-versus-rest is one-versus-one. In this scheme, we train k(k −1)/2 binary classifiers, one for each pair of different classes. If a binary classifier treats the classes asymmetrically, as happens with certain models, it makes more sense to train two classifiers for each pair, leading to a total of k(k −1) classifiers.

A convenient way to describe all these and other schemes to decompose a k-class task into l binary classification tasks is by means of a so-called output code matrix. This is a k-by-l matrix whose entries are +1, 0 or −1. The following are output codes describing the two ways to transform a three-class task by means of one-versus-one:

$$\begin{pmatrix} +1 & +1 & 0 \\ -1 & 0 & +1 \\ 0 & -1 & -1 \end{pmatrix} \qquad \begin{pmatrix} +1 & -1 & +1 & -1 & 0 & 0 \\ -1 & +1 & 0 & 0 & +1 & -1 \\ 0 & 0 & -1 & +1 & -1 & +1 \end{pmatrix}$$

So, in the symmetric scheme on the left, we train three clas sifiers: one to distinguish between C1 (positive) and C2 (negative), one to distinguish betweenC1 (positive) andC3 (negative), and the remaining one to distinguish between C2 (positive) and C3 (negative). The asymmetric scheme on the right learns three more classifiers with the roles of positives and negatives swapped. The code matrices for the unordered and ordered version of the one-versus-rest scheme are as follows:

$$\begin{pmatrix} +1 & -1 & -1 \\ -1 & +1 & -1 \\ -1 & -1 & +1 \end{pmatrix} \qquad \begin{pmatrix} +1 & 0 \\ -1 & +1 \\ -1 & -1 \end{pmatrix}$$

7 (b)	**With example explain how the Concept Learning task determines the Hypothesis for given target concept.** **For explaining about concept learning -4 Marks** **For example – 2 Marks**	**6M**	**20AI603.1**

In concept learning we only learn a description for the positive class, and label everything that doesn't satisfy that description as negative.

The hypothesis space:

The simplest concept learning setting is where we restrict the logical expressions de scribing concepts to conjunctions of literals.

Example:

Example 4.1 (Learning conjunctive concepts).

Suppose you come across a number of sea animals that you suspect belong to the same species. You observe their length in metres, whether they have gills, whether they have a prominent beak, and whether they have few or many teeth.

Using these features, the first animal can described by the following conjunction:

Length = 3 ∧ Gills = no ∧ Beak = yes ∧ Teeth = many

The next one has the same characteristics but is a metre longer, so you drop the length condition and generalise the conjunction to

Gills = no ∧ Beak = yes ∧ Teeth = many

The third animal is again 3 metres long, has a beak, no gills and few teeth, so your description becomes

Gills = no ∧ Beak = yes

All remaining animals satisfy this conjunction, and you finally decide they are some kind of dolphin

For every concept between the least general one and one of the most general ones is also a possible hypothesis, i.e., covers all the positives and none of the negatives. Mathematically speaking we say that the set of hypotheses that agree with the data is a convex set.

A concept is complete if it covers all positive examples. A concept is consistent if it covers none of the negative examples. The version space is the set of all complete and consistent concepts. This set is convex and is fully defined by its least and most general elements.

Algorithm 4.3: LGG-Conj-ID(x, y) – find least general conjunctive generalisation of two conjunctions, employing internal disjunction.

Input : conjunctions x, y.
Output : conjunction z.

$z \leftarrow$ true;
for each feature f **do**
  if $f = v_x$ is a conjunct in x and $f = v_y$ is a conjunct in y **then**
    add f = Combine-ID(v_x, v_y) to z ; // Combine-ID: see text
  end
end
return z

8 (a)	**Explain about decision tree based learning? How it is represented. Give some problems for which decision tree learning is appropriate.** **For explaining about decision trees – 3marks** **For exaplining about 3 function – 3marks**	20AI603.2	L2

Decision trees As already indicated, for a classification task we can simply define a set of instances D to be homogenous if they are all from the same class, and the function Label(D) will then obviously return that class. Notice that we may be calling Label(D) with a non-homogeneous set of instances in case one of the Di is empty, so the general definition of Label(D) is that it returns the majority class of the instances in D. 2 This leaves us to decide how to define the function BestSplit(D,F).

Minority class $\min(\dot{p}, 1-\dot{p})$ – this is sometimes referred to as the error rate, as it measures the proportion of misclassified examples if the leaf was labelled with the majority class; the purer the set of examples, the fewer errors this will make. This impurity measure can equivalently be written as $1/2 - |\dot{p} - 1/2|$.

Gini index $2\dot{p}(1-\dot{p})$ – this is the expected error if we label examples in the leaf randomly: positive with probability $\dot{p}$ and negative with probability $1-\dot{p}$. The probability of a false positive is then $\dot{p}(1-\dot{p})$ and the probability of a false negative $(1-\dot{p})\dot{p}$.[3]

entropy $-\dot{p}\log_2\dot{p} - (1-\dot{p})\log_2(1-\dot{p})$ – this is the expected information, in bits, conveyed by somebody telling you the class of a randomly drawn example; the purer the set of examples, the more predictable this message becomes and the smaller the expected information.

8 (b)	**Discuss in detail about Learning Ordered Rule Lists.**	**6M**	**20AI603.2**

Learning ordered rule lists The key idea of this kind of rule learning algorithm is to keep growing a conjunctive rule body by adding the literal that most improves its homogeneity. That is, we construct a downward path through the hypothesis space.

Algorithm 6.1: LearnRuleList(D) – learn an ordered list of rules.

Input : labelled training data D.
Output : rule list R.

$R \leftarrow \emptyset$;
while $D \neq \emptyset$ **do**
    $r \leftarrow$ LearnRule(D) ; // LearnRule: see Algorithm 6.2
    append r to the end of R;
    $D \leftarrow D \setminus \{x \in D \mid x \text{ is covered by } r\}$;
end
return R

Algorithm 6.1 specifies the separate-and-conquer rule learning strategy in more detail. While there are still training examples left, the algorithm learns another rule and removes all examples covered by the rule from the data set. This algorithm, which is the basis for the majority of rule learning systems, is also called the covering algo rithm. The algorithm for learning a single rule is given in Algorithm 6.2. Similar to decision trees, it uses the functions Homogeneous(D) and Label(D) to decide whether further specialisation is needed and what class to put in the head of the rule, respec tively. It also employs a function BestLiteral(D,L) that selects the best literal to add to the rule from the candidates in L given data D; in our example above, this literal would be selected on purity.

Algorithm 6.2: LearnRule(D) – learn a single rule.

Input : labelled training data D.
Output : rule r.

$b \leftarrow$ true;
$L \leftarrow$ set of available literals;
while not Homogeneous(D) **do**
  $l \leftarrow$ BestLiteral(D, L) ; // e.g., highest purity; see text
  $b \leftarrow b \wedge l$;
  $D \leftarrow \{x \in D \mid x$ is covered by $b\}$;
  $L \leftarrow L \setminus \{l' \in L \mid l'$ uses same feature as $l\}$;
end
$C \leftarrow$ Label(D) ; // e.g., majority class
$r \leftarrow$ ·if b then Class $= C$·;
return r

9 (a)	**Explain Tree Learning as variance reduction.** **regression trees – 3 marks** **clustering trees – 3 marks**	**6M**	**20AI603.2**

Tree learning as variance reduction:
We will now consider how to adapt decision trees to regression and clustering tasks.

Regression trees

In regression problems the target variable is continuous rather than binary, and in that case we can define the variance of a set Y of target values as the average squared distance from the mean:

$$\mathrm{Var}(Y) = \frac{1}{|Y|}\sum_{y\in Y}(y-\overline{y})^2$$

where $\overline{y} = \frac{1}{|Y|}\sum_{y\in Y} y$ is the mean of the target values in Y; see Background 5.1 for some useful properties of variance. If a split partitions the set of target values Y into mutually exclusive sets $\{Y_1, \ldots, Y_l\}$, the weighted average variance is then

$$\mathrm{Var}(\{Y_1,\ldots,Y_l\}) = \sum_{j=1}^{l}\frac{|Y_j|}{|Y|}\mathrm{Var}(Y_j) = \sum_{j=1}^{l}\frac{|Y_j|}{|Y|}\left(\frac{1}{|Y_j|}\sum_{y\in Y_j} y^2 - \overline{y}_j^2\right) = \frac{1}{|Y|}\sum_{y\in Y} y^2 - \sum_{j=1}^{l}\frac{|Y_j|}{|Y|}\overline{y}_j^2$$

So, in order to obtain a regression tree learning algorithm, we replace the impurity measure Imp in Algorithm 5.2 with the function Var. Notice that 1 |Y | y∈Y y2 is con stant for a given set Y , and so minimising variance over all possible splits of a given parent is the same as maximising the weighted average of squared means in the children.

Clustering trees :

The simple kind of regression tree considered here also suggests a way to learn cluster ing trees. This is perhaps surprising, since regression is a supervised learning problem while clustering is unsupervised. The key insight is that regression trees find instance space segments whose target values are tightly clustered around the mean value in the segment – indeed, the variance of a set of target values is simply the (univariate) average squared Euclidean distance to the mean. An immediate generalisation is to use a vector of target values, as this doesn't change the mathematics in an essential way. More generally yet, we can introduce an abstract function Dis : X ×X → R that measures the distance or dissimilarity of any two instances x,x ∈ X , such that the higher Dis(x,x) is, the less similar x and x are. The cluster dissimilarity of a set of instances D is then calculated as

$$\mathrm{Dis}(D) = \frac{1}{|D|^2} \sum_{x \in D} \sum_{x' \in D} \mathrm{Dis}(x, x') \tag{5.5}$$

The weighted average cluster dissimilarity over all children of a split then gives the *split dissimilarity*, which can be used to inform BestSplit(D, F) in the *GrowTree algorithm*

9 (b)	**Describe in detail about descriptive rule learning.** **for explaining about descriptive rule learning. – 3 marks** **for explaining about algorithm – 3 marks**	**6M**	**20AI603.2**

Descriptive rule learning:
When learning classification models it is natural to look for rules that identify pure subsets of the training examples: i.e., sets of examples that are all of the same class and that all satisfy the same conjunctive concept. However, as we have seen in Sec tion 3.3, sometimes we are less interested in predicting a class and more interested in finding interesting patterns. We defined subgroups as mappings gˆ : X → {true,false} – or alternatively, subsets of the instance space – that are learned from a set of labelled examples (xi ,l(xi)), where l : X → C is the true labelling function. A good subgroup is one whose class distribution is significantly different from the overall population. This is by definition true for pure subgroups, but these are not the only interesting ones. For instance, one could argue that the complement of a subgroup is as interesting as the subgroup itself: in our dolphin example, the concept Gills = yes, which covers four neg atives and no positives, could be considered as interesting as its complementGills = no, which covers one negative and all positives. This means that we need to move away from impurity-based evaluation measures.

Algorithm 6.5: WeightedCovering(D) – learn overlapping rules by weighting examples.

```
Input  : labelled training data D with instance weights initialised to 1.
Output : rule list R.
R ← ∅;
while some examples in D have weight 1 do
    r ← LearnRule(D) ;                    // LearnRule: see Algorithm 6.2
    append r to the end of R;
    decrease the weights of examples covered by r;
end
return R
```

Algorithm 6.6: FrequentItems(D, f_0) – find all maximal item sets exceeding a given support threshold.

```
Input  : data D ⊆ X; support threshold f0.
Output : set of maximal frequent item sets M.
M ← ∅;
initialise priority queue Q to contain the empty item set;
while Q is not empty do
    I ← next item set deleted from front of Q;
    max ← true ;                        // flag to indicate whether I is maximal
    for each possible extension I' of I do
        if Supp(I') ≥ f0 then
            max ← false ;               // frequent extension found, so I is not maximal
            add I' to back of Q;
        end
    end
    if max = true then M ← M ∪ {I};
end
return M
```

10 (a)	**Describe the procedure used by Least Square Methods for predicting the target classes.**	**6M**	**20AI603.3**

The least-squares method We start by introducing a method that can be used to learn linear models for classifica tion and regression.

Recall that the regression problem is to learn a function estimator ˆf : X → R from examples (xi , f (xi)), where in this chapter we assume X = Rd .

The differences between the actual and estimated function values on the training examples are called residuals i = f (xi) − ˆf (xi).

The least-squares method, introduced by Carl Friedrich Gauss in the late eighteenth century, consists in finding ˆf such that n i=1 2 i is minimised.

The following example illustrates the method in the simple case of a single feature, which is called univariate regression.

Example 7.1 (Univariate linear regression). Suppose we want to investigate the relationship between people's height and weight. We collect n height and weight measurements $(h_i, w_i), 1 \le i \le n$. Univariate linear regression assumes a linear equation $w = a + bh$, with parameters a and b chosen such that the sum of squared residuals $\sum_{i=1}^{n}(w_i - (a + bh_i))^2$ is minimised. In order to find the parameters we take partial derivatives of this expression, set the partial derivatives to 0 and solve for a and b:

$$\frac{\partial}{\partial a}\sum_{i=1}^{n}(w_i - (a+bh_i))^2 = -2\sum_{i=1}^{n}(w_i - (a+bh_i)) = 0 \qquad \Rightarrow \hat{a} = \overline{w} - \hat{b}\overline{h}$$

$$\frac{\partial}{\partial b}\sum_{i=1}^{n}(w_i - (a+bh_i))^2 = -2\sum_{i=1}^{n}(w_i - (a+bh_i))h_i = 0$$

$$\Rightarrow \hat{b} = \frac{\sum_{i=1}^{n}(h_i - \overline{h})(w_i - \overline{w})}{\sum_{i=1}^{n}(h_i - \overline{h})^2}$$

univariate linear regression can be understood as consisting of two steps:

1. normalisation of the feature by dividing its values by the feature's variance;
2. calculating the covariance of the target variable and the normalised feature.

Another important point to note is that the sum of the residuals of the least-squares solution is zero:

$$\sum_{i=1}^{n}(y_i - (\hat{a} + \hat{b}x_i)) = n(\overline{y} - \hat{a} - \hat{b}\overline{x}) = 0$$

The result follows because $\hat{a} = \overline{y} - \hat{b}\overline{x}$, as derived in Example 7.1. While this property is intuitively appealing, it is worth keeping in mind that it also makes linear regression susceptible to *outliers*: points that are far removed from the regression line, often because of measurement errors.

Multivariate linear regression

In order to deal with an arbitrary number of features it will be useful to employ matrix notation (see Background 7.2). We can write univariate linear regression in matrix form as

$$\begin{pmatrix} y_1 \\ \vdots \\ y_n \end{pmatrix} = \begin{pmatrix} 1 \\ \vdots \\ 1 \end{pmatrix} a + \begin{pmatrix} x_1 \\ \vdots \\ x_n \end{pmatrix} b + \begin{pmatrix} \epsilon_1 \\ \vdots \\ \epsilon_n \end{pmatrix}$$

$$\mathbf{y} = \mathbf{a} + \mathbf{X}\mathbf{b} + \epsilon$$

In the second form of this equation, $\mathbf{y}$, $\mathbf{a}$, $\mathbf{X}$ and ϵ are n-vectors, and $\mathbf{b}$ is a scalar. In case of d features, all that changes is that $\mathbf{X}$ becomes an n-by-d matrix, and $\mathbf{b}$ becomes a d-vector of regression coefficients.

We can apply the by now familiar trick of using homogeneous coordinates to simplify these equations as follows:

$$\begin{pmatrix} y_1 \\ \vdots \\ y_n \end{pmatrix} = \begin{pmatrix} 1 & x_1 \\ \vdots & \vdots \\ 1 & x_n \end{pmatrix} \begin{pmatrix} a \\ b \end{pmatrix} + \begin{pmatrix} \epsilon_1 \\ \vdots \\ \epsilon_n \end{pmatrix}$$

$$\mathbf{y} = \mathbf{X}^{\circ}\mathbf{w} + \epsilon$$

Regularised regression We have just seen a situation in which least-squares regression can become unstable: i.e., highly dependent on the training data. Instability is a manifestation of a tendency to overfit. Regularisation is a general method to avoid such overfitting by applying additional constraints to the weight vector. A common approach is to make sure the weights are, on average, small in magnitude: this is referred to as shrinkage. To show how this can be achieved, we first write down the least-squares regression problem as an optimisation problem:

w∗ = argmin w (y−Xw) T(y−Xw)

In the general case, the least-squares classifier learns the decision boundary w·x = t with

w = (XTX) −1(Pos μ⊕ −Neg μ⊖)

10 (b)	**Explain about nearest neighbor classification.** **For explaining nearest neighbor classification. the --- 5 Marks** **With an any one example – 2 Marks**	20AI603.3	**L2**

Nearest-neighbour classification:

Consequently, 'training' this classifier requires nothing more than memorising the training data. This extremely simple classifier is known as the nearest neighbour classifier. Its decision regions are made up of the cells of a Voronoi tessela tion, with piecewise linear decision boundaries selected from the Voronoi boundaries.

From an algorithmic point of view, training the nearest-neighbour classifier is very fast, taking only O(n) time for storing n exemplars. The downside is that classifying a single instance also takes O(n) time, as the instance will need to be compared with every exemplar to determine which one is the nearest. It is possible to reduce clas sification time at the expense of increased training time by storing the exemplars in a more elaborate data structure, but this tends not to scale well to large numbers of features.

In its simplest form, the k-nearest neighbour classifier takes a vote between the $k \geq 1$ nearest exemplars of the instance to be classified, and predicts the majority class. We can easily turn this into a probability estimator by returning the normalised class counts as a probability distribution over classes.

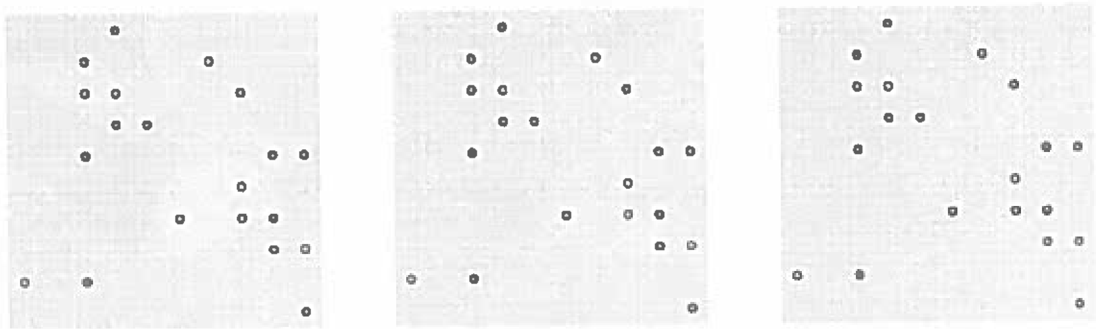


Figure 8.9. **(left)** Decision regions of a 3-nearest neighbour classifier; the shading represents the predicted probability distribution over the five classes. **(middle)** 5-nearest neighbour. **(right)** 7-nearest neighbour.

Figure 8.9 illustrates this on a small data set of 20 exemplars from five different classes, for $k = 3,5,7$. The class distribution is visualised by assigning each test point the class of a uniformly sampled neighbour: so, in a region where two of $k = 3$ neighbours are red and one is orange, the shading is a mix of two-thirds red and one-third orange. While for $k = 3$ the decision regions are still mostly discernible, this is much less so for $k = 5$ and $k = 7$. This may seem at odds with our earlier demonstration of the increase in the number of decision regions with increasing k in Example 8.2. However, this increase is countered by the fact that the probability vectors become more similar to each other. To take an extreme example: if k is equal to the number of exemplars n, every test instance will have the same number of neighbours and will receive the same probability vector which is equal to the prior distribution over the exemplars. If

If k-nearest neighbour is used for regression problems, the obvious way to aggre gate the predictions from the k neighbours is by taking the mean value, which can again be distance-weighted. This would lend the model additional predictive power by predicting values that aren't observed among the stored exemplars. More generally, we can apply k-means to any learning problem where we have an appropriate 'aggre gator' for multiple target values.

11 (a)	**Explain SVM Algorithm and its Kernel methods.**	6M	20AI603.3

Support vector machines Linearly separable data admits infinitely many decision boundaries that in early separable data admits infinitely many decision boundaries that separate the classes, but

intuitively some of these are better than others. For example, the left and middle decision boundaries in Figure 7.5 seem to be unnecessarily close to some of the positives; while the one on the right leaves a bit more space on either side, it doesn't seem particularly good either.
To make this a bit more precise, recall that in Section 2.2 we defined the ▯ margin of an example assigned by a scoring classifier as c(x)sˆ(x), where c(x) is +1 for positive examples and −1 for negative examples and sˆ(x) is the score of example x.

If we take sˆ(x) = w· x − t, then a true positive xi has margin w · xi − t > 0 and a true negative xj has margin −(w · xj − t) > 0. For a given training set and decision boundary, let m⊕ be the smallest margin of any positive, and m▯ the smallest margin of any negative, then we want the sum of these to be as large as possible.

This sum is independent of the decision threshold t, as long as we keep the nearest positives and negatives at the right sides of the decision boundary, and so we re-adjust t such that m⊕ and m▯ become equal.
Figure 7.7 depicts this graphically in a two dimensional instance space.

The training examples nearest to the decision boundary are called support vectors: as we shall see, the decision boundary of a support vector machine (SVM) is defined as a linear combination of the support vectors
The margin is thus defined as m/||w||, where m is the distance between the deci sion boundary and the nearest training instance measured along w. Since we are free to rescale t, ||w|| and m, it is customary to choose m = 1. Maximising the margin then corresponds to minimising ||w|| or, more con veniently, 1 2 ||w||2, provided of course that none of the training points fall inside the margin

Algorithm 7.3: PerceptronRegression(D, T) – train a perceptron for regression.

Input : labelled training data D in homogeneous coordinates;
maximum number of training epochs T.
Output : weight vector $\mathbf{w}$ defining function approximator $\hat{y} = \mathbf{w} \cdot \mathbf{x}$.

```
w ← 0; t ← 0;
while t < T do
    for i = 1 to |D| do
        w ← w + (y_i − ŷ_i)² x_i;
    end
    t ← t + 1;
end
```

Soft margin SVM If the data is not linearly separable, then the constraints w · xi − t ≥ 1 posed by the examples are not jointly satisfiable. However, there is a very elegant way of adapting the optimisation problem such that it admits a solution even in this case. The idea is to introduce slack variables ξi , one for each example, which allow some of them to be inside the margin or even at the wrong side of the decision boundary – we will call these margin errors. Thus, we change the constraints to w· xi − t ≥ 1−ξi and add the sum of all slack variables to the objective function to be minimised, resulting in the following soft margin optimisation problem:

$$\mathbf{w}^*, t^* = \underset{\mathbf{w}, t}{\arg\min} \frac{1}{2}||\mathbf{w}||^2$$

$$\text{subject to } y_i(\mathbf{w} \cdot \mathbf{x}_i - t) \geq 1 \quad \text{and} \quad , 1 \leq i \leq n$$

C is a user-defined parameter trading off margin maximisation against slack variable minimisation: a high value of C means that margin errors incur a high penalty, while a low value permits more margin errors (possibly including misclassifications) in or der to achieve a large margin. If we allow more margin errors we need fewer support vectors, hence C controls to some extent the 'complexity' of the SVM and hence is of ten referred to as the complexity parameter. It can be seen as a form of regularisation similar to that discussed in the context of least-squares regression.

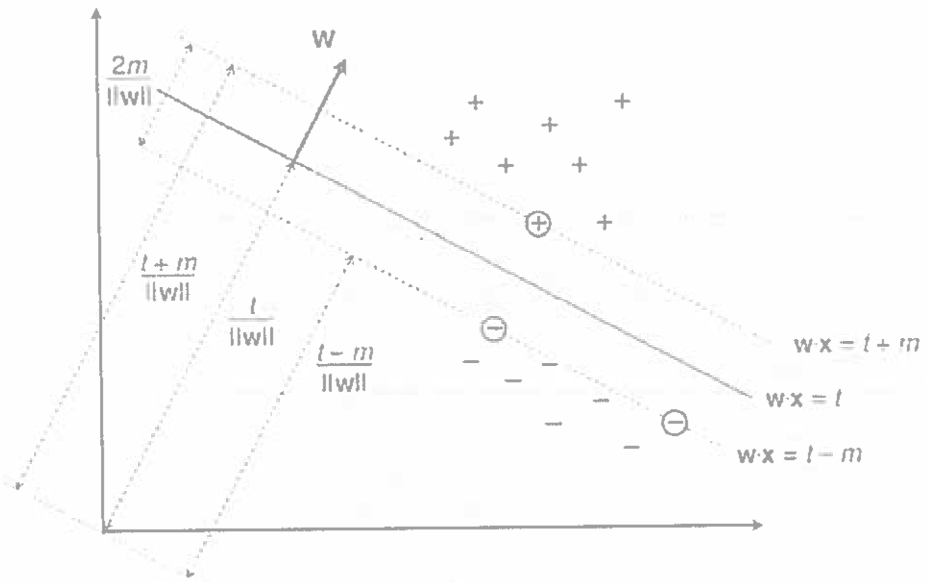


Figure 7.7. The geometry of a support vector classifier. The circled data points are the support vectors, which are the training examples nearest to the decision boundary. The support vector machine finds the decision boundary that maximises the margin $m/||w||$.

11 (b)	**Explain hierarchical clustering with an example.**	**6M**	**20AI603.3**

Hierarchical clustering The clustering methods discussed in the previous section use exemplars to represent a predictive clustering: a partition of the entire instance space. In this section we take a look at methods that represent clusters using trees. We previously encountered clustering trees in Section 5.3: those trees use features to navigate the instance space, similar to decision trees, and aren't distance-based as such. Here we consider trees called dendrograms, which are purely defined in terms of a distance measure. Because dendrograms use features only indirectly, as the basis on which the distance measure is calculated, they partition the given data rather than the entire instance space, and hence represent a descriptive clustering rather than a predictive one.

A precise definition of a dendrogram is as follows. Definition 8.4 (Dendrogram). Given a data set D, a dendrogram is a binary tree with the elements of D at its leaves. An internal node of the tree represents the subset of elements in the leaves of the subtree rooted at that node. The level of a node is the distance between the two clusters represented by the children of the node. Leaves have level 0. For this definition to work, we need a way to measure how close two clusters are. You might think that this is straightforward: just calculate the distance between the two cluster means. However, this occasionally leads to problems, as discussed later in this section. Furthermore, taking cluster means as exemplars assumes Euclidean distance, and we may want to use one of the other distance metrics discussed earlier. This has led to the introduction of the so-called linkage function, which is a general way to turn pairwise point distances into pairwise cluster distances

Definition 8.5 (Linkage function). *A linkage function $L: 2^{\mathscr{X}} \times 2^{\mathscr{X}} \to \mathbb{R}$ calculates the distance between arbitrary subsets of the instance space, given a distance metric* $\mathrm{Dis}: \mathscr{X} \times \mathscr{X} \to \mathbb{R}$.

The most common linkage functions are as follows:

Single linkage defines the distance between two clusters as the *smallest* pairwise distance between elements from each cluster.

Complete linkage defines the distance between two clusters as the *largest* pointwise distance.

Average linkage defines the cluster distance as the *average* pointwise distance.

Centroid linkage defines the cluster distance as the point distance between the cluster means.

These linkage functions can be defined mathematically as follows:

$$L_{\text{single}}(A,B) = \min_{x \in A, y \in B} \mathrm{Dis}(x,y)$$

$$L_{\text{complete}}(A,B) = \max_{x \in A, y \in B} \mathrm{Dis}(x,y)$$

$$L_{\text{average}}(A,B) = \frac{\sum_{x \in A, y \in B} \mathrm{Dis}(x,y)}{|A| \cdot |B|}$$

$$L_{\text{centroid}}(A,B) = \mathrm{Dis}\left(\frac{\sum_{x \in A} x}{|A|}, \frac{\sum_{y \in B} y}{|B|}\right)$$

Algorithm 8.4: HAC(D, L) – Hierarchical agglomerative clustering.

Input : data $D \subseteq \mathscr{X}$; linkage function $L: 2^{\mathscr{X}} \times 2^{\mathscr{X}} \to \mathbb{R}$ defined in terms of distance metric.

Output : a dendrogram representing a descriptive clustering of D.

initialise clusters to singleton data points;
create a leaf at level 0 for every singleton cluster;
repeat
    find the pair of clusters X, Y with lowest linkage l, and merge;
    create a parent of X, Y at level l;
until all data points are in one cluster;
return the constructed binary tree with linkage levels;

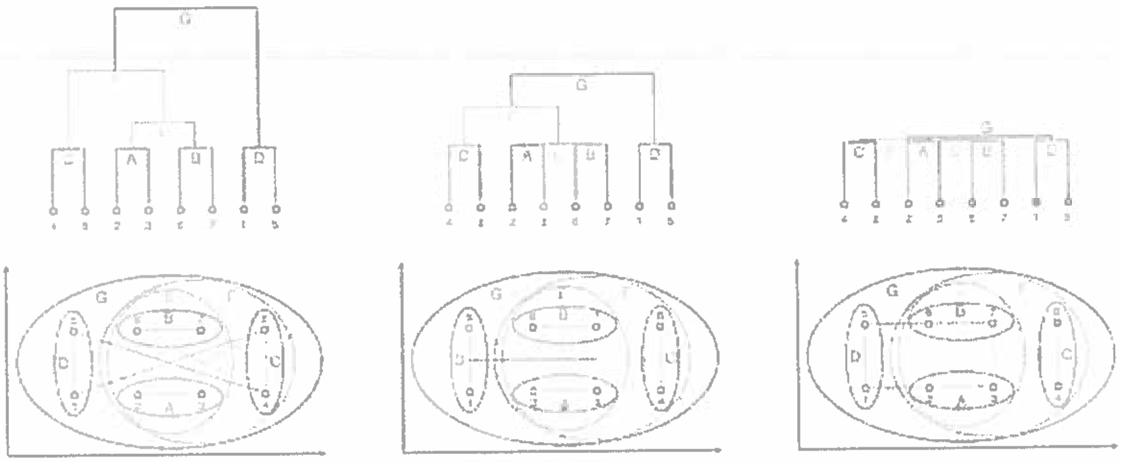


Figure 8.16. (left) Complete linkage defines cluster distance as the largest pairwise distance between elements from each cluster, indicated by the coloured lines between data points. The clustering found can be represented as nested partitions (bottom) or a dendrogram (top); the level of a horizontal connection between clusters in the dendrogram corresponds to the length of a linkage line. The example assumes that ties are broken by small irregularities in the grid. **(middle)** Centroid linkage defines the distance between clusters as the distance between their means. Notice that E obtains the same linkage as A and B, and so the latter clusters effectively disappear. **(right)** Single linkage defines the distance between clusters as the smallest pairwise distance. The dendrogram all but collapses, which means that no meaningful clusters are found in the given grid configuration.

12 a) **Demonstrate Normal or Gaussian distribution with an example ? -- 6 Marks**

Normal Distribution is an important concept in statistics and the backbone of Machine Learning.

As discovered by **Carl Friedrich Gauss, Normal Distribution/Gaussian Distribution** is a continuous probability distribution. It has a bell-shaped curve that is symmetrical from the mean point to both halves of the curve.

Mathematical Definition:

A continuous random variable "x" is said to follow a normal distribution with parameter μ(mean) and σ(standard deviation), if it's probability density function is given by,

$$y = \frac{1}{\sqrt{2\pi}} e^{-(x-\mu)^2/2\sigma}$$

μ = Mean
σ = Standard Deviation
$\pi \approx 3.14159$
$e \approx 2.71828$

We can draw a connection between probabilistic and geometric models by considering probability distributions defined over Euclidean spaces.

The most common such distributions are normal distributions, also called Gaussians; here recalls the most important facts concerning univariate and multivariate normal distributions.

We start by considering the univariate, two-class case. Suppose the values of x ∈ R follow a mixture model: i.e., each class has its own probability distribution (a component of the mixture model).

We will assume a Gaussian mixture model, which means that the components of the mixture are both Gaussians. We thus have

$$P(x|\oplus) = \frac{1}{\sqrt{2\pi}\sigma^{\oplus}} \exp\left(-\frac{1}{2}\left[\frac{x-\mu^{\oplus}}{\sigma^{\oplus}}\right]^2\right) \qquad P(x|\ ) = \frac{1}{\sqrt{2\pi}\sigma} \exp\left(-\frac{1}{2}\left[\frac{x-\mu}{\sigma}\right]^2\right)$$

where $\mu^{\oplus}$ and $\sigma^{\oplus}$ are the mean and standard deviation for the positive class, and μ and σ are the mean and standard deviation for the negative class. This gives the following likelihood ratio:

$$\mathrm{LR}(x) = \frac{P(x|\ominus)}{P(x|\ )} = \frac{\sigma}{\sigma} \exp\left(-\frac{1}{2}\left[\left(\frac{x-\mu^{\oplus}}{\sigma^{\oplus}}\right)^2 - \left(\frac{x-\mu}{\sigma}\right)^2\right]\right)$$

The univariate normal or Gaussian distribution has the following probability density function:

$$P(x|\mu,\sigma) = \frac{1}{\sqrt{2\pi}\sigma} \exp\left(-\frac{(x-\mu)^2}{2\sigma^2}\right) = \frac{1}{E}\exp\left(-\frac{1}{2}\left[\frac{x-\mu}{\sigma}\right]^2\right) = \frac{1}{E}\exp\left(-z^2/2\right), \quad E = \sqrt{2\pi}\sigma$$

The distribution has two parameters:

1) μ, which is the mean or expected value, as well as the median (i.e., the point where the area under the density function is split in half) and the mode (i.e., the point where the density function reaches its maximum); and

2) σ, which is the standard deviation and determines the width of the bell-shaped curve.

z = (x − μ)/σ is the z-score associated with x; it measures the number of standard deviations between x and the mean (it has itself mean 0 and standard deviation 1). It follows that P(x|μ,σ) = 1 σP(z|0, 1), where P(z|0, 1) denotes the standard normal distribution.

Note: In other words, any normal distribution can be obtained from the standard normal distribution by scaling the x-axis with a factor σ, scaling the y-axis with a factor 1/σ (so the area under the curve remains 1), and translating the origin over μ

The *multivariate normal distribution* over d-vectors $\mathbf{x} = (x_1, \ldots, x_d)^{\mathrm{T}} \in \mathbb{R}^d$ is

$$P(\mathbf{x}|\mu, \Sigma) = \frac{1}{E_d} \exp\left(-\frac{1}{2}(\mathbf{x}-\mu)^{\mathrm{T}}\Sigma^{-1}(\mathbf{x}-\mu)\right), \quad E_d = (2\pi)^{d/2}\sqrt{|\Sigma|}$$

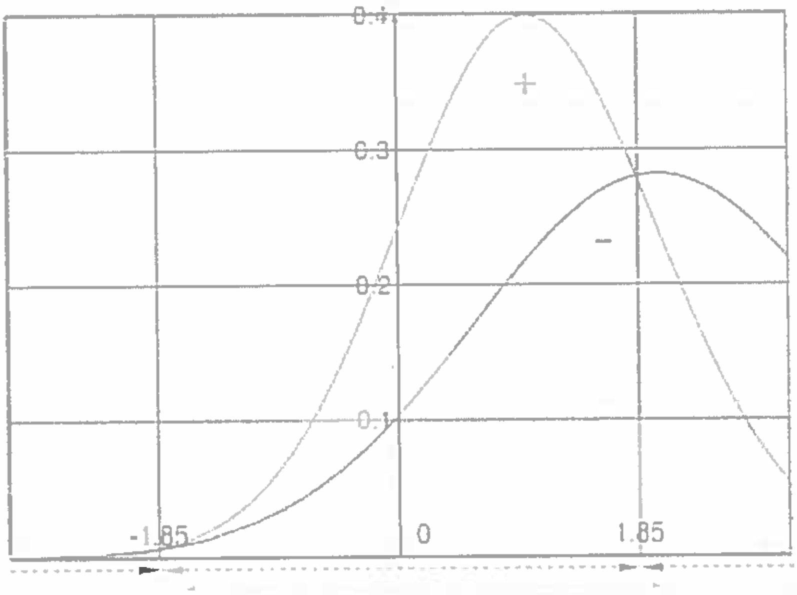
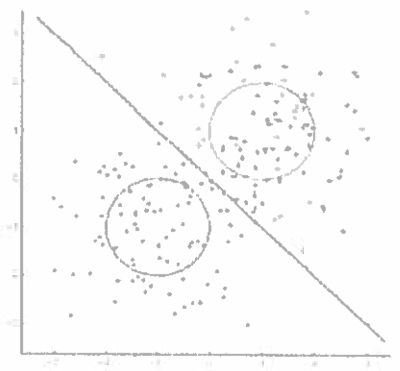


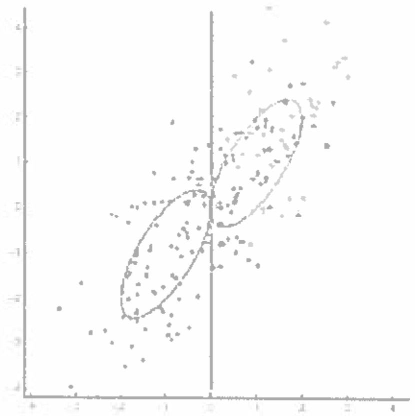
Figure 9.2. If positive examples are drawn from a Gaussian with mean and standard deviation 1 and negatives from a Gaussian with mean and standard deviation 2, then the two distributions cross at x = ±1.85. This means that the maximum-likelihood region for positives is the closed interval [−1.85, 1.85], and hence the negative region is non-contiguous.

If the features are uncorrelated and have the same variance,

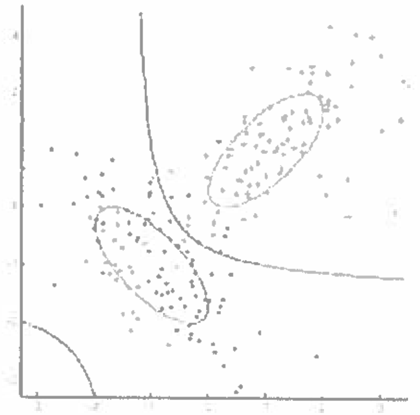
maximumlikelihood classification leads to the basic linear classifier,

whose decision boundary is orthogonal to the line connecting the means. (





As long as the per-class covariance matrices are identical, the Bayes-optimal decision boundary is linear – if we were to decorrelate the features by rotation and scaling, we would again obtain the basic linear classifier.



Unequal covariance matrices lead to hyperbolic decision boundaries, which means that one of the decision regions is non-contiguous.

Non-contiguous decision regions can also occur in higher-dimensional spaces.

Notice the circles and ellipses in Figure 9.3, which provide a visual summary of the covariance matrix. By projecting the shape for the positive class down to x-axis we obtain the interval $[\mu 1 \oplus -\sigma 1 \oplus, \mu 1 \oplus +\sigma 1 \oplus]$ – i.e., one standard deviation around the mean – and similar for the negative class and the y-axis.

Three cases can be distinguished:

(i) both x and y standard deviations are equal and the correlation coefficient is zero, in which case the shape is a circle;

(ii) the standard deviations are different and the correlation coefficient is zero, which means the shape is an ellipse parallel to the axis with the largest standard deviation;

(iii) the correlation coefficient is non-zero: the orientation of the ellipse gives the sign of the correlation coefficient, and its width varies with the magnitude of the correlation coefficient.

Note: for uncorrelated, unit-variance Gaussian features, the basic linear classifier is Bayes-optimal

the standard normal distribution translates Euclidean distances into probabilities:

$$P(\mathbf{x}|\mathbf{0},\mathbf{I}) = \frac{1}{(2\pi)^{d/2}} \exp\left(-\frac{1}{2}(\mathrm{Dis}_2(\mathbf{x},\mathbf{0}))^2\right)$$

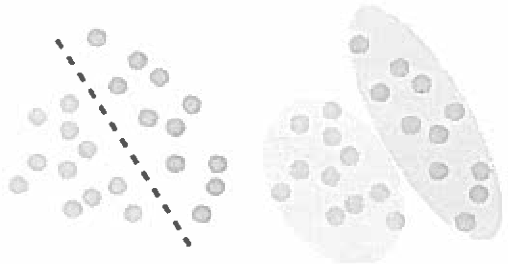
12(b) Explain how Discriminative Learning is used to perform pattern

For explaining the Discriminative Learning is used to perform pattern – 6 Marks

Discriminative Learning:

- Machine learning models can be classified into two types of models – Discriminative and Generative models.
- Discriminative models, also referred to as conditional models, are a class of logistical models used for classification or regression.
- Typical discriminative models include logistic regression (LR), conditional random fields (CRFs) (specified over an undirected graph), decision trees, and many others.
- In simple words, a discriminative model makes predictions on the unseen data based on conditional probability and can be used either for classification or regression problem statements.
- On the contrary, a generative model focuses on the distribution of a dataset to return a probability for a given example.
- Naive Bayes models are generative: after training they can be used to generate data.
- The most commonly used discriminative models: logistic regression.
- In generative models the decision boundary is a by-product of modelling the distributions of each class, logistic regression models the decision boundary directly.

Discriminative Generative



For example, if the classes are overlapping then logistic regression will tend to locate the decision boundary in an area where classes are maximally overlapping, regardless of the 'shapes' of the samples of each class. This results in decision boundaries that are noticeably different from those learned by generative classifiers

The logistic regression model is simply given by

$$\hat{p}(\mathbf{x}) = \frac{\exp(\mathbf{w}\cdot\mathbf{x} - t)}{\exp(\mathbf{w}\cdot\mathbf{x} - t) + 1} = \frac{1}{1 + \exp(-(\mathbf{w}\cdot\mathbf{x} - t))}$$

Assuming the class labels are y = 1 for positives and y = 0 for negatives, this defines a Bernoulli distribution for each training example:

$$P(y_i|\mathbf{x}_i) = \hat{p}(\mathbf{x}_i)^{y_i}(1 - \hat{p}(\mathbf{x}_i))^{(1-y_i)}$$

It is important to note that the parameters of these Bernoulli distributions are linked through w and t, and consequently there is one parameter for every feature dimension, rather than for every training instance.

The likelihood function is

$$\mathrm{CL}(\mathbf{w}, t) = \prod_i P(y_i|\mathbf{x}_i) = \prod_i \hat{p}(\mathbf{x}_i)^{y_i}(1 - \hat{p}(\mathbf{x}_i))^{(1-y_i)}$$

This is called conditional likelihood to stress that it gives us the conditional probability P(yi |xi) rather than P(xi) as in a generative model.
the logarithm of the likelihood function is easier to work with:

$$\mathrm{LCL}(\mathbf{w}, t) = \sum_i y_i \ln \hat{p}(\mathbf{x}_i) + (1 - y_i)\ln(1 - \hat{p}(\mathbf{x}_i)) = \sum_{\mathbf{x} \in Tr} \ln \hat{p}(\mathbf{x}) + \sum_{\mathbf{x} \in Tr} \ln(1 -$$

13 a) Explain Naïve Bayes Classifier with an example .
For Explaining Naïve Bayes Classifier -6 Marks

- A **Bernoulli trial** is an experiment that results in two outcomes: ***success*** and ***failure.***
- One example of a Bernoulli trial is the coin tossing experiment, which results in heads or tails.
- In a Bernoulli trial we define the probability of success and probability of failure as follows:

$$P[\text{success}] = p \qquad 0 \leq p \leq 1$$
$$P[\text{failure}] = 1 - p$$

- The **Binomial Distribution** arises when counting the number of successes S in n independent Bernoulli trials with the same parameter θ. It is described by

$$P(S = s) = \binom{n}{s}\theta^s(1-\theta)^{n-s} \text{ for } s \in \{0, \ldots, n\}$$

- The **categorical distribution** generalises the Bernoulli distribution to k ≥ 2 outcomes. The parameter of the distribution is a k-vector $\theta = (\theta_1, \ldots, \theta_k)$ such that $\sum_{i=1}^{k} \theta_i = 1$.
- the **multinomial distribution** tabulates the outcomes of n independent and identically distributed (i.i.d.) categorical trials. That is, X = X1,...,Xk is a

k-vector of integer counts, and

$$P(\mathbf{X} = (x_1, \ldots, x_k)) = n! \frac{\theta_1^{x_1}}{x_1!} \cdots \frac{\theta_k^{x_k}}{x_k!}$$

Using A Naive Bayes Model For Classification :

- Assume that we have chosen one of the possible distributions to model our data X.
- In a classification context, we furthermore assume that the distribution depends on the class, so that P(X|Y = spam) and P(X|Y = ham) are different distributions.
- The more different these two distributions are, the more useful the features X are for classification.
- Thus, for a specific e-mail x we calculate both P(X = x|Y = spam) and P(X = x|Y = ham), and apply one of several possible decision rules:

$$\text{maximum likelihood (ML)} \qquad - \text{predict } \arg\max_y P(X = x|Y = y);$$

$$\text{maximum a posteriori (MAP)} \qquad - \text{predict } \arg\max_y P(X = x|Y = y)P(Y = y);$$

$$\text{recalibrated likelihood} \qquad - \text{predict } \arg\max_y w_y P(X = x|Y = y).$$

Working of Naïve Bayes' Classifier:

Working of Naïve Bayes' Classifier can be understood with the help of the below example:

Suppose we have a dataset of **weather conditions** and corresponding target variable **"Play"**. So using this dataset we need to decide that whether we should play or not on a particular day according to the weather conditions. So to solve this problem, we need to follow the below steps:

1. Convert the given dataset into frequency tables.
2. Generate Likelihood table by finding the probabilities of given features.
3. Now, use Bayes theorem to calculate the posterior probability.

Example 9.4 (Prediction using a naive Bayes model):

Suppose our vocabulary contains three words a, b and c, and we use a multivariate Bernoulli model for our e-mails, with parameters $\theta\oplus = (0.5, 0.67, 0.33)$ $\theta = (0.67, 0.33, 0.33)$

This means, for example, that the presence of b is twice as likely in spam (+), compared with ham.

The e-mail to be classified contains words a and b but not c, and hence is described by the bit vector x = (1, 1, 0).

We obtain likelihoods

$P(x|\oplus) = 0.5 \cdot 0.67 \cdot (1-0.33) = 0.222$ $P(x|) = 0.67 \cdot 0.33 \cdot (1-0.33) = 0.148$

The ML classification of x is thus spam.

In the case of two classes it is often convenient to work with likelihood ratios and odds.

The likelihood ratio can be calculated as:

$$\frac{P(\mathbf{x}|\oplus)}{P(\mathbf{x}|\ )} = \frac{0.5}{0.67}\frac{0.67}{0.33}\frac{1-0.33}{1-0.33} = 3/2 > 1.$$

This means that the MAP classification of x is also spam if the prior odds are more than 2/3, but ham if they are less than that.

Advantages of Naïve Bayes Classifier:

- Naïve Bayes is one of the fast and easy ML algorithms to predict a class of datasets.
- It can be used for Binary as well as Multi-class Classifications.
- It performs well in Multi-class predictions as compared to the other Algorithms.
- It is the most popular choice for **text classification problems**.

Applications of Naïve Bayes Classifier:

- It is used for **Credit Scoring.**
- It is used in **medical data classification.**
- It can be used in **real-time predictions** because Naïve Bayes Classifier is an eager learner.

13(b)

Demonstrate in detail about Compression based models.. ---- 6 Marks

For writing the components –3 Marks

For writing the script for restaurant – 3 Marks

Compression-based models We end this chapter with a brief discussion of an approach to machine learning that is both closely related to and quite distinct from the probabilistic approach. Consider the maximum a posteriori decision rule again:

yMAP = argmax y P(X = x|Y = y)P(Y = y)

Y	$P(\text{Viagra} = 1\|Y)$	$IC(\text{Viagra} = 1\|Y)$	$P(\text{Viagra} = 0\|Y)$	$IC(\text{Viagra} = 0\|Y)$
spam	0.40	**1.32 bits**	0.60	**0.74 bits**
ham	0.12	**3.06 bits**	0.88	**0.18 bits**

Table 9.2. Example marginal likelihoods.

Taking negative logarithms, we can turn this into an equivalent minimisation:

$$y_{\text{MAP}} = \arg\min_{y} -\log P(X = x|Y = y) - \log P(Y = y) \tag{9.16}$$

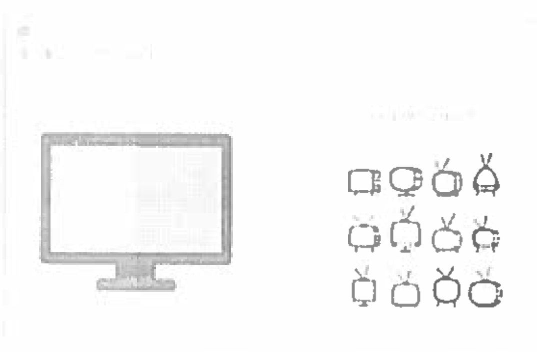
This follows because for any two probabilities $0 < p < p' < 1$ we have $\infty > -\log p > -\log p' > 0$. If an event has probability p of happening, the negative logarithm of p quantifies the *information content* of the message that the event has indeed happened. This makes intuitive sense, as the less expected an event is, the more information an announcement of the event contains. The unit of information depends on the base of the logarithm: it is customary to take logarithms to the base 2, in which case information is measured in bits. For example, if you toss a fair coin once and tell me it came up heads, this contains $-\log_2 1/2 = 1$ bit of information; if you roll a fair die once and let me know it came up six, the information content of your message is $-\log_2 1/6 = 2.6$ bits. Equation 9.16 tells us that the MAP decision rule chooses the least surprising or the most expected class for an instance x given particular prior distributions and likelihoods. We write $IC(X|Y) = -\log_2 P(X|Y)$ and $IC(Y) = -\log_2 P(Y)$.

14 a)	**Define Q learning? Explain with an example about Q-learning.** **About Q learning – 2 Marks** **For Writing Example -- 4 Marks**

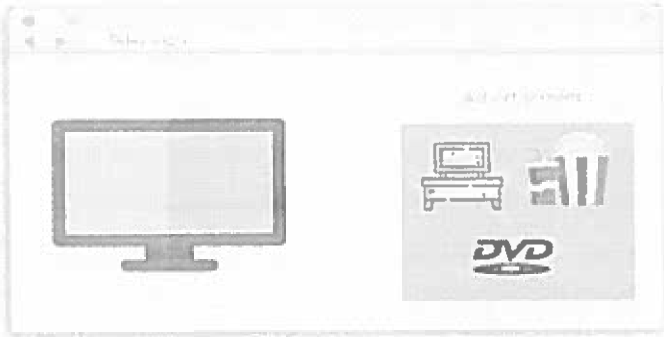
Q-Learning is a Reinforcement learning policy that will find the next best action, given a current state. It chooses this action at random and aims to maximize the reward.

- Q-learning is a model-free, off-policy reinforcement learning that will find the best course of action, given the current state of the agent.
- Depending on where the agent is in the environment, it will decide the next action to be taken.
- The objective of the model is to find the best course of action given its current state.
- To do this, it may come up with rules of its own or it may operate outside the policy given to it to follow.
- This means that there is no actual need for a policy, hence we call it off-policy.
- Model-free means that the agent uses predictions of the environment's expected response to move forward.
- It does not use the reward system to learn, but rather, trial and error.

An example of Q-learning is an Advertisement recommendation system. In a normal ad recommendation system, the ads you get are based on your previous purchases or websites you may have visited. If you've bought a TV, you will get recommended TVs of different brands.



Using Q-learning, we can optimize the ad recommendation system to recommend products that are frequently bought together. The reward will be if the user clicks on the suggested product.



Important Terms in Q-Learning

- **States:** The State, S, represents the current position of an agent in an environment.
- **Action:** The Action, A, is the step taken by the agent when it is in a particular state.
- **Rewards:** For every action, the agent will get a positive or negative reward.
- **Episodes:** When an agent ends up in a terminating state and can't take a new action.
- **Q-Values:** Used to determine how good an Action, A, taken at a particular state, S, is. Q (A, S).
- **Temporal Difference:** A formula used to find the Q-Value by using the value of current state and action and previous state and action.

The Bellman Equation is used to determine the value of a particular state and deduce how good it is to be in/take that state. The optimal state will give us the highest optimal value.

The equation is given below. It uses the current state, and the reward associated with that state, along with the maximum expected reward and a discount rate, which determines its importance to the current state, to find the next state of our agent. The learning rate determines how fast or slow, the model will be learning.

$$\text{New } Q(S,A) = \overbrace{Q(S,A)}^{\text{Current Q Value}} + \overbrace{\alpha}^{\text{Learning Rate}} [\overbrace{R(S,A)}^{\text{Reward}} + \underbrace{\gamma}_{\text{Discount Rate}} \underbrace{\text{Max } Q'(S',A')}_{\text{Maximum Expected Future Reward}} - Q(S,A)]$$

While running our algorithm, we will come across various solutions and the agent will take multiple paths. How do we find out the best among them? This is done by tabulating our findings in a table called a Q-Table.

A Q-Table helps us to find the best action for each state in the environment. We use the Bellman Equation at each state to get the expected future state and reward and save it in a table to compare with other states.

Lets us create a q-table for an agent that has to learn to run, fetch and sit on command. The steps taken to construct a q-table are :

Step 1: Create an initial Q Table with all values initialized to 0

When we initially start, the values of all states and rewards will be 0.

Consider the Q-Table shown below which shows a dog simulator learning to perform actions :

Action	Fetching	Sitting	Running
Start	0	0	0
Idle	0	0	0
Wrong Action	0	0	0
Correct Action	0	0	0
End	0	0	0

Step 2: Choose an action and perform it. Update values in the table

This is the starting point. We have performed no other action as of yet. Let us say that we want the agent to sit initially, which it does. The table will change to:

Action	Fetching	Sitting	Running
Start	0	1	0
Idle	0	0	0
Wrong Action	0	0	0
Correct Action	0	0	0
End	0	0	0

Step 3: Get the value of the reward and calculate the value Q-Value using Bellman Equation

For the action performed, we need to calculate the value of the actual reward and the Q(S, A) value

Action	Fetching	Sitting	Running
Start	0	1	0
Idle	0	0	0
Wrong Action	0	0	0
Correct Action	0	3.1	0
End	0	0	0

Figure 9: Updating Q-Table with Bellman Equation

Step 4: Continue the same until the table is filled or an episode ends

The agent continues taking actions and for each action. the reward and Q-value are calculated and it updates the table.

Action	Fetching	Sitting	Running
Start	5	7	10
Idle	2	5	3
Wrong Action	2	6	1
Correct Action	54	34	17
End	3	1	4

Figure 10: Final Q-Table at end of an episode

14 b) **DESCRIBE THE RANDOM FOREST ALGORITHM TO IMPROVE CLASSIFIER ACCURACY .**

For defining random forest – 1 Mark

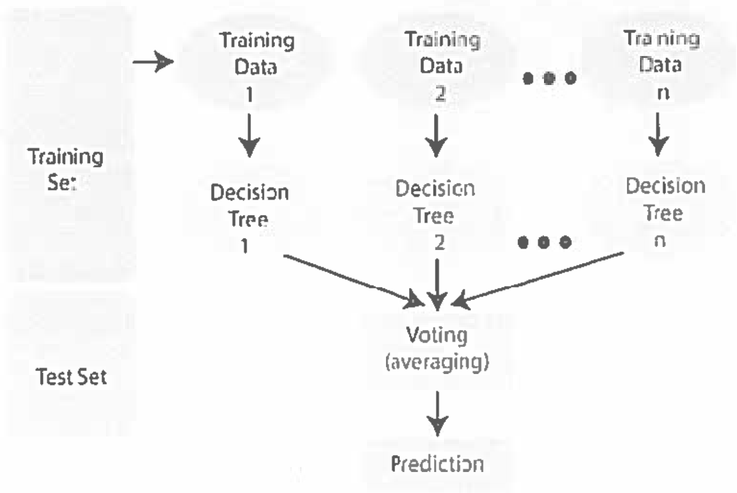
For Explaining algorithm – 5 Marks

Random Forest is a popular machine learning algorithm that belongs to the supervised learning technique. It can be used for both Classification and Regression problems in ML. It is based on the concept of **ensemble learning,** which is a process of *combining multiple classifiers to solve a complex problem and to improve the performance of the model.*

As the name suggests, ***"Random Forest is a classifier that contains a number of decision trees on various subsets of the given dataset and takes the average to improve the predictive accuracy of that dataset."*** Instead of relying on one decision tree, the random forest takes the prediction from each tree and based on the majority votes of predictions, and it predicts the final output.

The greater number of trees in the forest leads to higher accuracy and prevents the problem of overfitting.

The below diagram explains the working of the Random Forest algorithm:



Since the random forest combines multiple trees to predict the class of the dataset, it is possible that some decision trees may predict the correct output, while others may not. But together, all the trees predict the correct output. Therefore, below are two assumptions for a better Random forest classifier:

- There should be some actual values in the feature variable of the dataset so that the classifier can predict accurate results rather than a guessed result.
- The predictions from each tree must have very low correlations.

Random Forest works in two-phase first is to create the random forest by combining N decision tree, and second is to make predictions for each tree created in the first phase.
The Working process can be explained in the below steps and diagram:
Step-1: Select random K data points from the training set.
Step-2: Build the decision trees associated with the selected data points (Subsets).
Step-3: Choose the number N for decision trees that you want to build.
Step-4: Repeat Step 1 & 2.
Step-5: For new data points, find the predictions of each decision tree, and assign the new data points to the category that wins the majority votes.

(OR)

15 a) **Compare and Contrast Bagging and Boosting ensemble techniques .**

For Writing at least 6 differences ---- 6 Marks

Bagging Vs Boosting

We all use the Decision Tree Technique on day to day life to make the decision. Organizations use these supervised machine learning techniques like Decision trees to make a better decision and to generate more surplus and profit.

Ensemble methods combine different decision trees to deliver better predictive results, afterward utilizing a single decision tree. The primary principle behind the ensemble model is that a group of weak learners come together to form an active learner.

There are two techniques given below that are used to perform ensemble decision tree.

Bagging

Bagging is used when our objective is to reduce the variance of a decision tree. Here the concept is to create a few subsets of data from the training sample, which is chosen randomly with replacement. Now each collection of subset data is used to prepare their decision trees thus, we end up with an ensemble of various models. The average of all the assumptions from numerous tress is used, which is more powerful than a single decision tree.

Random Forest is an expansion over bagging. It takes one additional step to predict a random subset of data. It also makes the random selection of features rather than using all features to develop trees. When we have numerous random trees, it is called the Random

Forest.

These are the following steps which are taken to implement a Random forest:

- Let us consider **X** observations **Y** features in the training data set. First, a model from the training data set is taken randomly with substitution.
- The tree is developed to the largest.
- The given steps are repeated, and prediction is given, which is based on the collection of predictions from n number of trees.

Advantages of using Random Forest technique:

- It manages a higher dimension data set very well.
- It manages missing quantities and keeps accuracy for missing data.

Disadvantages of using Random Forest technique:

Since the last prediction depends on the mean predictions from subset trees, it won't give precise value for the regression model.

Boosting:

Boosting is another ensemble procedure to make a collection of predictors. In other words, we fit consecutive trees, usually random samples, and at each step, the objective is to solve net error from the prior trees.

If a given input is misclassified by theory, then its weight is increased so that the upcoming hypothesis is more likely to classify it correctly by consolidating the entire set at last converts weak learners into better performing models.

Gradient Boosting is an expansion of the boosting procedure.

Bagging	Boosting
Various training data subsets are randomly drawn with replacement from the whole training dataset.	Each new subset contains the components that were misclassified by previous models.
Bagging attempts to tackle the over-fitting issue.	Boosting tries to reduce bias.
If the classifier is unstable (high variance), then we need to apply bagging.	If the classifier is steady and straightforward (high bias), then we need to apply boosting.
Every model receives an equal weight.	Models are weighted by their performance.

Objective to decrease variance, not bias.	Objective to decrease bias, not variance.
It is the easiest way of connecting predictions that belong to the same type.	It is a way of connecting predictions that belong to the different types.
Every model is constructed independently.	New models are affected by the performance of the previously developed model.

15 b) **Explain in detail the main components of reinforcement learning. Is RNN reinforcement learning algorithm**

About main components – 3 Marks

RNN algorithm – 3 Marks

a reinforcement learning model has four essential components: a policy, a reward, a value function, and an environment model.

- A policy determines how an agent behaves at a specific point in time. In broad terms, it is a mapping between environmental conditions to actions, to the activities that the agent conducts in the environment. The policy might be as basic as a function or as sophisticated as function calculations. The policy is at the heart of everything the agent discovers.

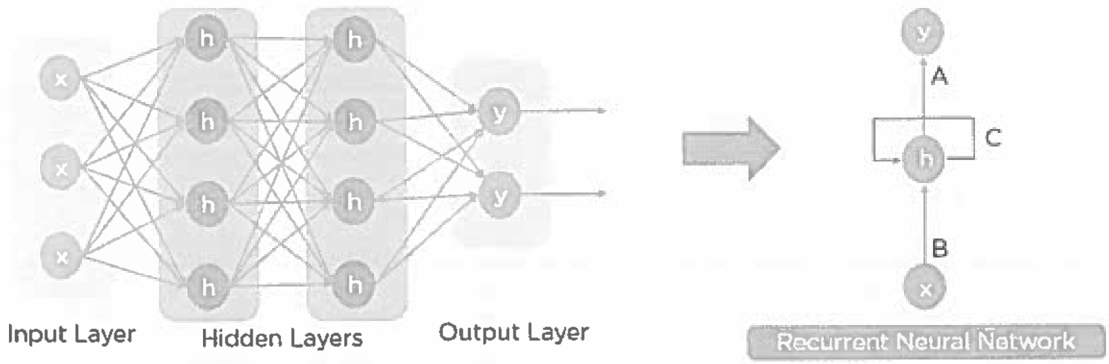
- A reward defines the goal of an RL issue. At each time step, the agent's behaviors result in a reward. The ultimate purpose of the agent is to optimize the overall reward earned. As a consequence, the reward differentiates between positive and bad action outcomes for the agent. In a natural system, we may attribute rewards and punishments as delightful and unpleasant experiences.

- A state's value is the total accumulated quantity of prizes that the agent may anticipate receiving in the future if it begins in that condition. Values represent the long-term attractiveness of a collection of states based on the expected future states and the benefits produced by those states. Even though a state produces a modest immediate reward, it might still be valuable since it is frequently followed by additional states that produce bigger benefits.

- The environment model is another significant component of several reinforcement learning systems. This is a mechanism that mimics environmental behavior and enables predictions about how the environment will respond. This model will let the agent forecast the next reward if an action is taken, enabling the agent to rely on developing action on future environmental responses.

RNN works on the principle of saving the output of a particular layer and feeding this back to the input in order to predict the output of the layer.

Below is how you can convert a Feed-Forward Neural Network into a Recurrent Neural Network:



The nodes in different layers of the neural network are compressed to form a single layer of recurrent neural networks. A, B, and C are the parameters of the network.

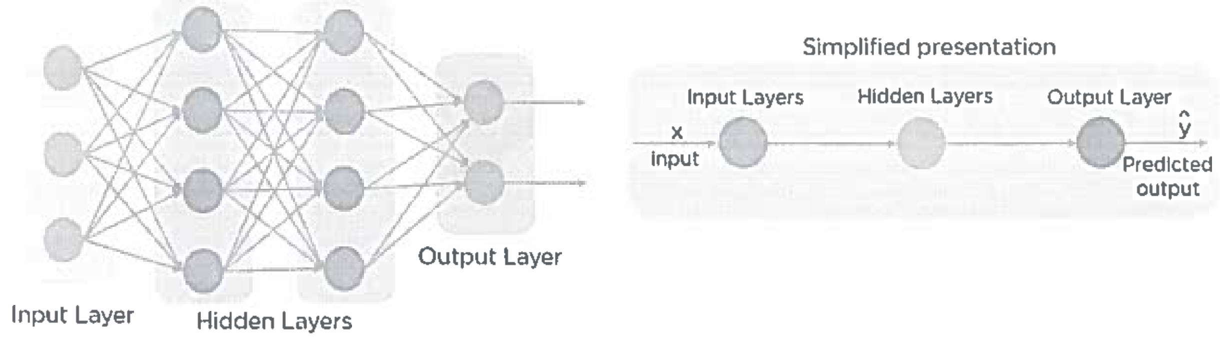
RNN were created because there were a few issues in the feed-forward neural network:

- Cannot handle sequential data
- Considers only the current input
- Cannot memorize previous inputs

The solution to these issues is the RNN. An RNN can handle sequential data, accepting the current input data, and previously received inputs. RNNs can memorize previous inputs due to their internal memory.

A feed-forward neural network allows information to flow only in the forward direction, from the input nodes, through the hidden layers, and to the output nodes. There are no cycles or loops in the network.

Below is how a simplified presentation of a feed-forward neural network looks like:



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3/12/2024

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Civil Engineering	Academic Year	2022 - 2023
Course Code	20CE005	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Construction Equipment Automation				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the unique features of construction equipment?	20CE005.1	L1
2	What is the need of construction management?	20CE005.2	L1
3	List any four selection factors for rear dump trucks.	20CE005.3	L1
4	List any two advantages of automation in concrete technology.	20CE005.4	L1
5	List any two benefits of robots in construction industry.	20CE005.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Classify the construction equipments on different basis with suitable examples.	12 M	20CE005.1	L2
	OR			
7 (a)	Discuss the applications of grouting equipment.	6M	20CE005.1	L2
7 (b)	Explain about plastering machines in construction industry.	6M	20CE005.1	L2
8 (a)	Discuss objectives of construction management.	6M	20CE005.2	L2
8 (b)	What are the different specifications to be followed while ordering construction equipment?	6M	20CE005.2	L2
	OR			
9 (a)	What are the advantages and disadvantages of mechanization in construction?	8M	20CE005.2	L2
9 (b)	Explain about forward planning.	4M	20CE005.2	L2
10 (a)	Explain the steps to be followed while transportation of concrete mix.	6M	20CE005.3	L2
10 (b)	Explain about construction compaction equipment.	6M	20CE005.3	L2
	OR			
11	Explain the concrete production steps in ready mix concrete plant.	12M	20CE005.3	L2
12	Explain the adoption of photogrammetry in construction industry with the help of drones.	12M	20CE005.4	L2
	OR			
13	Explain about structural health monitoring.	12M	20CE005.4	L2
14	Explain automation in production of steel components.	12M	20CE005.5	L2
	OR			
15	List various applications of automation in timber construction.	12M	20CE005.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Degree	B. Tech.	Program	Civil Engineering			Academic Year	2022 - 2023
Course Code	20CE005	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Construction Equipment Automation						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the unique features of construction equipment?	20CE005.1	L1

Ans: The selection of the appropriate type and size of construction equipments often affects the required amount of time and effort and thus the job-site productivity of a project. It is therefore important for site managers and construction planners to be familiar with the characteristics of the major types of equipment most commonly used in construction.

The activities involved in Construction Projects where the magnitude of the work is on a large scale, Speedy Work and Timely Completion of Work with Quality Control are very vital. In order to achieve this, Mechanization of Work has to be done, where Construction Machinery & Equipment play a pivotal role.

No.	Questions	Learning Outcome (s)	DoK
2	What is the need of construction management?	20CE005.2	L1

Ans: 1.Effective project management 2. Improving efficiency, reduces delay 3. Improves communication 4. Improve safety at sites 5. Ensure quality control

No.	Questions	Learning Outcome (s)	DoK
3	List any four selection factors for rear dump trucks.	20CE005.3	L1

Ans: 1. The Proper Engine 2. Appropriate Axle Requirements 3. Body material 4. Company specifications 5. Site specifications 6. Equipment specifications 7. Clint and project specifications 8. Manufacturer specifications

No.	Questions	Learning Outcome (s)	DoK
4	List any two advantages of automation in concrete technology.	20CE005.4	L1

Ans: 1. Automation and robotics in construction sector and precast concrete industry
2. Automation and robotics in prefabrication of masonry and on site masonry construction

No.	Questions	Learning Outcome (s)	DoK
5	List any two benefits of robots in construction industry.	20CE005.5	L1

Ans: Increased accuracy, significantly increased productivity, reduced errors, meeting deadlines, reduced number of accidents and reduced costs are just some of the improvements that robotics brings to the construction industry, reduce the labor, increase the efficiency

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Classify the construction equipments on different basis with suitable examples.	12 M	20CE005.1	L2

Ans:

- **Earth work equipment:**
- Backhoe - Backhoes are mainly used to clean up construction areas, to dig holes in the ground, to smooth

uneven ground, to make trenches, ditches and to help remove deep roots from trees

- Front shovel- Front shovel are mainly used for excavation purposes above its own track or wheel level they are suitable for heavy positive cutting in all types of dry soils.
- Dragline- They are used for bulk excavation below its track level in loose soils, marshy land and areas containing water
- Clamshell- It consists of a hydraulically controlled bucket suspended from a lifting arm. It is mainly used for deep confined cutting in pits and trenches..
- Dozers- They are used for moving earth up to distance of about 100m and act as a towing tractor and pusher to scraper machines. They can be track-mounted or wheel-mounted.
- Roller compactor- Roller compactor is mainly used to for compaction of earth and other materials in large works of highways, canals and airports.
- Scraper- They are used for site levelling, loading, hauling over distances varying between 150m-900m. They may be towed, two-axle or three-axle type.
- CONCRETING EQUIPMENTS
- Concrete batching and mixing plant- They are mainly used for weighing and mixing large quantity of concrete. Constituents capacity:- 20cum/hr-250cum/hr
- Concrete mixers- They are mainly used for mixing small quantities of concrete constituents. capacity:- 200lt/batch (small mixers), 200-750l/batch (large mixers)
- Concrete transit mixers- They are mainly used for transporting concrete from batching point. capacity:- 3cum-9cum
- Concrete pumps
- HOISTING EQUIPMENTS
- Boom hoists-

  Boom hoists are used to lift weights on the hooks that are attached to the special metal ropes designed to bear maximum loads. Boom hoist is mostly used as industrial machine where it loads the weight on containers.
- Chain hoists- Chain hoists are quite common example of hoist system and it can be seen at most of the construction and industrial purposes. Basically, chain hoist consists of chain rope and pulley that is used to move the load from up to down.
- Electric hoists- A powerful motor is thus required in order to power up this structure. The function of this chain hoist or electric hoist is to lift or the lower the materials by means of a drum
- Tractor hoists

OR

7 (a)	Discuss the applications of grouting equipment.	6M.	20CE005.1	L2

- Ans: repair machine foundations, base plates, load-bearing, and pillar joints in prefabricated structures.
- Grouting is used to fill gaps, cracks in concrete structures.
- Used for repairing footpaths and the ground under foundations.
- Defects in masonry and cracks in concrete are repaired by grouting.
- Used in soil stabilization.
- Used to control water leakage in mines, tunnels, dams, underground structures.
- Grouting is used to repair unusual and difficult geotechnical and structural problems
- Used to aid in the excavation process.

DRILLING EQUIPMENT.: Percussion drilling produces acceptable grout holes and, generally, is the most economical method of drilling shallow holes.

Grout Mixers.: Many types of grout mixers have been used, including hand-turned d ough mixers, concrete mixers of various sizes, and especially designed grout mixers. Any machine is suitable that has the desired capacity and that mixes the grout mechanically to a uniform consistency.

PUMPS: Pumps for cement grouting should be sufficiently flexible to permit close control of pressure and to provide for a variable rate of injection without clogging of valves and feed lines. With constant speed pumps, special arrangements of the supply piping systems and valves are needed to provide close control of the grouting operation.

7 (b)	Explain about plastering machines in construction industry.	6M	20CE005.1	L2

Ans: There are four different types of plastering machines. Each has its specification which determines the most suitable use of such equipment:

- Pneumatic sprayer. This type of sprayer is the simplest and easiest to use among all four types of plastering machines. The mechanism consists of a hopper connected to an air compressor. You pour the plaster into the hopper and air from the compressor moves the plaster out on to the wall.
- Worm-drive pump. The worm-drive pump within the machine pushes plaster material on to the spray gun for application. However, solvent-based plastering materials are not compatible with this type of plastering machine.
- Peristaltic pump. A peristaltic pump consists of a tube squeezed by rollers. Like other plastering equipment, compressed air produces the spray trajectory. A peristaltic pump has the advantage of spraying textured material as long as it is not solvent-based.
- Piston pump. Piston pumps are the latest type of plastering machine used today. Some do not require using compressed air to pump plaster. Piston pumps that use compressed air produce excellent output speed and coverage area which are suitable for commercial plaster applications.
  - After spraying plaster onto a surface, workers usually smooth it with a wide spatula. In some cases, the finished plaster is textured to prepare for decorative application.

8 (a)	Discuss objectives of construction management.	6M	20CE005.2	L2

Ans: An important part of any project, quality construction management helps ensure construction projects remain on time, on budget, and meet all goals for safety, scope, function, and quality. In order to ensure success, the Construction Manager must implement a variety of specialized project management methods.

- CONSTRUCTION MANAGEMENT GOALS
- After commissioning the project, owners hire Construction Managers to ensure everything goes as planned. Regardless of the type of project or the scope of work, Construction Managers provide an invaluable service, taking much of the stress off of the owner.
- The main goal of construction management is to manage and control the progress of construction projects. The Construction Manager plans, coordinates, budgets, and supervises the project from start to finish. They act on behalf of the owner, overseeing every stage of the project.
- IMPORTANCE OF HIRING AN EXPERIENCED CONSTRUCTION MANAGER
- In order to ensure the greatest success, owners of construction projects must consider the prior experience and skills of the Construction Manager they choose. It takes a wide range of skills and competencies to ensure the sufficient progress and successful completion of even simple construction projects. Construction Managers must deal with a multitude of changes including regulatory, design, as well as unforeseen conditions such as extreme weather. It is important the Construction Manager can maintain stability, regardless of the situation.
- CONSTRUCTION MANAGEMENT GIVES OWNERS MORE CONTROL
- Traditional methods do not allow the owner to take an active role in the construction project. This can create problems, especially for owners responsible for projects paid for with taxpayer dollars. While most people assume owners want nothing to do with the project until completion, it is becoming more common

for owners to want and/or need access to their project. Properly implemented construction management ensures owners maintain access and control, while concurrently ensuring projects run as smoothly as possible.

- During the initiation, design, and planning phase, the Construction Manager creates the project plan and develops a comprehensive strategy for completing each additional phase of the project on time and on budget. This is extremely important to the project's overall success.
- The Construction Manager maintains communication and provides guidance to the project team and ensures the smooth and uninterrupted execution of each phase. This aspect of construction management is extremely important, as success depends on the collaboration of all members of the team. Construction Managers must communicate clearly and in ways that everyone on the team can understand

8 (b)	What are the different specifications to be followed while ordering construction equipment?	6M	20CE005.2	L2

Ans: Equipment specifications are written documents or manuals that stipulate the method of production capacity, power requirement, fabrication methods and other finer details of the equipment that makes it apt for use

An engineer must have a clear understanding of equipment Ordering specifications to avoid equipment breakages. Here are five reasons that assert why an adept understanding of equipment specifications is important for an engineer

Six Essential Factors to Consider Before Buying/Ordering Construction Equipment

- There is an increase in infrastructure projects, which is a good reason for construction businesses to have equipment that will help them finish the projects on time, without compromising the integrity of the built structure. If the equipment is no longer helping you achieve that goal, it is time to replace it.
- However, this is not an easy task as heavy equipment is costly and needs significance investment. Here are six factors that you must consider before buying:
- Quality: There will be times when you will have to work in a remote location, where the weather conditions could be unpredictable, unfamiliar or harsh. For example, you could experience constant rain, snow or hail. These conditions can weaken and damage the heavy-lifting equipment if they remain exposed to the harsh elements for a prolonged period of time on a regular basis. The compromised equipment can prove hazardous to the employees working on the site and impact the integrity of the structure being constructed.
- Technology: Embrace technology as it is an ally you want on your side. If you have heavy equipment that has the latest technology, it will surely impact and enhance the overall performance of your business. It also helps in attracting and retaining more business to the contractors. The work would be smoother, helping them complete the projects faster and on time. Industrial weighing scales are an example of such technology
- Fuel efficiency: Heavy construction equipment does not come cheap. Not only is it expensive initially, you will have to shell out high maintenance costs down the line. Cost: Generally, construction projects span over a long time — ranging from a few months to even years or decades. Not planning and allocating assets and investments smartly will affect the overall project and the business.
- Dealer: Ensure that you always buy from a reputable dealer. Take your time and check out numerous dealers before making the purchase.
- Knowledge of using equipment: Efficient and reliable lifting equipment will do you no good if you do not have the skill or dexterity to use it. Working with heavy machinery is quite challenging and poses a workplace safety hazard if not handled carefully.

OR

9 (a)	What are the advantages and disadvantages of mechanization in	8M	20CE005.2	L2

	construction?			

Ans: The Mechanization is the process of shifting from working largely or exclusively by hand to do that work using machines. The construction projects are becoming more demanding and complicated in construction and delay of projects would arise if conventional const method is used.

Mechanization is based on rented construction equipment is cost effective. Construction equipment when rented can be exactly match the requirement. For rented equipment, time to make the equipment ready for operation is important.

Advantages of Construction Mechanization

- Economical
- Improve construction quality
- Increase safety of construction conditions
- Enhance speed of construction
- Feasibility

Disadvantages of Construction Mechanization

- Loss of Skill- The craftsman with the superior skill had disappeared. Such skill is no longer necessary. The only type of skill that is needed now run the machines
- Dependence- Machinery has increased our dependence on others
- Insanitary Surroundings- Big factories pollute their surroundings and make them filthy and insanitary. This has led to moral degradation and physical deterioration
- Over-specialization- Machinery leads to too much specialization. This over- specialization increases the risk of unemployment
- Unemployment- It creates unemployment because one machine can take the place of several men.

9 (b)	Explain about forward planning.	4M	20CE005.2	L2

Ans: As project managers, it can be all too easy to get caught up in the fancy & complicated parts of our job. You know, things like product release maps, feature planning, metrics, crafting a complete product development definition, etc. However, sometimes it is worth it to take a step back and make sure that we still have a firm grasp on the basics of project planning, or forward planning. One key part of this is doing a review of the process that your customers go through when they decide to use your product.

The solid thought behind the project planning to co-ordinates all parts of the project include:

- Reviews plans
- Monitors progress against plans
- Monitors risks, issues, & Change Control
- Reviews quality of the projects
- Manages & how to manage project governance
- Maintains financial control in advance.

10 (a)	Explain the steps to be followed while transportation of concrete mix.	6M	20CE005.3	L2

Ans: Transportation of concrete must be well thought out and organized efficiently. Normally thirty minutes of transportation time is acceptable for small jobs. For a central or portable plant like a ready-mix plant, concrete should be discharged from agitating transporting equipment within two hours. If the non-agitating transporting equipment is used, this time is reduced to one hour. All delays must be avoided to prevent honeycombing or cold joints. There are many factors that determine which type of transportation is most suitable for concreting and taken into consideration when choosing the mode of transportation of concrete such as,

- Type and constituents of the concrete mix,

- **Weather conditions such as humidity, temperature, wind speed etc.,**
- Size of construction,
- Type of construction,
- Topography,
- Location of the batching plant,
- Cost of transportation

If you choose the wrong mode of transportation, concrete might get segregated and useless. Therefore, it is most important that the adequate mode is selected for transportation of concrete as per requirements.

1. Mortar Pan

- Mortar pan is a commonly used method in our country, especially for small-scale works, where the concrete is carried in small quantities. This method is more labor-intensive.
- There are significantly fewer possibilities of concrete segregation, but this method exposes larger surface areas of concrete for dry conditions in thick members.
- This results in a substantial water loss due to more concrete exposure to an environment in hot weather.
- This method is mostly adopted for ground level, below ground level, and above ground level construction work without more difficulties.
- **Precaution**
- The mortar pan must be wetted to start with, and it must be clean during the entire operation of concreting.
- **Mortar Pan**

2. Crane, Bucket, And Ropeway

- Crane and Bucket are mostly correct transportation of concrete equipment, especially for above-ground level. Cranes give more preference to high-rise construction projects.
- Cranes are quick and versatile to move concrete horizontally and vertically along the boom and enable concrete placement at the exact point. Those cranes carry buckets or skips containing concrete. Skips have a discharge door at the bottom, whereas buckets are tilted for emptying.
- Excessive free fall of concrete should be avoided to minimize the segregation defect in concrete.
- **Precaution**
- The freefall of concrete should not be at a high level, and that concrete should be discharged from the smallest height.
- **Crane, Bucket, And Ropeway**

3. Wheel Barrow or Hand Cart

- It is generally used on the ground level (i.e., road construction and other similar structures. Segregation can occur if transportation of concrete is done on rough roads over a long distance. However, this vibration problem can be minimized if pneumatic tires are used.
- **Wheel Barrow or Hand Cart**

4. Belt Conveyor

- It has limited application in the concrete construction sector due to chances of segregation on steep slopes, roller points, and changes in the direction of the belt. Segregation also takes place due to the vibration of a rubber belt.
- Modern belt conveyors can have adjustable reach, traveling diverter, and variable speed both forward and reverse. Conveyors can quickly place a large volume of concrete where access is limited, and portable belt conveyors are used for short distances or lifts. The end discharge arrangements must prevent segregation and remove all the mortar on the belt's return. In adverse, It also involves over-exposure of concrete to the environment.

- **Precaution:** Concrete should be remixed at the end of delivery before placing the final position.
- **Belt Conveyor for Transportation of Concrete**

  5. Truck Mixer And Dumper
- For large-scale concrete works, mostly concrete to be placed at ground level, trucks and dumpers or ordinary open steel-body tipping lorries can be used.
- It is a most improved and better method for long lead concreting. The concrete is covered with a tarpaulin if transported in open trucks. If long-distance is involved, agitators should be used, which prevents segregation and stiffening, and it also helps the mixing process at a slow speed.
- For road construction using slip form paver, a large quantity of concrete must be supplied continuously.
- Another name of small dumper is tough riders, and it is used for factory floor construction.
- **Precaution**
- Before loading with concrete, the body's inside should be just wetted with water.
- **Truck Mixer And Dumper**

  6. Skip and Hoist
- High-rise structures for transporting concrete vertically up are a widely used method.
- Then concrete is fed into the skip, which travels vertically on rails like a lift. After discharge, it is better to turn over the concrete before avoiding segregation. Transportation of concrete is practically impossible with the mortar pan and ladder for more than 3 to 4 number floors; in that case, the skip and hoist method is adopted.
- The mixer feeds the skip, and the skip travels up over the rail up to the level where concrete is needed, and at that point, the skip discharges the concrete automatically or manually.
- The quality of concrete, i.e., freedom from segregation, will depend upon the extent of travel and rolling over the rails. If the concrete has traveled a considerable height, concrete on discharge must be required to be turned over before being placed finally.
- **Skip & Hoist**

  7. Chute
- It is usually used for concreting in deep locations.
- The workability should not be changed to suit the delivery by a chute.
- This method is extensively used in the field, though it is technically not a very good method.
- **Precaution**
- The slope should not be lower than 1V:2.5H; otherwise, concrete will not slide down smoothly.

  8. Transit Mixer
- In the ready-mix concrete plant, the transit mixer is one of the most popular concrete transporting equipment types.
- **Transit Mixer**
- They are truck-mounted transit mixers with 4 to 7 cubic meters.
- **There are two variations; they are as follows;**
- 1) Mixed concrete is transported to the site by keeping it agitated at a speed varying between 2 to 6 revolutions per minute.
- 2) At the central batching plant, the concrete is batched, and in the truck mixer, mixing is done either in transit or immediately before discharging concrete at the site.
- **9. Pump and Pipe-Line Method**
- It is the most complicated method uniquely proper for limited space or when a large quantity of concrete can be poured except cold joints.

10 (b)	Explain about construction compaction equipment.	6M	20CE005.3	L2

- Ans: Smooth Wheeled Roller- It is an important equipment for compaction. It consist large steel drum in front and one or two wheel on rear end.
- Sheepsfoot Roller- Sheepsfoot rollers also known as a tamping roller. Steel drum of sheepsfoot roller consist of many rectangular shaped boots of equal sizes fixed in hexagonal pattern.
- Pneumatic Roller- Pneumatic roller is also called rubber tyres roller. Pneumatic roller has number of tyres at the front and at the rear end
- Vibratory Roller- Vibratory roller consist two smooth drums with the vibrator. One is fixed at front and other one on rear side of vibratory roller. Both drums are of the same diameter, length and same weight.
- Rammer- Rammer compactor is used for compacting small area and providing impact load to soil. This equipment is light weight and can be hand or machine operated.
- Vibratory Plate Compactor- Vibratory plate compactor is used for compacting different types of soils in narrow and congested area where it is not possible to use large equipment

OR

11	Explain the concrete production steps in ready mix concrete plant.	12M	20CE005.3	L2

- **Ans: Concrete manufacturing steps:**

1. **Procurement of Ingredients of Concrete:** Procuring all the ingredients like Cement, Sand, Aggregate and Water in required quantity.
2. **Storage and Handling:** The procured material is stored in dry and damp free spaces so that they will not get moisture.
3. **Batching:** To measurethe materials required concrete is known as batching.
   There are two methods
   A. Volumetric Batching
   B. Weigh Batching
4. **Mixing**
   The cement is then mixed with the other ingredients: aggregates (sand, gravel, or crushed stone), admixtures, fibers, and water. Aggregates are pre-blended or added at the ready-mix concrete plant under normal operating conditions. The mixing operation uses rotation or stirring to coat the surface of the aggregate with cement paste and to blend the other ingredients uniformly. A variety of batch or continuous mixers are used.
   Fibers, if desired, can be added by a variety of methods including direct spraying, premixing, impregnating, or hand laying-up. Silica fume is often used as a dispersing or densifying agent.
5. **Placing and compacting**
   Once at the site, the concrete must be placed and compacted. These two operations are performed almost simultaneously. Placing must be done so that segregation of the various ingredients is avoided and full compaction—with all air bubbles eliminated—can be achieved. Whether chutes or buggies are used, position is important in achieving these goals. The rates of placing and of compaction should be equal; the latter is usually accomplished using internal or external vibrators. An internal vibrator uses a poker housing a motor-driven shaft. When the poker is inserted into the concrete, controlled vibration occurs to compact the concrete. External vibrators are used for precast or thin in situ sections having a shape or thickness unsuitable for internal vibrators. These type of vibrators are rigidly clamped to the formwork, which rests on an elastic support. Both the form and the concrete are vibrated. Vibrating tables are also used, where a table produces vertical vibration by using two shafts rotating in opposite directions
6. ***Curing***
   Once it is placed and compacted, the concrete must cured before it is finished to make sure that it doesn't dry too quickly. Concrete's strength is influenced by its moisture level during the hardening process: as the

	cement solidifies, the concrete shrinks. If site constraints prevent the concrete from contracting, tensile stresses will develop, weakening the concrete. To minimize this problem, concrete must be kept damp during the several days it requires to set and harden.			
12	Explain the adoption of photogrammetry in construction industry with the help of drones.	12M	20CE005.4	L2

- Ans: Drones. You've heard about them and seen them everywhere. Whether flying around a room or capturing unique aerial footage, drones offer vast possibilities and applications — even in the world of construction.
- In the past few years, drones have become one of the most compelling construction trends. The industry has experienced a 239% growth in drone use year over year, higher than any other commercial sector. Their aerial vantage point and data collecting abilities make them a viable tool, offering benefits ranging from on-site safety to remote monitoring.
- n particular, the benefits of drone technology have revolutionized the entire project life cycle, from inception through project closeout. Drone photos, videos, and imagery are used to scope out projects, track building progress and provide real-time updates.
- As the industry grows and construction projects become more complex, the use of drones in construction will continue to skyrocket. Read on to learn more about how these futuristic devices are transforming the industry.

  How Are Drones Used in Construction?
- With their real-time data recording and unique aerial advantage, drones can improve efficiency, cut costs and streamline workflow. Here are some of the ways drones are used in construction.
- Topographic Mapping and Land Surveys
- Equipment Tracking
- Remote Monitoring and Progress Reports
- Security Surveillance
- Personnel Safety
- Structure Inspection and Photography

OR

13	Explain about structural health monitoring.	12M	20CE005.4	L2

- Ans: Structural health monitoring (SHM) system is a method of evaluating and monitoring structural health.
- It has been widely applied in various engineering sectors due to its ability to respond to adverse structural changes, improving structural reliability and life cycle management.
- Various methods in SHM are presented in this chapter such as acoustic emission, ultrasonic, and thermal imaging. The advantages of SHM and nondestructive testing method are compared.
- In addition, SHM applications in the aerospace, civil, and energy sectors are discussed thoroughly.
- Furthermore, special attention is given to composite material that employs the SHM system in recent research investigations.
- Anisotropic composite materials can be divided into three different fiber reinforcements, namely synthetic, natural, and hybrid fibers. SHM applications in composites and the recent trends are explored.
- In approaching the fourth industrial revolution, SHM as a sensor for civil engineering applications is proposed due to its capability of interacting with the environment.
- Furthermore, the intention to enhance sensor capability to handle more functions is also discussed. At the end of this chapter, recent issues are highlighted to emphasize the technology and SHM potential as an energy harvester in the near future.

14	Explain automation in production of steel components.	12M	20CE005.5	L2

Ans : From a technical point of view, in any case, it is hardly possible to explain the difference of the development between Germany and Japan in prefabrication automation in steel housing market.

- The current situation in steel construction and assembly can be characterised as follows:

The building market mainly demands solutions from the steel construction companies which fulfil the clients' individual needs and therefore only conditionally allow rational standardisation with regard to production and assembly

- This applies to all fields of steel construction, e.g. bridge construction, multi-storey building and hall construction, container construction, compound construction and steel machine and plant construction. As the percentage of steel in the housing sector is low, the steel frames applied in prefabrication for room cells are to a large extent welded or screwed manually.
- If we wonder as to how the acceptance of steel in the housing industry can be enhanced in Germany, then Japan could be given as a good example. We see a possibility to learn from the experience made by Japan in the way the building material steel has been supported consistently and with perseverance by direct marketing with united forces. Today the material steel offers a variety of new possibilities in comparison with the first steel enterprises. Material and production technology have gone through enormous developments, whereby technological developments were in the majority of cases initiated by other branches (automobile industry).
- It can, however, be imagined that as a result of a new intensified use of steel in the housing industry innovation potential for the material will arise. By research, experiments and applications, steel can be improved in its capacities and characteristics so that any possible objections raised against steel in the housing sector will lose their validity.
- CAD/CAM solutions are the state-of-the-art in steel construction companies to ensure the flexibility required from projecting via CNC production to delivery (logistics) to the construction site and, if applicable, to assembly organisation. The aim is to produce constructions tuned to manufacturing and assembly requirements to a large extent without reworking at the construction site (e.g. adapting resp. cutting operations) enabling short assembly or construction operations.
- The construction parts are cut by laser, gas burner cutting, sawing, drilling, before undergoing straightening including metallic cleaning, interim and end coating and complete corrosive protection which are normally processes applied in pre-fabrication.

OR

15	List various applications of automation in timber construction.	12M	20CE005.5	L2

Ans: The prefabrication degree in wood construction can be characterised as favourable according to the current state-of-the-art prevailing in technology in Germany in comparison with other European countries. All conventional wood construction systems are applied.

- In the recent past focus has particularly been on the „novel block construction" system (glued laminated wood, bulk wood and log wood construction). Perhaps also because this - in the form of massive constructions normally made of bonded two-dimensional wood - associates wood construction more intensively with massive construction („knock test", wood/massive compound constructions).
- The processing technology in wood construction is developping continuously from manual processing with small machines to full-scope processing on CNC machines. The requirements with regard to flexibility in processing are noticeably rising.
- The division between raw construction and interior design no longer exists. Wood constructions are transformed into pieces of furniture. The standards required with regard to precision in production exceed the general level of a carpenter by far.

- In production there is an enormous difference whether raw wood constructions, construction parts for prefabricated houses, staircases or winter gardens or even all together have to be processed on one machine. In serial production the aim is to manufacture the largest possible quantity of identical or similar parts within the shortest possible space of time.
- For the wood construction worker the most important aspect is traditionally bonding construction wood. For these operations optimally functioning and reliable bonding systems have been on the market for many years. They are characterised by high performance and relatively low programming requirements.

6/12/22

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Mechanical Engineering	Academic Year	2022 - 2023
Course Code	20ME002	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Unconventional Machining Process				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	State the need for unconventional machining process.	20ME002.1	L1
2	List any two applications of Water Jet Machining.	20ME002.2	L1
3	What are the electrolytes commonly used in ECM?	20ME002.3	L1
4	What the functions of dielectric fluid in EDM process?	20ME002.4	L1
5	Write the various types of torches used in plasma arc machining.	20ME002.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What do you understand by the term unconventional machining methods? What is their importance?	6M	20ME002.1	L1
6 (b)	What is ultrasonic machining? Why is it recommended for brittle materials? What are various tool materials and abrasives used? Compare the abrasives based on cutting ability, life and cost.	6M	20ME002.1	L2
	OR			
7	Explain the working principle of USM with a neat sketch.	12M	20ME002.1	L2
8 (a)	Write the factors that affect the performance of water jet machining (WJM) process. Discuss their effects.	6M	20ME002.2	L2
8 (b)	List any four advantages, disadvantages and applications of abrasive water jet machining (AWJM) process.	6M	20ME002.2	L1
	OR			
9	Explain the working of an Abrasive Jet Machine with the help of a neat sketch.	12M	20ME002.2	L2
10 (a)	Explain the principle, working and advantages of electro chemical machining process.	6M	20ME002.3	L2
10 (b)	Explain the electrochemical deburring and honing processes.	6M	20ME002.3	L1
	OR			
11	Explain the mechanism of material removal during ECG and how is different from ECM	12M	20ME002.3	L2
12 (a)	Explain about R-C circuit used for pulse generation in EDM process.	6M	20ME002.4	L2
12 (b)	Explain the process of wire cut EDM with a neat sketch.	6M	20ME002.4	L2
	OR			
13	Explain the effect of different process parameters in EDM.	12M	20ME002.4	L2
14 (a)	Explain the principles and elements of EBM, also how the work table is protected from getting damaged by electron beam.	6M	20ME002.5	L2
14 (b)	Discuss the metal removal mechanism in Laser Beam Machining.	6M	20ME002.5	L2
	OR			
15	Explain the principle of plasma generation and mechanism of metal removal in plasma arch machining.	12M	20ME002.5	L2

NSRIT

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(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

SUBJECT AND SUBJECT CODE:- UCMP, 20ME002

Q.no.	Answers	Marks
1	a) Extremely hard and brittle materials or Difficult to machine material are difficult to machine by traditional machining processes. b) When the workpiece is too flexible or slender to support the cutting or grinding forces. c) When the shape of the part is too complex.	2
2	a) WJM is used to cut many nonmetallic materials like Keplar, glass epoxy, graphite, boron, leather and many other brittle materials. b) It is used mostly in shoe making industry and now has entered into steel plant to descale the chilled layer of steel ingots, in aircraft industries to profile cutting of FRP aircraft structures even glass windows.	2
3	Common electrolytes used are NaCl, $NaNO_3$, NaOH, potassium chloride, Sulphuric acid.	2
4	1.Insulates the gap between the tool and work, thus preventing a spark to form until the gap voltage are correct. 2.Cools the electrode, workpiece and solidifies the molten metal particles. 3.Flushes the metal particles out of the working gap to maintain ideal cutting conditions, increase metal removal rate.	2
5	Non-transferred and transferred DC torches.	2
6 (a)	There is a need for machine tools and processes which can accurately and easily machine the most difficult-to-machine materials to intricate and accurate shapes. The machine tools should be easily adaptable for automation as well. In order to meet this challenge, a number of newer material removal processes have now been developed to the level of commercial utilization. These newer methods are also called **unconventional** in the sense that conventional tools are not employed for metal cutting. Instead, the energy in its direct form is used to remove the materials from the workpiece. Unconventional machining processes are important where:- (i) Limitations of Conventional machining methods. (ii) Rapid improvements in material properties such as hardness. (iii) Need to machine complex shapes, low tolerances, high surface finish. (iv) Precision machining. (v) High production rate. (vi) Extremely hard and brittle materials or Difficult to machine materials are difficult to machine by traditional machining processes. (vii) When the work piece is too flexible or slender to support the cutting or grinding forces. (viii) Machining of composites.	6
6 (b)	Ultrasonic machining is a subtractive manufacturing process that removes material from the surface of a part through high frequency, low amplitude vibrations of a tool against the material surface in the presence of fine abrasive particles. This is an impact erosion process to machine materials where the work material is removed by repetitive impact of abrasive particles carried in a liquid medium in the form of slurry under the action of a 'shaped' vibrating tool attached to a vibrating mechanical system "Horn". It is employed to machine hard and brittle materials (both electrically conductive and non-conductive) having hardness usually greater than 40 RC. The word ultrasonic describes a vibratory wave having frequency larger than	6

waves and longitudinal waves. High velocity longitudinal waves can easily propagate in solids, liquids and gases. They are normally used in ultrasonic applications.

Tools are usually made of relatively ductile materials (brass, stainless steel, mild steel, etc) so that the tool wear rate (TWR) can be minimized. Value of the ratio of TWR and MRR depends upon the kind of abrasives, workpiece material, and tool material. Surface finish of the tool is important because it will affect the surface finish obtained on the workpiece. Tools should be properly designed to account for overcut. Silver brazing of the tool with tool holder minimizes the fatigue problem associated with screw attachment method.

Commonly used abrasives in the order of increasing hardness, are Al203, SiC and B4C. Hardness, particle size, usable lifetime and cost should be criteria for selecting abrasive grains to be used in USM. Coarser grains result in higher MRR and poorer surface finish while reverse is true with finer grains. Mesh sizes of grits generally used range from 240 to 800. Abrasive slurry consists of water and abrasives usually in the weight ratio 1:1. However, it varies according to the operations, such as low concentration mixtures are used while drilling deep holes, or machining complex cavities so that the slurry flow is more efficient.

OR

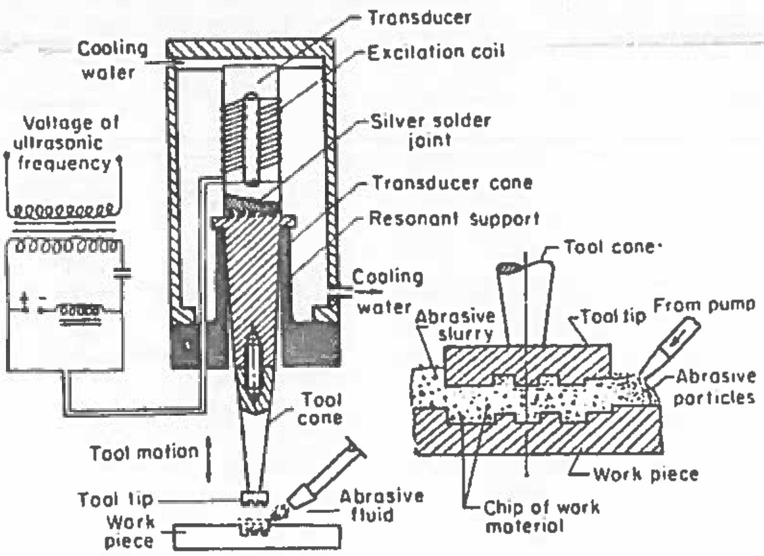
7 — **12**

This is an impact erosion process to machine materials where the work material is removed by repetitive impact of abrasive particles carried in a liquid medium in the form of slurry under the action of a 'shaped' vibrating tool attached to a vibrating mechanical system "Horn". The tool shape is made converse to the desired cavity. The tool is placed very near to the work surface, and the gap between the vibrating tool and the workpiece is flooded with abrasive slurry made up of fine abrasive particles and suspension medium (usually water). As the tool vibrates in its downward stroke, it strikes the abrasive particles. This impact from the tool propels the grains across the gap between the tool and the workpiece. These particles attain KE and strike the work surface with a force much higher than their own weight. This force is sufficient to remove material from the brittle workpiece surface and results in a crater on it. Each down stroke of the tool accelerates numerous abrasive particles resulting in the formation of thousands of tiny chips per second. A very small percentage (about 5 %) of material is also believed to be removed by a phenomenon known as cavitation erosion. To maintain a very low constant gap between the tool and the work, feed is usually given to the tool.

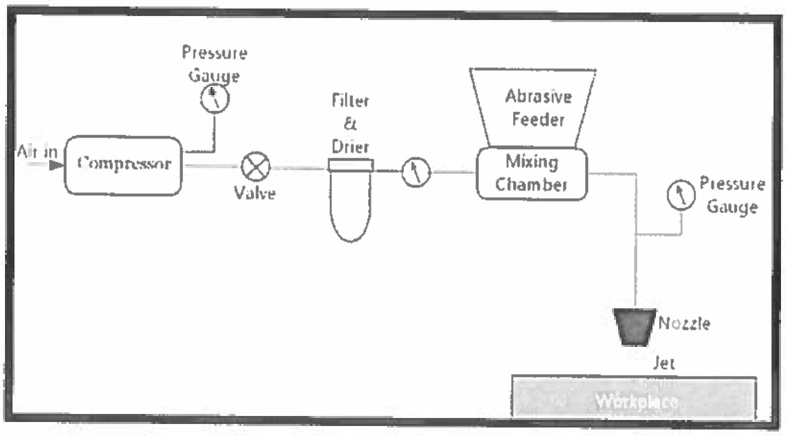
USM gives low **MRR** but it is capable to machine intricate cavities in single pass in fragile or /and hard materials.

In USM, there is no direct contact between the tool and workpiece hence it is a good process for machining very thin and fragile components.

A brittle material can be machined more easily than a ductile one. It is considered as a very safe process because it does not involve high voltage, chemicals, mechanical forces and heat.



A schematic diagram of ultrasonic machining

	(i) Many variables such as nozzle orifice diameter, water pressure, cutting feed rate and the stand distance affect the performance. (ii) Generally, high cutting quality would be the result of the conditions: high pressure, large nozzle orifice, low feed rate and narrow stand off distance. (iii) The equipment consist of three main units: (1) pump along with a intensifier to generate very high pressure (1-10 kbar); (2) cutting unit consisting nozzle and work table movement and (3) filtration unit to remove the debris from water after use. (iv) A polymer (glycerin, polyethylene oxide) is added to the working fluid to prevent freezing and provide lubricating action in the intensifier plunger type pump. (v) This employs a fine, high pressure (1500-4000 MN/cm2), high velocity (up to twice the speed of sound) jet of water, which when bombarded on the work piece erodes the material. (vi) High-pressure water jet has two properties which make it potentially useful in industries. They are, its destructive power and its application as a precision cutting tool.	
8 (b)	**Advantages** • No heat-affected zone • No cutter induced distortion • Eliminates thermal distortion • Low cutting forces on workpieces • Localises structural changes • No slag or cutting dross • Limited tooling requirements • Typical finish 125-250 microns **Disadvantages** • high capital cost and high noise levels during machining. • It cannot cut the materials that degrade quickly with moisture. • Surface finish loses at higher cut speeds. **Applications** • It is highly used in the automotive, aerospace, and electronics industries. • In aerospace industries, parts such as engine components (aluminium, titanium, and heat resistant alloys), aluminium body parts, titanium bodies for military aircraft, etc. are made using abrasive water jet machining process.	6
	OR	
9	Abrasive jet machining is the material removal process where the material is removed by high velocity stream of air/gas or water and abrasive mixture. A focused stream of the abrasive particles carried by high pressure air or gas is made to impinge on the work surface through a nozzle and the work material is removed by erosion by high velocity abrasive particles. AJM can be used to cut hard and brittle materials (e.g. germanium, silicon, mica, glass and ceramics) in a large variety cutting and deburring and the process is smooth and free from vibration. It differs from conventional sand blasting process for its fineness of particle size and controllable machining parameters. 	12

- Abrasive particles that are carried by carrier gas/air are made to impinge on the work material at high velocity.
- High velocity stream of the abrasives is generated by converting pressure energy of carrier gas or air to its kinetic energy and hence high velocity jet.
- Nozzle directs the abrasive jet in a controlled manner onto workpiece.
- Metal cutting action by micro-cutting as well as brittle fracture of the workpiece.
- Different from conventional sand blasting, finer abrasive grits are used and process parameters are easily controllable.

10(a) — 6

Electrochemical machining (ECM) is a machining process in which electrochemical process is used to remove materials from the workpiece. In the process, workpiece is taken as anode and tool is taken as cathode. The two electrodes workpiece and tool is immersed in an electrolyte (such as NaCl). When the voltage is applied across the two electrodes, the material removal from the workpiece starts. The workpiece and tool is placed very close to each other without touching. In ECM the material removal takes place at atomic level so it produces a mirror finish surface.

ECM working is opposite to the electrochemical or galvanic coating or deposition process.

During electrochemical machining process, the reactions take place at the electrodes i.e. at the anode (workpiece) and cathode (tool) and within the electrolyte.
Let's take an example of machining low carbon steel which is mainly composed of ferrous alloys (Fe). We generally use neutral salt solution of sodium chloride (NaCl) as the electrolyte to machine ferrous alloys. The ionic dissociation of NaCl and water takes place in the electrolyte as shown below.
As the potential difference is applied across the electrode, the movement of ions starts in between the tool and w/p. The positive ions move towards the tool (cathode) and negative ions move towards the workpiece.

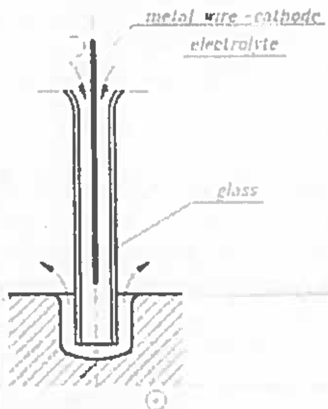


Fig. 1 Working Principle diagram

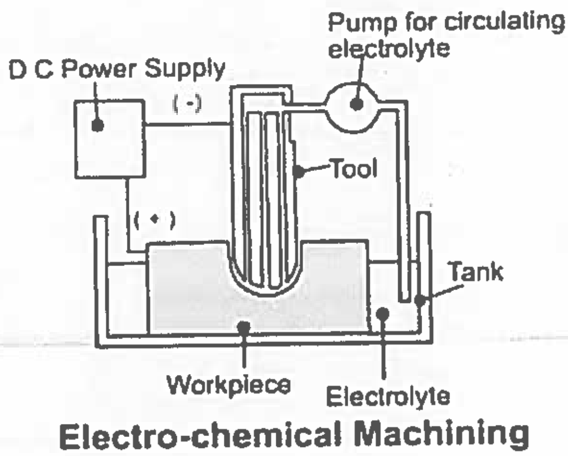


Fig. 2 ECM illustration

Advantages of ECM are

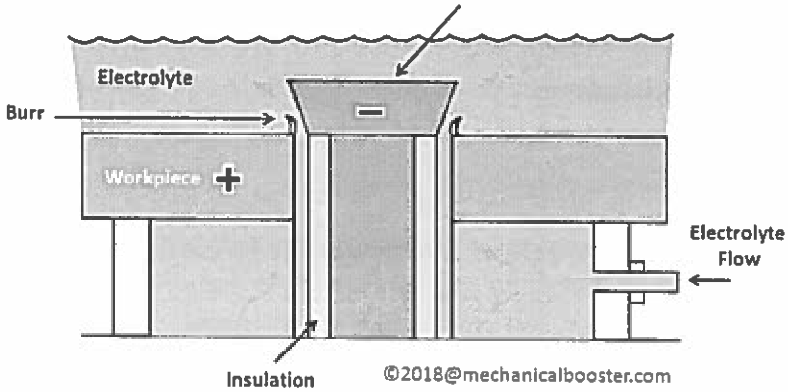
- There is no mechanical or cutting forces involved. Therefore clamping is not required except for controlled motion of the work piece.
- There is no heat affected zone.
- Very accurate.
- Relatively fast
- Can machine harder metals than the tool.

10(b) — 6

Electrochemical Deburring:-

Deburring is finishing process that is used to prepare metal or wood for finishing. This can be achieved by filing, using sand paper (wet and dry papering until the edges become smooth). Deburring is a standard deburring is performed using mechanical deburring process, but thermal deburring and electrochemical deburring processes are also used.

In electrochemical deburring process, burrs are dissolved by the action of a neutral salt electrolyte flowing through the gap between the tool or cathode and the workpiece which is the anode. This is a very fast, precision process. The amount of material removed is proportional to the amount of time and levels of current applied. The dissolved metal in the form of hydroxides is carried away by the controlled flow of the electrolyte which is the filtered for reuse. Electrochemical electrodes are made of copper- tungsten. The workpiece areas not being deburred are insulated by a non-conductive material.

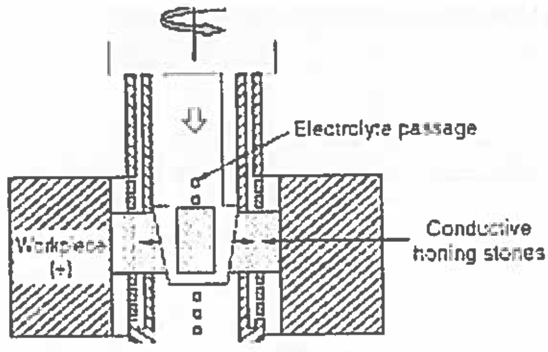
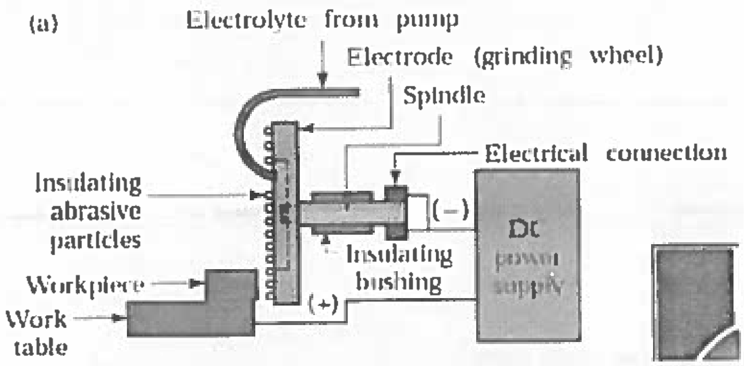


Electrochemical Deburring

Electrochemical Honing:-

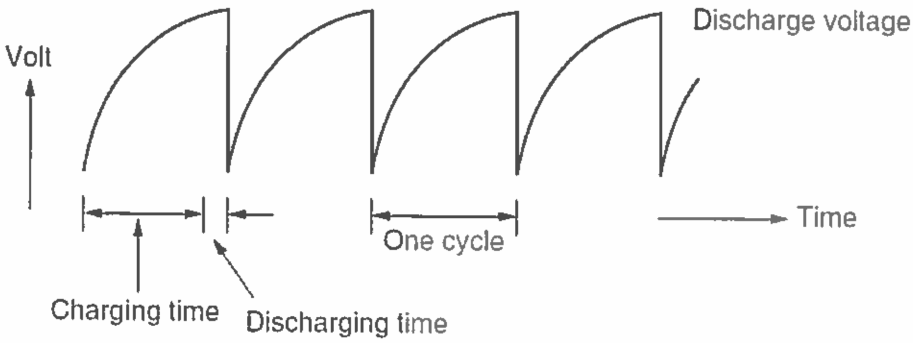
Electrochemical honing is a process in which the metal removal capabilities of ECM are combined with the accuracy capabilities of honing. The process consists of a rotating and reciprocating tool inside a cylindrical component. Material is removed through anodic dissolution and mechanical abrasion – 8% or more, of the material removal occurs through electrolytic action. As with conventional ECM, the workpiece is the anode and a stainless steel tool is the cathode. It is a process in which it combines the high removal characteristics of Electrochemical Dissolution(ECD) and Mechanical Abrasion(MA) of conventional Honing. It has much higher rates than either of honing & internal cylindrical grinding. Cathodic tool is similar to the conventional honing tool, with several rows of small holes so that electrolyte could enter directly into interelectrode gap. Electrolyte provides electron through the ionization process which acts as coolant and flushes away the chips that are formed off by mechanical abrasion and metal sludge that results from electrochemical dissolution action.

The tool is inserted inside the worked hole or a cylinder. Mechanical abrasion takes place first by the stones/hones. Oxides formed due to working from previous process will be removed by it and clean surface will be achieved. Now the clean surface will be in contact with electrolyte and then Electrochemical Dissolution will remove the desired material. Same procedure is continued till the required cut is made. To control surface roughness Mechanical Abrasion is allowed to continue for a few seconds after the current has been turned off.

	 Fig. Schematic of ECH process	
	OR	
11	• ECG also called electrolytic grinding is similar to ECM, except that the cathode is an electrically conductive abrasive grinding wheel instead of a tool shaped like the contour to be machined • Used primarily to machine difficult to cut alloys such as stainless steel, Hastelloy, Inconel, Monel, Waspalloy and tungsten carbide, heat treated workpeices, fragile or thermo-sensitive parts, or parts for which stress-free and burr-free results are required  Fig. 1 Schematic Illustration of ECG As the tool approaches the work piece it erodes the negative shape of it. Thus complex shapes are made from soft copper metal and used to produce negative duplicates of it. The abrasive grains on the ECG wheel serve three major purposes: 1. Act to wipe the oxide from the workpiece, exposing new metal and allowing the process to continue 2. Spacer to keep the conductive media in the wheel from making direct contact with the workpiece and generating a short circuit 3. The cavities between the grit are filled with electrolyte, and the grit acts as a carrier bringing the electrolyte to the work area between the workpiece and the wheel making the ECG process possible ECG wheels are made of an abrasive material, a bonding agent and a conductive medium. Most ECG wheels have aluminium oxide as the abrasive and contain copper impregnated resins for conductivity.	12
12 (a)	Two broad categories of generators (power supplies) are in use on EDM. • Commercially available: RC circuits based and transistor controlled pulses. • In the first category, the main parameters to choose from at setup time are the resistance(s) of the resistor(s) and the capacitance(s) of the capacitor(s). • In an ideal condition, these quantities would affect the maximum current delivered in a discharge. • Current delivery in a discharge is associated with the charge accumulated on the capacitors at a certain moment.	

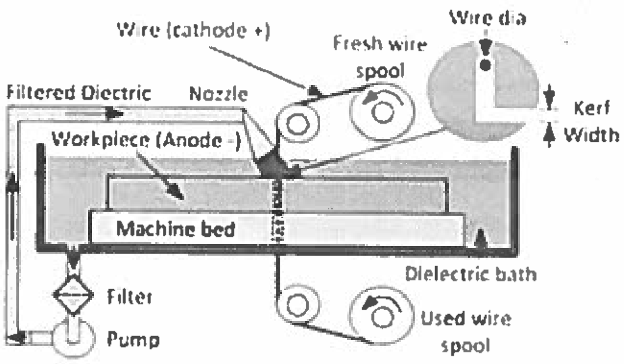
gap conditions.

- Advantage: RC circuit generator can allow the use of short discharge time more easily than the pulse-controlled generator.
- Also, the open circuit voltage (i.e. voltage between electrodes when dielectric is not broken) can be identified as steady state voltage of the RC circuit.
- When using RC generators, the voltage pulses, shown in Fig. are responsible for material removal.



12 (b)

WEDM consists of a thin single-strand metal wire (usually brass) which is fed through the workpiece submerged in a tank of dielectric fluid (typically deionized water). The electrode and workpiece are connected to a suitable power supply. The power supply generates an electrical potential between the two parts. As the electrode approaches the workpiece, dielectric breakdown occurs in the fluid, forming a plasma channel, and a small spark jumps. These sparks happen in huge numbers at seemingly random locations. As the base metal is eroded, and the spark gap subsequently increased, the electrode is lowered automatically so that the process can continue. Several hundred thousand sparks occur per second, with the actual duty cycle carefully controlled by the setup parameters.
WEDM is used to cut plates as thick as 300mm and to make punches, tools, and dies from hard metals that are difficult to machine with other methods. The water flushes the cut debris away from the cutting zone. Flushing is an important factor in determining the maximum feed rate for a given material thickness. It is commonly used when low residual stresses are desired, because it does not require high cutting forces for material removal.



OR

13

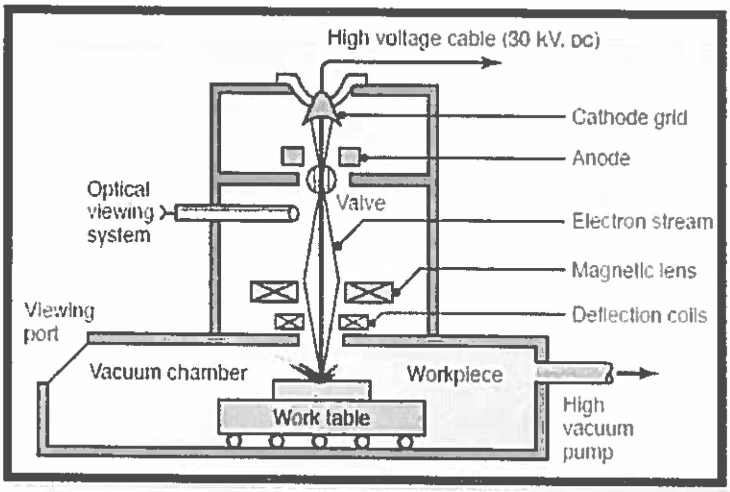
The EDM process parameters are:-

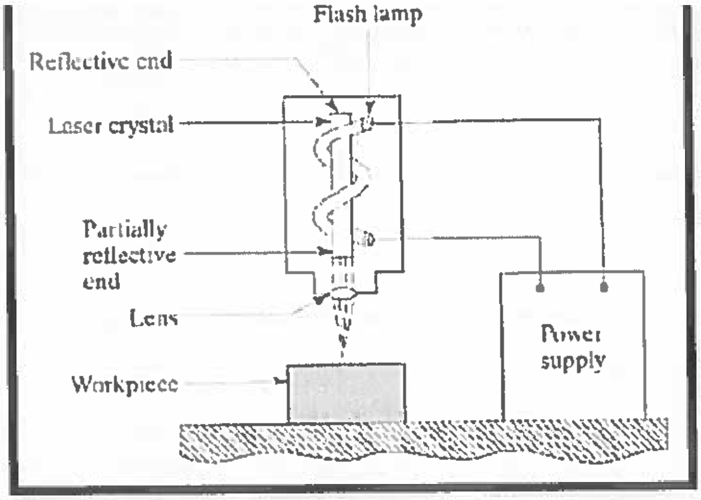
a) Open circuit voltage (Vo)- The circuit voltage is responsible for the spark created between the tool and the workpiece.

b) Working voltage (Vw)- Higher voltage will cause greater sparks that will lead to more material removal from the workpiece.

c) Maximum current (Io)- As the current and voltage are proportional, increase in voltage will increase current that will cause material removal from the workpiece.

	work is done. e) Pulse off time (toff) – It is the duration for which the voltage pulse is not applied. In this duration flushing occurs. f) Duty cycle- It is the summation of on-time and off-time. g) spark gap (δ)- The gap between the workpiece and the tool. This gap is responsible for material removal. More gap will cause less MRR where as smaller gap will cause more melting of the workpiece. Hence an optimum gap should be maintained. h) The polarity – Generally straight polarity is used where tool is -ve and workpiece is positive to get a better surface finish. i) The dielectric medium- Helps for cooling the tool and workpiece, ionization and external flushing through the spark gap.	
14 (a)	EBM equipment details 1. Cathode Cartridge • Tungsten/Tantalum • High voltage is applied • 2500 degree centigrade • Thermo-ionic emission of electrons (vacuum)- thermally induced flow of electrons from surface. • Negatively biased- repel the electrons. 2. Bias Grid • Highly negatively biased • Controls the flow of electrons • To avoid the divergence of the electrons and send them as a beam to the next step (anode). 3. Anode • Positively biased terminal. • Due to the potential difference between cathode and anode the electrons accelerate. • Velocity is approximately half the velocity of light- passing through the anode. 4. Magnetic Lens • Same function as that of any lens. • Concentrates the beam of electrons. • Shape the beam. • Reduce the divergence of the beam. 5. Deflection coils • Deflect the electron beam by small amount. • Correct the beam in case of not getting proper hold-ship. • Improve the shape of the machined holes. 6. Power Supply • The high-voltage power supply used for EBM systems generates voltages of up to 150 kv to accelerate the electrons. • The most powerful electron beam machining systems are capable of delivering enough power to operate guns at average power levels of up to 12 kw.	

	• To avoid the possibility of arcing and short circuits, the high voltage sections of the power supply are submerged in an insulating dielectric oil. 7. Electron beam Gun • Electron beam gun is the heart of EBM. • The basic functions of any electron beam gun are to generate free electrons at the cathode, accelerate them to a sufficiently high velocity and to focus them over a small spot size.  Schematic diagram of Electron Beam Machining	
14 (b)	Laser beam machining is based on the conversion of electrical energy into light energy and then to thermal energy. The electrons are charged particles, they carry some energy. The energy related with the orbit in which the electrons revolve. Generally the electrons are present in the outer most orbit of the atom take part in the process of energy absorption or emission. Ground state, the state with lowest energy is the most stable state for the electrons. After absorbing the energy electron jump to the higher energy state and staying there for some seconds jump to the ground state and release the absorbed energy. In the beginning in atom all the crystals are in ground state. When the light is flash over the crystal, most of the atoms are raised to the excited state. Some light waves incline to the axis of the crystal will leave the box either after only a few reflections or without strike on mirror. Some of the waves that travel parallel to the axis of the crystal, will spontaneously emit photon from chromium ions. These photon stimulate another atom to contribute a second photon. This process continues as the photons are reflected to and fro between the mirror. At the each reflection a certain loss occurs. It is very interesting that laser has to be used on materials where it absorbs laser energy. Upon absorption of the laser energy, there is rapid rise in the temperature leading once again to melting and vaporization and material removal. Although several types of laser exist, all lasers produce (emit) intense, coherent, highly collimated beam of single wavelength light. In material processing applications, this narrow beam is focused by an optical lens to produce a small, intense spot of light on the work piece surface. Optical energy is converted into heat energy upon impact and temperatures generated can be high to melt and/or vaporize any material.	

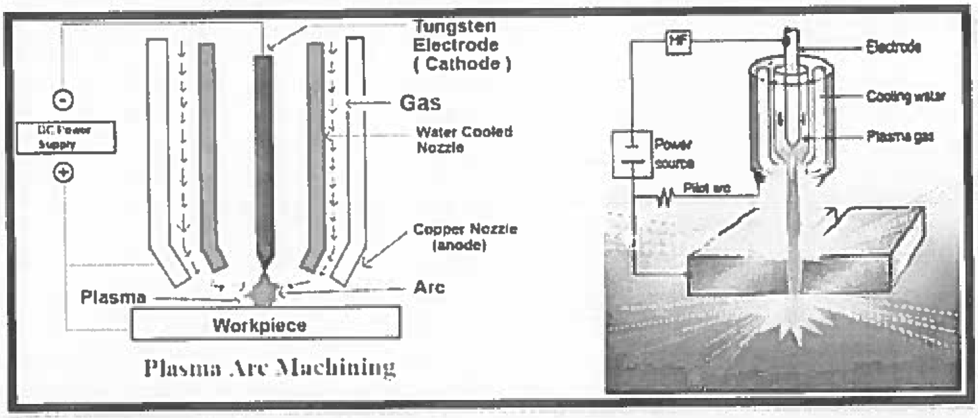


Laser Mechanism

OR

15

- Plasma-arc machining (PAM) employs a high-velocity jet of high temperature gas to melt and displace material in its path.
- Gases are heated and charged to plasma state.
- Plasma state is the superheated and electrically ionized gases at approximately 5000°C.
- These gases are directed on the workpiece in the form of high velocity stream.



Plasma Arc Machining

PAM consists of :-

1. Plasma gun

- The plasma gun consists of a tungsten electrode fitted in the chamber.
- The electrode is given negative polarity and nozzle of the gun is given positive polarity.
- A strong arc is established between the two terminals anode and cathode.
- There is a collision between molecules of gas and electrons of the established arc.
- Gas molecules get ionized and plasma state is formed.
- Plasma is directed to the workpiece with high velocity.

2. Power supply

- Power supply (DC) is used to develop two terminals in the plasma gun.
- A tungsten electrode is inserted to the gun and made cathode and nozzle of the gun is made anode.

	3. Cooling mechanism • Hot gases continuously comes out of nozzle so there are chances of its over heating. • A water jacket is used to surround the nozzle to avoid its overheating. 4. Work piece • Hot gases continuously comes out of nozzle so there are chances of its over heating. • A water jacket is used to surround the nozzle to avoid its overheating.	

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	EEE			**Academic Year**	2022-2023
Course Code	20EE002	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	V
Course	Digital Control Systems						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four merits of digital control system	20EE002.1	L1
2	Recall primary strips	20EE002.2	L1
3	Define Routh's stability criterion	20EE002.3	L1
4	State the space equation	20EE002.4	L1
5	List any two applications of state feedback controllers	20EE002.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain successive approximation type (Sample and Hold circuit) analog to digital converters with neat schematic diagram	8M	20EE002.1	L2
6 (b)	Explain digital control systems with an example	4M	20EE002.1	L2
	OR			
7	Explain sampling theorem with neat diagram	12M	20EE002.1	L2
8	Find the inverse z transform of $2z^3 + Z / (z-2)^2 (Z-1)$	12M	20EE002.2	L3
	OR			
9	Explain the Z transform method for solving difference equations	12M	20EE002.2	L2
10(a)	Determine the stability of the system using Jury's stability test for the characteristic equation $P(Z)= Z^4 - 1.2Z^3 + 0.07Z^2 - 0.08=0$	8M	20EE002.3	L3
10 (b)	Write short note on the stability analysis of digital control system using Routh Hurwitz criterion	4M	20EE002.3	L2
	OR			
11 (a)	Explain the mapping between the S-Plane and Z-Plane	8M	20EE002.3	L2
11 (b)	Explain about complimentary strips	4M	20EE002.3	L2
12	Given the pulse-transfer function, determine whether the system is completely state controllable and state observable. $\frac{Y(z)}{U(z)} = \frac{z^{-1}(1+0.8z^{-1})}{1+1.3z^{-1}+0.4z^{-2}}$	12M	20EE002.4	L3
	OR			
13 (a)	Discuss controllability and observability with an example using a 3X3 matrix	6M	20EE002.4	L3
13 (b)	Compare the stability analysis of discrete control system using (i) Routh stability criteria (ii) Jury's Stability criteria	6M	20EE002.4	L2
14 (a)	Explain the concept of state feedback controllers	4M	20EE002.5	L2
14 (b)	State and prove the necessary and sufficient condition for arbitrary pole-placement	8M	20EE002.5	L2
	OR			

15	Consider the system x(k+1) = Gx(k) +Hu(k) $G = \begin{bmatrix} 0 & 1 \\ -0.16 & -1 \end{bmatrix}, H = \begin{bmatrix} 0 \\ 1 \end{bmatrix}$ Determine a suitable state feedback gain matrix 'k' such that the system will have the closed loop poles at $Z = 0.5 \pm j0.5$	12M	20EE002.5	L3

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ANSWER KEY AND SCHEME OF EVALUATION A.Y : 2022-2023

Course code : 20EE002.
Course Title : Digital control systems.

Part A :

Q1. List any four Merits of DCS (2M.)

① Digital components are less suseptable to aging and environment variation.

② Digital processor are more compact and light weighted.

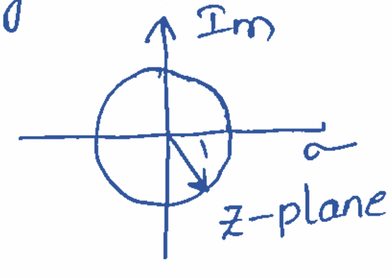
③ They are growing cheaper in cost.

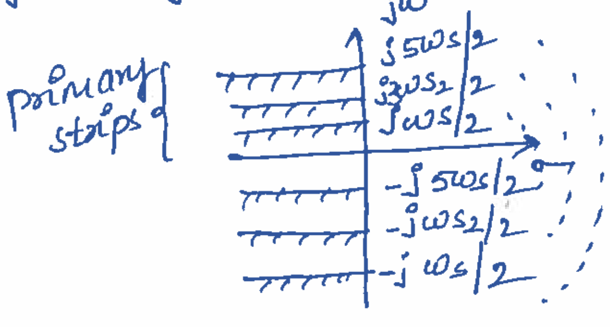
④ Allow more flexibility in programming.

Q2) Recall primary strips (2M)

We know that $|z| = e^{\sigma T}$ the angle of z varies from $-\pi$ to π from $-\infty$ to ∞. consider a respective point on the $j\omega$ axis in the s-plane. as the point moves from $-j\omega_s/2$ to $j\omega_s/2$. where $\omega_s = 2\pi/t$ is the sampling frequency we have $|z| = 1$ and $\angle z$ varies from $-\pi$ to π in the counter clockwise direction. It is clear that each strip of width ω_s in the left half of the s-plane maps into the inside of the unit circle.

The primary strips extend from $j\omega = -j\frac{1}{2}\omega_s$ to $j\frac{1}{2}\omega_s$.





(03) Define Routh's stability criterion — (2M)

Routh's - stability criterion is an algebric procedure for determining whether a polynomial has any zeros in the right half - plane.

$$a_n z^n + a_{n-1} z^{n-1} + a_{n-2} z^{n-2} + \cdots + a_z + a_0 = 0; \quad a_n > 0$$

	odd	even
w^3	a_n	a_{n-1}
w^2	a_{n-1}	a_{n-3}.
w^1	b_1	b_2
w^0	c_1	c_2.

(04) State the space equation — (2M)

The equation that represents the internal state of a system

$$\dot{x} = Ax(t) + Bu(t)$$
$$y = Cx(t) + Du(t)$$

where $\dot{x}$ is a state equation, y is an output equation, A is the system matrix, B is control matrix, D is the feed forward matrix.

05. List any two applications of state feed back controllers. → (2M)

* feed back controllers are used in the

* Boilers, power plants

*. Substations.

* control systems.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

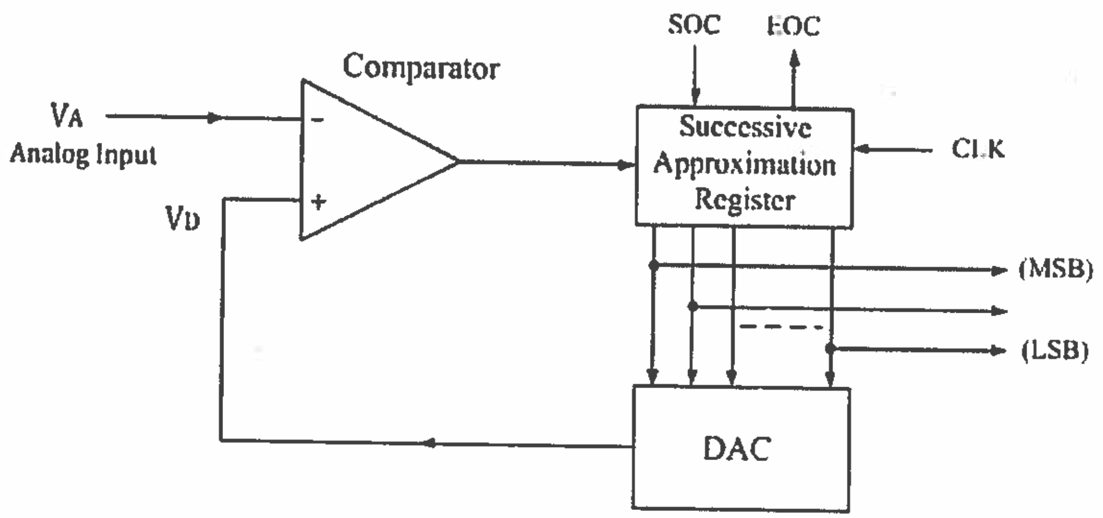
6 (a)Explain successive approximation type (Sample and Hold circuit) analog to digital converters with neat schematic diagram (8M)

Ans: Successive Approximation type ADC is the most widely used and popular ADC method. The conversion time is maintained constant in successive approximation type ADC, and is proportional to the number of bits in the digitaloutput, unlike the counter and continuous type A/D converters. The basic principle of this type of A/D converter is that the unknown analog input voltage is approximated against an n-bit digital value by trying one bit at a time, beginning with the MSB. The principle of successive approximation process for a 4-bit conversion is explained here. This type of ADC operates by successively dividing the voltage range by half, as explained in the following steps.

1) The MSB is initially set to 1 with the remaining three bits set as 000. The digital equivalent voltage is compared with the unknown analog input voltage.

(2) If the analog input voltage is higher than the digital equivalent voltage, the MSB is retained as 1 and the second MSB is set to 1. Otherwise, the MSB is set to 0 and the second MSB is set to 1. Comparison is made as given in step (1) to decide whether to retain or reset the second MSB

The functional block diagram of successive approximation type of ADC is shown below.



It consists of a successive approximation register (SAR), DAC and comparator. The output of SAR is given to n-bit DAC. The equivalent analog output voltage of DAC, VD is applied to the non-inverting input of the comparator. The second input to the comparator is the unknown analog input voltage VA. The output of the comparator is used to activate the successive approximation logic of SAR. When the start command is applied, the SAR sets the MSB to logic 1 and other bits are made logic 0, so that the trial code becomes 1000.

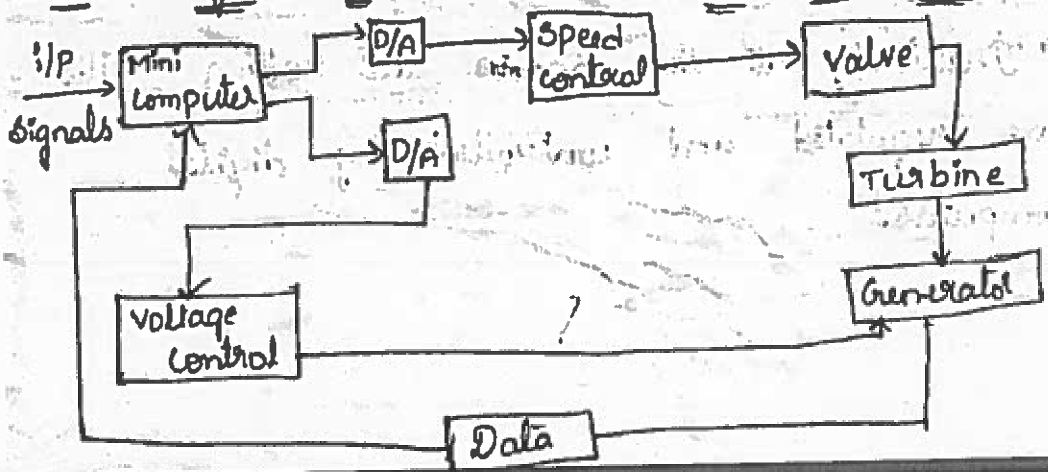
Advantages:
1 Conversion time is very small.
2 Conversion time is constant and independent of the amplitude of the analog input signal VA.
Disadvantages:
1 Circuit is complex.
2 The conversion time is more compared to flash type ADC

6 (b) Explain digital control systems with an example 4M

→ DCS is used in chemical plants, petro chemical and machinaries, pulp and paper mills, boiler controller, Power plant systems, Nuclear power plant systems.

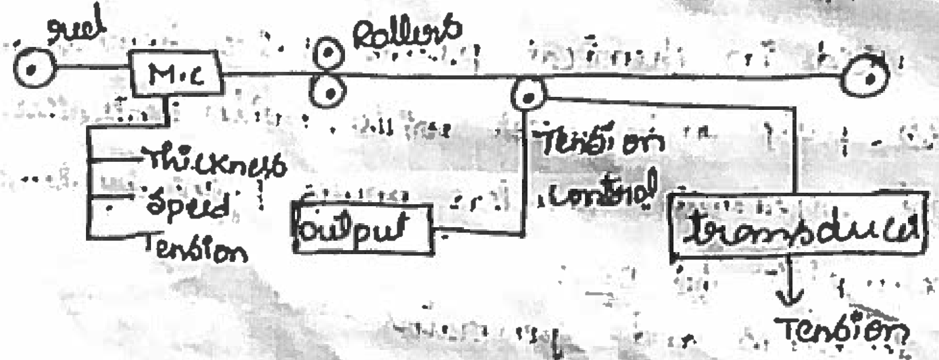
Typical Examples for DCS:
(i) DCS at turbines and generators
(ii) DCS at rolling mills.
(iii) DCS at single axis auto pilot control systems.

Block diagram for ~~turbine~~ and Boiler and Generator:



Here the block diagram mainly tells the working of boiler and generators where the mini computer is working as digital control system the input signal which is given is analog using D/A converter the signal is converted as digital and given to the next blocks where the spped, value of steam is been checked by using the reference digital signal and tha data is again feedback to the minicomputer in the form of analog signal.
The procedure is applicable for the auto-pilot system and the rolling mills.

Rolling Mill:



→ Many industrial processes are controlled and monitored by digital computers and digital transducer.

→ Practically all the modern steel rolling mills are regulated and controlled by digital computers.

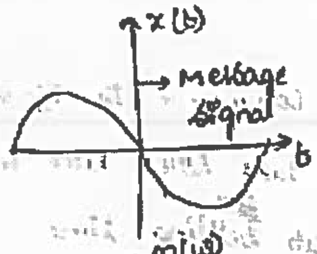
7 Explain sampling theorem with neat diagram (12)

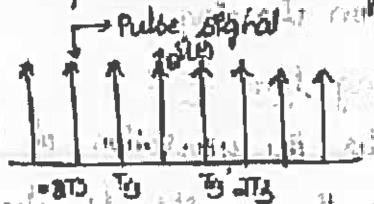
- Statement: A continuous time signal can be represented and its samples can be recovered back when sampling frequency fs is greater than equal to twice of highest frequency component of message signal i.e, $f_s \geq 2f_m$.

Proof: Consider a continuous time signal $x(t)$ the spectrum of $x(t)$ is a band limited to f_m i.e, spectrum of $x(t)$ is 0.

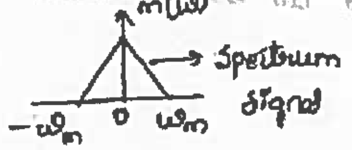
Sampling of input signal $x(t)$ can be obtained by multiplying $x(t)$ with an impulse $\delta(t)$ with a time period t_s.

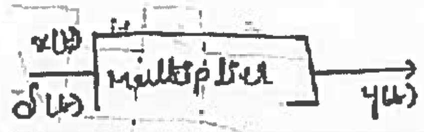
The output of multiplier is a discrete signal called sampled signal which is represented with $y(t)$.

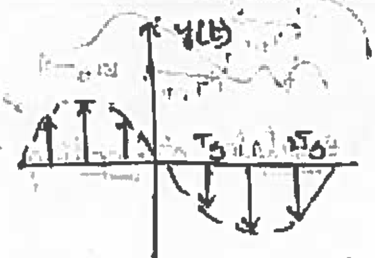




Sampling frequency $= \frac{2\pi}{T_s}$







Sampled Signal:

$$y(t) = x(t)\,\delta(t) \longrightarrow ①$$

The trignometric fourier series representation of $\delta(t)$ is

$$\delta(t) = a_0 + \sum_{n=1}^{\infty} (a_n \cos n\omega_s t + b_n \sin n\omega_s t) \longrightarrow ②$$

where, $a_0 = \frac{1}{T_s}\int_{-T/2}^{T/2} \delta(t)\,dt$,

$$= \frac{1}{T_s}\int_{-T/2}^{T/2} \delta(0)\,dt$$

$$= \frac{1}{T_s}\int_{-T/2}^{T/2} (1)\,dt = \frac{1}{T_s}$$

Even component :

$$a_n = \frac{2}{T_s}\int_{-T/2}^{T/2} \delta(t)\cos n\omega_s t$$

$$= \frac{2}{T_s}\int_{-T/2}^{T/2} \delta(0)\cdot\cos n\omega_s(0)\cdot dt = 2/T_s$$

Odd component :

$$b_n = \frac{2}{T_s}\int_{-T/2}^{T/2} \delta(t)\sin n\omega_s t$$

$$= \frac{2}{T_s}\int_{-T/2}^{T/2} \delta(t)\cdot\sin n\omega_s(0) = 0$$

Substitute the values in eq -②

$$\delta(t) = \frac{1}{T_s} + \sum_{n=1}^{\infty}\frac{2}{T_s}\cos n\omega_s t + b_n(0).$$

Substitute the value of $\delta(t)$ in eq ①

$$y(t) = x(t)\left[\frac{1}{T_s} + \sum_{n=1}^{\infty}\left(\frac{2}{T_s}\cos n\omega_s t\right)\right]$$

$$= \frac{1}{T_s}\left[x(t) + 2\sum_{n=1}^{\infty}(\cos n\omega_s t)\,x(t)\right]$$

$$y(t) = \frac{1}{T_s}\left[x(t) + 2\cos\omega_s t\,x(t) + 2\cos 2\omega_s t\,x(t) + 2\cos 3\omega_s t\,x(t)\right]$$

Take fourier series on both sides.

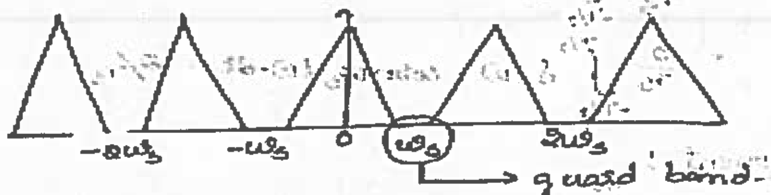
$$y(\omega) = \frac{1}{T_s}\left[x(\omega) + x(\omega-\omega_s) + x(\omega+\omega_s) + x(\omega-2\omega_s) + x(\omega+2\omega_s) + \cdots\right]$$

$$\boxed{Y(\omega) = \frac{1}{T_s}\sum_{n=-\infty}^{\infty} X(\omega - n\omega_s)}$$

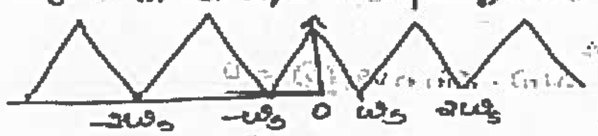
where $n = 0, 1, 2, 3$ —

To reconstruct $x(t)$ you must recover input signal spectrum $X(\omega)$ from sampled signal spectrum $Y(\omega)$.

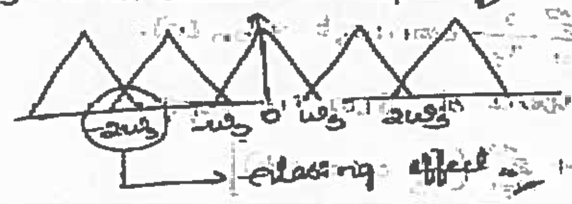
If $f_s > 2f_m$ (over sampling):



If $f_s = 2f_m$ (Perfect sampling):



If $f_s < 2f_m$ (under sampling)



9 Explain the Z transform method for solving difference equations (12M)

The Z-transform plays a vital role in the field of communication Engineering and control Engineering, especially in digital signal processing. Laplace transform and Fourier transform are the most effective tools in the study of continuous time signals, where as Z –transform is used in discrete time signal analysis. The application of Z – transform in discrete analysis is similar to that of the Laplace transform in continuous systems. Moreover, Z-transform has many properties similar to those of the Laplace transform. But, the main difference is Z-transform operates only on sequences of the discrete integer-valued arguments. This chapter gives concrete ideas about Z-transforms and their properties. The last section applies Z-transforms to the solution of difference equations.

Difference equations arise naturally in all situations in which sequential relation exists at various discrete values of the independent variables. These equations may be thought of as the discrete counterparts of the differential equations. Z-transform is a very useful tool to solve these equations.

Formula:

$$x(k) + a_1 x(k-1) + \cdots + a_n x(k-n) = b_0 u(k) + b_1 u(k-1) + \cdots + b_n u(k-n)$$

where $u(k)$ and $x(k)$ are sample of input respectively

Discrete function	Z-transform
$x(k+4)$	$z^4 X(z) - z^4 x(0) - z^3 x(1) - z^2 x(2) - z x(3)$
$x(k+3)$	$z^3 X(z) - z^3 x(0) - z^2 x(1) - z x(2)$
$x(k+2)$	$z^2 X(z) - z^2 x(0) - z x(1)$
$x(k+1)$	$z X(z) - z x(0)$
$x(k)$	$X(z)$
$x(k-1)$	$z^{-1} X(z)$
$x(k-2)$	$z^{-2} X(z)$
$x(k-3)$	$z^{-3} X(z)$

⑧ Find the inverse z-transform of $\frac{2z^3+z}{(z-2)^2(z-1)}$ → 12M.

Sol: Given $F(z) = \frac{2z^3+z}{(z-2)^2(z-1)}$.

$$\frac{F(z)}{z} = \frac{2z^2+1}{(z-2)^2(z-1)}$$

on taking partial fractions.

$$\frac{2z^2+1}{(z-2)^2(z-1)} = \frac{A}{(z-2)} + \frac{B}{(z-2)^2} + \frac{C}{(z-1)}$$

$$\frac{2z^2+1}{\cancel{(z-2)^2(z-1)}} = \frac{A(z-2)(z-1)+B(z-1)+C(z-2)^2}{\cancel{(z-2)^2(z-1)}}$$

$$2z^2+1 = A(z-2)(z-1)+B(z-1)+C(z-2)^2$$

$$2z^2+1 = A[z^2-z+2z+2]+B[z-1]+C(z-2)^2$$

$$\Rightarrow A[z^2+z+2]+B[z-1]+C(z^2+4-4z)$$

on comparing z^2 coefficent.

$$2 = A+C$$

on comparing z coefficent

$$0 = A+B+4C$$

on comparing constants.

$$1 = 2A-B+4C$$

on solving the equation.

$$A = -3/2, \quad B = -7/2, \quad C = -1/2.$$

on substituting in the above eq.

$$\frac{F(z)}{z} = \frac{A}{(z-2)} + \frac{B}{(z-2)^2} + \frac{C}{(z-1)}$$

$$F(z) = A\cdot\frac{z}{(z-2)} + B\cdot\frac{z}{(z-2)^2} + C\cdot\frac{z}{(z-1)}$$

$$F(z) = -3/2\cdot\frac{z}{(z-2)} - \frac{7}{2}\cdot\frac{z}{(z-2)^2} - 1/2\,\frac{z}{(z-1)}$$

Applying I.Z.T.

$$F(n) = -3/2\cdot(2)^n - 7/2\,(2)^{n-1}u(n) - \frac{1}{2}(-1)^n$$

10(a) Determine the stability of the system using Jury's stability test for the characteristic eq (6)

$P(z) = z^4 - 1.2z^3 + 0.07z^2 - 0.08 = 0$ —— [8M]

Sol: Given characteristic eq

$$P(z) = z^4 - 1.2z^3 + 0.07z^2 - 0.08 = 0 \longrightarrow (1)$$

Generalized ch. eq is $P(z) = a_0 z^n + a_1 z^{n-1} + a_2 z^{n-2} \cdots + a_{n-1} z + a_n = 0 \longrightarrow (2)$

→ Total no. of rows $2n-3 = 2(4)-3 = 5$.

Comparing eq-(1) and eq-(2)

$n=4$, $a_0 = 1$, $a_1 = -1.2$, $a_2 = 0.07$, $a_3 = 0$, $a_n = a_4 = -0.08$

→ Conditions :- (i) $|a_0| > |a_n| = |1| > |-0.08|$

condition satisfied.

(ii) $P(z)|_{z=1} > 0$

$P(1) = (1)^4 - 1.2(1)^3 + 0.07(1)^2 - 0.08$

$= -0.21 < 0$

Condition not satisfied.

(iii) $P(z)|_{z=-1} > 0$ n = even

$P(-1) = (-1)^4 - 1.2(-1)^3 + 0.07(-1)^2 - 0.08$

$= 1 + 1.2 + 0.07 - 0.08$

$= 2.19 > 0$ condition satisfied.

iv) $|b_{n-1}| > |b_0|$

row	z^0	z^1	z^2	z^3	z^4
1	-0.08	0	0.07	-1.2	1
2	1	-1.2	0.07	0	-0.08
3	b_3	b_2	b_1	b_0	
4	b_0	b_1	b_2	b_3	
5	q_2	q_1	q_0		

$$b_3 = \begin{vmatrix} -0.08 & 1 \\ 1 & -0.08 \end{vmatrix} = -0.99.$$

$$b_2 = \begin{vmatrix} -0.08 & -1.2 \\ 1 & 0 \end{vmatrix} = 1.2.$$

$$b_1 = \begin{vmatrix} -0.08 & 0.07 \\ 1 & 0.07 \end{vmatrix} = -0.075.$$

$$b_0 = \begin{vmatrix} -0.08 & 0 \\ 1 & -1.2 \end{vmatrix} = 0.096.$$

$$q_2 = \begin{vmatrix} b_3 & b_0 \\ b_0 & b_3 \end{vmatrix} = \begin{vmatrix} -0.99 & 0.096 \\ 0.096 & -0.99 \end{vmatrix} = 0.97$$

$$q_1 = \begin{vmatrix} b_3 & b_1 \\ b_0 & b_2 \end{vmatrix} = \begin{vmatrix} -0.99 & -0.075 \\ 0.096 & -1.2 \end{vmatrix} = -1.180$$

$$q_0 = \begin{vmatrix} b_3 & b_2 \\ b_0 & b_1 \end{vmatrix} = \begin{vmatrix} -0.99 & 1.2 \\ 0.096 & -0.075 \end{vmatrix} = 0.040.$$

Jury's table.

row	z^0	z^1	z^2	z^3	z^4
1	-0.08	0	0.07	-1.2	1
2	1	-1.2	0.07	0	-0.08
3	-0.99	1.2	-0.07	0.096	
4	0.096	-0.07	1.2	-0.99	
5	0.97	-1.18	-0.04		

(iv) $|b_3| > |b_0| = |-0.99| > |0.096|$
condition satisfied.

Since all the given conditions are not satisfied the given system is not stable.

7

10 (b) Write short note on the stability analysis of digital control system using Routh Hurwitz criterion

(4M)

Routh stability Criterion:- (or) Bilinear transformation:-

→ The method requires transformation from the z-plane to another complex plane (w-plane).

* The bilinear transformation defined by

$$z = \frac{w+1}{w-1}$$

which, when solved for w, gives $w = \frac{z+1}{z-1}$ maps the inside of the unit circle in the z-plane into the left half of the w-plane.

Let, $w = \sigma + j\omega$ σ - real part, ω - imaginary part

since, inside of the unit circle in the z-plane is

$$|z| < 1$$

$$\left|\frac{w+1}{w-1}\right| = \left|\frac{\sigma + j\omega + 1}{\sigma + j\omega - 1}\right| < 1$$

$$\frac{(\sigma+1)^2 + \omega^2}{(\sigma-1)^2 + \omega^2} < 1$$

$$(\sigma+1)^2 + \omega^2 < (\sigma-1)^2 + \omega^2 \rightarrow (\sigma+1)^2 < (\sigma-1)^2$$

$$\sigma^2 + 2\sigma + 1 < \sigma^2 - 2\sigma + 1$$

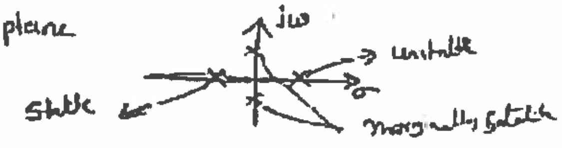
$$2\sigma < -2\sigma \rightarrow \sigma < -\sigma$$

$$2\sigma < 0$$

$$\boxed{\sigma < 0}$$

* Thus, the inside of the unit circle in the z-plane ($|z| < 1$) corresponds to the left half of the w-plane.

* The unit circle in the z-plane is mapped into the imaginary axis in w-plane



Characteristic equation :-

Substitute $z = \frac{w+1}{w-1}$ in $P(z) = 0$

$$P(z) = a_0 z^n + a_1 z^{n-1} + \dots + a_{n-1} z + a_n = 0$$

$$= a_0\left(\frac{w+1}{w-1}\right)^n + a_1\left(\frac{w+1}{w-1}\right)^{n-1} + \dots + a_{n-1}\left(\frac{w+1}{w-1}\right) + a_n = 0$$

multiply the both side with $(w-1)^n$,

$Q(w) = \cancel{P(z)}$

$$a_0(w+1)^n + a_1(w+1)^{n-1}(w-1)^n + \dots + a_n(w-1)^n = 0$$

$$\boxed{Q(w) = b_0 w^n + b_1 w^{n-1} + \dots b_{n-1} w + b_n = 0}$$

apply Routh stability criterion to the $Q(w)$ same manner as in Continuous-time systems; ~~for example 5th order system~~ ~~$Q(w) = b_0w^5 + b_1w^4 + b_2w^3$~~

5th order system

$Q(w) = b_0w^5 + b_1w^4 + b_2w^3 + b_3w^2 + b_4w + b_5 = 0$

Routh array :

w^5	b_0	b_2	b_4
w^4	b_1	b_3	b_5
w^3	c_1	c_2	0
w^2	d_1	d_2	0
w	e_1	0	0
1	f_1		

note :

$f_1 = d_2 = b_5 \dots$

$$c_1 = \frac{-1}{b_1}\begin{vmatrix} b_0 & b_2 \\ b_1 & b_3 \end{vmatrix} ; \quad c_2 = \frac{-1}{b_1}\begin{vmatrix} b_0 & b_4 \\ b_1 & b_5 \end{vmatrix}$$

$$d_1 = \frac{-1}{c_1}\begin{vmatrix} b_1 & b_3 \\ c_1 & c_2 \end{vmatrix} ; \quad d_2 = \frac{-1}{c_1}\begin{vmatrix} b_1 & b_5 \\ c_1 & 0 \end{vmatrix} ; \quad e_1 = \frac{-1}{d_1}\begin{vmatrix} c_1 & c_2 \\ d_1 & d_2 \end{vmatrix}$$

$$f_1 = \frac{-1}{e_1}\begin{vmatrix} d_1 & d_2 \\ e_1 & 0 \end{vmatrix}$$

(8)

11 (a) Explain the mapping between the S-Plane and Z-Plane 8M

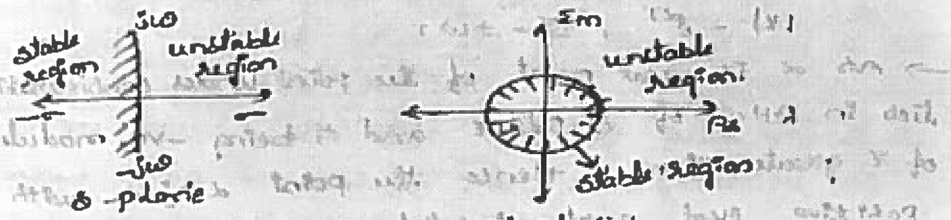
* Maping between Z-plane and S-plane:

In the design of a continuous time control system the location of poles and zeros are very important in predicting the dynamic behaviour of the system. Similarly in design of discrete time control system the location of poles and zeros in the z-plane are very important. Since the complex variables s & z are related by $z = e^{Ts}$. The pole and zero location in the z-plane are related to the poles and zeros in the s-plane. Therefore, the stability of the linear time invarient discrete closed loop system can be determined in terms of the location of poles.

When impulse sampling is incorporated into the process the complex variable z and s are related by the equation $z = e^{Ts}$ ——→ ① with reference to location of roots of the characteristic equation in the z-domain the imaginary axis i.e., jω axis in the s-plane divides stable and unstable region and the corresponding regions in the z-domain can be obtained by puting $s = \pm j\omega t$ in eq ① and plotting the values of z thus obtained in another complex plane called z-plane.

$$Z = e^{\pm j\omega t} = \cos\omega t \pm j\sin\omega t$$

$$|Z| = 1, \quad \angle Z = \pm\omega t \longrightarrow (2)$$



z-plane.

→ the L.H.S of s-plane is maped inside of the unit circle in z-plane.

→ In this region system is stable.

→ The R.H.S of s-plane is maped as the outside of unit circle in z-plane. In this region the system is unstable.

These conditions are verified as follows:

Let $s = -\alpha \pm j\omega$ a point in the LHS of the s-plane the corresponding point in the z-plane is given by

$$Z = e^{(-\alpha + j\omega)T}$$

$$= e^{-\alpha T}(\cos\omega T \pm j\sin\omega T)$$

$$|Z| = e^{-\alpha T}, \quad \angle Z = \pm\omega T.$$

→ As α is real part of the point under consideration lies in L.H.S of s-plane and T being +ve, modulus of Z is less than 1. Hence the point $(-\alpha \pm j\omega)$ with negative real part located in s plane lies inside the unit circle when mapped into z plane.

→ ~~as α is real part of the point~~ under consideration lies

case (ii): Let $s = \alpha \pm j\omega$ a point in the RHS of the s-plane the corresponding point in the z-plane is given by

$$Z = e^{(\alpha \pm j\omega T)}$$

$$= e^{\alpha T}(\cos\omega T \pm j\sin\omega T)$$

$|z| = e^{\alpha T}$, $\angle Z = \pm\omega T$

→ as α is real part of the point under consideration lies in RHS of s-plane and T being -ve, modulus of z greater than 1. Hence the point $\alpha \pm j\omega$ with positive real part located in 's' plane lies outside the unit circle when mapped into z-plane.

→ If the roots of characteristic equation are lies in inside the unit circle. z-plane. It is stable system.

→ If the roots of characteristic equation lies outside the unit circle z-plane. It is unstable system.

11 (b) Explain about complimentary strips 4M

* Complimentary strip and primary strip:

W.K.T

$|z| = \omega T$ the angle of z varies from $-\pi$ to π from $-\infty$ to ∞. consider a respective point on the $j\omega$ axis. In the s-plane as the point moves from $-j\omega_s/2$ to $j\omega_s/2$. where $\omega_s = 2\pi/t$ is a sampling frequency we have $|z| = 1$ and $\angle z$ varies from $-\pi$ to π in the counter clock wise direction in the z-plane.

→ As the respective point moves from $j\frac{1}{2}\omega_s$ to $j\frac{3}{2}\omega_s$ on the $j\omega$ axis. The corresponding point in the z-plane traces out the unit circle in the counter clock wise direction. Thus as the point in the

13 (a) Discuss controllability and observability with an example using a 3X3 matrix 6M

Controllability:

A linear invarient system with state vector 'x' is said to be controllable if the exist signal that will change the state of system from an initial state to some other desired state. For a given state model the controllability can be accessed by using the matrix A and B the necessary and sufficient condition for controlability is given by

$$C_c = [B : AB : A^2B \ldots].$$

Rank must be n and the matrix must be non-singular i.e $|A| \neq 0$

Observability:

A linear invarient system is said to be observable if and only the intial state can be obtained from:

$$C_o = [C^T : A^TC^T : A^T(A^TC^T) \ldots]$$

Rank must be n and the matrix must be non-singular

EXAMPLE:

③ Investigate the controllability and observability of the given system $\dot{x} = \begin{bmatrix} -3 & 1 & 1 \\ -1 & 0 & 1 \\ 0 & 0 & 1 \end{bmatrix} x + \begin{bmatrix} 0 & 1 \\ 0 & 0 \\ 2 & 1 \end{bmatrix} u$, $y = \begin{bmatrix} 0 & 0 & 1 \\ 1 & 1 & 0 \end{bmatrix} x$

Sol: Test for contrability

$$C_c = [B : AB : A^2B]$$

$$AB = \begin{bmatrix} -3 & 1 & 1 \\ -1 & 0 & 1 \\ 0 & 0 & 1 \end{bmatrix} \begin{bmatrix} 0 & 1 \\ 0 & 0 \\ 2 & 1 \end{bmatrix}$$

$$= \begin{bmatrix} 0+0+2 & -3+0+1 \\ 0+0+2 & -1+0+1 \\ 0+0+2 & 0+0+1 \end{bmatrix} = \begin{bmatrix} 2 & -2 \\ 2 & 0 \\ 2 & 1 \end{bmatrix}$$

$$A^2 = \begin{bmatrix} -3 & 1 & 1 \\ -1 & 0 & 1 \\ 0 & 0 & 1 \end{bmatrix} \begin{bmatrix} -3 & 1 & 1 \\ -1 & 0 & 1 \\ 0 & 0 & 1 \end{bmatrix} \Rightarrow \begin{bmatrix} 9+1+0 & -3+0+0 & -3+1+1 \\ 3+0+0 & -1+0+0 & -1+0+1 \\ 0+0+0 & 0+0+0 & 0+0+1 \end{bmatrix}$$

$$\Rightarrow \begin{bmatrix} 8 & -3 & -1 \\ 3 & -1 & 0 \\ 0 & 0 & 1 \end{bmatrix}$$

$$A^2B = \begin{bmatrix} 8 & -3 & -1 \\ 3 & -1 & 0 \\ 0 & 0 & 1 \end{bmatrix} \begin{bmatrix} 0 & 1 \\ 0 & 0 \\ 2 & 1 \end{bmatrix}$$

$$= \begin{bmatrix} 0+0-2 & 8+0-1 \\ 0+0+0 & 3+0+0 \\ 0+0+2 & 0+0+1 \end{bmatrix} = \begin{bmatrix} -2 & 7 \\ 0 & 3 \\ 2 & 1 \end{bmatrix}$$

$$C_c = \begin{bmatrix} 0 & 1 & 8 & -2 & -2 & 7 \\ 0 & 0 & 2 & 0 & 0 & 3 \\ 2 & 1 & 2 & 1 & 2 & 1 \end{bmatrix}_{3\times 6}$$

$$|C_c| = \begin{vmatrix} \overset{+}{1} & \overset{-}{2} & \overset{+}{7} \\ 0 & 2 & 3 \\ 1 & 2 & 1 \end{vmatrix}$$

$$= 1[2-6] - 2[0-3] + 7[0-2]$$

$$= -4 + 6 + 7(-2)$$

$$= -2 - 14 \Rightarrow -16.$$

$$|C_c| \neq 0$$

So it is non singular, the system is controllable

Test for observability

$$C_0 = [C^T : A^TC^T : A^T(A^TC^T)]$$

$$A = \begin{bmatrix} -3 & 1 & 1 \\ -1 & 0 & 1 \\ 0 & 0 & 1 \end{bmatrix}, \quad A^T = \begin{bmatrix} -3 & -1 & 0 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \end{bmatrix}, \quad C^T = \begin{bmatrix} 0 & 1 \\ 0 & 1 \\ 1 & 0 \end{bmatrix}$$

$$A^TC^T = \begin{bmatrix} -3 & -1 & 0 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \end{bmatrix} \begin{bmatrix} 0 & 1 \\ 0 & 1 \\ 1 & 0 \end{bmatrix}$$

$$= \begin{bmatrix} 0+0+0 & -3-1+0 \\ 0+0+0 & 1+0+0 \\ 0+0+1 & 1+1+0 \end{bmatrix} = \begin{bmatrix} 0 & -4 \\ 0 & 1 \\ 1 & 2 \end{bmatrix}$$

$$A^T(A^TC^T) = \begin{bmatrix} -3 & -1 & 0 \\ 1 & 0 & 0 \\ 1 & 1 & 1 \end{bmatrix} \begin{bmatrix} 0 & -4 \\ 0 & 1 \\ 1 & 2 \end{bmatrix} = \begin{bmatrix} 0+0+0 & 12-1+0 \\ 0+0+0 & -4+0+0 \\ 0+0+1 & -4+1+2 \end{bmatrix} = \begin{bmatrix} 0 & 11 \\ 0 & -4 \\ 1 & -1 \end{bmatrix}$$

$$C_0 = \begin{bmatrix} 0 & 1 & 0 & -4 & 0 & 11 \\ 0 & 1 & 0 & 1 & 0 & -4 \\ 1 & 0 & 1 & 2 & 1 & -1 \end{bmatrix}$$

$$= \begin{vmatrix} \overset{+}{1} & \overset{-}{-4} & \overset{+}{11} \\ 1 & 1 & -4 \\ 0 & 2 & -1 \end{vmatrix}$$

$$= 1\,[-1+8] + 4\,[-1+0] + 11\,[2+0]$$

$$= 1\,[7] + 4(-1) + 11(2)$$

$$= 7 - 4 + 22$$

$$= 25$$

$|C_0| \neq 0.$

It is non singular, the system is observable

13 (b) **Compare the stability analysis of discrete control system using** **(6M)**

(i) Routh stability criteria (ii) Jury's Stability criteria

(ii) Jury's stability test :

→ For applying Jurys stability test to a given characteristic equation $P(z)=0$. we construct a table whose elements are based on the coefficients of $P(z)$.

→ assume that the characteristic equation $P(z)$ is polynomial in z-plane.

$$P(z)=a_0 z^n + a_1 z^{n-1} + a_2 z^{n-2} + \dots + a_{n-1} z^1 + a_n = 0$$

where a_0 is always greater than 0 $(a_0>0)$

→ the Jurys stable becomes as given below.

Row	z^0	z^1	z^2	z^3	....	z^{n-2}	z^{n-1}	z^n
1	a_n	a_{n-1}	a_{n-2}	a_{n-3}	....	a_2	a_1	a_0
2	a_0	a_1	a_2	a_3	....	a_{n-2}	a_{n-1}	a_n
3	b_{n-1}	b_{n-2}	b_{n-3}	b_{n-4}	....	b_1	b_0	
4	b_0	b_1	b_2	b_3	....	b_{n-2}	b_{n-1}	
5	c_{n-2}	c_{n-3}	c_{n-4}	c_{n-5}	....	c_0		
6	c_0	c_1	c_2	c_3	....	c_{n-2}		
⋮	⋮	⋮	⋮	⋮				
2n-5	p_3	p_2	p_1	p_0				
2n-4	p_0	p_1	p_2	p_3				
2n-3	q_2	q_1	q_0					

we will stop the jury's table at 2n-3

$$b_k = \begin{vmatrix} a_n & a_{n-1-k} \\ a_0 & a_{k+1} \end{vmatrix} \quad k=0,1,2,\dots,n-1$$

$$c_k = \begin{vmatrix} b_{n-1} & b_{n-2-k} \\ b_0 & b_{k+1} \end{vmatrix} \quad k=0,1,2,\dots,n-2$$

$$q_k = \begin{vmatrix} p_3 & p_{n-2} \\ p_0 & p_{k+1} \end{vmatrix} \quad k=0,1,2.$$

→ procedure is continue until ~~and~~ 2n-3 row is reached which contains exactly 3 elements.

→ According to jurys table the system is said to

Routh stability criterion:- (or) Bilinear transformation:-

→ The method requires transformation from the z-plane to another complex plane (w-plane).

* The bilinear transformation defined by

$$z = \frac{w+1}{w-1}$$

which, when solved for w, gives $w = \frac{z+1}{z-1}$

maps the inside of the unit circle in the z-plane into the left half of the w-plane.

Let, $w = \sigma + j\omega$ σ - real part, ω - imaginary part

Since, inside of the unit circle in the z-plane is

$$|z| < 1$$

$$\left|\frac{w+1}{w-1}\right| = \left|\frac{\sigma+j\omega+1}{\sigma+j\omega-1}\right| < 1$$

$$\frac{(\sigma+1)^2+\omega^2}{(\sigma-1)^2+\omega^2} < 1$$

$$(\sigma+1)^2+\omega^2 < (\sigma-1)^2+\omega^2 \rightarrow (\sigma+1)^2 < (\sigma-1)^2$$

$$\sigma^2+2\sigma+1 < \sigma^2-2\sigma+1$$

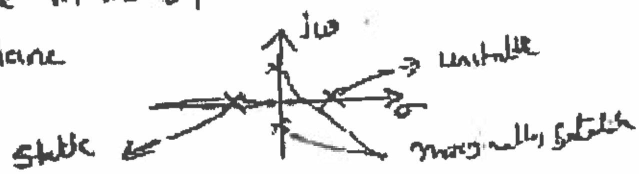
$$2\sigma < -2\sigma \rightarrow \sigma < -\sigma$$

$$2\sigma < 0$$

$$\boxed{\sigma < 0}$$

* Thus, the inside of the unit circle in the z-plane ($|z|<1$) corresponds to the left half of the w-plane.

* The unit circle in the z-plane is mapped into the imaginary axis in w-plane



14 (a) Explain the concept of state feedback controllers **4M**

*Concept of state FB controller (or) what is meant by state FB controller.

A system represented in the state space form can be designed by use of state F.B, where the state of system are considered. The state space model is represented as

$$x(K+1) = A x(K) + B u(K)$$
$$y(K) = C x(K) \longrightarrow ①$$

In a state feedback design all the state variables are measurable and they are feedback to the input of the system in order to place the closed loop poles. The eigen values of the system which controlls the behaviour of the system, for example let the control i/p which is feedback to the system is

$$u(K) = -K x(K) \longrightarrow ②$$

substitute eq ② in eq ①

$$x(K+1) = A x(K) - B K x(K)$$
$$= x(K) [A - BK]$$

where, K is the state feedback gain matrix.

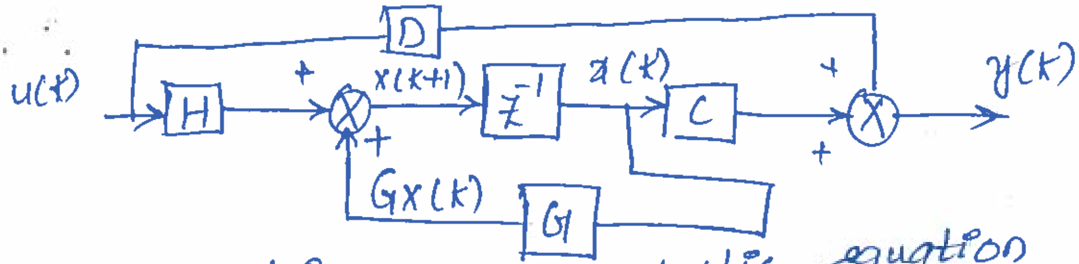
The value of K should be choose such that eigen values of [A-BK] are with in the unit circle. So, the internal stable of the system is achieved.

14(b) state and prove the necessary and sufficienct condition for arbitary pole placement. [8M]

A: System must be controllable.

state space equations with a state feed back controller.

$$x(K+1) = G x(K) + H U(K) \longrightarrow ①$$
$$y(K) = C x(K) + D U(K) \longrightarrow ②$$



System matrix = G characteristic equation $|zI - G| = 0$

Take state feed back

$$U(k) = -Kx(t) \longrightarrow (3)$$

$X(k)$ — state vector with state variables

$$X(k) = \begin{bmatrix} x_1(t) \\ x_2(t) \\ \vdots \\ x_n(k) \end{bmatrix}$$

$$u[(t)] = -[K][x(k)]$$

Substitute Eq-(3) in Eq-(1) $\quad u(t) = -Kx(k)$

$$x(k+1) = Gx(t) - HKx(k)$$
$$= (G - HK)x(k)$$

Modified system matrix $\hat{G} = (G - HK)$

The characterstic eq: $|zI - \hat{G}| = 0$

$$|zI - G + HK| = 0 \longrightarrow (3)$$

For desired location of poles.

Let $P_1, P_2 \cdots P_n$ are poles of n^{th} order.

ch. eq $\quad (z - P_1)(z - P_2) \cdots (z - P_n) = 0$

$$z^n + \alpha_1 z^{n-1} + \alpha_2 z^{n-2} + \cdots + \alpha_{n-1} z + \alpha_n = 0 \longrightarrow (4)$$

compare eq-(4) with specified equation

the eq (3) and eq-(4)

$$|zI - G + HK| = 0$$

$$K = [k_1 \quad k_2 \quad k_3 \cdots k_n]_{1 \times n}.$$

→ The K matrix is depending on the s/m matrix G.

(15) Consider the system $x(k+1) = Gx(k) + Hu(k)$

$$G = \begin{bmatrix} 0 & 1 \\ -0.16 & -1 \end{bmatrix} \quad H = \begin{bmatrix} 0 \\ 1 \end{bmatrix}$$

Determine a suitable state feedback matrix "k" such that the system will have closed loop poles at $0.5 \pm j0.5$. → [12M]

Sol: Given $G = \begin{bmatrix} 0 & 1 \\ -0.16 & -1 \end{bmatrix}$, $H = \begin{bmatrix} 0 \\ 1 \end{bmatrix}$

$$[SI - G] = s\begin{bmatrix} 1 & 0 \\ 0 & 1 \end{bmatrix} - \begin{bmatrix} 0 & 1 \\ -0.16 & -1 \end{bmatrix} \Rightarrow \begin{bmatrix} s & 0 \\ 0 & s \end{bmatrix} - \begin{bmatrix} 0 & 1 \\ -0.16 & -1 \end{bmatrix}$$

$$|SI - G| = \begin{bmatrix} s & -1 \\ 0.16 & s+1 \end{bmatrix}$$

$$= s(s+1) + 0.16 = s^2 + s + 0.16$$

Comparing eq with general eq $\lambda^n + a_1\lambda^{n-1} + a_2\lambda^{n-2} = 0$

$$n = 2, \ a_1 = 1, \ a_2 = 0.16$$

Given closed loop poles at $0.5 \pm j0.5$

$$(s - \mu_1)(s - \mu_2)$$

$$(s + 0.5 + j0.5)(s - 0.5 - j0.5)$$

$$s^2 - s - 0.18 = 0$$

Comparing equation with general eq.

$$\lambda^n + \alpha_1\lambda^{n-1} + \alpha_2\lambda^{n-2} = 0$$

$$n = 2, \ \alpha_1 = 1, \ \alpha_2 = -0.18$$

$$k = [\alpha_2 - a_2 \ \vdots \ \alpha_1 - a_1]\,T^{-1}$$

Here T^{-1} is 1 because given H Matrix is Canonical form

$$k = [\alpha_2 - a_2 \ \vdots \ \alpha_1 - a_1]$$

$$k = [-0.34 \ \vdots \ -2]$$

(12) Given the pulse-transfer function, determine whether the system is completely state controllable and state observable $\frac{y(z)}{u(z)} = \frac{z^{-1}(1+0.8z^{-1})}{1+1.3z^{-1}+0.4z^{-2}}$ [12 M]

Sol: Given transfer function $\frac{y(z)}{u(z)} = \frac{z^{-1}(1+0.8z^{-1})}{1+1.3z^{-1}+0.4z^{-2}}$

$$= \frac{\frac{1}{z}\left(1+\frac{0.8}{z}\right)}{1+\frac{1.3}{z}+\frac{0.4}{z^2}} = \frac{\frac{(z+0.8)}{z}}{\frac{z^2+1.3z+0.4}{z^2}} = \frac{z(z+0.8)}{z^2+1.3z+0.4}$$

$$\frac{y(z)}{u(z)} = \frac{z^2+0.8z}{z^2+1.3z+0.4} \longrightarrow (1)$$

Equation (1) is in the form of $\frac{y(z)}{u(z)} = \frac{b_0 z^n + b_1 z^{n-1} + \dots + b_n}{z^n + a_1 z^{n-1} + \dots + a_n}$ $\longrightarrow (2)$

on comparing (1) & (2) we get

$n=2$, $a_1 = 1.3$, $a_2 = 0.4$.

$b_0 = 1$, $b_1 = 0.8$, $b_1 = 0$

Test for controllability.

on substituting $n=2$ we get

$$\begin{bmatrix} x_1(k+1) \\ x_2(k+2) \end{bmatrix} = \begin{bmatrix} 0 & 1 \\ -a_2 & -a_1 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + \begin{bmatrix} 0 \\ 1 \end{bmatrix} x(k)$$

$$y(k) = \begin{bmatrix} b_2 - a_2 b_0 & b_1 - a_1 b_0 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + b_0 u(k)$$

on substituting the values.

$$\begin{bmatrix} x_1(k+1) \\ x_2(k+2) \end{bmatrix} = \begin{bmatrix} 0 & 1 \\ -0.4 & -1.3 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + \begin{bmatrix} 0 \\ 1 \end{bmatrix} u(k)$$

$$y(k) = \begin{bmatrix} -0.4 & -0.5 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + u(k)$$

Test for observability

Representation.

$$\begin{bmatrix} x_1(k+1) \\ x_2(k+1) \\ \vdots \\ x_n(k+1) \end{bmatrix} = \begin{bmatrix} 0 & 0 & \cdots & -a_n \\ 1 & 0 & \cdots & -a_{n-1} \\ \vdots & \vdots & & \\ 0 & 0 & \cdots & -a_1 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \\ \vdots \\ x_n(k) \end{bmatrix} + \begin{bmatrix} b_n - a_n b_0 \\ b_{n-1} - a_{n-1} b_0 \\ b_1 - a_1 b_0 \end{bmatrix} u(k)$$

$$y(k) = [0 \; 0 \; \cdots \; 1] \begin{bmatrix} x_1(k) \\ x_2(k) \\ \vdots \\ x_n(k) \end{bmatrix} + b_0 u(k)$$

for $n = 2$

$$\begin{bmatrix} x_1(k+1) \\ x_2(k+1) \end{bmatrix} = \begin{bmatrix} 0 & -a_2 \\ 1 & -a_1 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + \begin{bmatrix} b_2 & -a_2 b_0 \\ b_1 & -a_1 b_0 \end{bmatrix} u(k)$$

$$y(k) = [0 \; 1] \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + b_0 u(k)$$

on substituting the values.

$$\begin{bmatrix} x_1(k+1) \\ x_2(k+1) \end{bmatrix} = \begin{bmatrix} 0 & -0.4 \\ 1 & -1.3 \end{bmatrix} \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + \begin{bmatrix} -0.4 \\ -0.5 \end{bmatrix} u(k)$$

$$y(k) = [0 \quad 1] \begin{bmatrix} x_1(k) \\ x_2(k) \end{bmatrix} + [1] u(k),$$

Hence the given system is observable as well as controllable.

7/12/22

HOD-ECE 7/12/22

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	ECE	Academic Year	2022 - 2023
Course Code	20EC006	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Electronic Measurements & Instrumentation				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define precision	20EC006.1	L1
2	List any four applications of Spectrum Analyzer	20EC006.2	L1
3	Mention any two difference between general CRO and special purpose CRO	20EC006.3	L1
4	Write any four applications of bridges	20EC006.4	L1
5	Differentiate active and passive transducers	20EC006.5	L2

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Classify and explain thermo-couple type ammeters	6M	20EC006.1	L2
6 (b)	Describe the operation of shunt type ohmmeter	6M	20EC006.1	L2
	OR			
7 (a)	Define fidelity, lag and resolution	6M	20EC006.1	L2
7 (b)	Describe the operation of series type ohmmeter	6M	20EC006.1	L2
8 (a)	Describe the working of AF sine and square wave generator	6M	20EC006.2	L2
8 (b)	Explain the working of function generator with a neat block diagram	6M	20EC006.2	L2
	OR			
9 (a)	Explain working of harmonic distortion analyzer with neat sketch	6M	20EC006.2	L2
9 (b)	Explain the working of RF Spectrum Analyzer with neat block diagram	6M	20EC006.2	L2
10 (a)	Explain the operation of vertical amplifier section	6M	20EC006.3	L2
10 (b)	Draw the block diagram of Sampling oscilloscope and explain its working	6M	20EC006.3	L2
	OR			
11 (a)	Discuss horizontal deflection system	4M	20EC006.3	L2
11 (b)	Explain storage oscilloscope with neat sketch	8M	20EC006.3	L2
12 (a)	Derive the expression for unknown inductance using Maxwell Bridge	6M	20EC006.4	L3
12 (b)	Explain the working principle of schering bridge with necessary equations	6M	20EC006.4	L2
	OR			
13 (a)	Explain Q meter	6M	20EC006.4	L2
13 (b)	Derive the equation for unknown resistance using wheat stone bridge	6M	20EC006.4	L3

14 (a)	Discuss semi-conductor type strain gauge	3M	20EC006.5	L2
14 (b)	Explain the principle, working, construction, characteristics and applications of LVDTs	9M	20EC006.5	L2
	OR			
15	Discuss the measurement of physical parameters i) Force - 2M ii) Pressure (Using resistive pressure transducer) - 4M iii) Velocity (Using moving coil and moving magnet type transducer) - 6M	12M	20EC006.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	ECE		**Academic Year**	2022 - 2023
Course Code	20EC006	**Test Duration**	3 Hrs.	Max. Marks 70	**Semester**	V
Course	**Electronic Measurements & Instrumentation**					

1	Definition for precision	2M
2	Four applications of Spectrum Analyzer	2M
3	Two difference between general CRO and special purpose CRO	2M
4	Four applications of bridges	2M
5	Difference between active and passive transducers	2M
6 (a)	Diagrams and Classification Explanation of thermo-couple type ammeters	2M 4M
6 (b)	Diagram Operation of shunt type ohmmeter	2M 4M
7 (a)	Definition for fidelity Definition for lag Definition for resolution	2M 2M 2M
7 (b)	Diagram Operation of series type ohmmeter	2M 4M
8 (a)	Diagram Working of AF sine and square wave generator	2M 4M
8 (b)	Neat block diagram Working of function generator	2M 4M
9 (a)	Diagram Explanation of harmonic distortion analyzer	2M 4M
9 (b)	Neat block diagram	2M

	Explanation of RF Spectrum Analyzer	4M
10 (a)	Diagram Operation of vertical amplifier section	2M 4M
10 (b)	Block diagram of Sampling oscilloscope Explanation of Sampling oscilloscope	2M 4M
11 (a)	Diagram Discussion on horizontal deflection system	2M 2M
11 (b)	Neat sketches Explanation of storage oscilloscope	3M 5M
12 (a)	Maxwell's Bridge circuit digram Derivation for the expression for unknown inductance	2M 4M
12 (b)	Schering bridge circuit diagram Working principle with necessary equations	2M 4M
13 (a)	Diagram Explanation of Q meter	2M 4M
13 (b)	Circuit diagram of Wheat stone's bridge Derivation for the equation for unknown resistance	2M 4M
14 (a)	Diagram Discuss on semi-conductor type strain gauge	1M 2M
14 (b)	Constructional diagrams Explanation of working of LVDT Characteristics and applications of LVDTs	3M 4M 2M
15	i) Force -Measurement ii) Diagram for resistive pressure transducer Explanation of Pressure measurement iii) Diagrams of moving coil and moving magnet type transducer Explanation of Velocity measurement	2M 1M 3M 2M 4M

PART-A (Short Answer Questions)

1) Define precision.

A measure of the Consistency or repeatability of measurements i.e, Successive reading donot differ is Called precision.

2) List any four applications of Spectrum Analyzer.

(1) Measuring frequency response
(2) Occupied bandwidth.
(3) noise distortion
(4) Interference Sources

3) Mention any two difference between general CRO and Special purpose CRO.

general CRO	Special purpose CRO
(1) They Cannot Storage Signals.	(1) They Can Store Signals.
(2) They donot have memory.	(2) They have memory.
(3) manipulations are not Possible at any time	(3) manipulations are Possible at any time.

4) Write any four applications of bridges.

(1) Instrumentation

(2) filtering

(3) Power Consumption.

(4) finding resistance Values.

5) Differentiate active and passive transducers.

Active Transducers	Passive transducers.
(1) They donot require. additional energy Source.	(1) They require additional energy Source.
(2) Design is Simple	(2) Design is Complicated.

PART-B (Long Answer Questions)

6) a) Classify and explain thermo-Couple type ammeters.

Types of Thermocouples:-

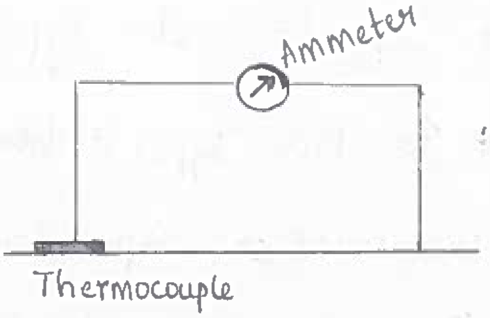
(1) Mutual type thermocouple

(2) Contack type thermocouple

(3) Seperate heater type thermocouple

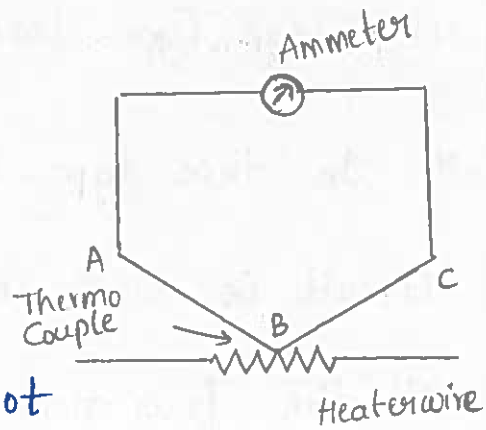
(4) Bridge type thermocouple.

1) Mutual Type Thermocouple :-



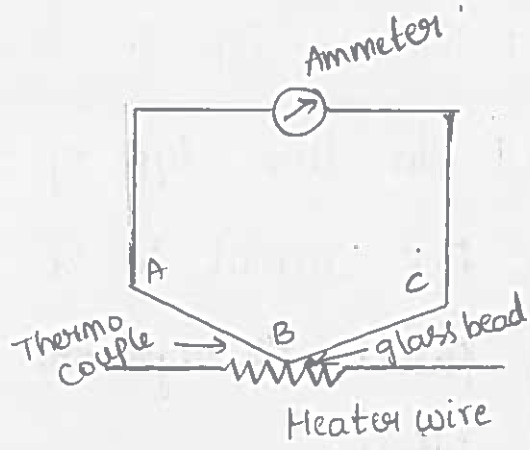
* In this type of thermocouple, the R.F Current to be measured is passed through the thermocouple itself.
* In this type, a Seperate heater wire is not required.
* But the major drawback of the mutual type thermocouple is that ammeter Shunts the thermocouple.

2) Contact type Ammeter / Direct Contact Ammeter :-



* In this type of thermocouple, a Seperate heaterwire is Used.
* So, that R.F Current to be measured is passed through heating wire and not the thermocouple.
* Two thermocouple leads are taken Out which are Used to Conduct heat away from the heating wire.
* It is less Senstive Compared to Mutual - Type Ammeter.

3) Seperate heater type thermocouple :-

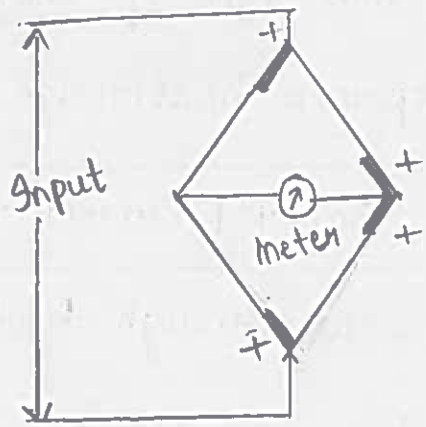


* In this type, a heater wire is Insulated or Seperated from a thermocouple with the help of glass bead.

* Due to this the Insulatument becomes less Sensitive.

* To Increase Sensitivity, the Instrument is placed in Vaccum which avoid losses in the heat radiation.

4) Bridge Type Thermocouple Ammeter :-

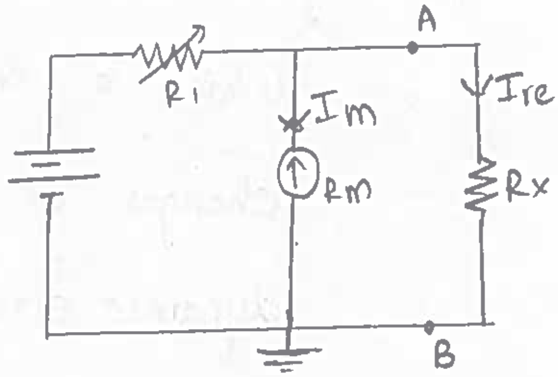


* In this type, a bridge Circuit is Used in which all the four arms Consists Similar thermocouples arranged as shown in the fig.

* This type Shows Sensitivity Same as basic mutual type thermocouple but in bridge type thermocouple Shunting effects are nullified.

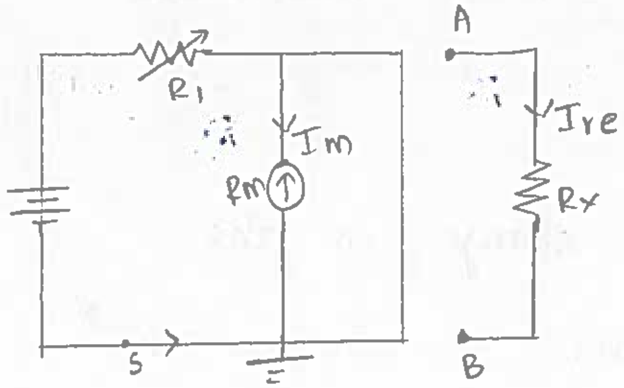
* To Increase the Sensitivity it is placed in Vaccum.

6)
b) Describe the Operation of shunt type ohmmeter.



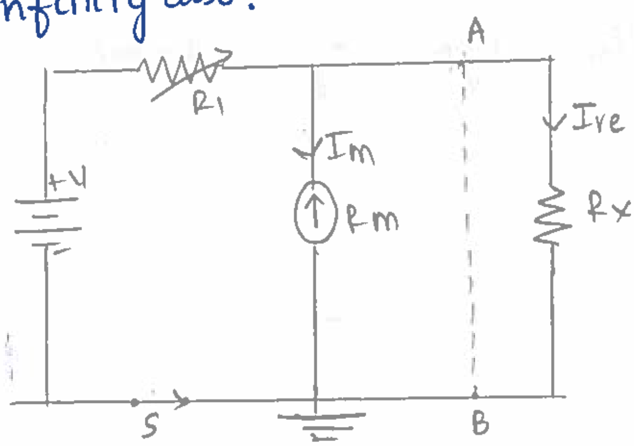
* The Shunt type Ohmmeter Consists of a battery in Series with an adjustable resistance R_1 and a meter movement.

Case(i):- If we Short the terminals of A and B then resistance will be zero.



* The 'I_{re} = 0' as the A & B terminals are Short circuited.

Case(ii): If the A and B terminals are Open circuited. then the resistance will be high and it may be Infinity also.



7)
a) Define fidelity, lag and resolution.

(i) fidelity:- It is defined as the degree to which a measurement System Indicates Changes in measured quantity without dynamic error.

lag :- It is the retardation (or) delay in the response of a measured system to the changes in the measured quantity.

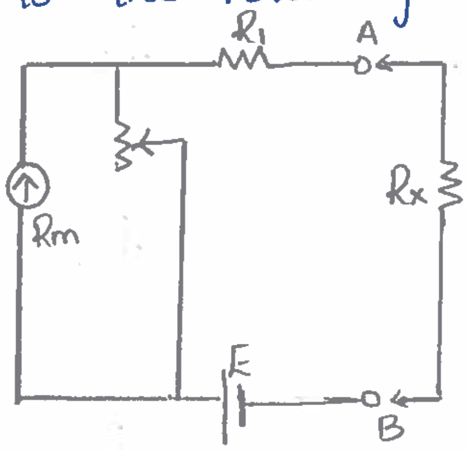
Resolution :- The Smallest change in the measured Variable to which an Instrument will respond is Called as Resolution.

These are the types of Static characteristics.

7(b) Describe the operation of Series type ohmmeter?

1. The Series type ohmmeter consists of a D'Arsenal movement connected in Series with a resistance and a battery to a pair of terminals to which unknown is connected.

2. The current through the movement then depends on the magnitude of the unknown resistor, and the meter indicator is propotional to the value of the unknown provided.



3. Fig shows Simple Single-range Series ohmmeter.

4. In figure
   R_1 = current-limiting resistor
   R_2 = Zero adjust resistor
   E = internal battery
   R_m = internal resistance of the d'Arsenal movement
   R_x = unknown resistor.

5. When the $R_x = 0$, max current flows in circuit, under this condition, shunt resistor R_a is adjusted untill movement indicator full-scale current I_{fsd}.

6. Although the Series type ohmmeter is a popular design and is extensively in portable Instruments for general Service work, it has certain disadvantage

7. Adjustment by R_2 is a superior solution.

8. In the circuit, does not compensate completely for aging of the battery.

The voltage across the shunt is equal to the voltage across the movement and

$$E_{sh} = E_m$$

$$I_2 R_2 = I_{fsd} R_m$$

$$R_2 = \frac{I_{fsd} R_m}{I_2}$$

$$R_2 = \frac{I_{fsd} R_m}{I_t - I_{fsd}}$$

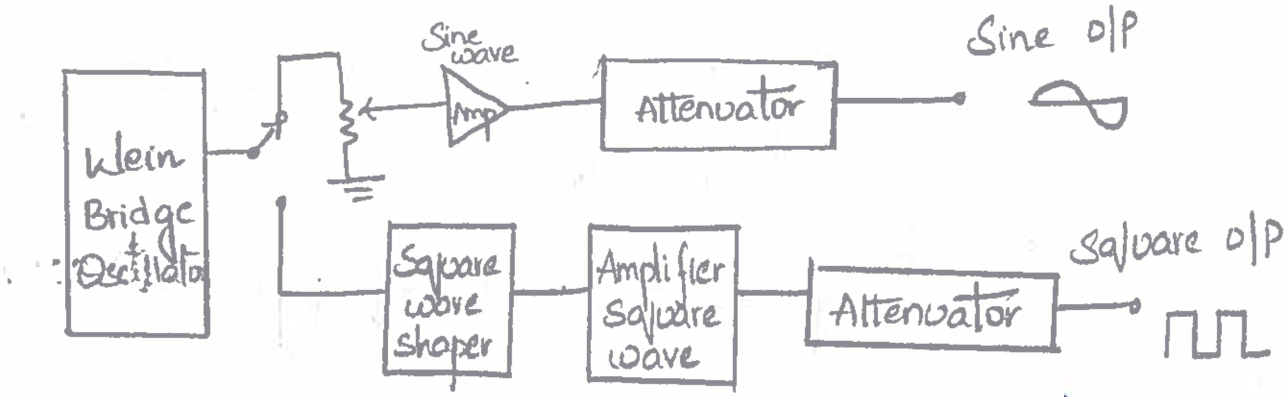
$$R_2 = \frac{I_{fsd} R_h R_m}{E - R_h I_{fsd}}$$

Eq-(1) $$R_1 = R_h - \frac{R_2 R_m}{R_2 + R_m}$$

$$R_1 = R_h - \frac{I_{fsd} R_m R_h}{E}$$

8(a) Describe the working of AF Sine and Square wave generator

A: The block diagram of an AF Sine - Square wave audio oscillator is illustrated in figure



1) The Signal generator is called an oscillator
2) A wein Bridge Oscillator is used in the generator
3) The wein Bridge Oscillator is the best for the audio frequency range.
4) The frequency of oscillations can be changed by varying the capacitance in the oscillator.
5) The frequency can also be controlled in steps by Switching in resistor of different values.
6) At output, we get either a Square (or) Sine wave
7) The output is Varied by means of an attenuator
8) The instrument generates a frequency ranging from 10 Hz to 1MHz continuously Variable in 5 decades with overlapping range.
9) The output Sine wave amplitude can be Varied from 5mv to 5V (rms).

10) The Square wave amplitude can be manged by 0 - 20V

11) It is possible to adjust the symmetry of the Square wave from 30 - 70%

12) The instrument requires only 7W of power at 220V - 50Hz.

→ The front panel of a signal generator consists of the following.

1. Frequency Selector :- It selects the frequency in different ranges and varies it continuously in a ratio of 1:11
2. Frequency Multiplier :- It selects the frequency range over 5 decades from 10 Hz to 1 MHz.
3. Amplitude Multiplier :- It attenuates the sine wave in 3 decades x1, x 0.1 and x 0.01.
4. Variable Amplitude :- It attenuates the Sine wave amplitude continuously.
5. Symmetry control :- It varies the Symmetry of the Square wave from 30% - 70%
6. Amplitude :- It attenuates the Square wave output continuously
7. Function Switch :- It selects either Sine wave (or)

Square wave output.

8) Output available :- This produces Sine wave (or) Square wave output.

9) Sync :- It is used to provide Synchronization of the internal Signal with an external Signal.

8(b) Explain the working of function generator with a neat block diagram?

1) A function generator produces different waveforms of adjustable frequency.

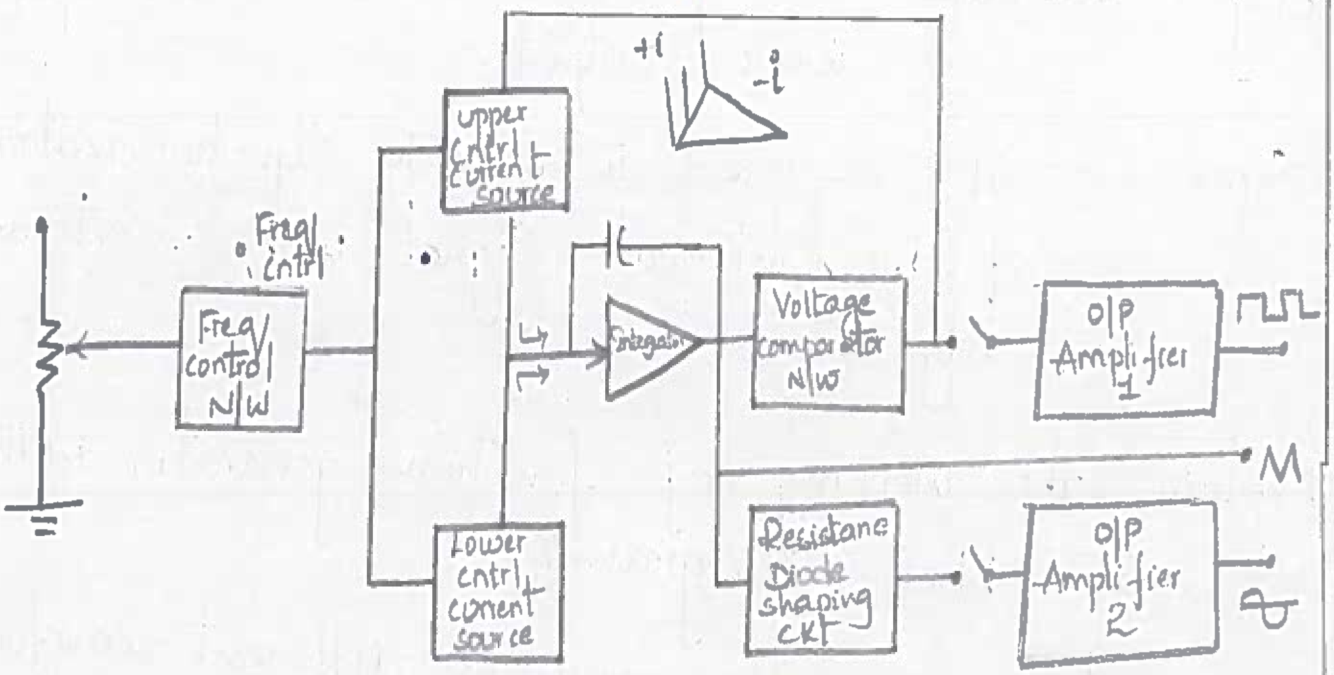
2) The common output wave-form are the Sine, Square, triangle and Sawtooth wave.

3) The frequency may be adjusted from a function of Hertzs to Several hundred KHz.

4) The various outputs can be made available at the Same time.

For Eg:- the generator can provide a Squarewave to test the linearity for an amplifier and Simultaneously provide a Sawtooth to drive the horizontal deflection amplifier of the CRO to provide a visual display.

The block diagram of a function generator is



5. Usually the frequency is controlled by varying the capacitor in the LC (or) RC circuit.

6. In the instrument the frequency is controlled by varying the magnitude of current which drives the integrator.

7. It produces a frequency range of 0.01Hz to 100KHz

8. The frequency controlled voltage regulates two current Sources.

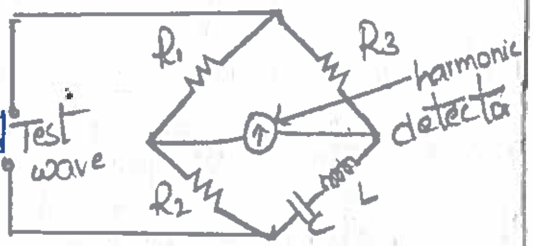
9. The upper current source Supplies constant current to the integrator whose output voltage increases linearly with time

10. The output of integrator is triangular waveform and resistance diode network produces a sine wave of less than % distortion

7(a) Explain working of Harmonic distortion analyzer with neat sketch?

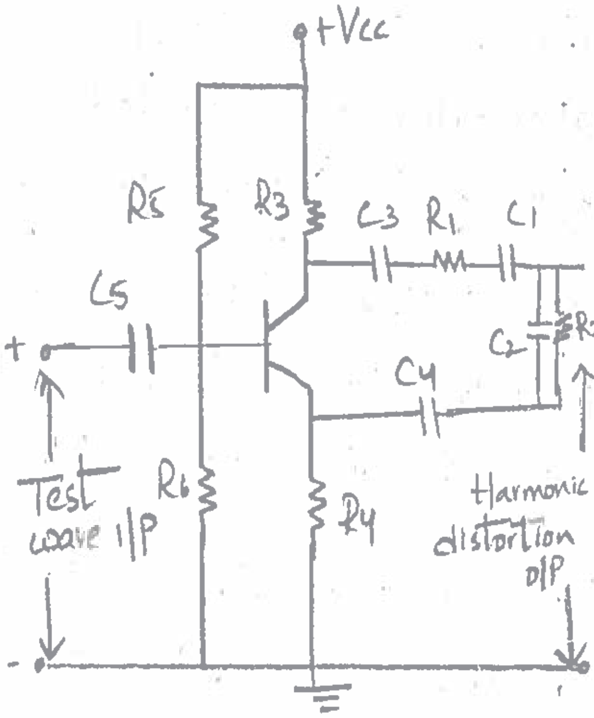
A:- To analyze the Harmonics in an instrument we need harmonic distortion analyzer. It measures total harmonic power present in test wave.

1) Employing a Resonance Bridge



* The Bridge shown in fig is balanced for the fundamental frequency i.e L and C are tuned to fundamental frequency
* The bridge is unbalanced for the Harmonics i.e only Harmonic power will be available at O/P terminal
* If fundamental frequency is changed the bridge has to be balanced again, if L and C are fixed components then this method is Suitable only when test wave has fixed frequency.

2) Wein's Bridge Method



The bridge is balanced for the fundamental frequency. The fundamental energy is dissipated in Bridge circuit elements.

for a Balance fundamental freq

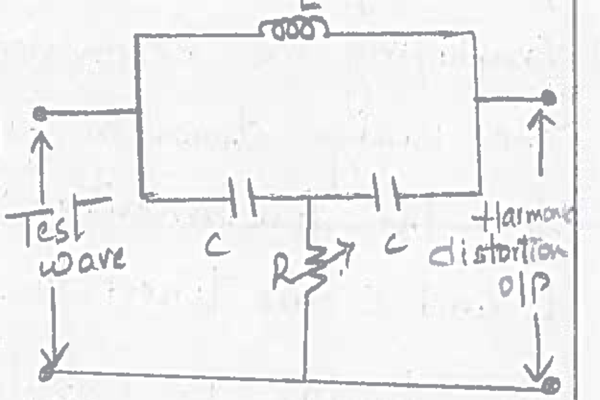
$C_1 = C_2 = C$

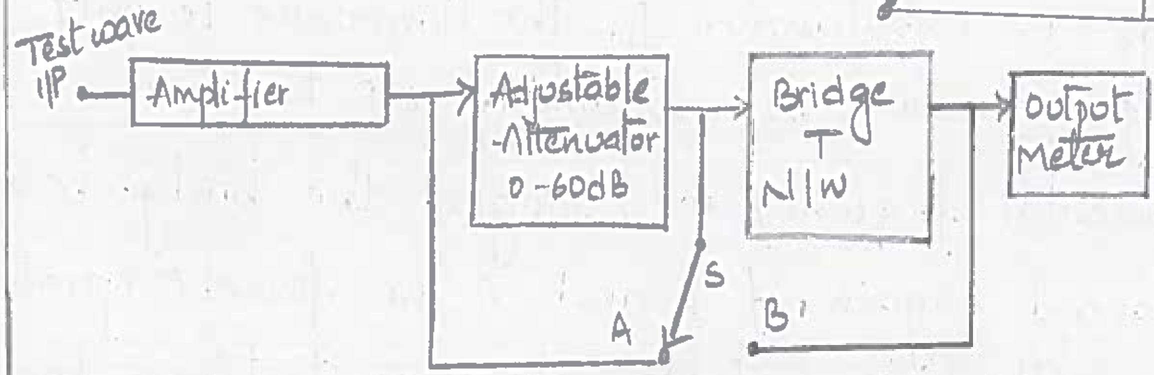
$R_1 = R_2 = R$

$R_3 = 2R_4$

3) Bridged T- Network Method - The Fig shows the L and C are tuned to the fundamental frequency and R is adjusted to bypass fundamental frequency.

* The Tank ckt being tuned to the fundamental frequency and this frequency energy will circulate in the tank. Only harmonic components will reach the output terminals and the distorted output can be measured by the meter.





* The Switch 'S' is first connected to point A, so that the attenuator is excluded and bridge T Network is adjusted for full suppression of fundamental frequency.

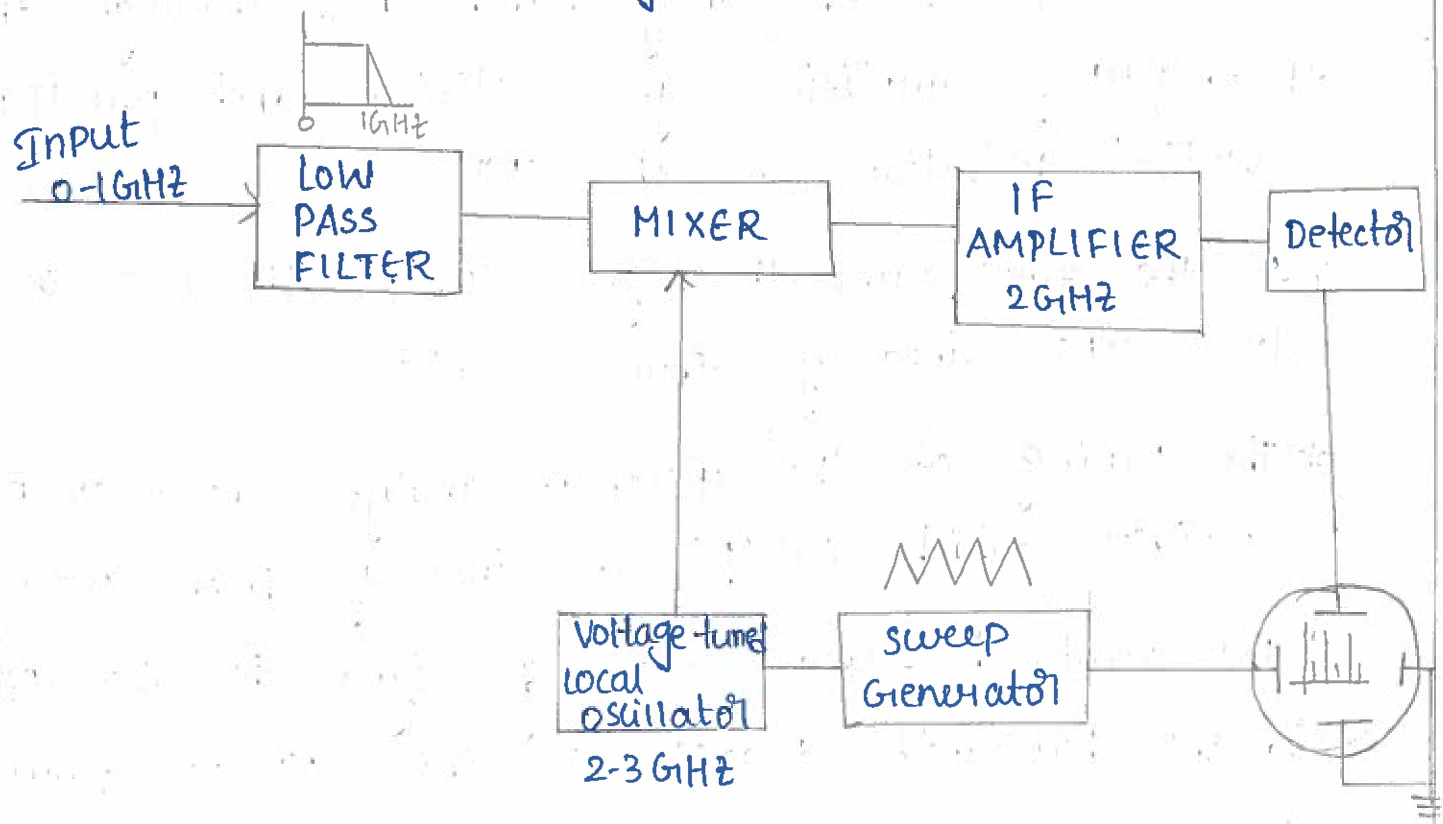
* If Switch 'S' is connected to terminal B i.e the bridge T N/w is excluded. Attenuation is adjusted untill same reading is obtained on the meter.

* The Disadvantage is that they give only the total distortion and not the amplitude of individual distortion components.

9 (b) Explain the working of RF spectrum Analyzer with neat block diagram.

A:- RF Spectrum Analyzer:-

1) The basic block diagram of spectrum analyzer of superheterodyne type covering the range 500 KHz to 1GHz is shown in figure.



RF Spectrum Analyzer

* The input signal is fed into a mixer which is ~~done by~~ driven by local oscillator. This oscillator is linearly tunable electrically over the range 2-3GHz.

* The mixer provides two signals at its output that are proportional in amplitude to the input signal

but of frequencies which are sum and difference of the input signal and local oscillator frequency.

* The IF amplifier is tuned to a narrow band around 2GHz, since the local oscillator is tuned over the range of 2-3 GHz. only inputs that are seperated from the local oscillator frequency by 2GHz will be converted to IF frequency band, pass through the IF frequency amplifier, get rectified and practice a vertical deflection on the CRT.

* As the ~~swit~~ sawtooth signals sweeps, the local oscillator also sweeps linearly from 2-3 GHz.

* The tuning of the spectrum analyzer is a swept receiver which sweeps ~~to~~ linearly from 0 to 1GHz.

* The sawtooth ~~signal~~ scanning signal is also applied to the horizontal plates of the CRT to form the frequency axis.

* The spectrum analyzer are widely used in radars, oceanography, and bio-medical fields.

Q(a) Explain the operation of vertical amplifier section.

A:- VERTICAL AMPLIFIER:-

The sensitivity and frequency bandwidth response characteristics of the oscilloscope are mainly determined by the vertical amplifier. Since the gain BW product is constant, to obtain a greater sensitivity the BW is measured, or vice-versa.

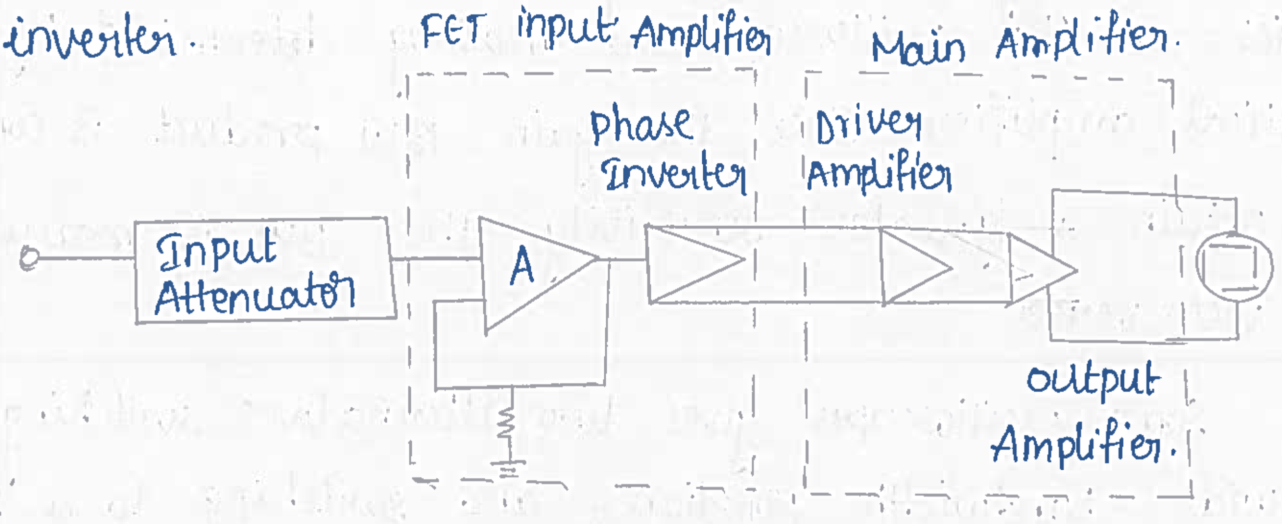
Some oscilloscopes gives two alternatives switching to a wide bandwidth position, and switching to a high sensitivity position.

Block diagram of a vertical Amplifier:-

The vertical amplifier consists of several stages, with fixed overall sensitivity or gain expressed in v/div. The advantage of fixed gain is that the amplifier can be more easily designed to meet the requirements of stability and BW. The vertical amplifier is kept within its signal handling capability of proper selection of the input attenuator switch.

The first element of the pre-amplifier is the input stage. often consisting of a FET source follower whose high input impedance indicates the amplifier from the attenuator.

This FET input stage is followed by a BJT emitter follower, to match the medium impedance of FET output with the low impedance input of the phase inverter.



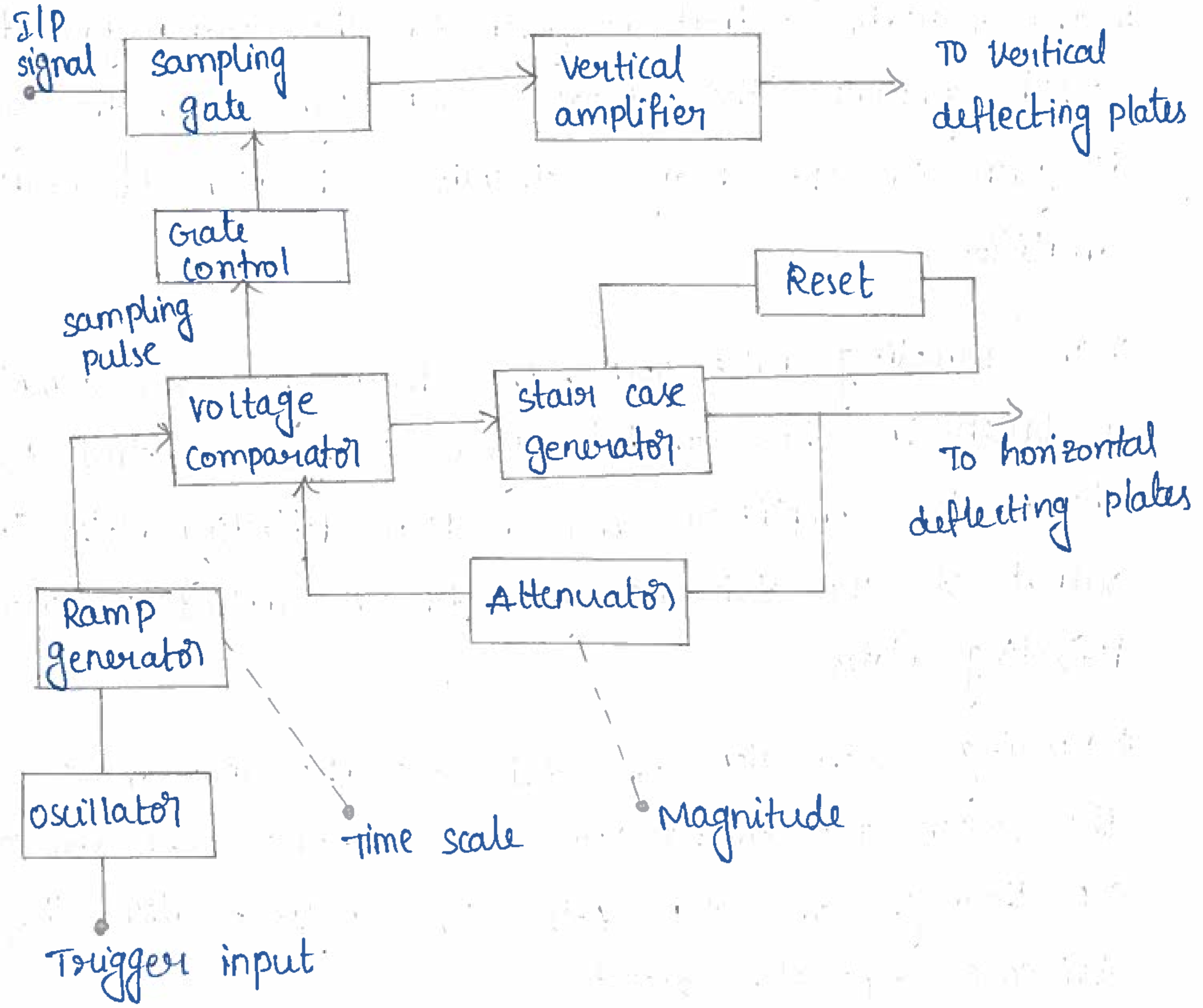
Vertical Amplifier.

The large 2nd harmonic is cancelled out, and greater power output per tube as a result of even harmonic cancellation. In addition, a number of defocussing and non-linear affects are reduced because neither plate is at ground potential.

10 (b) Draw the block diagram of sampling oscilloscope and explain its working.

A: * The staircase generator produces a staircase wave forms which is applied to an attenuator.

* The attenuator controls the magnitude of the staircase and then it is applied to the voltage comparator.



* Another input to the voltage generator is the ramp signal which is the output of ramp generator.

* The voltage comparator compares the two signals and produces the output pulse when two voltages are equal.

* At the start of each sampling cycle a trigger input pulse is generates which activates the blocking oscillator.

* The oscillator output is given to the ramp generator which generates the linear ramp signal and the signal is given to the delayed circuit and then to vertical amplifiers.

* The sampling pulse opens the diode gate and sample is taken in. The sampled signal is then applied to the vertical amplifier and vertical deflecting plates. The output of the staircase generator is applied to horizontal deflecting plates.

* During each step of stair case the spot moves on the screen. The comparator o/p advances the staircase o/p through one step. After 1000 pulses also, the staircase generator resets.

* The step size decreases then input samples are high. So, the resolution of the image is high.

* For best resolution 1000 samples are required.

* It will ~~take~~ measure the voltage, frequency, time period, current, speed, distance etc.

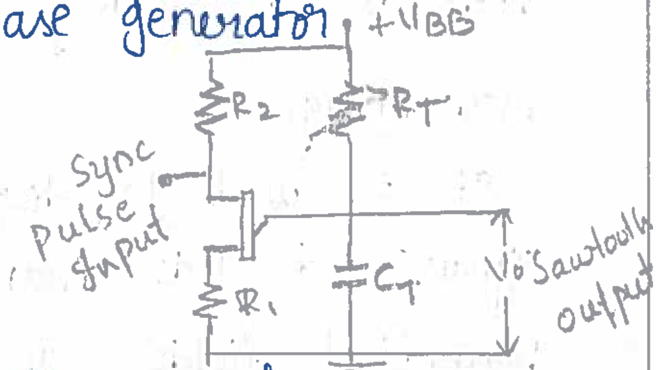
11 (a) Discuss horizontal deflection system?

A:- The horizontal deflecting system of the cathode Ray oscilloscope (CRO) consists of the following stages.

1) sweep generator or time base generator

2) The horizontal amplifier.

3) The trigger circuit



For an oscilloscope to display the waveform under study a voltage which is linearly increasing with time is required. Also the voltage has to fall in its amplitudes at time periods ~~as~~ that are equal or multiples of the input signal time periods. The voltage which is of the sawtooth waveform is applied to the horizontal deflection plates to sweep the beam horizontally. Hence the name sweep voltage or time base voltage.

11(b) Explain storage oscilloscope with neat sketch?

A:- Storage oscilloscope:-

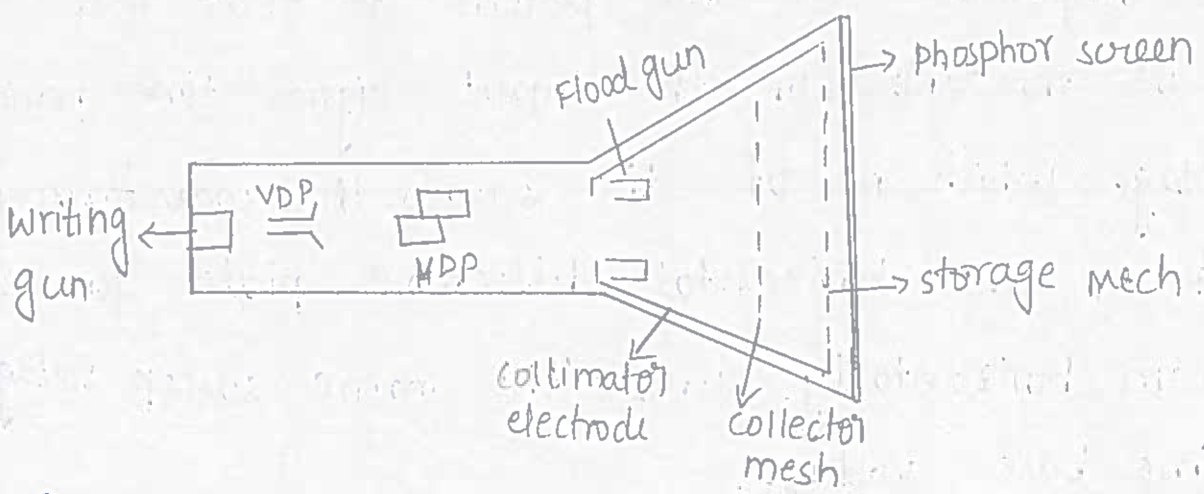
Storage targets can be distinguished from standard Phosphor targets by their ability to retain a wave form pattern for a long time, independent of

Phosphor persistence.

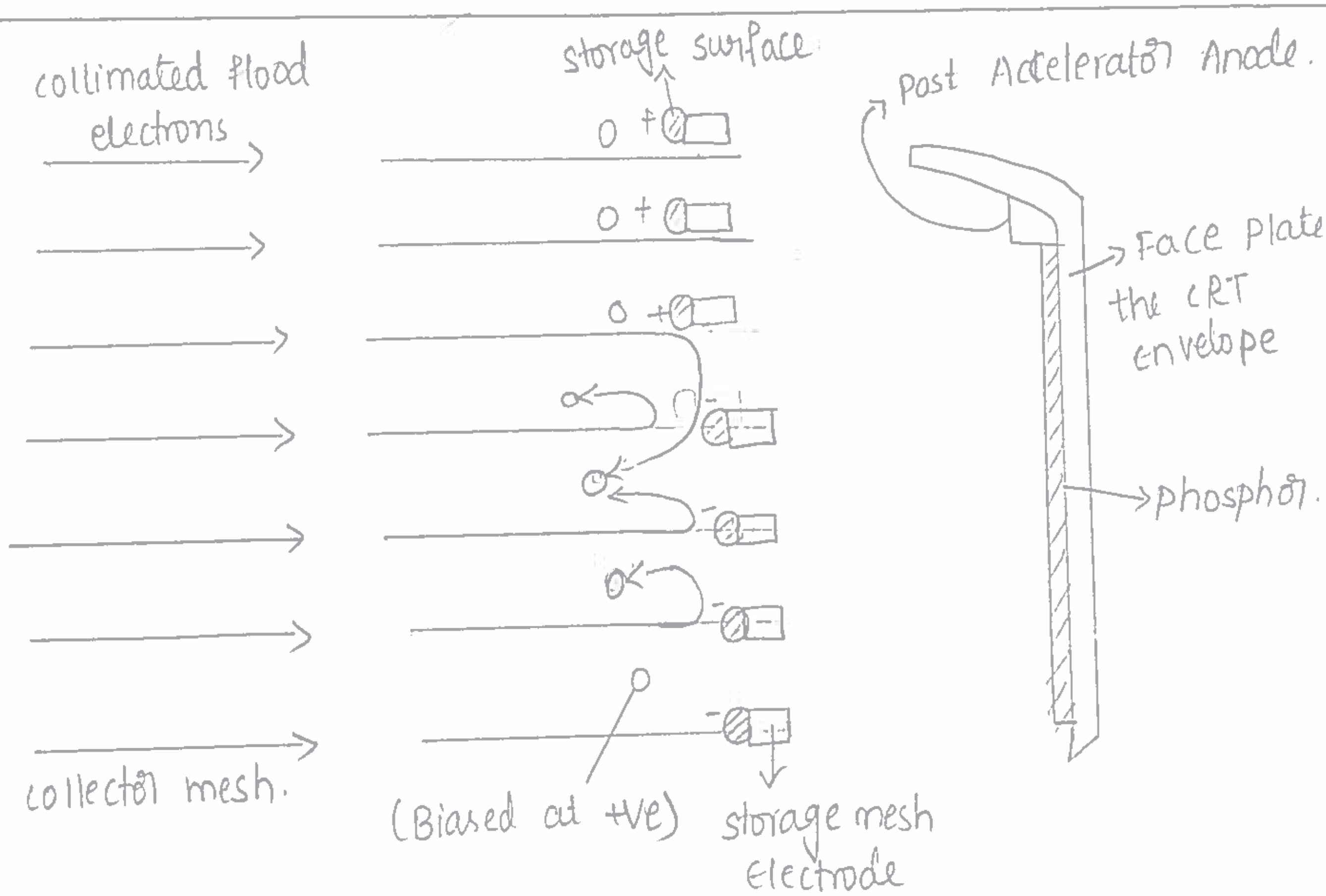
A mesh-storage CRT uses a dielectric material deposit~~ed~~-ion a storage mesh as the storage target. The phosphor storage CRT uses a thin layer of phosphor to serve both as the storage and the display element.

Mesh Storage:-

It is used to display very low frequencies (VLF) signals and finds many applications in mechanical and biomedical fields. The ~~commer~~ conventional scope has a display with a phosphor pesistance ranging from a few micro seconds to a few seconds. The persistence can be increased to a few hours from a few seconds.

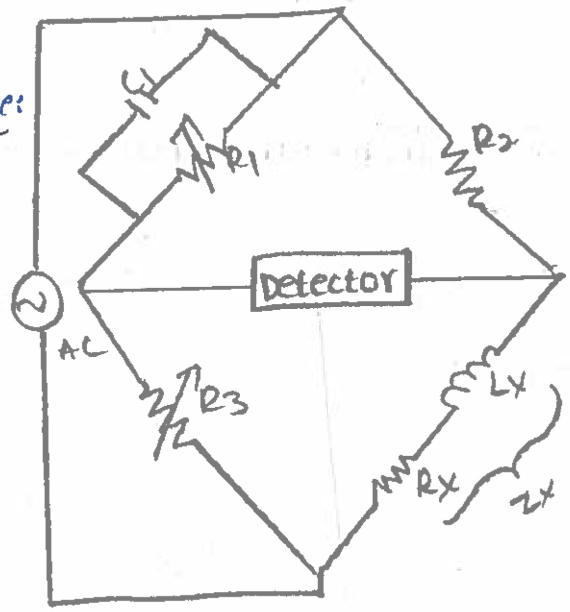


The CRT will now display the signal and it will remains as be as long as the flood guns operate. To erase the pattern on the storage mesh, a -ve voltage is applied to neutralise ~~them~~ the stored positive charge.



Display of stored charged pattern of MESH storage.

12(a) Derive the expression for unknown inductance using Maxwell Bridge:



Expression for unknown Inductance:

For Balance;

$$Z_1 Z_x = Z_2 Z_3$$

$$Z_x = \frac{Z_2 Z_3}{Z_1} = Z_2 Z_3 Y_1 \quad \text{—} (1)$$

$$Z_2 = R_2$$

$$Z_3 = R_3$$

$$Z_x = R_x + j\omega L_x$$

$$Y_1 = \frac{1}{R_1} + j\omega C_1$$

Sub values of Z_2, Z_3 Z_x & Y_1 in eq (1)

$$R_x + j\omega L_x = R_2 R_3 \left(\frac{1}{R_1} + j\omega C_1\right)$$

$$R_x + j\omega L_x = \frac{R_2 R_3}{R_1} + j\omega R_2 R_3 C_1 \quad \text{—} (2)$$

Equating Real & Imaginary terms of eq (2)

Real term:

$$R_x = \frac{R_2 R_3}{R_1}$$

Imaginary:

$$j\omega L_x = j\omega R_2 R_3 C_1$$

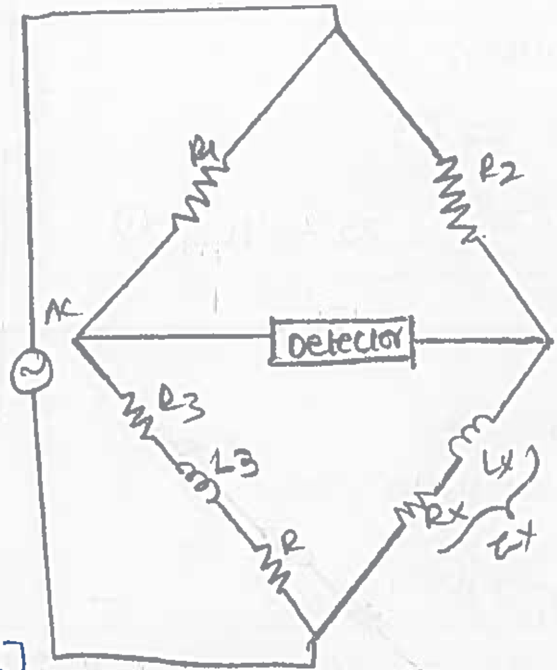
$$\therefore L_x = R_2 R_3 C_1$$

$$Q = \text{Ratio of } \frac{L_x}{R_x}$$

$$= \frac{\omega R_2 R_3 C_1}{R_2 R_3} \times R_1$$

$$\boxed{Q = \omega R_1 C_1}$$

Maxwell's Inductance Bridge :



For Balance;

$R_1 R_x = R_2 R_3$

(1) R_1

(2) R_2

(3) $R_3 + r + j\omega L_3$

(4) $R_x + j\omega L_x$

$$R_1 [R_x + j\omega L_x] = R_2 [R_3 + r + j\omega L_3]$$

$$R_1 R_x + j\omega R_1 L_x = R_2 R_3 + R_2 r + R_2 j\omega L_3$$

Equate Real and Imaginary terms

Real :-

$R_1 R_x = R_2 (R_3 + r)$

$$\boxed{R_x = \frac{R_2 (R_3 + r)}{R_1}}$$

Imaginary :-

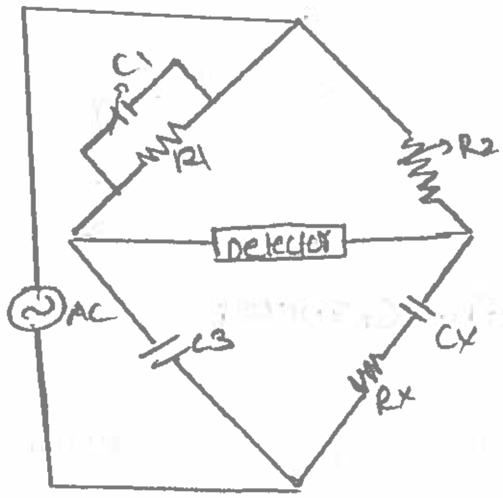
$j\omega R_1 L_x = R_2 j\omega L_3$

$$\boxed{L_x = \frac{R_2 L_3}{R_1}}$$

12(b) Explain the working principle of Schering Bridge with necessary Equation.

A

For Bridge Balance



$$Z_1 Z_x = Z_2 Z_3$$

$$Z_x = \frac{Z_2 Z_3}{Z_1} = Z_2 Z_3 Y_1 \quad \text{—} (1)$$

$Z_2 = R_3$ | $Z_x = R_x \frac{-j}{\omega C_x}$

$Z_3 = -\frac{j}{\omega C_3}$ | $Y_1 = \frac{1}{R_1} + j\omega C_1$

Substitute the Z_2, Z_3, Z_x & Y_1 in eq (1)

$$R_x \frac{-j}{\omega C_x} = R_2 \left(\frac{-j}{\omega C_3}\right) \left[\frac{1}{R_1} + j\omega C_1\right]$$

$$R_x - \frac{j}{\omega C_x} = -\frac{R_2 j}{\omega C_3}\left[\frac{1}{R_1} + j\omega C_1\right]$$

$$R_x - \frac{j}{\omega C_x} = \frac{-jR_2}{R_1(\omega C_3)} + \frac{\cancel{\omega} R_2 C_1}{\cancel{\omega} C_3}$$

$$R_x - \frac{j}{\omega C_x} = \frac{-jR_2}{\omega R_1 C_3} + \frac{R_2 C_1}{C_3} \quad \text{—} (2)$$

Equating Real and Imaginary terms of eq (2)

Real :

$$R_x = \frac{R_2 C_1}{C_3}$$

$$\frac{1}{C_x} = \frac{R_2}{R_1 C_3}$$

$$\therefore C_x = \frac{R_1 C_3}{R_2}$$

Imaginary :-

$$\frac{-j}{\omega C_x} = \frac{-jR_2}{\omega R_{L3}}$$

$\boxed{D = R_x \omega C_x}$ $\boxed{D = \frac{1}{Q}}$

13 (a) Explain Q meter.

Q-meter is also known as RLC meter for measuring the quality of the coil, capacitance and inductance values and resistance value.

→ The ratio of inductance reactance to the resistor will provide the quality

$$\boxed{Q = \frac{XL}{R} = \frac{EC}{E}}$$

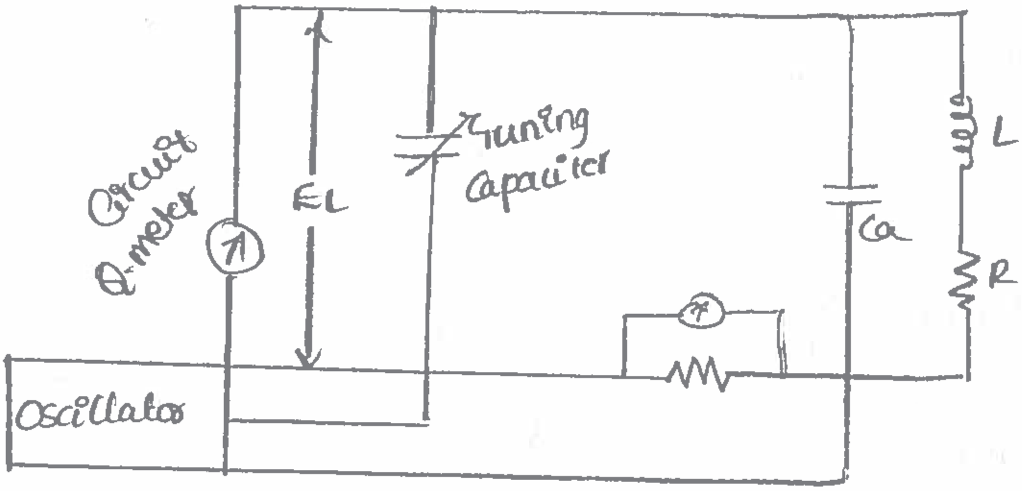
there;

$E_L = IX_L$

$E_C = IX_C$

$E = IR$

$$\therefore \frac{EC}{E} = \frac{X_C}{R} = Q$$



- If we choose Rsh = 0.02 to 0.04
- The voltage at the output can be obtained by using Q-meter
- The voltmeter across the capacitor is Ec
- The tuning capacitor will tune the voltage across the capacitor.
- Unit $X_L = X_C$ doesn't come as constant we have to change capacitance value and by varying the value at oscillator.

So;

$$X_C = X_L$$

$$f = \frac{1}{2\pi\sqrt{LC}}$$

$$L = \frac{1}{(2\pi f)^2 C}$$

(or)

$$C = \frac{1}{(2\pi f)^2 L}$$

⇒ factors may causes errors in Q meter :

- If value of Rsh > 0.4 then error will increase.

$$Q_{act} = \frac{\omega L}{R}$$

$$Q_{obs} = \frac{\omega C}{R + Rsh}$$

$$\frac{Q_{act}}{Q_{obs}} = \frac{\omega C / R}{\omega L / R + Rsh}$$

$$= \frac{R + Rsh}{R} = 1 + \frac{Rsh}{R}$$

$$\boxed{\frac{Q_{act}}{Q_{obs}} = 1 + \frac{Rsh}{R}}$$

Q_{obs} - observed Q factor

⇒ stray capacitance, shunt capacitance, wire capacitance Now $f_2 = 2f_1$

$$f_1 = \frac{1}{2\pi\sqrt{L_1(C_1+C_3)}} \qquad f_2 = \frac{1}{2\pi\sqrt{L(C_2+C_3)}}$$

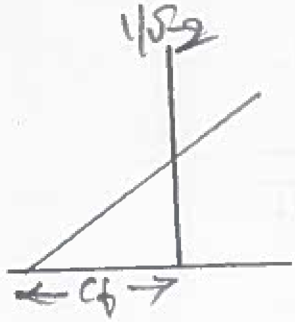
W.K.T $f_2 = 2f_1$

$$\frac{1}{2\pi\sqrt{L_1(C_1+C_3)}} = \frac{1}{2\pi\sqrt{L(C_2+C_3)}}$$

$$\frac{1}{L(C_2+C_3)} = \frac{4}{L(C_1+C_3)}$$

$$C_1 + C_3 = 4C_2 + 4C_3$$

$$C_3 = \frac{C_1 - 4C_2}{3}$$



$$f = \frac{1}{2\pi\sqrt{L(C+C_3)}}$$

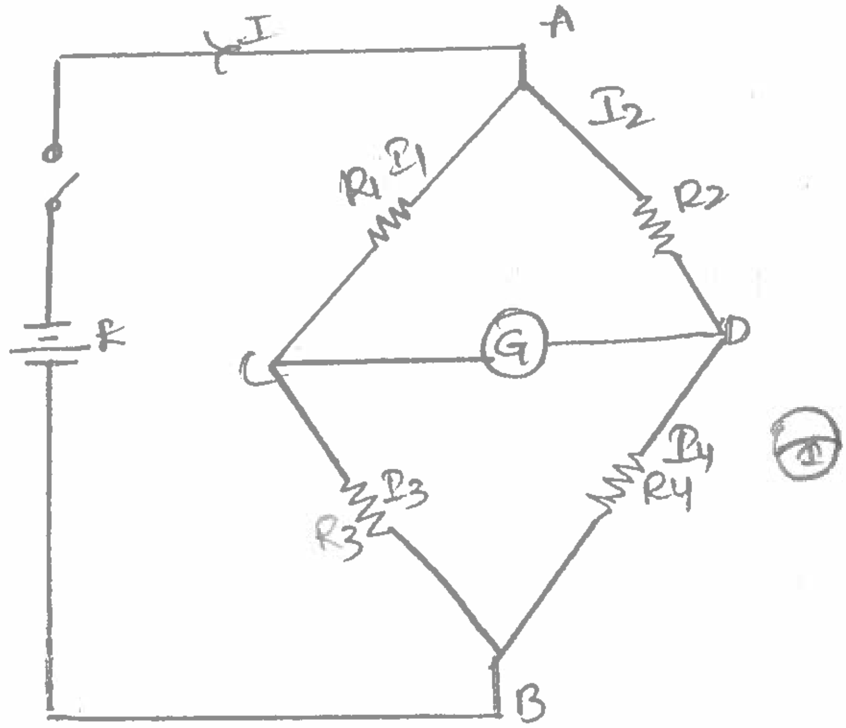
$$f^2 = \frac{1}{4\pi^2 L(C+C_3)}$$

$$1/f^2 = 4\pi^2 L(C+C_3)$$

if $1/f^2 \Rightarrow C = -C_3$

13b) Derive the equation for unknown resistance using wheat stone bridge.

A. Wheat stone Bridge is Balanced :-



→ These are used for measuring the unknown resistance

→ when the switch is closed the voltage at C = voltage at D so it shows null indication.

under Balance at Bridge

$$I_1 R_1 = I_2 R_2 \quad \text{—} (1)$$

⇒ for the Galvanometer current to be an zero to follow the conditions

$$I_1 = I_3 = \frac{E}{R_1+R_3} \quad \text{—— (2)}$$

$$I_2 = I_4 = \frac{E}{R_2+R_4} \quad \text{—— (3)}$$

Now substituting the value of I_1 and I_2 from eq (2) and eq (3) in eq (1)

$$I_1 R_1 = I_2 R_2$$

$$\frac{\cancel{E}}{R_1+R_3} \times R_1 = \frac{\cancel{E}}{R_2+R_4} \times R_2$$

$$R_1(R_2+R_4) = R_2(R_1+R_3)$$

$$R_1R_2 + R_1R_4 = R_1R_2 + R_2R_3$$

$$R_1R_4 = R_2R_3$$

$$R_4 = \frac{R_2R_3}{R_1}$$

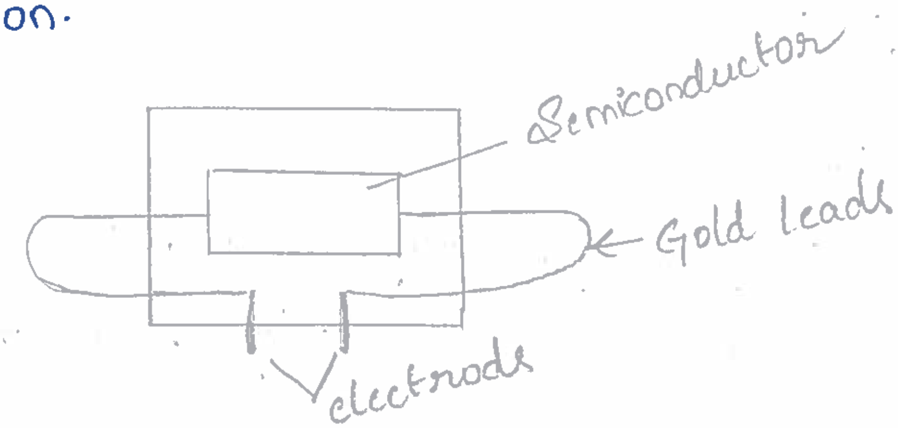
$$\boxed{R_x = \frac{R_2R_3}{R_1}}$$

R_x = unknown Resistance of the wheat-stone bridge.

14
Q) Discuss Semi-conductor type Strain gauge?

Semi Conductor type Strain gauge:

* These are used when a high value of gauge factor is desired.
* The basic principle is "piezo-resistive effect" i.e; the change in the value of resistance due to change in resistivity of the Semiconductor because of strain applied.
* The Semiconductor materials used are germanium and Silicon.



* The gold leads are used because of the Sensitivity of the Semiconductor and they are used for making contacts.
* The Size is very Small i.e., 0.7mm to 7 mm.
* The gauge factor is 130.
* These are used to convert the pressure into electrical Signal.

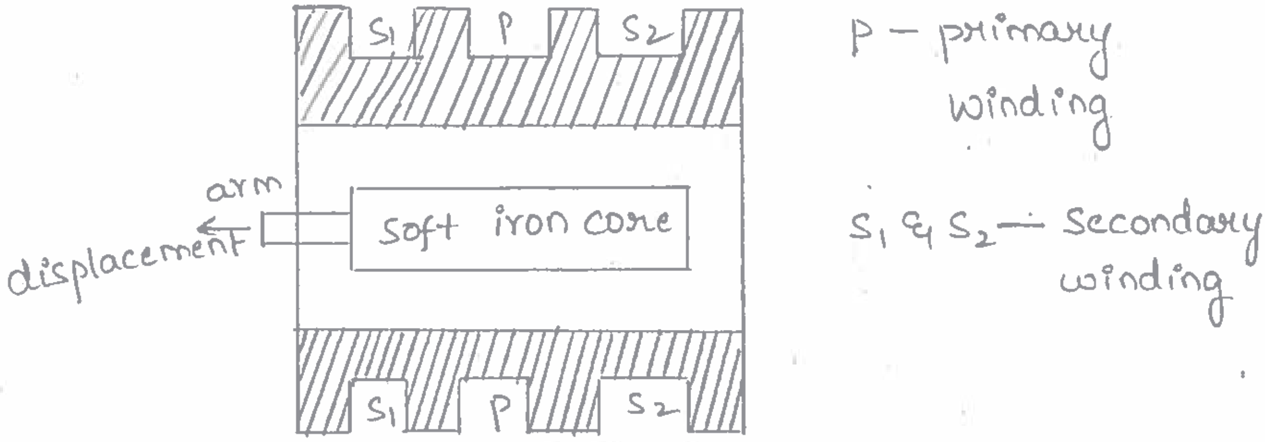
* It can measure ~~Signals~~ & Small strains from 0·1 to 500μ strains (micro strains).
* The hysteresis characteristics are excellent i·e; less than 0·05%.

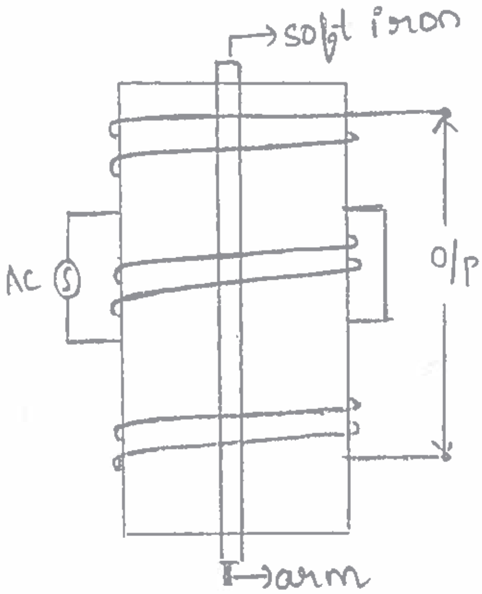
(4)
(b) Explain the principle, working, construction, characteristics and applications of LVDTs.

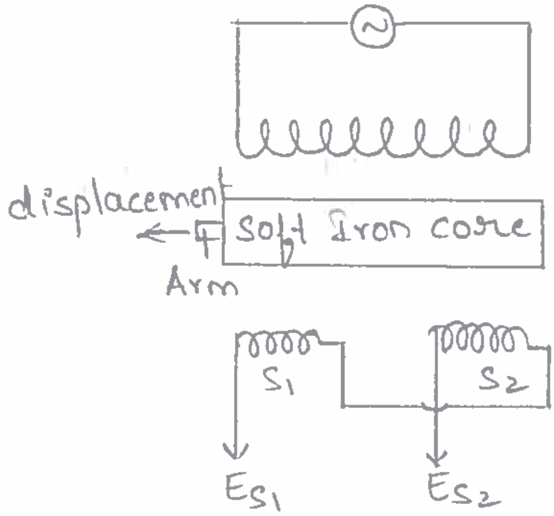
<u>LVDT</u>: A Linear Variable differential transformer is an electromagnetic device that produces an electrical voltage proportional to the displacement of a movable magnetic core.

<u>Construction</u>:

* It consists of one primary coil and two Secondary coils which are Symmetrically Spaced on a tabular center.
* The primary winding is connected to an AC Source which produces a flux in the air gap and voltages are induced in Secondary winding.
* A movable iron core is placed inside the former and when the displacement is applied to the arm it will move the iron core, hence the displacement is measured.







Principle & Working:

* Based upon the difference of the voltage (E_{S_1} & E_{S_2}) we will find the displacement.
* If the flux is induced b/w soft iron & E_{S_1} & E_{S_2} and they are equal then.

$$\boxed{E_0 = E_{S_1} \sim E_{S_2}}$$

* The Value of E_0 changes & according to the position of iron core.

Case (i): If the iron Core is in Center then

$$E_{S_1} = E_{S_2}$$

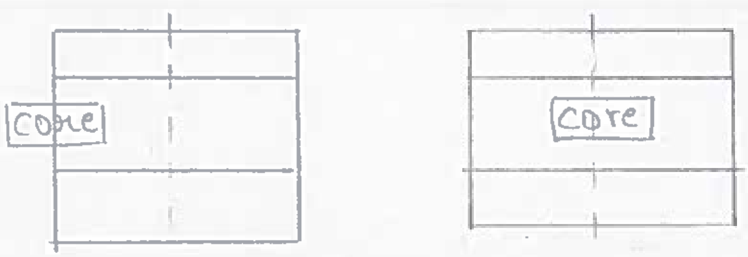
$$\boxed{E_0 = 0V}$$

<u>case(ii)</u>: If the position of the iron core is left side then the flux induced is E_{S_2} is low so,

$$\boxed{E_0 = E_{S_1} - E_{S_2}}$$

<u>case(iii)</u>: If the position of the iron core is right side then the flux induced is E_{S_1} is low so,

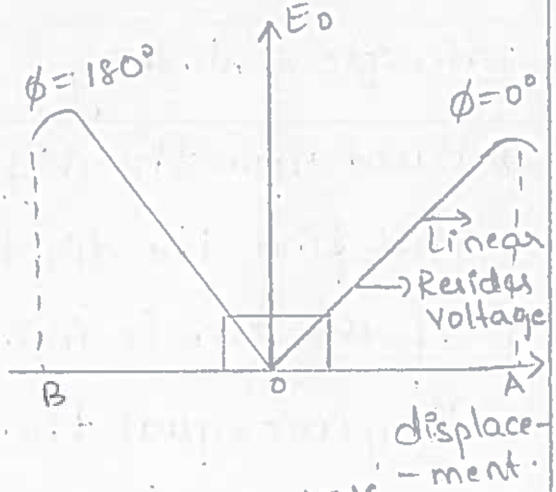
$$\boxed{E_0 = E_{S_2} - E_{S_1}}$$



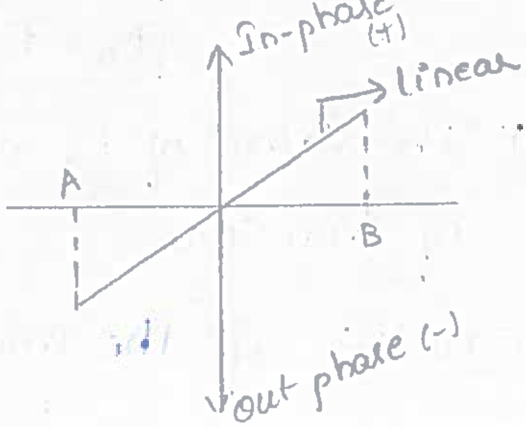
<u>Characteristics:</u>

The graph of output voltage Vs displacement is:

* A linear curve shows that o/p voltage varies linearly with the displacement of the core



* The graph b/w the in-phase and out-phase will be linear wrt to the core position & output voltage.



Applications:

(i) It is used in aircrafts.

(ii) It is used in satellites and nuclear reactors

(iii) It is used in power turbines and hydraulises.

Advantages:

(i) The resolution is high

(ii) It requires less amount of power

Dis-advantages:

(i) Temperature affects the transducer.

(ii) The dynamic response is limited by mass of the core

15) Discuss the measurement of physical parameters.

(i) Force

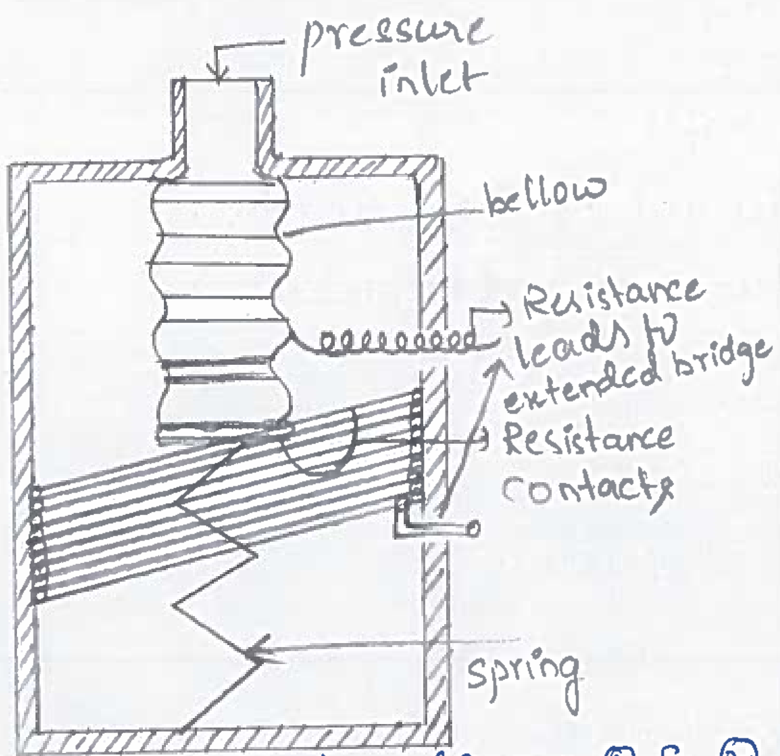
(ii) pressure (using resistive pressure transducer)

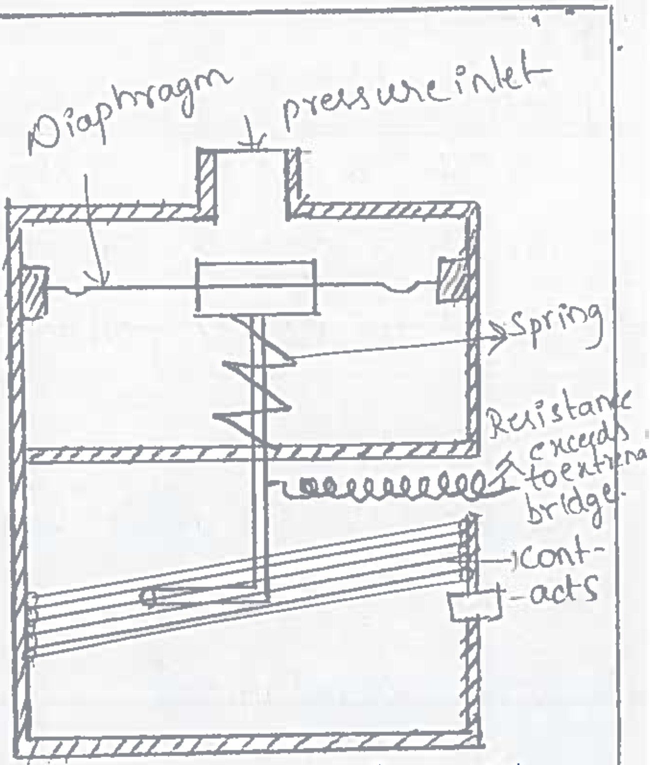
(iii) velocity (using moving coil and moving magnet type transducer).

(i)
Force: According to Newton's second law, force is given as the product of mass and acceleration.

(ii) pressure (using resistive pressure transducer):

In resistive pressure transducer, when ever the pressure changes then it results in the change of resistance of sensing element.



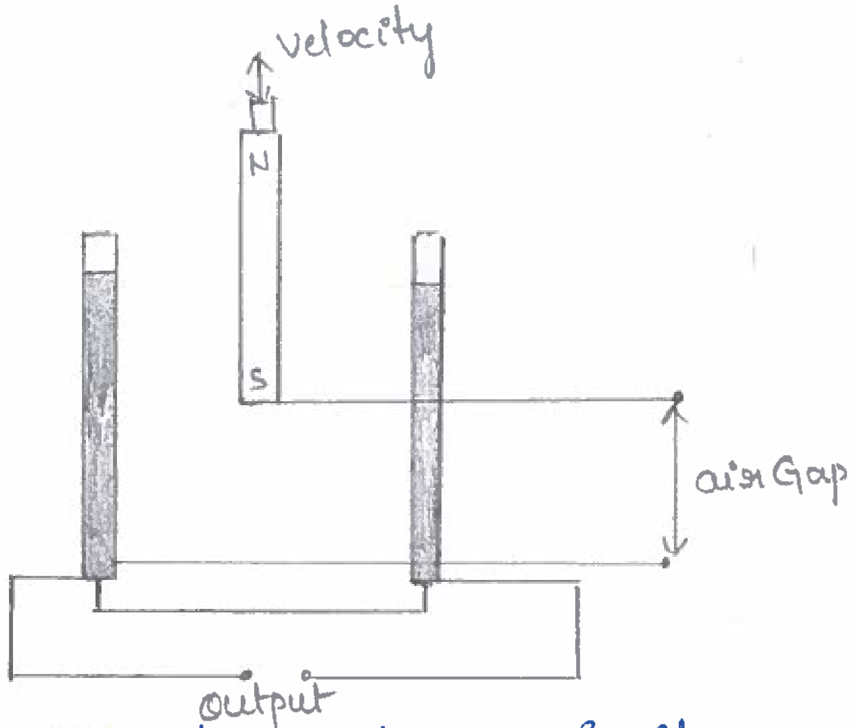


In the above two figures ① & ② shows us the two ways by which the pressure changes.

* Due to the pressure change the resistance changes.
* That means the pressure influences the change in resi-stance.
* In case (i) when we apply the pressure from the inlet of the bellows, then the bellow position changes and position of the variable contact will be slides on the resistance and the resistance will be changes.
* In case (ii), with the help of the applied pressure the contact is moved up (or) down. through the pressure applied from the diaphragm. Based on the applied pressu-re the resistance variations and voltage variations takes place.

Like this we will measure the pressure through resistive transducers.

(iii) Moving Magnet type velocity Transducer.



* When a permanent magnet moves insides a coil, the change in the length of the air gap varies the reluctance.

* Hence the o/p voltage is directly proportional to the rate of change of the length of the air gap (change in length produced by velocity). Thus the output voltage becomes a measure of velocity when calibrated.

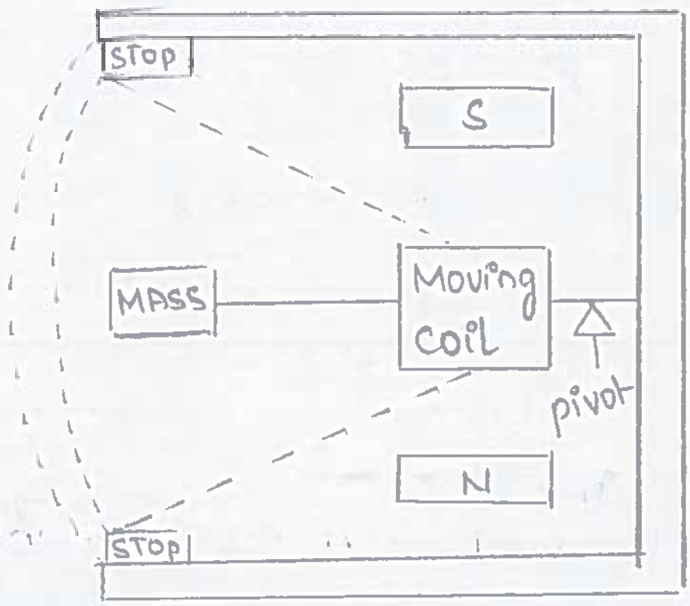
Operation:

* The Instrument is fixed to the device whose velocity is to measured.

* Due to the application of the velocity, the permanent magnet moves in or out of the coil. Due to its motion, the length of the air gap values.

* The polarity of the o/p voltages determines the direction of the velocity.

Moving Coil type Velocity transducer:



* A coil moves in a magnetic field according to the velocity applied. The voltage in the coil becomes a measure of velocity when calculated.

Operation:

* The Instrument has permanent pole pieces which generate the magnetic field.
* There is a pivoted arm on which a coil is mounted.
* There is a mass attached at the end of the coil.
* The whole device is contained in an antimagnetic case.

(Dr. K. Ravikumar)

HOD-ECE

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE			Academic Year	2022 – 2023
Course Code	20CS001	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Object Oriented Analysis and Design						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List two methods of analysis and design.	20CS001.1	L1
2	Compare links and aggregation.	20CS001.2	L2
3	What is the purpose of state-chart diagram?	20CS001.3	L1
4	'A system must be loosely coupled and highly cohesive'. Justify.	20CS001.4	L2
5	What is component?	20CS001.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What are the Attributes of Complex Systems? Explain.	6M	20CS001.1	L2
6 (b)	Explain Designing Complex System.	6M	20CS001.1	L2
	OR			
7 (a)	Explain major elements of object model.	6M	20CS001.1	L2
7 (b)	What are the benefits and issues in applying object model.	6M	20CS001.1	L1
8 (a)	Describe the strategies used to identify relationship among the classes.	6M	20CS001.2	L2
8 (b)	Discuss how to build quality classes and objects.	6M	20CS001.2	L2
	OR			
9 (a)	Explain how you identify Classes and Objects.	6M	20CS001.2	L2
9 (b)	Discuss Key Abstractions and Mechanisms.	6M	20CS001.2	L2
10 (a)	What is the importance of modeling and why you need Model?	6M	20CS001.3	L2
10 (b)	Describe building blocks of the UML.	6M	20CS001.3	L2
	OR			
11 (a)	Discuss the essential Class Relationships.	6M	20CS001.3	L2
11 (b)	Draw the Class Diagram for Online Shopping Management System.	6M	20CS001.3	L2
12 (a)	How do you use Interaction diagram when you model dynamic aspects of system. Explain with an example.	6M	20CS001.4	L2
12 (b)	Draw and explain the activity diagram for an Online Railway Management System.	6M	20CS001.4	L2
	OR			
13 (a)	What is use case? Explain use cases for ATM with a diagram.	6M	20CS001.4	L2
13 (b)	Difference between collaboration diagram and sequence diagram.	6M	20CS001.4	L2
14 (a)	What is component diagram? Draw the component diagram for Online Reservation System.	6M	20CS001.5	L2
14(b)	Explain deployment diagram.	6M	20CS001.5	L2
	OR			
15 (a)	Draw the deployment diagram for Bank Management System.	6M	20CS001.5	L2
15 (b)	Draw the Component and Deployment Diagram for ATM system.	6M	20CS001.5	L2

OBJECT ORIENTED ANALYSIS AND DESIGN

Part A - Short Notes

Note: The diagrams in answer sheet or one type of model. The student is expected to UML Notations and domain terminology only, he need not draw exactly.

1) List two methods of Analysis and Design

Answer:

a) Structured System Analysis and Design (SSAD) or Procedural oriented or Traditional Method....
b) Object oriented Analysis and Design

2) Compare links and Aggregation

Answer:

Link and Aggregation

1) Link is connection among objects.
2) Link Establish relationship between two objects.

AGGREGATION:

1) Aggregation relationships provide the whole / part relationships among classes.
2) Aggregation relationships among classes have direct parallel aggregation relationships among objects corresponding to classes.
3) Multiple inheritances is often confused with aggregation.
4) If, we cannot affirm that there "is a" relationship between classes then we must use Aggregation or some other relationship instead of Inheritance.

3) What is the purpose of state-chart diagram?

Answer:

1) In UML, we model the event order behavior of an object by using the state chart diagram.
2) State chart diagram is one of the 5 diagrams in UML for modeling the dynamic aspects of the system.
3) A state chart diagram shows the state machine.
4) Both activity and state chart diagrams are useful in modeling the life time of an object.
5) In activity diagram shows the flow of control from activity to activity.
6) A state chart diagram shows the flow of control from state to state.
7) A state chart diagram is simply a presentation of state machine emphasizing the flow of control from State to state.

4) "A system must be loosely coupled and highly cohesive". Justify.

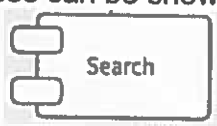
Answer:

1) The coupling and cohesiveness are relevant for various modules in a system
2) The loose coupling makes each Module independent that is less dependent on other modules.
3) The cohesive of a module implies that all relevant data and methods related to the functionality Of the module or within it and so less dependency.

5) What is component

Answer:

1) A component is a physical and replaceable part of a system that confirms to and provides the realization of a set of interfaces.
2) A component in UML represents a modular part of a system.
3) We use components to model the physical things that may reside on a node such as executables, libraries, tables, files and documents.
4) Object libraries, executables, Enterprise Java Beans all are examples of components.
5) Graphically, a component is represented as a rectangle with Tabs.
6) The components and Classes can be show in the following Diagram.



Part – B (Long Answers)

6) A) What are the attributes of Complex System? Explain

1) Depending on the nature of complexity there are Five attributes common to all complex systems.
   A) Hierarchical Structure B) Relative Primitives C) Separation of Concerns
   D) Common Pattern E) Static Intermediate pattern.

A) Hierarchical Structure:

1) Complexity takes the form of hierarchy.
2) A Complex system is composed of inter related sub-systems that have in turn own Sub-systems and so on until some lowest level of component is reached.
3) Complex systems have nearly decomposable, hierarchal structure and this factor enable us to understand, describe and even "see" such systems and parts.
4) We can understand only those systems that have hierarchal structures.
5) Architecture of complex system is a function of its components as well as their hierarchal relations among those components.
6) All systems have sub-systems and all systems are parts of large systems.
7) The value added by a system must come from their relationship between the parts.

B) Relative Primitives:

1) The choice of what components in a system are primitive is relatively arbitrary and largely up to the discretion of the observer of the system.
2) What is primitive for one observer may be a much higher level of Abstraction for another.

C) Separation of Concerns:

1) Hierarchal systems are decomposable because they can be divided into identifiable parts.
2) They are nearly decomposable because their parts are not completely independent.
3) Intra component linkages are generally stronger than inter component linkages.
4) The difference between intra and inter component interactions provides a clear separation of concerns among various parts of system.
5) This makes us possible to study each part in isolation.

D) Common Pattern:

1) Hierarchal systems usually composed of only few different kinds of sub-systems in various combinations and arrangements.
2) i.e. complex systems have common patterns.
3) The pattern may involve the re-use of small components.
   Example: Cells found both in plants and animals or larger structures such as Vascular systems.

E) Static Intermediate pattern:

1) We have noticed the complex systems tend to evolve over time.
2) Complex systems will evolve from simple systems .....Work....stable....grow.
3) A complex system designed from scratch never works and cannot be patched up to make it work.
4) You have to start over beginning with working simple systems.
5) As systems evolve, objects that were once considered complex become the primitive objects on which more complex systems are built.
6) We must use them in context first, and then improve them overtime as we know the real behavior of the system.

6) B) Explain Designing Complex System.

Answer:

1) The practice of every engineering discipline be it mechanical, civil, electrical , software engineering involves elements of both art and science.

2) The conception of a design for new structure can involve
   a) A Leap of Imagination b) Synthesis of experience c) Knowledge
3) Once the design is articulated by the engineer as an artist, then it must be analyzed by an Engineer as scientist in rigorous application testing.

Meaning of Design:

1) Design encompasses the disciplined approach used to invent solution to a problem.
2) The purpose of design is:
   a) Satisfies given functional specification
   b) Confirms to limitations of target medium.
   c) Meets implicit and explicit requirements
   d) Satisfies the restrictions on the design process like cost, time and tools available for design
3) The purpose of design is to create a clean and relatively simple internal structures, sometimes called "Architecture".

4) The products of design are models that enable us to reason and provide blue print for implementation.

Importance of Model Building:

1) Model building appeals to the principles of Decomposition, Abstraction and Hierarchy.
2) We build new models sometimes based on old models.
3) We evaluate each model in both expected and unusual situations,
4) The engineers have to view models both from static and dynamic perspectives.

The Elements of Software Design Methodologies:

1) The software design of systems involves incremental and iterative process.
2) Software Engineering has different design methodologies but all these include:

   a) Notation : Language for expressing model
   b) Process : Activities concern to orderly construction of system's model.
   c) Tools : 1) Are the artifacts that eliminate tediousness of model building.
   2) Enforce Rules on models.
   3) Errors and inconsistencies can be removed.

The Models of Object Oriented Development:

1) Object Oriented Analysis and design is the method that leads us to Object Oriented Decomposition.

1) By applying object oriented design, we create software that is resilient to change in economy or less cost.
3) We reduce the risk inherent in developing complex software systems.
4) Thus we are making the case for Object Oriented Analysis and Design.

The Role of Abstraction:

1) We have developed an exceptionally powerful technique for dealing with complexity. i.e. we abstract it.
2) Abstraction refers to the act of representing essential features without including details.
3)) Abstraction solves the problem on the design side.
4) Unable to master the entirety of a complex object, we choose to ignore inessential details.
5) We take or deal with the generalized or most essential or ideal model of object.
6) An example a) Engine of a Car (it may contain several elements in it)

Role of Hierarchy:

1) One way for information or domain representation is by explicitly recognizing class and object hierarchies.
2) Object structure illustrates how different objects collaborate.
3) Class structure highlights common attribute structure and behavior within a system.
4) By classifying objects into groups or related abstractions....we distinguish between common and distinct properties of objects.
5) To identify hierarchies in system....we has to discover patterns.

7) A) Explain major elements of Object Model.

Answer:

1) There are broadly five main programming styles. They are:

a) Procedure Oriented	-------	Algorithms
b) Object Oriented	-------	Classes and Objects
c) Logical Oriented	-------	set of sentences in logical form, goals, Expressing facts, rules example : PROLOG.
d) Rule Bases Oriented	-------	If then rules example AI or C, C++
e) Constraint Oriented	--------	Invariant relationships. Constraint logic programming is Prolog III,

2) Each of these styles require different conceptual frameworks --- different thinking.
3) There are four major elements of Object Oriented Model. They are:
   a) Abstraction b) Encapsulation c) Modularity d) Hierarchy.

4) Major we mean without these it is not object oriented.
5) There are three minor elements. These are useful but not essential part of object oriented model.
   A) Typing b) Concurrency c) Persistence

ABSTRACTION:

1) Abstraction denotes the essential characteristics of an object that distinguish it from all other objects.
2) Abstraction focuses on outside view of an object and takes entire behavior of object.
3) Abstraction includes:
   A) Entity Abstraction → Object representing problem domain.
   B) Action Abstraction → Object that provide generalized set of operations.
   C) Vitual Machine Abstraction → Object group operations used by superior level
   d) Coincidental Abstractions → An object that packages a set of operations that has no reaction for each other.

4) A Client is an object that uses the resources of another object (known as Server)
5) All Abstractions have static as well as dynamic properties.

Example: A file object, amount of space in on memory device, its name and it has contents t.
These are all static.
The names of these properties are dynamic. Ex: name and contents change.

Examples of Abstractions:

a) Automobile: Steering, Engine and Break system
c) Bank : Account s, Loan and Deposits

Meaning of Encapsulation:

1) Abstraction and Encapsulation are complementary concepts.
2) Abstraction focuses on the observable behavior of an object where as Encapsulation focuses on implementation that give rise to this behavior.
3) Encapsulation is most often achieved through information hiding (Not Data Hiding) i.e. hiding all secrets of which are not essential.
4) For Abstraction to work implementation must be encapsulated.
5) Each class has two parts
   a) Interface (is outside view) and implementation.
6) Encapsulation is the process of compartmentalizing elements of Abstraction that is, its structure and behavior.
7) Hiding is relative, what is hidden at one level of Abstraction may be outside view in another level of abstraction.
8) Hiding is for prevention of accidents.

Modularity:

1) Modularization consists of dividing program into parts or modules which can be compiled separately but have connection to other modules.
2) Modularity and Encapsulation go hand in hand.
3) Modules serve as physical containers in which we declare classes and objects in our logical design.
4) The developer might decide to declare every class and object in the same package.
5) But, better way is to group logically related classes and objects in the same module.
6) The task is to decide where to physically package the classes and objects. Example: Java has packages contain classes.
7) Each module should be simple.
8) Modularity is the property of a system that has been decomposed into a set of cohesive and loosely coupled modules.
9) The technical issues are:
   a) Re-use across application
   b) Practical limits in size of modules (Compiled Segments)
10) We much define some main program from which we invoke application. In object oriented design this is least important decision.

Hierarchy:

1) Hierarchy is a ranking or ordering of Abstraction.
   Examples: Computer -> CPU -> ALU -> Registers
   Two Wheeler -> Tire -> Tube -> Air
2) Two most important hierarchies in complex system are its:
   a) Class structure ("is a" hierarchy)
   b) Object Structure ("part of" hierarchy)
3) Inheritance is most important "is a" hierarchy.

4) Inheritance defines a relationship among classes, where in one class shares the structure and behavior defined in one or more classes.
5) Inheritance represents hierarchy of Abstraction.
6) There can exist Single and Multiple Inheritances (hierarchies)
7) "is a" hierarchies denote generalization / specialization relationships. (Classes)
8) "part of" hierarchies describe aggregation relationships. (Objects)

7) B) what are the benefits and issues in applying object model?

Answer:

1) Object model is fundamentally different from traditional methods of Structured Analysis, Structured Design and Structured Programming.

2) Object Model offers a number of benefits:

   a) Object model helps us in use of more powerful features in languages like C++, Java and Small Talk.
   b) Object Model encourages Re-Use not only software but of entire design leading the Creation of Frameworks.
   c) User of object model helps build systems on stable intermediate forms.
   d) Object Model reduces risks inherent in complex systems.
   e) Object Model appeals to the workings of human Cognition.
   f) Many people who have no idea how a computer works finds the ides of Object Oriented Systems quite natural.

8) A) Describe the strategies used to identify relationship among the classes.

Answer:

1) There are various ways and strategies for identifying the relationships among classes.
2) Relationships depend upon type of class.
3) Majority times the functionality of the problem decides the type of relationships.
4) Student can write about various types of class relationships with examples.
5) The answer can include something like the following:

   1) Class, like objects, do not exist in Isolation.
   2) A class relationship might indicate some sort of sharing.
   3) A class relationship might indicate some kind of semantic connection.
   4) There are Three Basic kinds of class relationships.

   5) The First of these is GENERALIZATION/SPECILIZATION , denoting "is a" relationship.
      Example: A Rose is a kind of flower, meaning that a rose is a specialized sub-class of the more general class i.e. Flower.

   6) The second is a WHOLE / PART, which denotes a "Part of" relationship.
      Example: A petal is not a kind of a flower, it is a part of flower.

   7) The third is ASSOCIATION, which denotes some semantic dependency among otherwise unrelated classes.
      Example: Roses and candles are largely independent classes, but both Represent things that we might use to decorate a dinner plate.

8) B) Discuss how to build quality classes and objects.

Answer:

Measuring the Quality of classes and objects:

As a system Analyst upon identifying classes and objects, we have to check their quality that Is how far we have achieved the abstraction while identifying classes and objects?

The quality of classes and objects can be measurable by using some metrics.

1) There are Five important metrics to know, if, given classes and objects are well Designed or not.

a) Coupling: Complexity can be reduced with weakest possible coupling between Modules. But for classes it increases.
b) Cohesion Degree of Connectivity among elements in single Module.
c) Sufficiency: (Does the class capture enough details for model to be useful?)
d) Completeness: Extent of re-usability.
e) Primitiveness: (Primitive classes are small, easier to understand, efficiency etc.

9) A) Explain how you identify classes and objects

Answer:

1) There are several approaches for identifying classes and objects they include:

a) Classical Categorization
b) Conceptual Categorization
c) Prototype Theory.

2) In Analysis, we focus to fully analyze the problem at hand or domain.
3) We model the world by discovering classes and objects from vocabulary of problem domain.
4) There are number of methodologies for deriving classes and objects from requirements of the problem domain.
5) We call these as Classical approaches because they are derived from principles of classical Categorization.
6) As per MELLOR, he suggested that classes and objects usually come from one of the following sources:

a) Tangible Things ---------> Cars, Sensors,
b) Roles ---------> Mother, Teacher, Politician
c) Events ----------> Landing, Request, Interrupt.
d) Interactions ----------> Loan, Meeting and interactions.

7) From Database Modeling ROSS offers list

a) People b) Places c) Things d) Organizations e) Events

8) COAD and YOURDON says all the potential objects are :

a) Structures – "is a" and "part of" relationships.
b) Other Systems
c) Devices
d) Roles
e) Locations
f) Organization Units

Behavior Analysis:

1) The classical approaches focus on tangible things in the problem domain, object orientation Analysis focus on dynamic behaviour as the primary source for classes and objects.
2) This is more close / akin to conceptual Clustering.
3) We form classes based on groups of objects that exhibit similar behavior.

Domain Analysis:

1) Domain Analysis Seeks to identify classes and objects that are common to all Applications within a given Domain.
2) Examples: Patient Record Tracking, Bonds / Shares Trading, Compilers.
3) We define Domain Analysis as "An attempt to identify objects, operations and Relationships those domain experts perceive to be important in that domain".

Steps in Domain Analysis:

1) Construct a strawman generic model of the domain consulting with domain expert.
2) Examine existing system within the domain and represent the understanding in common format.
3) Identify the similarities and differences between system by consulting with experts.
4) Refine the generic model to accommodate existing system.
5) When starting to design a new patient monitoring system, it is reasonable to survey the architecture of existing system to understand what key Abstractions and Mechanisms were previously employed and to evaluate which were useful and which are not.
6) Sometimes direct communication between the developer and end users is much useful.
7) To clear up any design problem a meeting between Domain Expert and an Architect or Software Engineer or Software Developer is mandatory.

USE CASE ANALYSIS:

1) In, isolation of classical Analysis, behaviour analysis and Domain Analysis all depend upon personal experience.
2) There is one more practice coupled with all three earlier approaches to drive the process of analysis. That is called "Use Case Analysis".
3) Use Case as " A behaviorally related sequence of transactions performed by an Actor in a dialogue with the system to provide some measurable value to the actor.
4) We can apply use case analysis as early as Requirement Analysis at which time end users, domain experts and software development team enumerate the scenarios that are fundamental to system.

Steps For Use Case Analysis:

a) Initially prepare scenarios.
b) Analysis then proceeds by a study of each scenario using storyboard Technique. Similar to practices in Television or Movie Industry.

c) As they go thru the scenario's, they must identify objects and responsibility of objects.
d) Initial scenarios are expanded to consider exceptions and new Abstractions, add , Modify.
e) Scenarios serve as the basis for system tests.

CRC CARDS:

1) 3 X 5 ` index card on which analyst write with pencil the name of class, its Responsibility and its collaborations.
2) One card for each class identified as a relevant scenario.
3) Team members walk thru the scenario and may assign new responsibilities.
4) Cards are arranged for generalization, specialization and Aggregation hierarchies among classes.

INFORMAL ENGLISH DESCRIPTION:

1) Writing an English description of the problem then underlining Nouns and Verbs.
2) The nouns represent candidate objects and Verbs represent candidate operations on them.
3) This forces the Analyst to work on Vocabulary of Domain.
4) Sometimes a Noun can be Verb and a Verb can be Noun. So.........

9) B) Discuss key Abstractions and Mechanisms

Answer:

1) A Key abstraction is a class or object that forms part of the vocabulary of the problem domain.
2) The primary value of identifying such object is that they give boundaries to our problem.
3) The term mechanism is to describe any structure whereby object collect , collaborate to provide some behavior that satisfies a requirement of the problem.
4) Mechanisms represent patterns of behavior.
5) Now we study the identification and refinement of these key abstractions and mechanisms.

Identifying Key Abstractions:

1) Identification of key abstractions is highly domain specific.
2) The identification of key abstractions involves two processes
   a) Discovery b) Invention
3) Thru discovery, we come to recognize the abstractions used by domain experts.
4) Thru invention, we create new classes and objects that are not necessarily part of problem domain but useful in design and implementation.
5) For Example, Customer using an automated teller speaks the terms like Customer, deposits and withdraws these words are part of problem domain. A developer of such systems uses same abstraction by may also introduce new ones such as database, screen managers, lists, queue's.
6) These key abstractions are artifacts of the particular design not of the problem domain.

REFINING Key Abstractions:

1) Given a new abstraction, we must place it in the context of the existing class and object hierarchies that we have designed.
2) The most common re-organizations of class hierarchy are factoring the common part of two classes into a new class and splitting a class into two classes.
3) Sometimes we may find a general subclass and so may choose to move it up in the class structure. i.e. increasing the sharing and we call it promotion.
4) We may find a class to be too general, thus making inheritance by subclass difficult, because of large semantic gap. This we call grain size conflict.

Naming Key Abstractions:

1) Objects should be named with proper noun phrases such as the Image Sensor, Sphere etc.,
2) Classes should be named with common noun phrases such as sensor or shape.
3) The names chosen should reflect the names used and recognized by the domain experts.
4) Modifier operations should be named with active verb phrases such as Draw or MoveLeft.
5) Selector operators should imply a query such as extentOf or isOpen.
6) Use of and styles of capitalization are personal and whatever cosmetics of design, they must be consistent.

Identifying Mechanisms:

1) Mechanisms represent strategic design decisions.
2) Mechanisms denote strategic decisions regarding the collaborative activity of many different kinds of objects.
3) Normally, several mechanisms may exist to deliver the required behaviour

   Example: Push Accelerator ----- Engine Runs faster.
   Releasing Accelerator ----- Cause Engine to run slow.
   There are many ways to achieve.
4) The choice of mechanism from alternatives is influenced by cost, reliability, safety and manufacturability.
5) Mechanisms represent a level of reuse that is higher than the reuse of individual classes.
6) Example: MVC (Model View Control) is a paradigm used in GUI's i.e. most of the Object Oriented GUI Frameworks.
7) Frameworks exports a number of individual classes and mechanisms that client can use or adopt.

8) Example of Frameworks are
   a) MS .Net b) Apache Software Foundations
   c) Struts Frameworks d) JUnit Testing Frameworks.

9) Mechanisms and patterns are found in every domain.
10) CODD has identified a number of Common mechanisms in Object Oriented Systems:
   a) Patterns of time association
   b) Event Logging
   c) Broadcasting

11) All the above mechanisms manifest themselves as not one class but as structure of collaborative classes.

10) A) what is the importance of modeling and why you need model?

Answer:

1) A successful software organization is one that consistently deploys quality software that meets the needs of end users.
2) For achieving to deploy quality software we must have good and sound (solid) development process.
3) Modeling is the central part for deployment of good software.
4) Modeling is proven and well accepted engineering technique.
5) We build models of Houses, High rise building and Math Models for analyzing winds or earthquakes.
6) Aircrafts, Automobiles, Computers and in other fields too we build Models.
7) A Model is a simplification of reality.
8) The main reason why we model is "We build models so that we can better understand the system that we are developing".
9) From Models we achieve Four Aims:
    a) Models help us to visualize a system
    b) Models permit us to specify structure and behavior of the system.
    c) Models give us template.
    d) Models document the decision we made.
10) There are four principles of Modeling.
    a) The choice of what models to create has profound influence on how a problem is attacked and how solution is shaped.
    b) Every model may be expressed at different levels of precisions.
    c) The best models are connected to Reality.
    d) No single model is sufficient; every non-trivial system is approached thru a small set of independent models.
11) In software the two most common ways of modeling are
    a) Algorithmic Perspective b) Object Oriented Perspective
12) UML is used to visualize, specify, constructing and documenting Object Oriented systems.
13) The purpose of UML (Unified Modeling Language) is to build the artifacts of system.

10) B) describe building blocks of the UML

Answer:

Building Blocks of UML:

1) The vocabulary of UML encompasses three kinds of building blocks
    a) Things b) Relationships c) Diagrams
2) Things are the Abstractions (Classes)
3) Relationships tie these things together
4) Diagrams group interesting collection of things

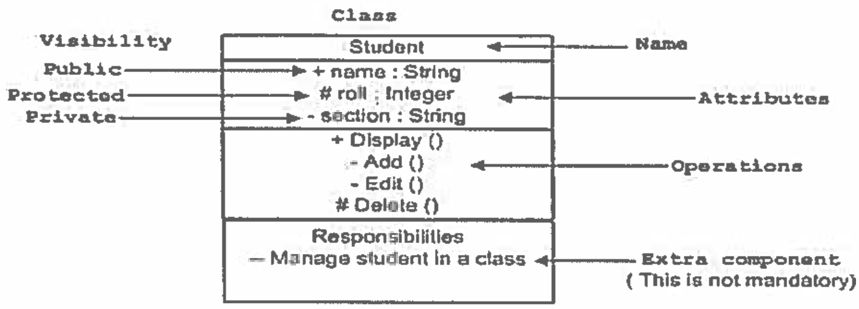
5) Things in UML are Four kinds
   a) Structural Things
   b) Behavioral things
   c) Grouping things
   d) Annotation Things

A) Structural Things:

1) Structural things are nouns of UML.
2) These are mostly static parts of model either physical or conceptual.
3) Class is a description of set of objects that share same attributes, operations, relationships and semantics.
4) The structural things can be
   a) Class b) Interface c) Collaborations d) Component e) Node

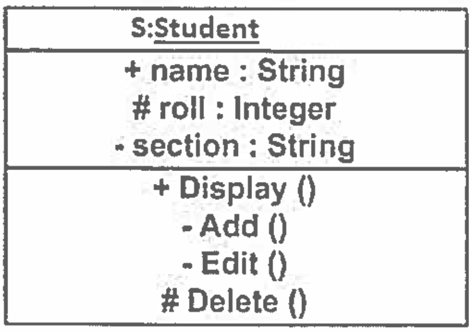
CLASS:

1) A Class implements one or more interfaces.
2) Class is represented as Rectangle with 3 divisions or sections with Name, Attributes and operations.
3) Example: UML Notation



Object :

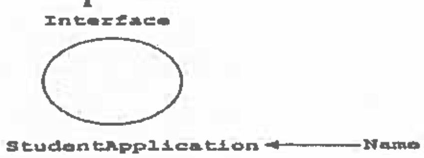
a) The object is an Instance of a class
b) The object is represented in the same way as the class.
c) The only difference is the name which is underlined
d) Example: UML Notation



e) In the Above Diagram "S" is object and "Student" is class.

INTERFACE:

1) Interface is a collection of operations that specify a service of class or component.
2) Interface describes the external visible behavior of an element.
3) An Interface is represented by Circle together with its name.
4) Rarely standalone mostly attached to class or component.
5) Example:



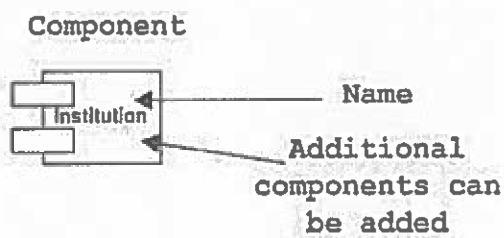
COLLABORATION:

1) Collaboration defines an interaction and is a society of roles and other elements that work together to provide higher behavior that is bigger than the sum of all the elements.
2) Following is the representation of Collaborations:

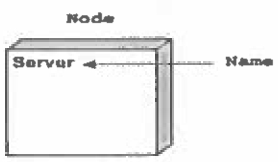
3) Collaborations have structural as well as behavior dimensions.
4) A Given class might participate in several collaborations.
5) Ellipse with dashed lines includes that only the name of collaboration.

COMPONENT:

1) Component is physical or replaceable part of a system that conforms and provides the realization of set of interfaces.
2) Component here is a software component.
3) Com++ components or Java Beans
4) Example :



NODE: Physical element that exists at run time.



BEHAVIOURAL THINGS:

1) The behavior things are dynamic parts of UML
2) These are the verbs of the model.
3) There are two primary kinds of behavioral things
   a) <u>Interactions</u> -- is a behavior that comprises of set of messages exchanged among set of objects.
   An Interaction involves a number of elements including messages, action sequences, and links. Graphically a message is rendered as directed line.

   Display

   Example:

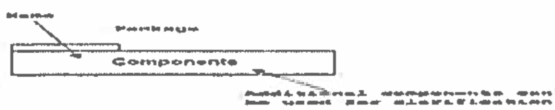
   Messages

   b) <u>State Machine:</u> Is a behavior that specifies the sequence of states an object or interaction goes through during the life time in response to events.
   State Machine involves a number of elements including states, transitions, events and activities.
   A State machine is rendered as a rounded rectangle.
   Example:

State

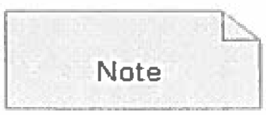
GROUPING THINGS:

1) Grouping things are organizational part of UML i.e. Package
2) A Package is a general purpose mechanism.
3) A Package contains structural things, behavior things and even other grouping things.
4) Package is rendered as Tabbed folder with name.
5) Example:



Annotational Things:

1) Annotational things are explanatory part of UML Model
2) Note is an important an notational thing.
3) These are the comments we apply to describe.
4) Annotational Things are rendered as dog-eared corner
5) Example



RELATIONSHIPS IN THE UML:

1) There are four kinds of relationships in the UML. They are
   a) Dependency b) Association c) Generalization d) Realization

DEPENDENCY:

1) Dependency is a semantic relationship between two things in which a change to one independent thing may affect the other thing (Dependent thing).
2) Dependency is rendered as a dashed line.
3) Possibly directed.
4) Example : Dependent------------------->Independent

ASSOCIATION:

1) Association is a structural relationship that describes a set of links.
2) Link is connection among objects
3) Association is rendered as a solid line.
4) Example:

Employer ———Association——— Employee

GENERALIZATION:

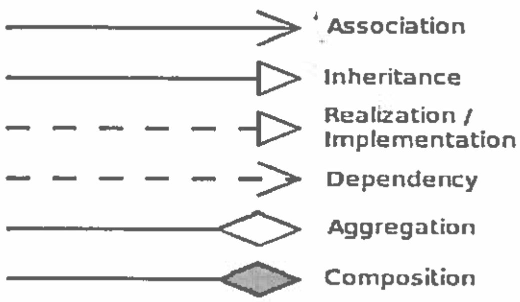
1) Generalization is a specialization – generalization relationship in which objects of specialized element(child) are substitutable for the objects of generalized element (parent)
2) Child shares the structure and behavior of parent.
3) Generalization is rendered as Solid line with arrow pointing to parent.
4) Example :

Child ———Generalization———▷ Parent

REALIZATION:

1) Realization is a semantic relationship between classifiers, where in one classifier specifies a contract that another classifier guarantees to carry out.
2) We encounter realization in two places.
   a) Between interfaces and classes that realizes them.
   b) Uses cases and the collaboration that realizes them.
3) Realization relationship is rendered as a cross between generalization and dependency.
4) Example: - - - - - -▷

Summing up Relationship Notation:



DIAGRAMS IN THE UML:

1) A Diagram is a graphical representation of set of elements.
2) Diagram is a connected graph of vertices(things) and arcs (relationships)
3) Diagrams are used to visualize a system from different perspectives.
4) The various UML Diagrams include:

a) Class Diagram	Static Diagram
b) Object Diagram	Static Diagram
c) Use Case Diagram	Dynamic Diagram
d) Sequence Diagram	Dynamic Diagram
e) Collaboration Diagram	Dynamic Diagram
f) State Chart Diagram	Dynamic Diagram
g) Activity Diagram	Dynamic Diagram
h) Component Diagram	Static Diagram
i) Deployment Diagram	Static Diagram

11) A) Discuss the essential Class Relationships.

Answer:

8) A Class, like objects, do not exist in Isolation.
9) A class relationship might indicate some sort of sharing.
10) A class relationship might indicate some kind of semantic connection.
11) There are Three Basic kinds of class relationships.

12) The First of these is GENERALIZATION/SPECILIZATION , denoting "is a" relationship.
Example: A Rose is a kind of flower, meaning that a rose is a specialized sub-class of the more general class i.e. Flower.

13) The second is a WHOLE / PART, which denotes a "Part of" relationship.
Example: A petal is not a kind of a flower, it is a part of flower.

14) The third is ASSOCIATION, which denotes some semantic dependency among otherwise unrelated classes.
Example: Roses and candles are largely independent classes, but both Represent things that we might use to decorate a dinner plate.

ASSOCIATION:

1) The identification of associations among classes is often an activity of Analysis and early design.
2) At this time we begin to discover the general dependencies among our Abstractions.
3) An Association only denotes a semantic dependency and , does not state the direction of this dependency nor does it relates the exact way in which once class relates to another.
4) Unless otherwise stated an Association implies bidirectional Navigation.

5) Here we show one-to-Many Association.

6) Each instance of wheel relates to one vehicle and each instance of Vehicle may have many wheels.

MULTIPLICITY:

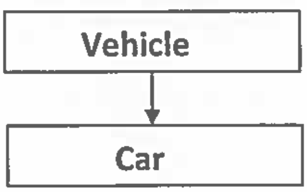
1) One-to-Many Association, meaning that for each instance of the class vehicle, There are zero (a boat which is a vehicle has no wheels) or more instances of the class Wheel, and for each wheel, there is exactly one Vehicle.

2) This denotes multiplicity of Association and there are 3 kinds of multiplicity of Associations.
   a) One-to-ONE -----> Retail Market ----> Class Sale.
      Has exactly one credit card transaction.
   b) One-to-Many ----> Automobile -------> Classes Vehicle and wheels
   c) Many-to-Many --> Retail Market ---> Classes Customer and Salesperson
      A Sales person can interact with many customers and A customer can Interact with many Sales people.

INHERITANCE:

1) Inheritance is a relationship where in one class shares the structures and / or behaviour defined in one (Single inheritance) or more (multiple inheritance) of other class.
2) The class from which another class inherits is called Super Class.
3) The class that inherits from one or more classes is Sub-class.

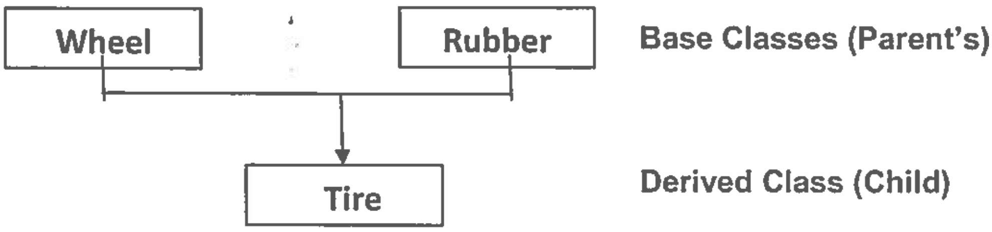
Single Inheritance: When a single class is derived from a single parent class.
It is called single inheritance.
Animal is derived from Living Things.
We can re-use attributes and methods existing in parent class.
Re-usability is important property of inheritance.



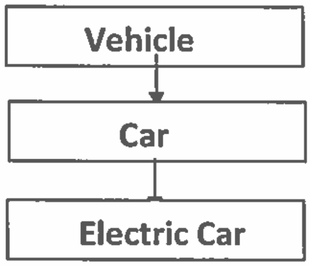
Base Class (Parent)

Derived Class (Child)

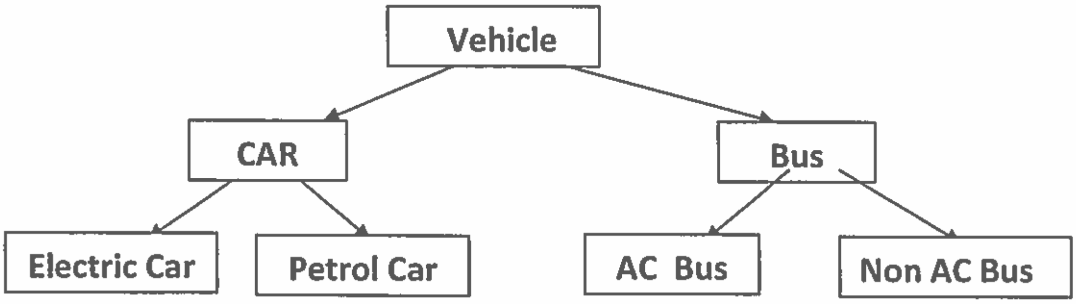
Multiple Inheritances: A Feature in OOAD in which a class can inherit characteristics and features from more than one class.



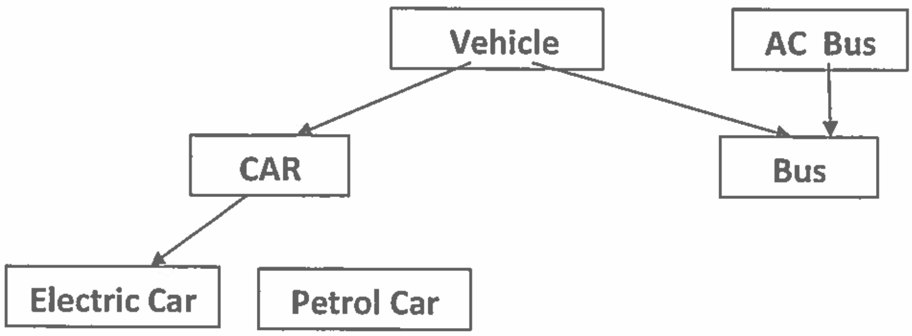
Multi Level Inheritance: A derived class is created from another derived class.



Hierarchical Inheritance: More than one sub-class is inherited from single base Base class. i.e. more than one derived class created From one base class.



Hybrid Inheritance : The combination of any two type of inheritance is used.
Example: Hybrid inheritance Plus Multi Level Inheritance.



POLYMORPHISM:

1) Polymorphism means the ability to take multiple forms.
2) It implies using operations in different ways, depending upon instance of object they are working on.
3) Allows objects with internal structures to have a common external interface.

4) Polymorphism is effective while implementing Inheritance.

5) draw() Method in a Class named Shapes may have as follows:

Public void draw int a); This draws Square by accepting one parameter.

Public void draw (int a, int b); this draws Rectangle by accepting two parameters.

AGGREGATION:

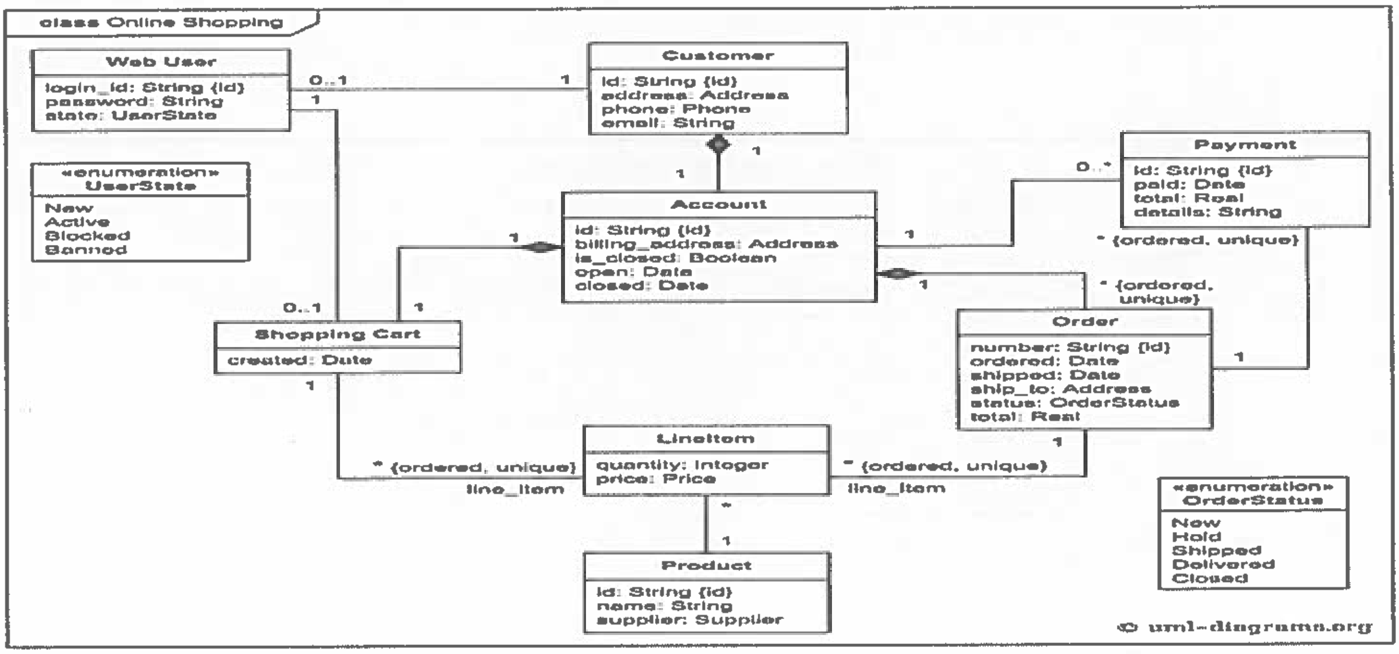
5) Aggregation relationships provide the whole / part relationships among classes.
6) Aggregation relationships among classes have direct parallel aggregation relationships among objects corresponding to classes.
7) Multiple inheritances is often confused with aggregation.
8) If, we cannot affirm that there "is a" relationship between classes then we must use Aggregation or some other relationship instead of Inheritance.

DEPNDENCIES:

1) Aside from Inheritance, Aggregation and Association there is another group of relationships called Dependencies.
2) A dependency indicates that an element on one end of the relationship, in some manner depends on the element on other end of the relationship.
3) If, one of these element changes, there could be impact to other.
   Example: One Module depends on another Module.

11) B) Draw the class Diagram for Online Shopping Management System

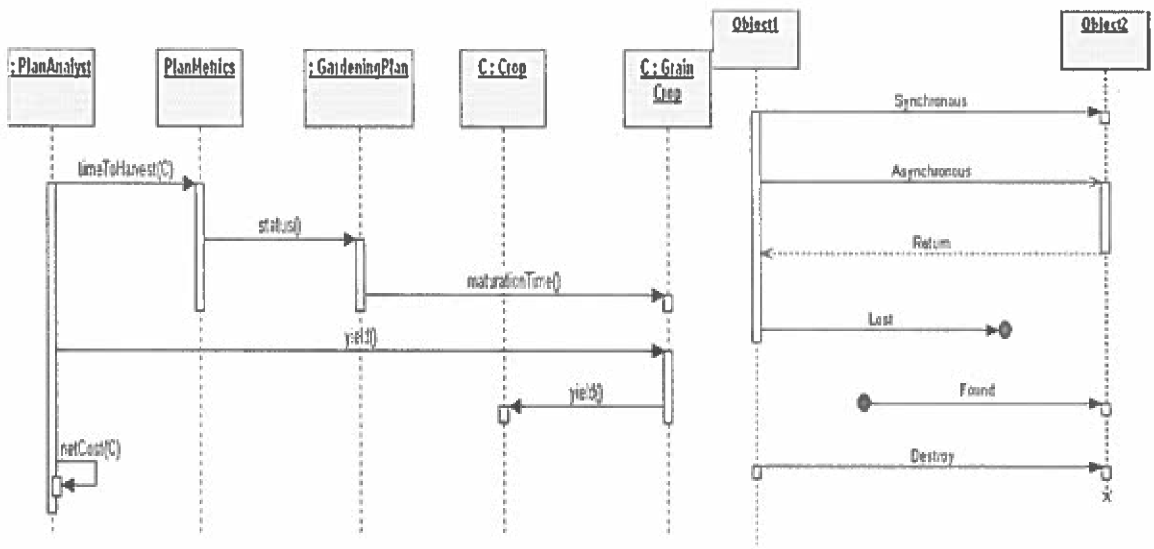
Answer:



12) A) How do you use interaction diagram when you model dynamic aspects of system. Explain with example.

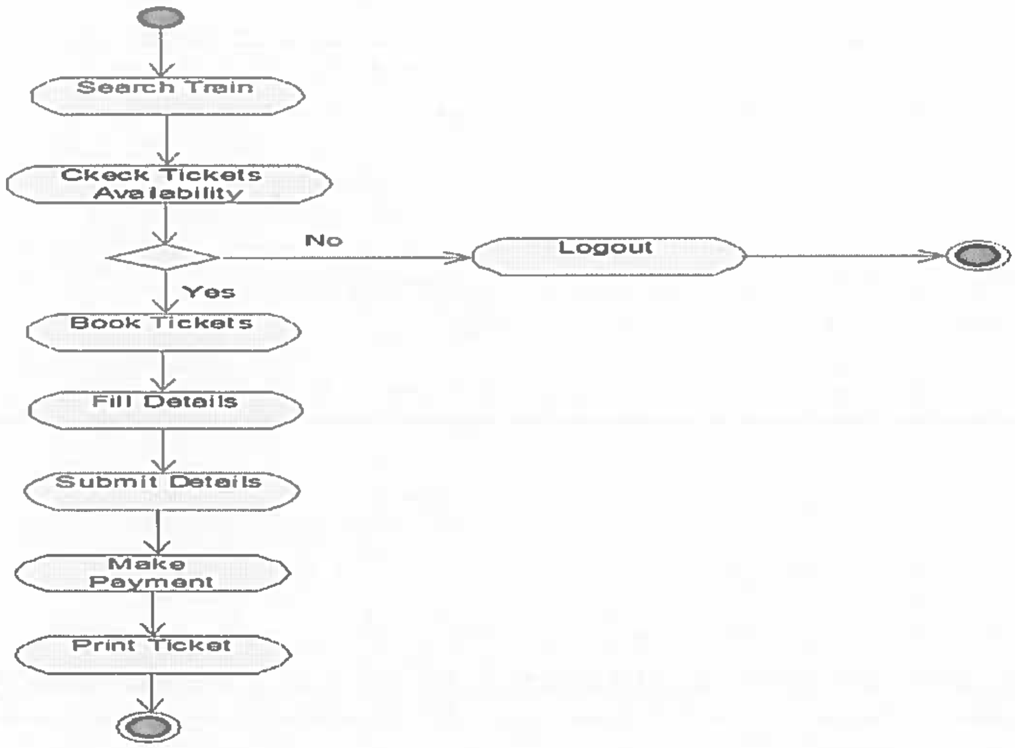
Answer:

1. There are two interaction Diagrams,
   a) Sequence diagram b) Collaboration diagram
2. There are 5 Dynamic diagrams used in UML and out of which 2 are interaction diagrams.
3. An interaction diagram shows an interaction, consisting of set of objects, and their relationships, include the messages dispatched among them.
4. A sequence diagram is an interaction diagram that emphasizes the time ordering of message.
5. A collaboration diagram is an interaction diagram that emphasizes the structural organizations of the object that send and receive message.
6. Interaction diagrams used for the constructing executable systems through forward and reverse engineering.
7. To draw these diagrams we have to visualize a running system.
8. Interaction diagrams commonly contains
   a) Objects b) Links c) Messages
9. Interaction diagrams may also contain notes and constraints.
10. Following is the example of Interaction Diagram for Gardening and Agriculture related activities.



12) B) Draw and explain the activity diagram for an online Railway Management System.

Answer:



13) A) What is use case? Explain use cases for ATM with a Diagram.

Answer:

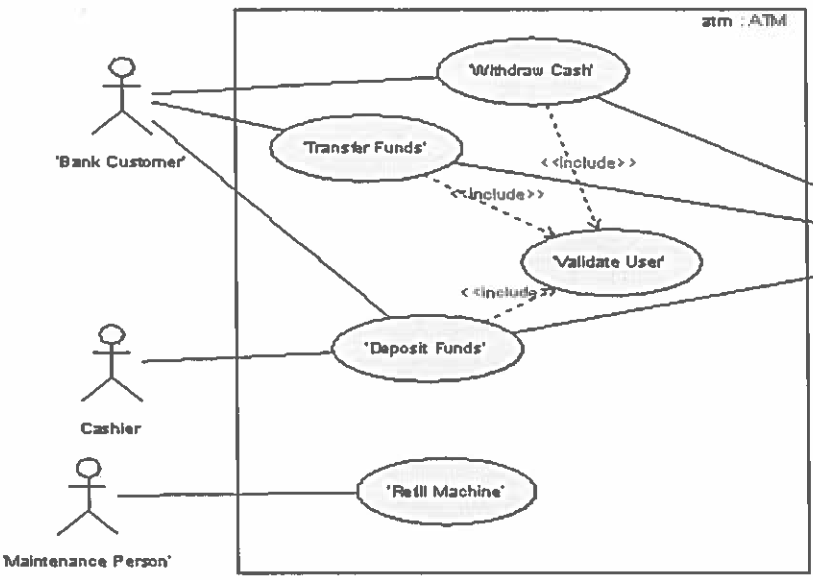
USE CASES

1. A use case specifies the behavior of a system or part of a system and is description of set of sequences of actions, including variants that a system performs, to yield an observable result of value to an actor.
2. We apply use cases to capture the intended behavior of the system that we are developing.
3. Use cases help us to understand with system end users and domain experts.
4. Use Cases serve to help validate architecture.
5. A use case represents functional requirements of system. Example, Bank processing of Loans.
6. A use case involves interaction of actors and system.

7) Use case diagram is one of the five diagrams in UML for modeling the dynamic aspect of systems.
8) The use case diagram commonly contains

a) Use cases b) Actors c) Dependencies, generalization and association relationships.

9) Following can be one of the model of USE CASE for ATM:



10) The student should explain above type of diagram in his own words.

13) B) Difference between collaboration diagram and sequence diagram.

Answer:

1) The sequence diagram are used to represent the sequence of messages that are flowing from one object to another. The collaboration diagram are used to represent the structural organization of the system and the messages that are sent and received.

2) The key difference between sequence diagram and collaboration diagram is that the sequence diagram is used when the time sequence is more important while the collaboration diagram is used when the object organization is more important.

3) In collaboration diagram method call sequence number is indicated while in sequence diagram we do no show method call sequence number.

4) The sequence diagrams are better suited for analysis of activities while the collaboration diagrams are better suited for depicting simpler interactions of the small number of objects.

14) A) what is component diagram? Draw the component diagram for online reservation system.

Answer:

1) Component diagrams are one of the two kinds of diagrams found in modeling the physical aspects of OO systems.
2) We use component diagram to model static implementation view of a system
3) Component diagrams are essentially class diagrams that focus on system's components
4) Components diagrams involves in modeling the physical things that reside on a node, such as executables, libraries, files and documents

5) Component diagrams are used to visualize the static aspects of these physical components, their relationships to specify details of construction.

6) A component diagram commonly contains

   a) Components b) Interfaces

   c) Dependency, generalization association and realization relationships

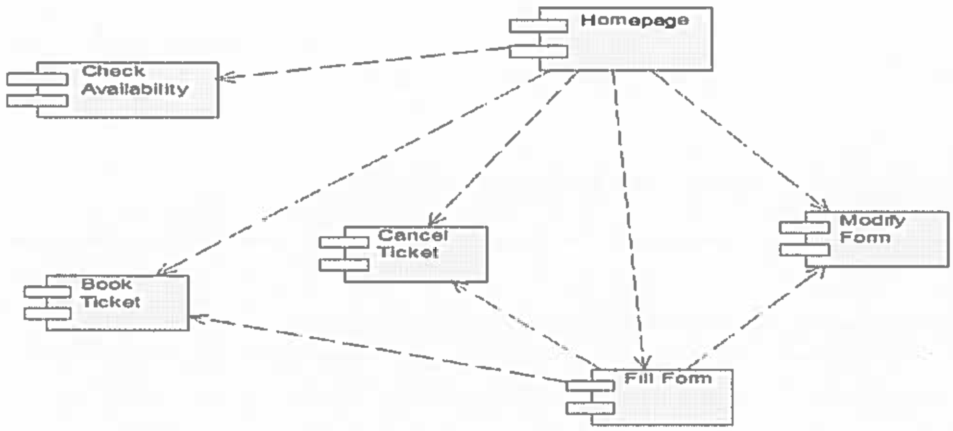
7) Components diagrams contain notes, constants, packages, subsystems

8) Components diagrams describe the organization and wiring of the physical components in a system.

9) The component diagram common uses include

a) To model source code
b) To model executables releases
c) To model physical databases
d) To model adaptable systems

10) Following is one of the model's of component diagram for **online reservation system.**



14) B) Explain Deployment Diagram.

Answer:

Deployment Diagrams

1) Deployment diagram is one of the two kinds of diagrams used in modeling the physical aspects of an Object Oriented System.

2) A Deployment diagram shows the configuration of run time processing nodes and the components that live on them.

3) We use Deployment diagram to model the static deployment view of a system.

4) Modeling the topology of the Hardware on which the system executes.

5) Deployment diagrams are essentially class diagrams that focus on system nodes.

6) Deployment diagrams are not only important for visualizing, specifying , documenting the embedded , client server and distributed systems but also for managing the executable systems thru forward and reverse engineering.

7) Graphically a deployment diagram is a collection of Vertices and Arcs.

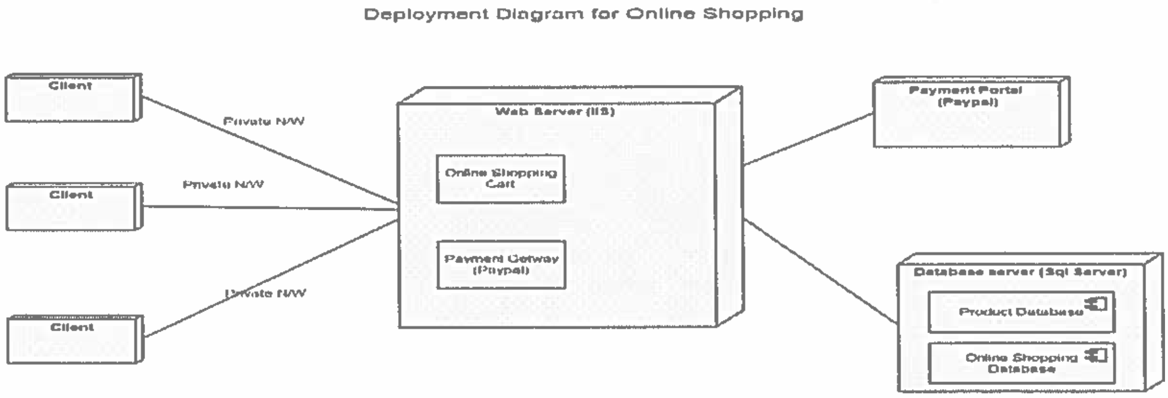
8) Deployment diagram commonly contains :

   a) Nodes b) Dependency and Association relationships
   a) May also contain notes and constraints

9) Deployment diagrams may also contain components, each of which may live in some Node.

10)May also contain packages or sub-programs.

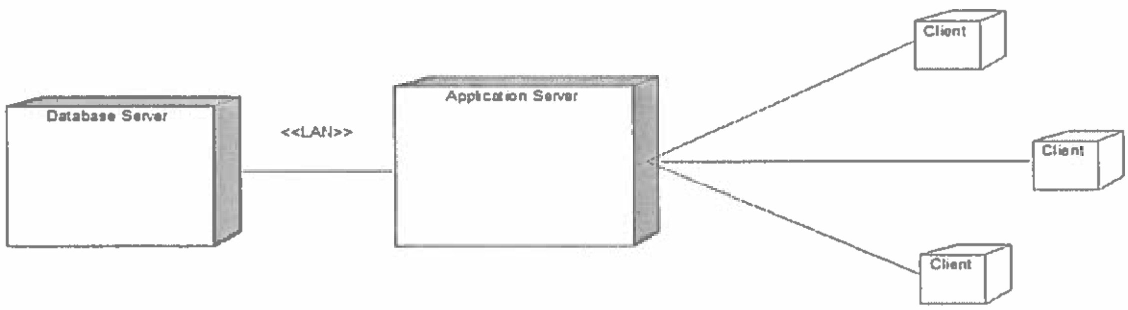
11)Following is the depict or diagram of deployment Diagram for **online reservation system**



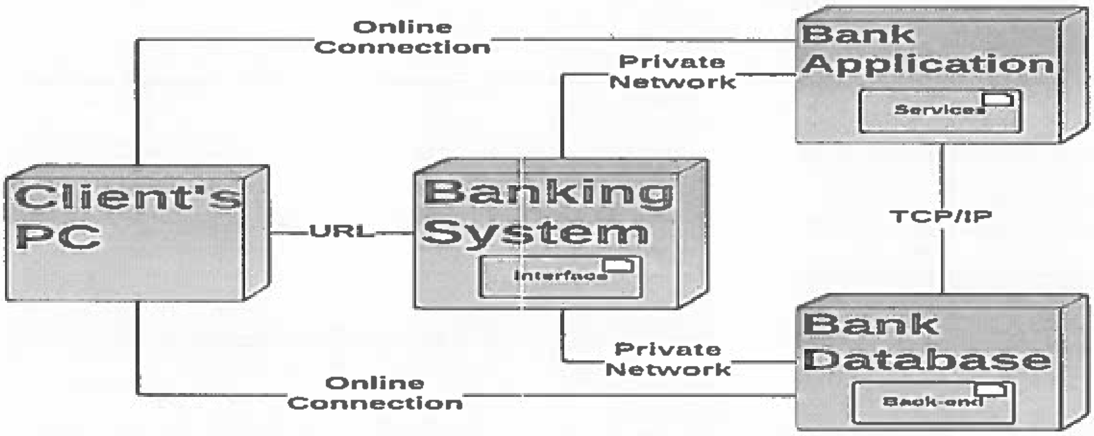
Note: Deployment diagram is not mandatory

15) A) Draw the deployment diagram for Bank Management System

Answer:



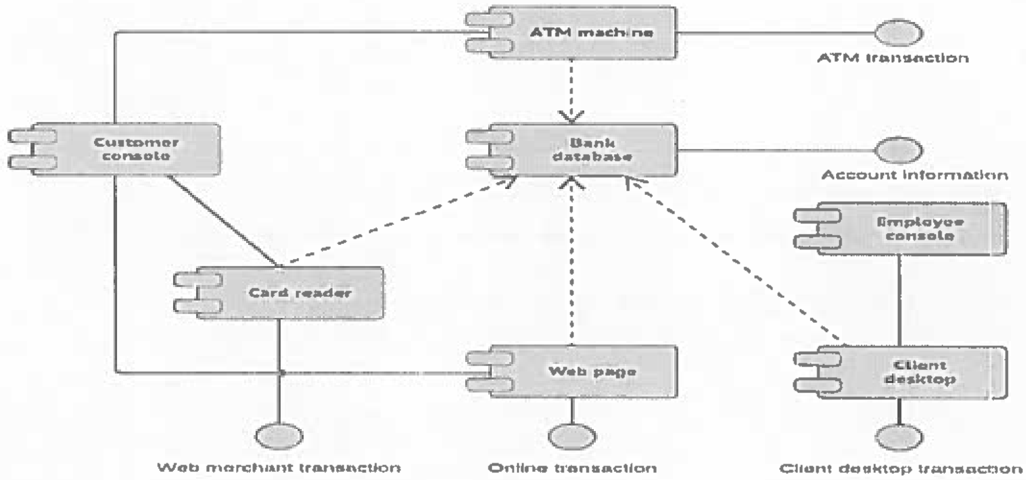
OR



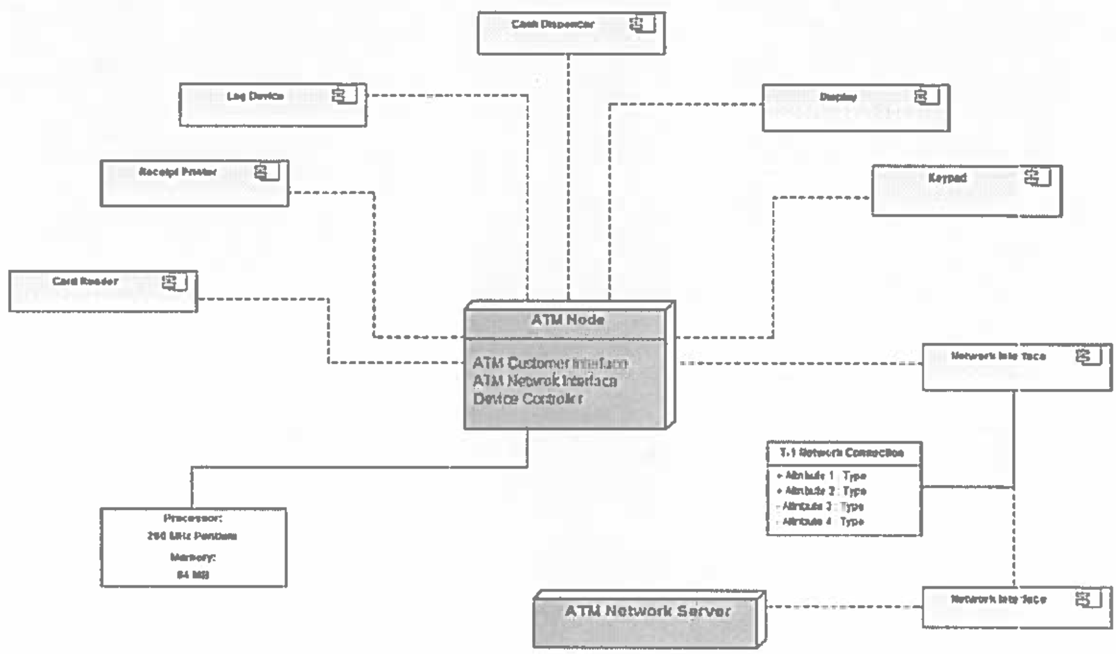
15) B) Draw the component and deployment diagram for ATM

Answer:

Following is the model of Component diagram for ATM:



Following is the Deployment diagram for ATM:



NSRIT

Semester End Regular Examination, Nov./Dec., 2022
Scheme of Evaluation

Degree	B. Tech.	**Program**	CSE (AI & ML)			**Academic Year**	2022 - 2023
Course Code	20AI003	**Test Duration**	3 Hrs.	**Max. Marks**	70	**Semester**	V
Course	**Cloud Computing Essentials**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)		
No.	Questions (1 through 5)	Marks
1	Any four characteristics of cloud computing.	2 M
2	Virtualization definition	2M
3	Any four advantages of cloud storage.	2M
4	2 types of resource provisioning	2M
5	Hadoop definition	2M
Part B (Long Answer Questions 5 x 12 = 60 Marks)		
No.	Questions (6 through 15)	Marks
6 (a)	Evolution of cloud computing. (6 M)	6M
6 (b)	principles of distributed computing. (6M)	6M
	OR	
7 (a)	principles of parallel computing. (6M)	6M
7 (b)	on demand provisioning 2 types explanation (3 M +3M)	6M
8 (a)	Any 5 types of virtualization.	5M
8 (b)	Service Oriented Architecture explanation (4M) SOA diagram (3M)	7M
	OR	
9 (a)	virtualization of CPU (2 M) virtualization of Memory (2M) Virtualization of I/O devices. (2M)	6M
9 (b)	Tools and mechanisms of cloud virtualization (6M)	6M
10 (a)	NIST cloud computing Reference Architecture diagram. (3 M) Explanation (4 M)	7M
10 (b)	Cloud deployment models any three models. (5M)	5M
	OR	
11 (a)	Layers of cloud computing architecture (3M) Explanation (4M)	7M
11 (b)	Storage as a service explanation. (5M)	5M
12 (a)	Definition of resource provisioning (2M) Explanation resource provisioning 3 methods. (3M)	5M
12 (b)	security governance (7M)	7M
	OR	
13 (a)	cloud security challenges any 6 (6 M)	6M
13 (b)	Virtual machine security (6 M)	6M
14 (a)	Map Reduce explanation. (3M) Map Reduce example (3M)	6M
14 (b)	Google App Engine explanation and diagram (3M +3M)	6M
	OR	
15 (a)	Explanation of Open Nebula and Eucalyptus. (3M+3M)	6M
15 (b)	Explanation of all four levels of federation. (6M)	6M

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Answer Key

Degree	B. Tech.	**Program**	CSE (AI & ML)			**Academic Year**	2022 - 2023
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Course	**Cloud Computing Essentials**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four characteristics of cloud computing.	20AI003.1	L1
	1. On-demand self-services: The Cloud computing services does not require any human administrators, user themselves are able to provision, monitor and manage computing resources as needed. 2. Broad network access: The Computing services are generally provided over standard networks and heterogeneous devices. 3. Rapid elasticity: The Computing services should have IT resources that are able to scale out and in quickly and on as needed basis. 4. Resource pooling: The IT resource (e.g., networks, servers, storage, applications, and services) present are shared across multiple applications and occupant in an uncommitted manner. 5. Measured service: The resource utilization is tracked for each application and occupant; it will provide both the user and the resource provider with an account of what has been used.		
2	Define Virtualization.	20AI003.2	L1
	Virtualization is technology that lets you create useful IT services using resources that are traditionally bound to hardware. It allows you to use a physical machine's full capacity by distributing its capabilities among many users or environments.		
3	List any four advantages of cloud storage.	20AI003.3	L1
	1. Usability and accessibility 2. Security 3. Cost-efficient 4. Convenient sharing of files 5. Automation 6. Multiple users		
4	List the resource provisioning types.	20AI003.4	L1
	Resource Provisioning means the selection, deployment, and run-time management of software (e.g., database server management systems, load balancers) and hardware resources (e.g., CPU, storage, and network) for ensuring guaranteed performance for applications. There are two types: 1. Static Provisioning 2. Dynamic Provisioning Static Provisioning: For applications that have predictable and generally unchanging demands / workloads, it is possible to use "static provisioning" effectively. With advance provisioning, the customer contracts with the provider for services. Dynamic Provisioning: In cases where demand by applications may change or vary, "dynamic provisioning" techniques have been suggested whereby VMs may be migrated on-the-fly to new compute nodes within the cloud.		
5	Define Hadoop.	20AI003.5	L1
	Hadoop is an Apache open source framework written in java that allows distributed processing of large datasets across clusters of computers using simple programming models. Hadoop = MapReduce + HDFS (MapReduce-> Processing ; HDFS -> Storage)		

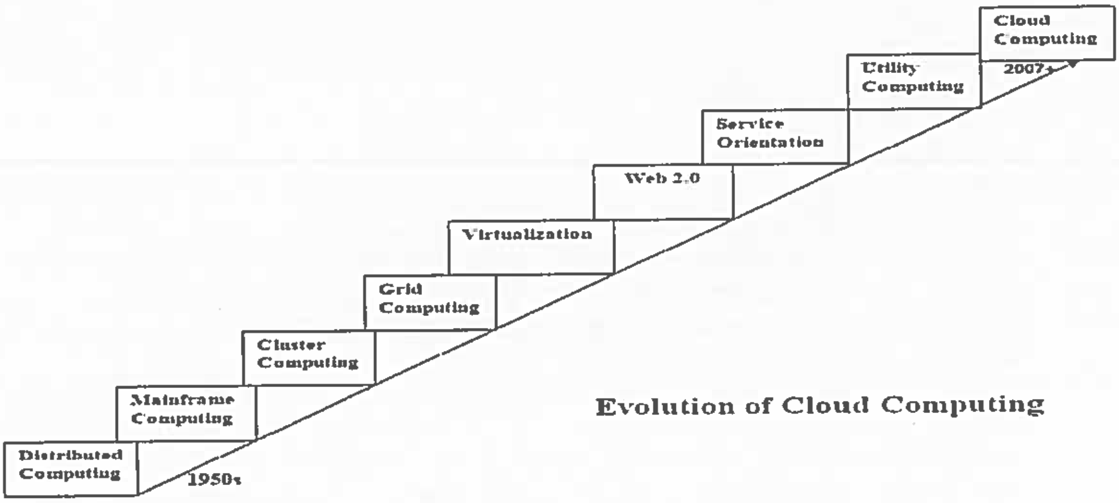
Part B (Long Answer Questions 5 x 12 = 60 Marks)				
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain Evolution of cloud computing.	6M	20AI003.1	L2

1. Distributed Systems: It is a composition of multiple independent systems but all of them are depicted as a single entity to the users. The purpose of distributed systems is to share resources and also use them effectively and efficiently. Distributed systems possess characteristics such as scalability, concurrency, continuous availability, heterogeneity, and independence in failures. But the main problem with this system was that all the systems were required to be present at the same geographical location. Thus to solve this problem, distributed computing led to three more types of computing and they were-Mainframe computing, cluster computing, and grid computing.

2. Mainframe computing: Mainframes which first came into existence in 1951 are highly powerful and reliable computing machines. These are responsible for handling large data such as massive input-output operations. Even today these are used for bulk processing tasks such as online transactions etc. These systems have almost no downtime with high fault tolerance. After distributed computing, these increased the processing capabilities of the system. But these were very expensive. To reduce this cost, cluster computing came as an alternative to mainframe technology.

3. Cluster computing: In 1980s, cluster computing came as an alternative to mainframe computing. Each machine in the cluster was connected to each other by a network with high bandwidth. These were way cheaper than those mainframe systems. These were equally capable of high computations. Also, new nodes could easily be added to the cluster if it was required. Thus, the problem of the cost was solved to some extent but the problem related to geographical restrictions still pertained. To solve this, the concept of grid computing was introduced.

4. Grid computing: In 1990s, the concept of grid computing was introduced. It means that different systems were placed at entirely different geographical locations and these all were connected via the internet. These systems belonged to different organizations and thus the grid consisted of heterogeneous nodes. Although it solved some problems but new problems emerged as the distance between the nodes increased. The main problem which was encountered was the low availability of high bandwidth connectivity and with it other network associated issues. Thus cloud computing is often referred to as "Successor of grid computing".



Evolution of Cloud Computing

5. Virtualization: It was introduced nearly 40 years back. It refers to the process of creating a virtual layer over the hardware which allows the user to run multiple instances simultaneously on the hardware. It is a key technology used in cloud computing. It is the base on which major cloud computing services such as Amazon EC2, VMware

6. Web 2.0: It is the interface through which the cloud computing services interact with the clients. It is because of Web 2.0 that we have interactive and dynamic web pages. It also increases flexibility among web pages. Popular examples of web 2.0 include Google Maps, Facebook, Twitter, etc. Needless to say, social media is possible because of this technology only. In gained major popularity in 2004.

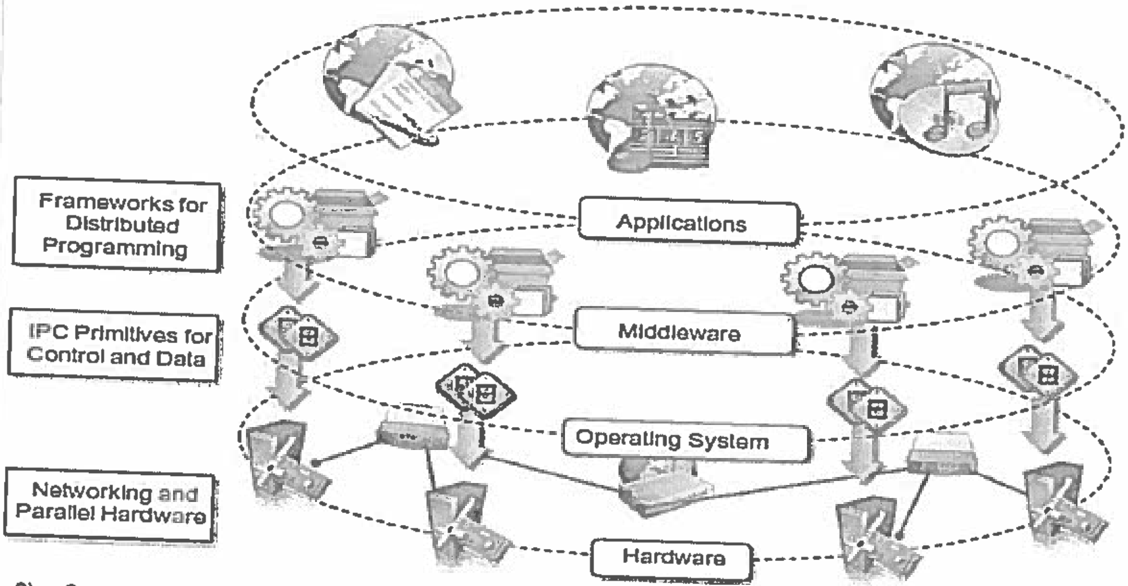
7. Service orientation: It acts as a reference model for cloud computing. It supports low-cost, flexible, and evolvable applications. Two important concepts were introduced in this computing model. These were Quality of Service (QoS) which also includes the SLA (Service Level Agreement) and Software as a Service (SaaS).
8. Utility computing: It is a computing model that defines service provisioning techniques for services such as compute services along with other major services such as storage, infrastructure, etc which are provisioned on a pay-per-use basis.

6 (b)	Explain the principles of distributed computing.	6M	20AI003.1	L2

Elements of Distributed Computing:

1) General concepts and definitions

- Distributed computing studies the models, architectures, and algorithms used for building and managing distributed systems.
- As general definition of the term distributed system, we use the one proposed by Tanenbaum
  - A distributed system is a collection of independent computers that appears to its users as a single coherent system.
- This definition is general enough to include various types of distributed computing systems that are especially focused on unified usage and aggregation of distributed resources.
- Communications is another fundamental aspect of distributed computing. Since distributed systems are composed of more than one computer that collaborate together, it is necessary to provide some sort of data and information exchange between them, which generally occurs through the network.
  - A distributed system is one in which components located at networked computers communicate and coordinate their action only by passing messages.
- As specified in this definition, the components of a distributed system communicate with some sort of message passing. This is a term that encompasses several communication models.



2) Components of distributed System

- A distributed system is the result of the interaction of several components that traverse the entire computing stack from hardware to software.
- It emerges from the collaboration of several elements that- by working together- give users the illusion of a single coherent system.
- The figure provides an overview of the different layers that are involved in providing the services of a distributed system.

3) Architectural styles for distributed computing

- At the very bottom layer, computer and network hardware constitute the physical infrastructure; these components are directly managed by the operating system, which provides the basic services for inter process communication (IPC), process scheduling and management, and resource management in terms of file system and local devices.

- Taken together these two layers become the platform on top of which specialized software is deployed to turn a set of networked computers into a distributed system
- Although a distributed system comprises the interaction of several layers, the middleware layer is the one that enables distributed computing, because it provides a coherent and uniform runtime environment for applications.
- There are many different ways to organize the components that, taken together, constitute such an environment.
- The interactions among these components and their responsibilities give structure to the middleware and characterize its type or, in other words, define its architecture.
- Architectural styles aid in understanding the classifying the organization of the software systems in general and distributed computing in particular.
- The use of well-known standards at the operating system level and even more at the hardware and network levels allows easy harnessing of heterogeneous components and their organization into a coherent and uniform system.

OR

7 (a)	Explain the principles of parallel computing.	6M	20AI003.1	L2

Parallel processing influencing factors:

The development of parallel processing is being influenced by many factors. The prominent among them include the following:

- Computational requirements are ever increasing in the areas of both scientific and business computing. The technical computing problems, which require high-speed computational power, are related to life sciences, aerospace, geographical information systems, mechanical design and analysis etc.
  - Sequential architectures are reaching mechanical physical limitations as they are constrained by the speed of light and thermodynamics laws.
- The speed which sequential CPUs can operated is reaching saturation point (no more vertical growth), and hence an alternative way to get high computation speed is to connect multiple CPUs (opportunity for horizontal growth).
  - Hardware improvements in pipelining , super scalar, and the like are non scalable and require sophisticated compiler technology.
- Developing such compiler technology is a difficult task.
  - Vector processing works well for certain kinds of problems. It is suitable mostly for scientific problems (involving lots of matrix operations) and graphical processing.
- It is not useful for other areas, such as databases.
  - The technology of parallel processing is mature and can be exploited commercially here is already significant R&D work on development tools and environments.
  - Significant development in networking technology is paving the way for heterogeneous computing.

b. Hardware architectures for parallel Processing

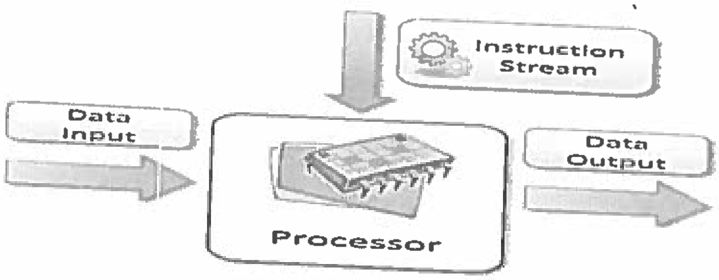
- The core elements of parallel processing are CPUs. Based on the number of instructions and data streams, that can be processed simultaneously, computing systems are classified into the following four categories:

i. Single-instruction, Single-data (SISD) systems
ii. Single-instruction, Multiple-data (SIMD) systems
iii. Multiple-instruction, Single-data (MISD) systems
iv. Multiple-instruction, Multiple-data (MIMD) systems

(i) Single – Instruction , Single Data (SISD) systems

- SISD computing system is a uni-processor machine capable of executing a singleinstruction, which operates on a single data stream.
- Machine instructions are processed sequentially, hence computers adopting this modelare popularly called sequential computers.
- Most conventional computers are built using SISD model.

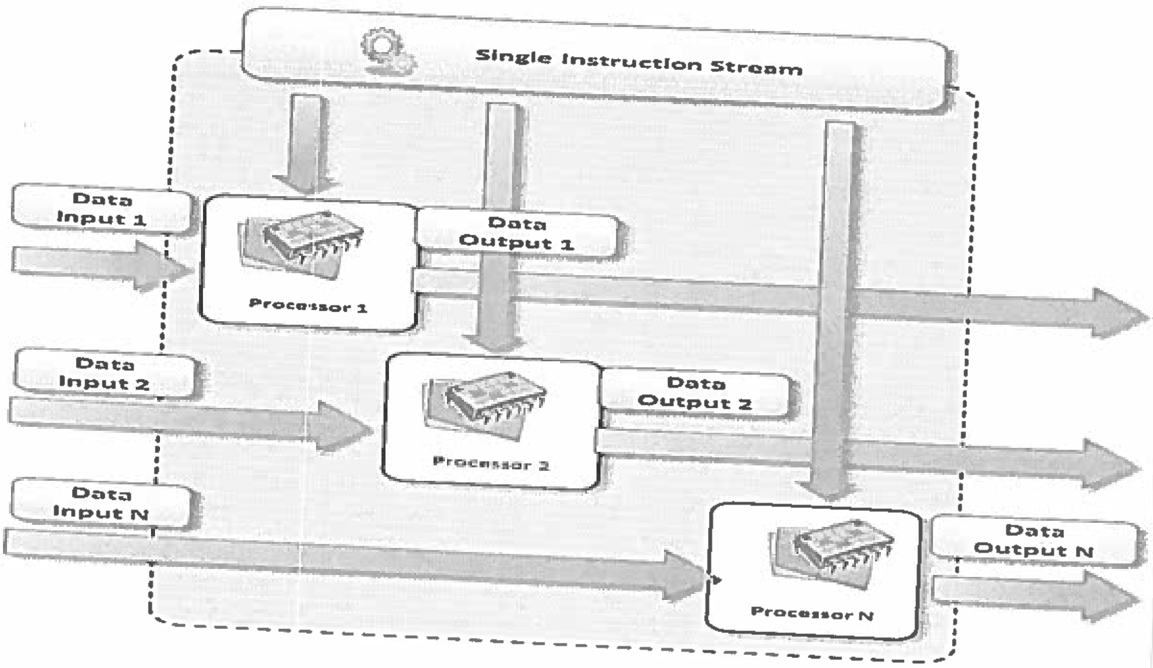
- All the instructions and data to be processed have to be stored in primary memory.
- The speed of processing element in the SISD model is limited by the rate at which thecomputer can transfer information internally.
- Dominant representative SISD systems are IBM PC, Macintosh, and workstations.



(ii) Single – Instruction, Multiple Data (SIMD) systems

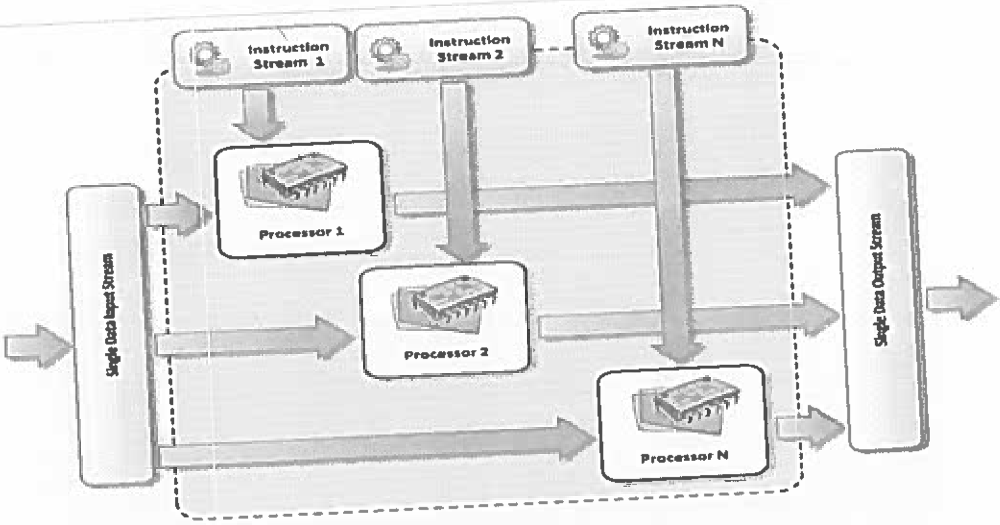
- SIMD computing system is a multiprocessor machine capable of executing the same instruction on all the CPUs but operating on different data streams.
- Machines based on this model are well suited for scientific computing since they involve lots of vector and matrix operations.
- For instance statement Ci = Ai * Bi, can be passed to all the processing elements (PEs), organized data elements of vectors A and B can be divided into multiple sets (N- sets forN PE systems), and each PE can process one data set.

Dominant representative SIMD systems are Cray's Vector processing machine and Thinking Machines Cm*, and GPGPU accelerators



(iii) Multiple – Instruction , Single Data (MISD) systems

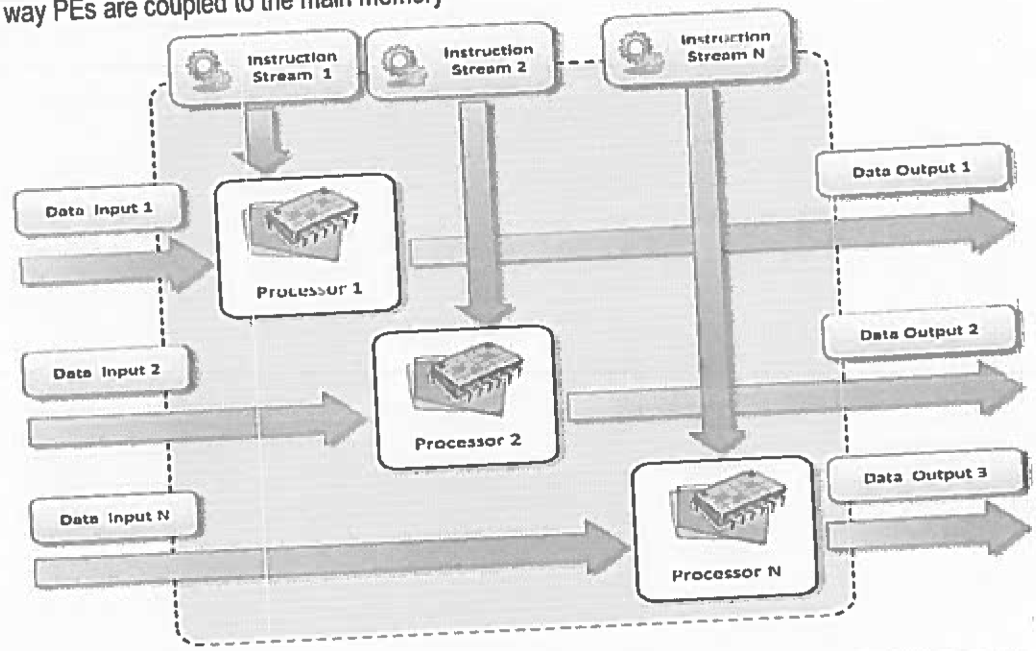
- MISD computing system is a multi processor machine capable of executing different instructions on different PEs all of them operating on the same data set.
- Machines built using MISD model are not useful in most of the applications.
- Few machines are built but none of them available commercially.
- This type of systems is more of an intellectual exercise than a practical configuration.



(iv) Multiple – Instruction, Multiple Data (MIMD) systems

- MIMD computing system is a multi processor machine capable of executing multiple instructions on multiple data sets.
- Each PE in the MIMD model has separate instruction and data streams; hence machines built using this model are well suited to any kind of application.
- Unlike SIMD, MISD machine, PEs in MIMD machines work asynchronously,

MIMD machines are broadly categorized into shared-memory MIMD and distributed memory MIMD based on the way PEs are coupled to the main memory



7 (b)	Explain on demand provisioning in cloud computing.	6M	20AI003.1	L2

On-demand Provisioning.

- Resource Provisioning means the selection, deployment, and run-time management of software (e.g., database server management systems, load balancers) and hardware resources (e.g., CPU, storage, and network) for ensuring guaranteed performance for applications.
- Resource Provisioning is an important and challenging problem in the large-scale distributed systems such as Cloud computing environments.
- There are many resource provisioning techniques, both static and dynamic each one having its own advantages and also some challenges.
- These resource provisioning techniques used must meet Quality of Service (QoS) parameters like availability, throughput, response time, security, reliability etc., and thereby avoiding Service Level

Agreement (SLA) violation.
- Over provisioning and under provisioning of resources must be avoided.
- Another important constraint is power consumption.
- The ultimate goal of the cloud user is to minimize cost by renting the resources and from the cloud service Provider's perspective to maximize profit by efficiently allocating the resources.
- In order to achieve the goal, the cloud user has to request cloud service provider to make a provision for the resources either statically or dynamically.
- So that the cloud service provider will know how many instances of the resources and what resources are required for a particular application.
- By provisioning the resources, the QoS parameters like availability, throughput, security, response time, reliability, performance etc must be achieved without violating SLA

There are two types

- Static Provisioning
- Dynamic Provisioning

Static Provisioning :

- For applications that have predictable and generally unchanging demands/workloads, it ispossible to use "static provisioning" effectively.
- With advance provisioning, the customer contracts with the provider for services.
- The provider prepares the appropriate resources in advance of start of service.
- The customer is charged a flat fee or is billed on a monthly basis.

Dynamic Provisioning:

- In cases where demand by applications may change or vary, "dynamic provisioning" techniques have been suggested whereby VMs may be migrated on-the-fly to new compute nodes within the cloud.
- The provider allocates more resources as they are needed and removes them when they are not.
- The customer is billed on a pay-per-use basis.
- When dynamic provisioning is used to create a hybrid cloud, it is sometimes referred to as cloud bursting.

8 (a)	Explain the types of virtualization.	5M	20AI003.2	L2

Types of virtualization:
There are 7 types of virtualization based on the component which is used.

1. Hardware Virtualization
2. Software Virtualization
3. Memory Virtualization
4. Storage Virtualization
5. Network Virtualization
6. Desktop Virtualization
7. Data Virtualization

1. Hardware Virtualization:
Hardware virtualization is a common type, often referred to the types of server virtualization, that runs multiple virtual machines in a single physical server. All the virtual servers being executed can share the resources (CPU, memory, storage, etc.) of the physical server on the hypervisor (the software layer between the underlying hardware and the virtual machines).
Hardware virtualization can also be divided into 3 different types as follows.
Full-Virtualization. Completely emulate the hardware with the software (including client operating systems and applications) unmodified.
Emulation Virtualization. The virtual machine simulates the hardware and becomes independent of it. The guest operating system does not require any modifications.
Para-Virtualization Virtualization. Guest applications perform in their own isolated domains, as if they were running on a separate system, that doesn't emulate the environment. This means that guest applications need to

be specially modified to run properly in this environment.

2 . Software Virtualization

Software virtualization is to create multiple virtualized environments on a physical host. This type of virtualization creates a complete computer system with hardware which allows guest operating systems to run.

Operating System Virtualization. This type (also be called a container) creates a virtualization layer of software on top of the host operating system and a separate operating system interface for the application, as if it were in a standalone operating system, even with a proprietary file system, system libraries, network configuration, etc.

Application Virtualization. This type essentially abstracts the application layer from the operating system so that applications are dependent on the underlying operating system, for example allowing Windows applications to run on Linux.

Service Virtualization. This type hosts specific processes and services related to a particular application such as API-driven applications, cloud-based applications, service-oriented architectures, etc.

3. Memory Virtualization:

This type of virtualization is to aggregate physical memory across different servers into a single virtualized memory pool. The common memory virtualization products and platforms are RNA Network Memory Virtualization Platform, IBM Websphere extremeScale, Oracle Coherence.

Application-level control. Applications running on the computer can access the memory pool directly through the API or the file system.

Operating system-level control. Provide access to the memory pool through the operating system.

4. Storage Virtualization

Storage virtualization is presenting a logical view of physical storage resources to the host system, which can be simply understood as combining multiple physical storage together and displaying them as a single storage device such as dividing the hard disk into multiple partitions. The common storage virtualization products are FalconStor Network Storage Server, IBM Storage Virtualization, QUADStor Systems.

Block Virtualization. Separate logical storage (partitioned area) from physical storage for direct access regardless of physical storage or heterogeneous organization.

File Virtualization. Eliminate the dependency between the file level data and the physical storage location (Network Attached Storage). You can simply understand this as the storage system granting access to files stored over multiple hosts.

5.Network Virtualization

Network virtualization combines all the physical network tools into a single software-based resource (virtual network). This type is often related to platform virtualization and generally used with resource virtualization. The common network virtualization products and platforms are 6WIND Virtual Accelerator, Cisco Nexus Virtual Services Appliance, JunosV App Engine.

Internal network: Provides network function for software on a single web server, enabling a single system to operate as a network.

External network: Combine multiple networks or parts of networks into a single unit.

6. Desktop Virtualization

Desktop virtualization also known as virtual desktop infrastructure (VDI) separates the desktop environment, software, and the physical client device used to access that software, and store the user's desktop on a remote server so that the user can access the desktop remotely from another location. In addition, the security is high because the transferred data associated with desktop virtualization is protected by secure protocols. The common desktop virtualization products and platforms include Amazon WorkSpaces, Citrix XenDesktop, VMware Workstation.

7.Data Virtualization:

Data virtualization is a data management approach that retrievals and manipulates data to provide customers with a holistic view of data when the format or physical location of the data source is not yet clear, which will effectively reduce the errors of data entry and formatting. The common data virtualization products are Denodo, JBoss Data Virtualization, TIBCO Data Virtualization.

8 (b)	Explain Service Oriented Architecture with neat sketch.	7M	20AI003.2	L2

SOA – Service Oriented Architecture:

- Service provider publishes service description (WSDL), e.g. on a service broker
- Service Requester finds service (on service broker) and dynamically binds to service
- Enables ad-hoc collaboration and Enterprise Application Integration (EAI) within web- based information

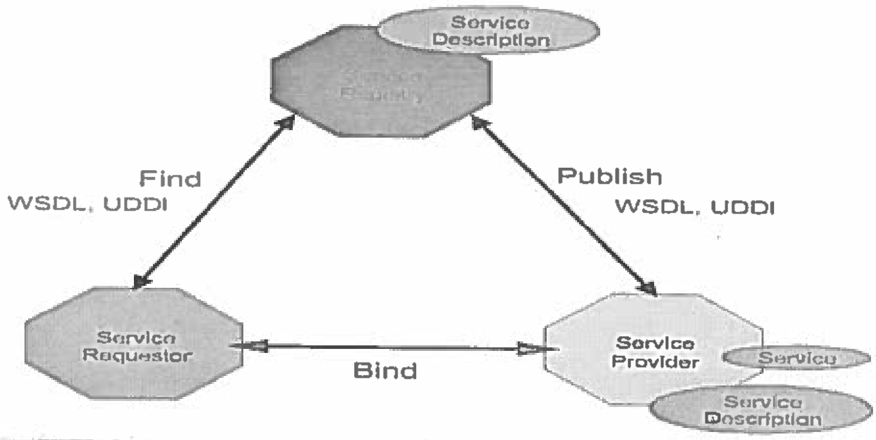
systems

- SOA is about how to design a software system that makes use of services of new or legacy applications through their published or discoverable interfaces.
- These applications are often distributed over the networks.
- SOA also aims to make service interoperability extensible and effective.
- It prompts architecture styles such as loose coupling, published interfaces and a standard communication model in order to support this goal.

Properties of SOA

1. Logical view
2. Message orientation
3. Description orientation

1. Logical view



- The SOA is an abstracted, logical view of actual programs, databases, business processes.
- Defined in terms of what it does, typically carrying out a business-level operation.
- The service is formally defined in terms of the messages exchanged between provider agents and requester agents.

2. Message Orientation

- The internal structure of providers and requesters include the implementation language, process structure, and even database structure.
- These features are deliberately abstracted away in the SOA
- Using the SOA discipline one does not and should not need to know how an agent implementing a service is constructed.
- The key benefit of this concerns legacy systems.
- By avoiding any knowledge of the internal structure of an agent, one can incorporate any software component or application to adhere to the formal service definition.

3. Description orientation

- A service is described by machine-executable metadata.
- The description supports the public nature of the SOA.
- Only those details that are exposed to the public and are important for the use of the service should be included in the description.
- The semantics of a service should be documented, either directly or indirectly, by its description.

SOA :

- XML - SOAP Based Web Services
- Extensible Markup Language (XML) is a markup language designed as a standard way to encode documents and data.
- SOAP is an acronym for Simple Object Access Protocol. It is an XML-based messaging protocol for exchanging information among computers.
- RESTful Web services :Web service is the terminology used everywhere

Service Oriented Architecture model implemented by XML Web Services

Model WSDL – Web Services Description Languages

- Provides a machine-readable description of how the service can be called, what parameters it expects, and what data structures it returns.
- Used in combination with SOAP and an XML Schema to provide Web services over the Internet.

UDDI – Universal Description, Discovery and Integration

- White Pages — address, contact, and known identifiers.
- Yellow Pages — industrial categorizations based on standard taxonomies;
- Green Pages — technical information about services exposed by the business.

OR

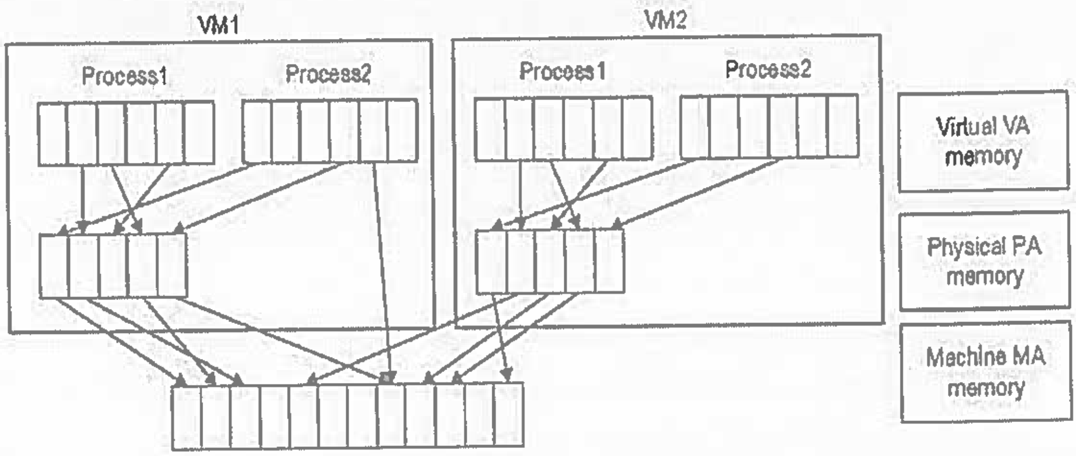
9 (a)	Explain virtualization of CPU, Memory and I/O devices.	6M	20AI003.2	L2

1. CPU Virtualization:
   A VM is a duplicate of an existing computer system in which a majority of the VM instructions are executed on the host processor in native mode. Thus, unprivileged instructions of VMs run directly on the host machine for higher efficiency. Other critical instructions should be handled carefully for correctness and stability.
   The critical instructions are divided into three categories:
   1. Privileged instructions - Privileged instructions execute in a privileged mode and will be trapped if executed outside this mode.
   2. Control sensitive instructions - Control-sensitive instructions attempt to change the configuration of resources used.
   3. Behavior-sensitive instructions - Behavior-sensitive instructions have different behaviors depending on the configuration of resources, including the load and store operations over the virtual memory.

A CPU architecture is virtualizable if it supports the ability to run the VM's privileged and privileged instructions in the CPU's user mode while the VMM runs in supervisor mode. When the privileged instructions including control- and behavior sensitive instructions of a VM are executed, they are trapped in the VMM. In this case, the VMM acts as a unified mediator for hardware access from different VMs to guarantee the correctness and stability of the whole system. RISC CPU architectures can be naturally virtualized because all control- and behavior- sensitive instructions are privileged instructions.

2 Memory Virtualization:

Virtual memory virtualization is similar to the virtual memory support provided by modern operating systems. In a traditional execution environment, the operating system maintains mappings of virtual memory to machine memory using page tables, which is a one- stage mapping from virtual memory to machine memory. All modern x86 CPUs include a memory management unit (MMU) and a translation lookaside buffer (TLB) to optimize virtual memory performance.



However, in a virtual execution environment, virtual memory virtualization involves sharing the physical system memory in RAM and dynamically allocating it to the physical memory of the VMs. That means a two- stage mapping process should be maintained by the guest OS and the VMM, respectively: virtual memory to physical memory and physical memory to machine memory. Furthermore, MMU virtualization should be supported, which is transparent to the guest OS. The guest OS continues to control the mapping of virtual addresses to the physical memory addresses of VMs. But the guest OS cannot directly access the actual machine memory. The VMM is responsible for mapping the guest physical memory to the actual machine memory. Figure 2.16 shows the two-level memory mapping procedure.

3. I/O Virtualization: I/O virtualization involves managing the routing of I/O requests between virtual devices and the hared physical hardware. There are three ways to implement I/O virtualization:

- Full device emulation
- Para virtualization
- Direct I/O

Full device emulation is the first approach for I/O virtualization. Generally, this approach emulates well known, real-world devices. All the functions of a device or bus infrastructure, such as device enumeration, identification, interrupts, and DMA, are replicated in software. This software is located in the VMM and acts as a virtual device. The I/O access requests of the guest OS are trapped in the VMM which interacts with the I/O devices.

A single hardware device can be shared by multiple VMs that run concurrently. However, software emulation runs much slower than the hardware it emulates. The para virtualization method of I/O virtualization is typically used in Xen. It is also known as the split driver model consisting of a frontend driver and a backend driver. The frontend driver is running in Domain U and the backend driver is running in Domain 0. They interact with each other via a block of shared memory. The frontend driver manages the I/O requests of the guest OSes and the backend driver is responsible for managing the real I/O devices and multiplexing the I/O data of different VMs. Although para I/O-virtualization achieves better device performance than full device emulation, it comes with a higher CPU overhead.

9 (b)	What are tools and mechanisms of cloud virtualization?	6M	20AI003.2	L2

Virtualization Tools and Mechanisms:

There are three typical classes of VM architecture. Before virtualization, the operating system manages the hardware. After virtualization, a virtualization layer is inserted between the hardware and the operating system. In such a case, the virtualization layer is responsible for converting portions of the real hardware into virtual hardware. Therefore, different operating systems such as Linux and Windows can run on the same physical machine, simultaneously.

Depending on the position of the virtualization layer, there are several classes of VM architectures, namely the hypervisor architecture, para-virtualization, and host based virtualization. The hypervisor is also known as the VMM (Virtual Machine Monitor). They both perform the same virtualization operations.

1. Hypervisor and Xen Architecture:

The hypervisor supports hardware-level virtualization on bare metal devices like CPU, memory, disk and network interfaces. The hypervisor software sits directly between the physical hardware and its OS. This virtualization layer is referred to as either the VMM or the hypervisor. The hypervisor provides hyper calls for the guest OS's and applications. Depending on the functionality, a hypervisor can assume micro-kernel architecture like the Microsoft Hyper-V. Or it can assume monolithic hypervisor architecture like the VMware ESX for server virtualization.

A micro-kernel hypervisor includes only the basic and unchanging functions (such as physical memory management and processor scheduling). The device drivers and other changeable components are outside the hypervisor. A monolithic hypervisor implements all the aforementioned functions, including those of the device drivers.

The Xen Architecture:

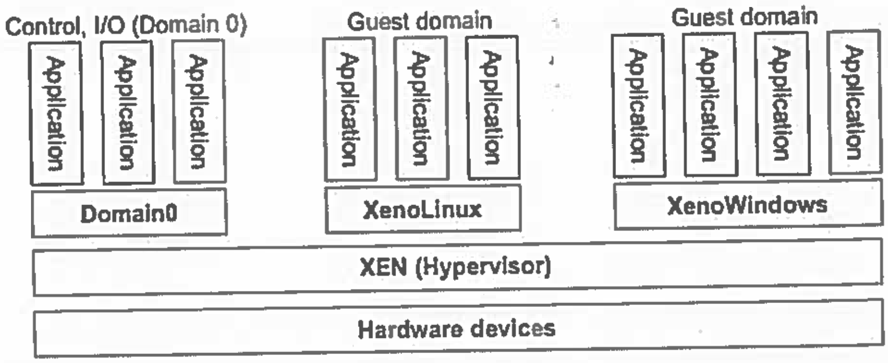


FIGURE 3.5

The Xen architecture's special domain 0 for control and I/O, and several guest domains for user applications.

The core components of a Xen system are the hypervisor, kernel, and applications. The organization of the three components is important. Like other virtualization systems, many guest OSes can run on top of the hypervisor. However, not all guest OSes are created equal, and one in particular controls the others. The guest OS, which has control ability, is called Domain 0, and the others are called Domain U. Domain 0 is a privileged guest OS of Xen. It is first loaded when Xen boots without any file system drivers being available. Domain 0 is designed to access hardware directly and manage devices. Therefore, one of the responsibilities of Domain 0 is to allocate and map hardware resources for the guest domains (the Domain U domains).

2 Binary Translation with Full Virtualization:

Depending on implementation technologies, hardware virtualization can be classified into two categories: full virtualization and host-based virtualization. Full virtualization does not need to modify the host OS. It relies on binary translation to trap and to virtualize the execution of certain sensitive, non virtualizable instructions. The guest OSes and their applications consist of noncritical and critical instructions. In a host-based system, both a host OS and a guest OS are used. A virtualization software layer is built between the host OS and guest OS.

2.1 Full Virtualization: With full virtualization, noncritical instructions run on the hardware directly while critical instructions are discovered and replaced with traps into the VMM to be emulated by software. Both the hypervisor and VMM approaches are considered full virtualization.

2.2 Binary Translation of Guest OS Requests Using a VMM :

VMware puts the VMM at Ring 0 and the guest OS at Ring 1. The VMM scans the instruction stream and identifies the privileged, control- and behavior-sensitive instructions. When these instructions are identified, they are trapped into the VMM, which emulates the behavior of these instructions.

The method used in this emulation is called binary translation. Therefore, full virtualization combines binary translation and direct execution. The guest OS is completely decoupled from the underlying hardware. Consequently, the guest OS is unaware that it is being virtualized. Binary translation employs a code cache to store translated hot instructions to improve performance, but it increases the cost of memory usage.

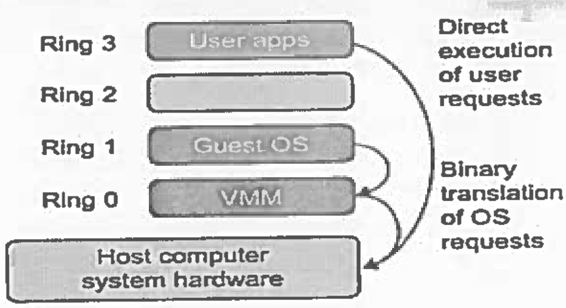


FIGURE 3.6

Indirect execution of complex instructions via binary translation of guest OS requests using the VMM plus direct execution of simple instructions on the same host.

2.3 Host-Based Virtualization:

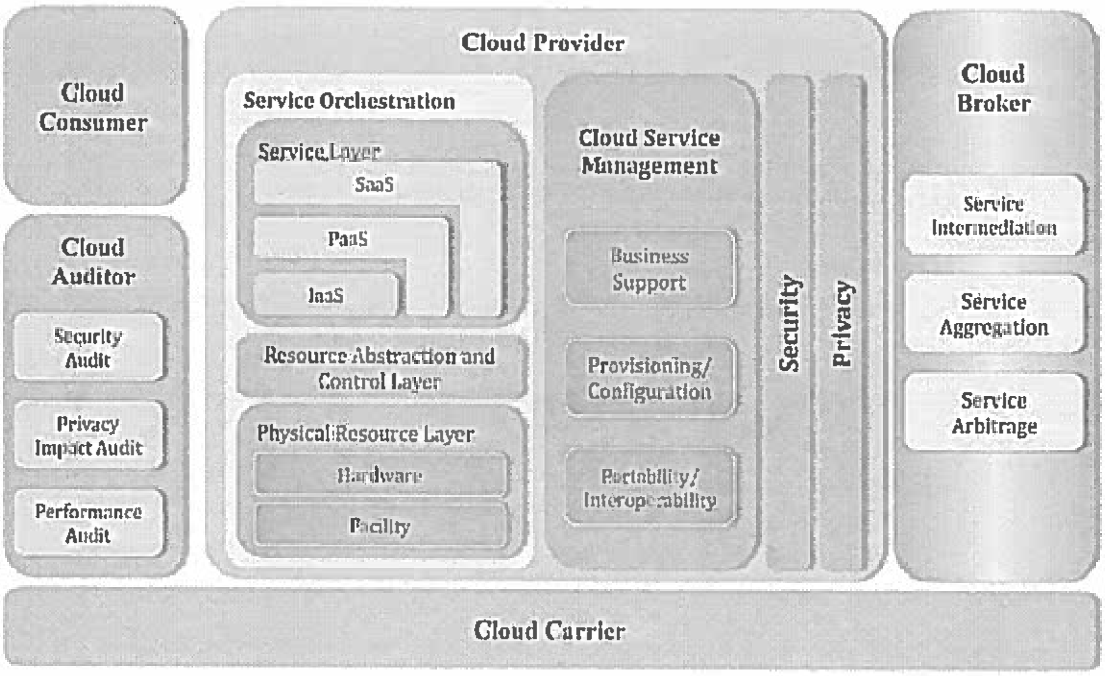
An alternative VM architecture is to install a virtualization layer on top of the host OS. This host OS is still

Cloud Carrier	An intermediary that provides connectivity and transport of cloud services from *Cloud Providers* to *Cloud Consumers*.

responsible for managing the hardware. The guest OSes are installed and run on top of the virtualization layer. Dedicated applications may run on the VMs. Certainly, some other applications can also run with the host OS directly. This host-based architecture has some distinct advantages, as enumerated next. First, the user can install this VM architecture without modifying the host OS. The vitalizing software can rely on the host OS to provide device drivers and other low level services. This will simplify the VM design and ease its deployment. Second, the host-based approach appeals to many host machine configurations. Compared to the hypervisor/VMM architecture, the performance of the host based architecture may also be low. When an application requests hardware access, it involves four layers of mapping which downgrades performance significantly.

10 (a)	Explain NIST cloud computing Reference Architecture.	7M	20AI003.3	L2

NIST cloud computing Reference Architecture:



Actor	Definition
Cloud Consumer	A person or organization that maintains a business relationship with, and uses service from, *Cloud Providers*.
Cloud Provider	A person, organization, or entity responsible for making a service available to interested parties.
Cloud Auditor	A party that can conduct independent assessment of cloud services, information system operations, performance and security of the cloud implementation.
Cloud Broker	An entity that manages the use, performance and delivery of cloud services, and negotiates relationships between *Cloud Providers* and *Cloud Consumers*.
Cloud Carrier	An intermediary that provides connectivity and transport of cloud services from *Cloud Providers* to *Cloud Consumers*.

- The cloud consumer is the principal stakeholder for the cloud computing service.
- A cloud consumer represents a person or organization that maintains a business relationship with, and uses the service from a cloud provider. The cloud consumer may be billed for the service provisioned, and needs to arrange payments accordingly.
- The consumers of SaaS can be organizations that provide their members with access to software applications, end users or software application administrators.
- SaaS consumers can be billed based on the number of end users, the time of use, the network bandwidth consumed, and the amount of data stored or duration of stored data.
- Cloud consumers of PaaS can employ the tools and execution resources provided by cloud providers to develop, test, deploy and manage the applications.
- PaaS consumers can be application developers or application testers who run and test applications in cloud-based environments,.
- PaaS consumers can be billed according to, processing, database storage and network resources consumed.
- Consumers of IaaS have access to virtual computers, network-accessible storage & network infrastructure components.
- The consumers of IaaS can be system developers, system administrators and IT managers.
- IaaS consumers are billed according to the amount or duration of the resources consumed, such as CPU hours used by virtual computers, volume and duration of data stored.

10 (b)	List the cloud deployment models and explain any three models.	5M	20AI003.3	L2

Cloud Deployment Model

1. Public Cloud
2. Private Cloud
3. Hybrid Cloud
4. Community Cloud

1. Public Cloud:

- A public cloud is one in which the cloud infrastructure and computing resources aremade available to the general public over a public network.
- A public cloud is meant to serve a multitude(huge number) of users, not a singlecustomer.
- A fundamental characteristic of public clouds is multitenancy.
- Multitenancy allows multiple users to work in a software environment at the sametime, each with their own resources.
- Built over the Internet (i.e., service provider offers resources, applications storage tothe customers over the internet) and can be accessed by any user.
- Owned by service providers and are accessible through a subscription.
- Best Option for small enterprises, which are able to start their businesses withoutlarge up-front(initial) investment.
- By renting the services, customers were able to dynamically upsize or downsize theirIT according to the demands of their business.
- Services are offered on a price-per-use basis.
- Promotes standardization, preserve capital investment
- Public clouds have geographically dispersed datacenters to share the load of users andbetter serve them

according to their locations

Provider is in control of the infrastructure Examples:

- o Amazon EC2 is a public cloud that provides Infrastructure as a Service
- o Google AppEngine is a public cloud that provides Platform as a Service
- o SalesForce.com is a public cloud that provides software as a service.

Advantage:

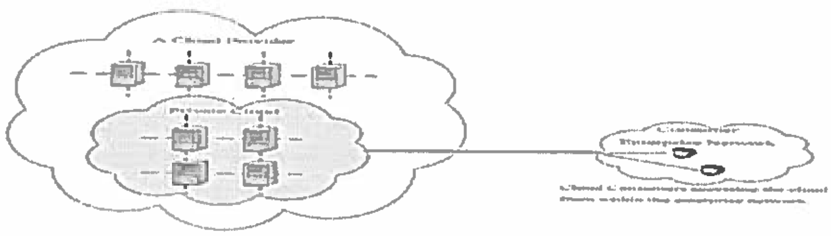
- Offers unlimited scalability – on demand resources are available to meet your business needs.
- Lower costs—no need to purchase hardware or software and you pay only for the service you use.
- No maintenance - Service provider provides the maintenance.
- Offers reliability: Vast number of resources are available so failure of a system will not interrupt service.
- Services like SaaS, PaaS, IaaS are easily available on Public Cloud platform as it canbe accessed from anywhere through any Internet enabled devices.
- Location independent – the services can be accessed from any location

Disadvantage:

- No control over privacy or security
- Cannot be used for use of sensitive applications
- Lacks complete flexibility(since dependent on provider)
- No stringent (strict) protocols regarding data management

2. Private Cloud

- Cloud services are used by a single organization, which are not exposed to the public
- Services are always maintained on a private network and the hardware and software are dedicated only to single organization
- Private cloud is physically located at
- Organization's premises [On-site private clouds] (or)
- Outsourced(Given) to a third party[Outsource private Clouds]
- It may be managed either by
  - Cloud Consumer organization (or)
  - By a third party
  - Private clouds are used by
  - government agencies
  - financial institutions
  - Mid size to large-size organisations.
  - On-site private clouds



Advantage

- Offers greater Security and Privacy
- Organization has control over resources
- Highly reliable

- Saves money by virtualizing the resources

Disadvantage

- Expensive when compared to public cloud
- Requires IT Expertise to maintain resources.

3. Hybrid Cloud

- Built with both public and private clouds
- It is a heterogeneous cloud resulting from private and public clouds.
- Private cloud are used for
  - sensitive applications are kept inside the organization's network
  - business-critical operations like financial reporting
- Public Cloud are used when
  - Other services are kept outside the organization's network
  - high-volume of data
  - Lower-security needs such as web-based email (gmail, yahoo mail etc)
- The resources or services are temporarily leased for the time required and then released. This practice is also known as cloud bursting.

Advantage

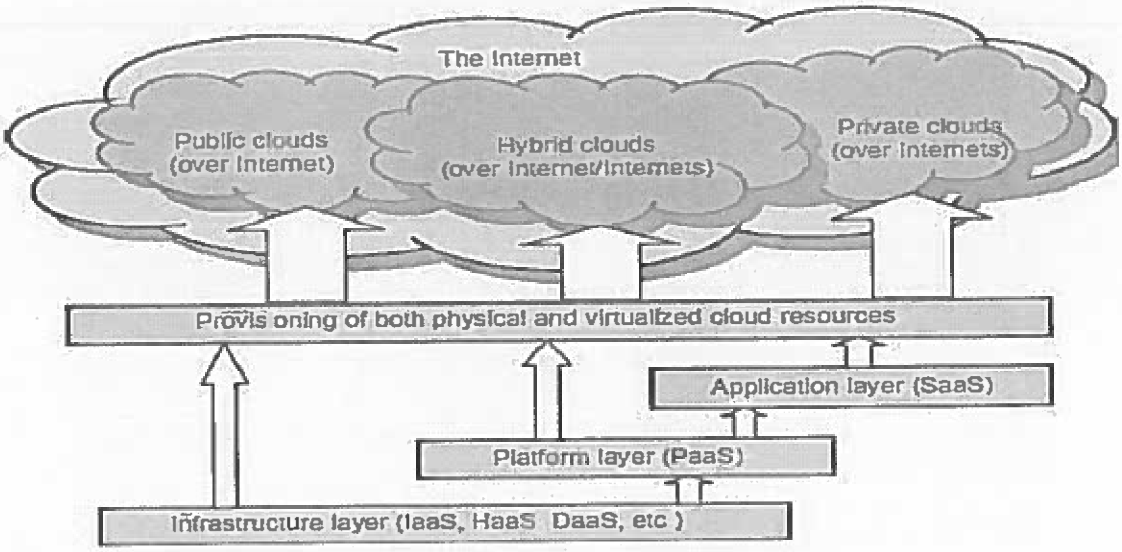
- It is scalable
- Offers better security
- Flexible-Additional resources are availed in public cloud when needed
- Cost-effectiveness—we have to pay for extra resources only when needed.
- Control - Organization can maintain a private infrastructure for sensitive application

Disadvantage

- Infrastructure Dependency
- Possibility of security breach(violate) through public cloud

OR

11 (a)	Explain about Layers of cloud computing.	7M	20AI003.3	L2



Layered Cloud Architectural Development

- The architecture of a cloud is developed at three layers
  - Infrastructure

- Platform
- Application

- Implemented with virtualization and standardization of hardware and software resources provisioned in the cloud.

The services to public, private and hybrid clouds are conveyed to users through networking support

Infrastructure Layer

- Foundation for building the platform layer.
- Built with virtualized compute, storage, and network resources.
- Provide the flexibility demanded by users.
- Virtualization realizes automated provisioning of resources and optimizes the infrastructure management process.

Platform Layer

- Foundation for implementing the application layer for SaaS applications.
- Used for general-purpose and repeated usage of the collection of software resources.
- Provides users with an environment to develop their applications, to test operation flows, and to monitor execution results and performance.

The platform should be able to assure users that they have scalability, dependability, and security protection.

Application Layer

- Collection of all needed software modules for SaaS applications.
- Service applications in this layer include daily office management work, such as information retrieval, document processing, and authentication services.
- The application layer is also heavily used by enterprises in business marketing and sales, consumer relationship management (CRM) and financial transactions.
- Not all cloud services are restricted to a single layer.
- Many applications may apply resources at mixed layers.
- Three layers are built from the bottom up with a dependence relationship.

11 (b)	Explain about storage as a service.	5M	20AI003.3	L2

Storage as a Service or STaaS is cloud storage that you rent from a Cloud Service Provider (CSP) and that provides basic ways to access that storage. Enterprises, small and medium businesses, home offices, and individuals can use the cloud for multimedia storage, data repositories, data backup and recovery, and disaster recovery. There are also higher-tier managed services that build on top of STaaS, such as Database as a Service, in which you can write data into tables that are hosted through CSP resources.

The key benefit to STaaS is that you are offloading the cost and effort to manage data storage infrastructure and technology to a third-party CSP. This makes it much more effective to scale up storage resources without investing in new hardware or taking on configuration costs. You can also respond to changing market conditions faster. With just a few clicks you can rent terabytes or more of storage, and you don't have to spin up new storage appliances on your own.

Another factor that influences cost is the type of data storage used. There are three main types of cloud storage: block storage, file storage, and object-based storage.

- **Block storage** breaks data into segmented pieces and distributes them to the storage environment wherever it is most efficient for the platform to do so. This simulates the same functionality as writing data to a standard hard disk drive or solid-state drive. Data remains available for quick access, but it is also costly to maintain and works best for warm or hot data storage.
- **File storage** lists data in a navigable hierarchy, usually a file directory. This is most like the file storage system that you would find on a PC or in cloud storage apps like Microsoft OneDrive. Because it is designed for humans to navigate, file storage is ideal anytime you need to collaborate on a project with other people or businesses. Whether the data is hot or cold doesn't matter as much. However, file storage does not scale well. The more files you add, the more complex the system becomes and the more difficult it is to navigate.
- **Object-based storage** organizes data by adding meta information to it, making it easy to recognize and retrieve at any time. This type of cloud storage scales up in the most cost-efficient manner, because you can keep adding to it. It is typically the least expensive type of STaaS and best suited for massive amounts of cold media or data files.

Data Center Storage Security

The nature of these risks is also different for STaaS providers than for businesses or individual users and extends from basic platform hardening and antimalware measures to things like multitenancy. Customers may also introduce vulnerabilities that a CSP has very little control over.

1. **Multitenancy** : In a cloud environment, compute and storage resources are abstracted from the hardware layer and made available in virtual pools, either through virtual machines (VMs) or containers. Multiple VMs and containers can run on the same physical server. Your data and applications are oftentimes sharing the same bare metal resources as the data and applications of other customers. This is called multitenancy, as there are multiple tenants or customers sharing the same physical resources. Vulnerabilities in another tenant's workloads can expose your workloads to risks.
2. **Customer Vulnerabilities:** For STaaS customers, the main security concern will be managing who within their organization has access to the data and what level of permissions they have, such as read-only vs. read and write–level access. CSPs don't control who can access their customers' devices, so it's important to be vigilant against threats such as email phishing schemes that can create vulnerability in your point of access. Using strong passwords, two-factor authentication, and following other best practices can add more layers of protection.
3. **Provider Practices:** Cloud security in STaaS is primarily the concern of the CSP who manages the cloud storage environment. It's up to the CSP to treat vulnerabilities in both the hardware and software layers and address the human element by ensuring that all personnel in charge of maintaining the cloud infrastructure are trustworthy and follow best practices. The customer in this case should become educated and learn how to ask pointed questions about security when deciding which CSP to choose.

12 (a)	What is resource provisioning? Explain resource provisioning methods.	5M	20AI003.4	L2

. Resource Provisioning (Providing) and Platform Deployment There are techniques to provision computer resources or VMs. Parallelism is exploited at the cluster node level.

1. Provisioning of Compute Resources (VMs)
   a. Providers supply cloud services by signing SLAs with end users.
   b. The SLAs must specify resources such as
      i. CPU
      ii. Memory
      iii. Bandwidth

   Users can use these for a preset (fixed) period.
   c. Under provisioning of resources will lead to broken SLAs and penalties.
   d. Over provisioning of resources will lead to resource underutilization, and consequently, a decrease in revenue for the provider.
   e. Provisioning of resources to users is a challenging problem. The difficulty comes from the following
      - Unpredictability of consumer demand

- Software and hardware failures
- Heterogeneity of services
- Power management
- Conflict in signed SLAs between consumers and service providers.

2. Resource Provisioning Methods

There are 3 Resource-provisioning methods are

A. Demand-driven method - Provides static resources and has been used in grid computing
B. Event-driven method - Based on predicted workload by time.
C. Popularity-Driven Resource Provisioning – Based on Internet traffic monitored

A.Demand Driven Methods:

- Provides Static resources
- This method adds or removes nodes (VM) based on the current utilization(Use) level of the allocated resources.
- When a resource has surpassed (exceeded) a threshold (Upperlimit) for a certain amount of time, the scheme increases the resource (nodes) based on demand.
- When a resource is below a threshold for a certain amount of time, then resources could be decreased accordingly.
- This method is easy to implement.
- The scheme does not work out properly if the workload changes abruptly.

B.Event-Driven Resource Provisioning

- This scheme adds or removes machine instances based on a specific time event.
- The scheme works better for seasonal or predicted events such as Christmastime in the West and the Lunar New Year in the East.
- During these events, the number of users grows before the event period and then decreases during the event period. This scheme anticipates peak traffic before it happens.

The method results in a minimal loss of QoS, if the event is predicted correctly

C.Popularity-Driven Resource Provisioning

- Internet searches for popularity of certain applications and allocates resources by popularity demand.
- This scheme has a minimal loss of QoS, if the predicted popularity is correct.
- Resources may be wasted if traffic does not occur as expected.
- Again, the scheme has a minimal loss of QoS, if the predicted popularity is correct.
- Resources may be wasted if traffic does not occur as expected.

12 (b)	Write short on security governance.	7M	20AI003.4	L2

Security governance

Cloud Security Governance: It is a set of rules you create to monitor and amend as necessary in order to control costs, improve efficiency, and mitigating the security risks.

Cloud Security Governance Challenges: Whether developing a governance model from the start or having to retrofit one on existing investments in cloud, these are some of the common challenges:

1. Lack of senior management participation and buy-in: The lack of a senior management influenced and endorsed security policy is one of the common challenges facing cloud customers. An enterprise security policy is intended to set the executive tone, principles and expectations for security management and operations in the cloud. However, many enterprises tend to author security policies that are often laden with tactical content, and lack executive input or influence. The result of this situation is the ineffective definition and communication of executive tone and expectations for security in the cloud.

2. Lack of embedded management operational controls: Another common cloud security governance challenge is lack of embedded management controls into cloud security operational processes and procedures. Controls are often interpreted as an auditor's checklist or repackaged as procedures, and as a result, are not effectively embedded into security operational processes and procedures as they should be, for purposes of optimizing value and reducing day-to-day operational risks. This lack of embedded controls may result in operational risks that may not be apparent to the enterprise. For example, the security configuration of a device may be modified (change event) by a staffer without proper analysis of the business impact (control) of the modification. The net result could be the introduction of exploitable security weaknesses that may not have been apparent with this modification.

3. Lack of operating model, roles, and responsibilities: Many enterprises moving into the cloud environment tend to lack a formal operating model for security, or do not have strategic and tactical roles and responsibilities properly defined and operationalized. This situation stifles the effectiveness of a security management and operational function/organization to support security in the cloud. Simply, establishing a hierarchy that includes designating an accountable official at the top, supported by a stakeholder committee, management team, operational staff, and third-party provider support (in that order) can help an enterprise to better manage and control security in the cloud, and protect associated investments in accordance with enterprise business goals.

4. Lack of metrics for measuring performance and risk: Another major challenge for cloud customers is the lack of defined metrics to measure security performance and risks – a problem that also stifles executive visibility into the real security risks in the cloud. This challenge is directly attributable to the combination of other challenges discussed above. For example, a metric that quantitatively measures the number of exploitable security vulnerabilities on host devices in the cloud over time can be leveraged as an indicator of risk in the host device environment. Similarly, a metric that measures the number of user-reported security incidents over a given period can be leveraged as a performance indicatorof staff awareness and training efforts. Metrics enable executive visibility into the extent to which security tone and expectations (per established policy) are being met within the enterprise and support prompt decision-making in reducing risks or rewarding performance as appropriate. The challenges described above clearly highlight the need for cloud customers to establish a framework to effectively manage and support security in cloud management, so that the pursuit of business targets are not potentially compromised. Unless tone and expectations for cloud security are established (via an enterprise policy) to drive operational processes and procedures with embedded management controls, it is very difficult to determine or evaluate business value, performance, resource effectiveness, and risks regarding security operations in the cloud. Cloud security governance facilitates the institution of a model that helps enterprises explicitly address the challenges described above.

OR

13 (a)	Write short note on cloud security challenges	6M	20AI003.4	L2

Cloud security challenges : Security Issues in Cloud Computing : There is no doubt that Cloud Computing provides various Advantages but there are also some security issues in cloud computing. Below are some following Security Issues in Cloud Computing as follows.

1. Data Leakage – Data Loss is one of the issues faced in Cloud Computing. This is also known as Data Leakage. As we know that our sensitive data is in the hands of Somebody else, and we don't have full control over our database. So if the security of cloud service is to break by hackers then it may be possible that hackers will get access to our sensitive data or personal file.
2. Interference of Hackers and Insecure API's – As we know if we are talking about the cloud and its services it

means we are talking about the Internet. Also, we know that the easiest way to communicate with Cloud is using API. So it is important to protect the Interface's and API's which are used by an external user. But also in cloud computing, few services are available in the public domain. An is the vulnerable part of Cloud Computing because it may be possible that these services are accessed by some third parties. So it may be possible that with the help of these services hackers can easily hack or harm our data.

3. User Account Hijacking – Account Hijacking is the most serious security issue in Cloud Computing. If some how the Account of User or an Organization is hijacked by Hacker. Then the hacker has full authority to perform Unauthorized Activities.
4. Changing Service Provider – Vendor lock In is also an important Security issue in Cloud Computing. Many organizations will face different problems while shifting from one vendor to another. For example, An Organization wants to shift from AWS Cloud to Google Cloud Services then they ace various problem's like shifting of all data, also both cloud services have different techniques and functions, so they also face problems regarding that. Also, it may be possible that the charges of AWS are different from Google Cloud, etc.
5. Lack of Skill – While working, shifting to another service provider, need an extra skill , how to use a feature, API etc. are the main problems caused in IT Company who doesn't have skilled Employee. So it requires a skilled person to work with cloud Computing.
6. Denial of Service (DoS) attack – This type of attack occurs when the system receives too much traffic and when the service is not available to the end users. Mostly DoS attacks occur in large organizations such as the banking sector, government sector, etc. When a DoS attack occurs data is lost. So in order to recover data, it requires a great amount of money as well as time to handle it.

13 (b)	Explain about virtual machine security	6M	20AI003.4	L2

Virtual machine security

Virtual security services are typically provided by the VMM. Another alternative is to have a dedicated security services VM. A secure trusted computing base (TCB) is a necessary condition for security in a virtual machine environment; if the TCB is compromised, the security of the entire system is affected.

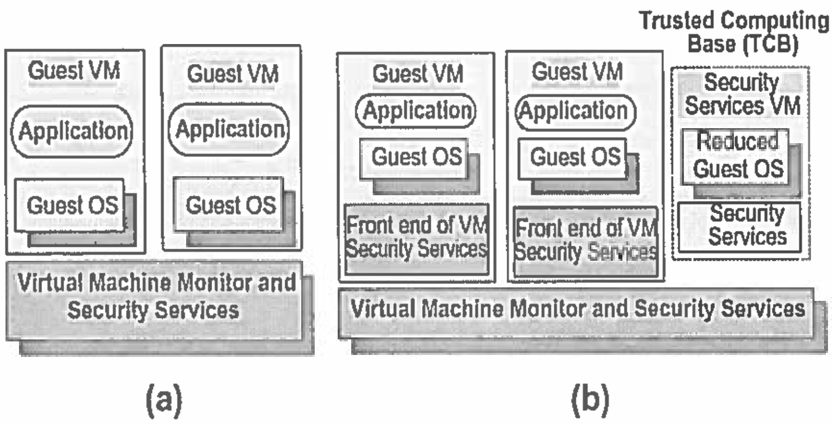


FIGURE 9.2

(a) Virtual security services provided by the VMM. (b) A dedicated security VM.

Following are deployed on virtual machines to ensure security

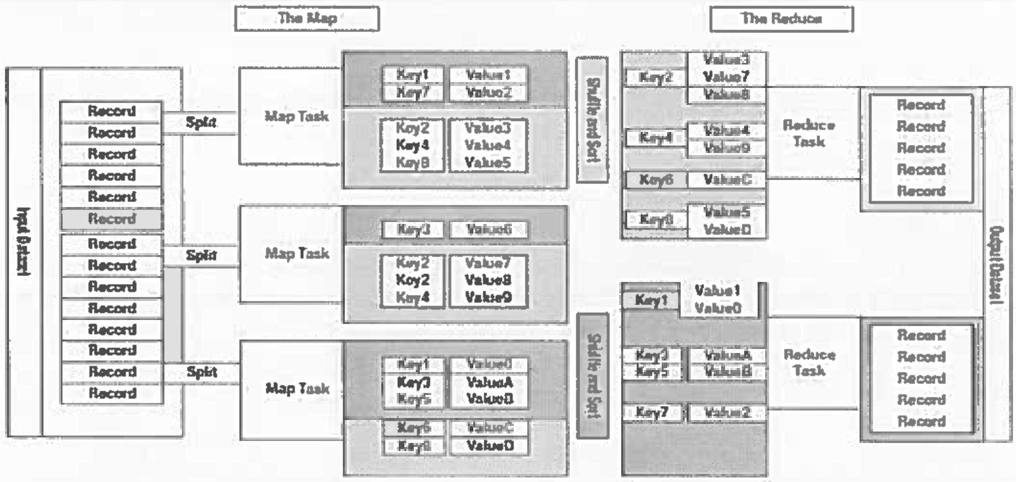
1. Firewalls: Isolate each virtual machine you have by installing a firewall. Only allow approved protocols to be deployed.
2. Intrusion detection and prevention: Ensure that antivirus programs are installed on the virtual machines and kept current with updates. Virtual machines, like physical machines are at risk for viruses and worms.
3. Integrity monitoring: Avoid sharing IP addresses. Again this is typical of sharing a resource and will attract problems and vulnerabilities.

4. Log inspection: Monitor the event log and security events on both the host machine and on the virtual machine. These logs need to be stored in your log vault for security and for auditing purposes at a later date.

14 (a)	Explain about Map Reduce.	6M	20AI003.5	L2

MAP REDUCE

- ❖ MapReduce is a programming model for data processing.
- ❖ MapReduce is designed to efficiently process large volumes of data by connecting many commodity computers together to work in parallel
- ❖ Hadoop can run MapReduce programs written in various languages like Java, Ruby, and Python
- ❖ MapReduce works by breaking the processing into two phases:
  - o The map phase and
  - o The reduce phase.

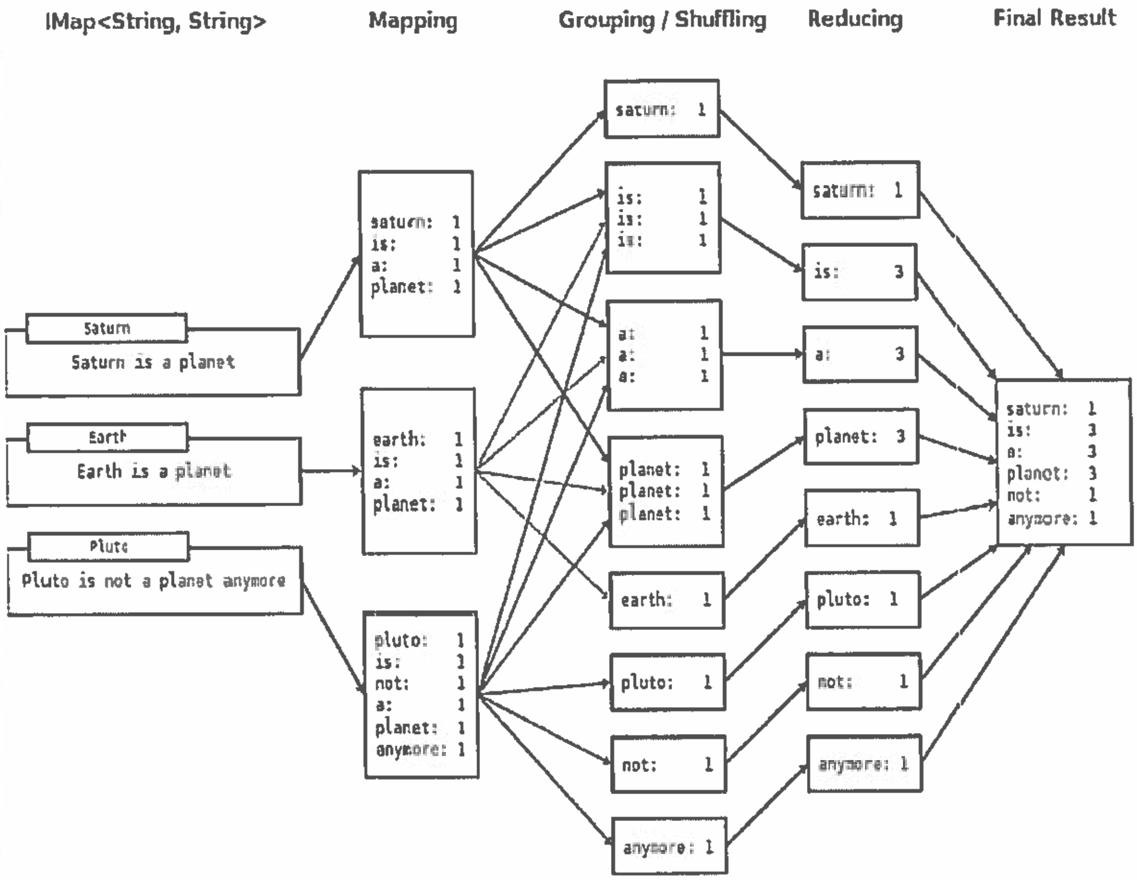


- ❖ Each phase has key-value pairs as input and output, the types of which may be chosen by the programmer.
- ❖ The programmer also specifies two functions:
  - o The map function and
  - o The reduce function.
- ❖ In MapReduce, chunks are processed in isolation by tasks called Mappers.
- ❖ The outputs from the mappers are denoted as intermediate outputs (IOs) and are brought into a second set of tasks called Reducers
- ❖ The process of bringing together IOs into a set of Reducers is known as shuffling process
- ❖ The Reducers produce the final outputs (FOs) Prepare the work list for the next application.

1. Input Splits: An input to a MapReduce in Big Data job is divided into fixed-size pieces called input splits Input split is a chunk of the input that is consumed by a single map
2. Mapping: This is the very first phase in the execution of map-reduce program. In this phase data in each split is passed to a mapping function to produce output values. In our example, a job of mapping phase is to count a number of occurrences of each word from input splits (more details about input-split is given below) and prepare a list in the form of <word, frequency>
3. Shuffling: This phase consumes the output of Mapping phase. Its task is to consolidate the relevant

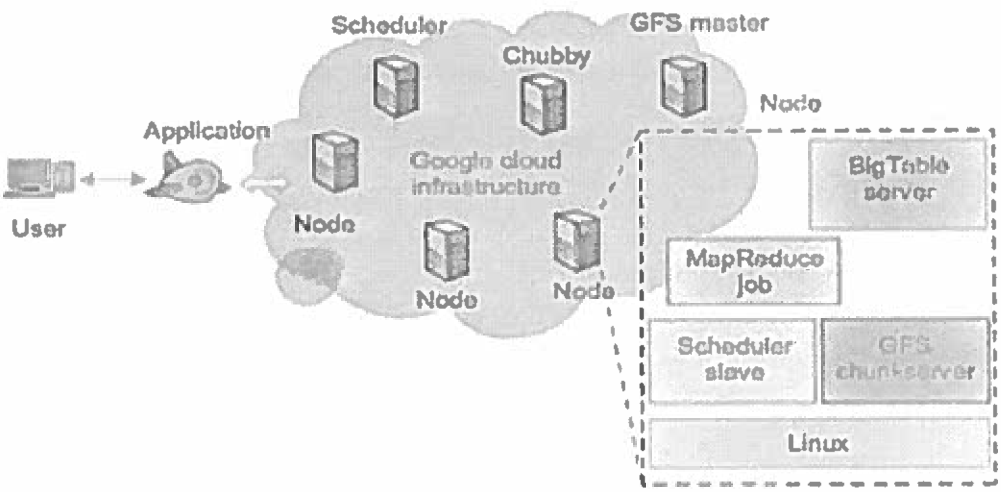
records from Mapping phase output. In our example, the same words are clubed together along with their respective frequency.

4. Reducing: In this phase, output values from the Shuffling phase are aggregated. This phase combines values from Shuffling phase and returns a single output value. In short, this phase summarizes the complete dataset.



14 (b)	Explain the architecture of Google App Engine.	6M	20AI003.5	L2

GAE Architecture:



TECHNOLOGIES USED BY GOOGLE ARE

- Google File System(GFS) ->for storing large amounts of data.
- MapReduce->for application program development.
- Chubby-> for distributed application lock services.
- BigTable-> offers a storage service.
- Third-party application providers can use GAE to build cloud applications for providing services.
- Inside each data center, there are thousands of servers forming different clusters.
- GAE runs the user program on Google's infrastructure.
- Application developers now do not need to worry about the maintenance of servers.
- GAE can be thought of as the combination of several software components.
- GAE supports Python and Java programming environments.

FUNCTIONAL MODULES OF GAE

- The GAE platform comprises the following five major components.
- Data Store: offers data storage services based on Big Table techniques.
- The Google App Engine (GAE) provides a powerful distributed data storage service.
- This provides a secure data Storage.

OR

15 (a)	Explain Open Nebula and Eucalyptus.	6M	20AI003.5	L2

1. **OpenNebula** is a cloud computing platform for managing heterogeneous distributed data center infrastructures. The Open Nebula platform manages a data center's virtual infrastructure to build private, public and hybrid implementations of Infrastructure as a Service. The two primary uses of the OpenNebula platform are data center virtualization and cloud deployments based on the KVM hypervisor, LXD/LXC system containers, and AWS Firecracker microVMs. The platform is also capable of offering the cloud infrastructure necessary to operate a cloud on top of existing VMware infrastructure. In early June 2020, OpenNebula announced the release of a new Enterprise Edition for corporate users, along with a Community Edition. OpenNebula CE is free and open-source software, released under the Apache License version-2. OpenNebula CE comes with free access to maintenance releases but with upgrades to new minor/major versions only available for users with non-commercial deployments or with significant contributions to the OpenNebula Community. OpenNebula EE is distributed under a closed-source license and requires a commercial Subscription.

2. **Eucalyptus** is a genus of over seven hundred species of flowering trees, shrubs or mallees in the myrtle family, Myrtaceae. Along with several other genera in the tribe Eucalypteae, including Corymbia, they are commonly known as eucalypts.[3] Plants in the genus Eucalyptus have bark that is either smooth, fibrous, hard or stringy, leaves with oil glands, and sepals and petals that are fused to form a "cap" or operculum over the stamens. The fruit is a woody capsule commonly referred to as a "gumnut".

   Most species of Eucalyptus are native to Australia, and every state and territory has representative species. About three-quarters of Australian forests are eucalypt forests. Wildfire is a feature of the Australian landscape and many eucalypt species are adapted to fire, and resprout after fire or have seeds which survive fire.

   A few species are native to islands north of Australia and a smaller number are only found outside the continent. Eucalypts have been grown in plantations in many other countries because they are fast growing and have valuable timber, or can be used for pulpwood, for honey production or essential oils. In some countries, however, they have been removed because of the danger of forest fires due to their high inflammability.

15 (b)	Describe about four levels of federation.	6M	20AI003.5	L2

Four Levels of Federation:

1. Permissive Federation:

   Permissive federation occurs when a server accepts a connection from a peer network server without verifying its identity using DNS lookups or certificate checking. The lack of verification or authentication may lead to domain spoofing. The unauthorized use of a third party domain name in an email message in order to (pretend to be someone else), which opens the door to widespread spam and other abuses.

2. Verified Federation:

   This type of federation occurs when a server accepts a connection from a peer after the identity of the peer has been verified. It uses information obtained via DNS and by means of domain-specific keys exchanged beforehand. The connection is not encrypted, and the use of identity verification effectively prevents domain spoofing. Federation requires proper DNS setup, and that is still subject to DNS poisoning attacks. Verified federation has been the default service policy on the open XMPP since the release of the open-source jabberd 1.2 server. XMPP-real time communication protocol uses XML. Prevent Address spoofing

3. Encrypted federation:

   Server accepts a connection from a peer if and only if the peer supports Transport Layer security (TLS) as defined. The peer must present a digital certificate. The certificate may be self-signed, but this prevents using mutual authentication. Server Dial back uses the DNS as the basis for verifying identity. The basic approach is that a receiving server receives a server-to- server connection request from an originating server. It does not accept the request until it has verified a key with an authoritative server for the domain asserted by the originating server. Server Dial back does not provide strong authentication or trusted federation Although it is subject to DNS attacks; it has effectively prevented most instances of address spoofing on the network.

4. Trusted federation:

   A server accepts a connection from a peer only under the stipulation that the peer supports TLS and the peer can present a digital certificate issued by a root certification authority (CA) that is trusted by the authenticating server. The list of trusted root CAs may be determined by one or more factors, such as the operating system, XMPP server software, or local service policy. The use of digital certificates results not only in a channel encryption but also in strong authentication. The use of trusted domain certificates effectively prevents DNS poisoning attacks. But makes federation more difficult, since such certificates have traditionally not been easy to obtain.

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	CSE (DS)	Academic Year	2022 - 2023
Course Code	20CS005	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Mobile Computing				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define GSM.	20CS005.1	L1
2	What is Medium Access Control?	20CS005.2	L1
3	What is the use of DHCP?	20CS005.3	L1
4	How does Client Server Computing work?	20CS005.4	L1
5	Define Protocol.	20CS005.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain mobile communication.	6M	20CS005.1	L2
6 (b)	Discuss the paradigms of Mobile Computing.	6M	20CS005.1	L2
	OR			
7 (a)	Explain GSM System Architecture.	6M	20CS005.1	L2
7 (b)	Explain Radio Interface.	6M	20CS005.1	L2
8 (a)	Explain Time Division Multiple Access.	6M	20CS005.2	L2
8 (b)	Explain Hidden and Exposed Terminals.	6M	20CS005.2	L2
	OR			
9 (a)	Compare CDMA with FDMA.	6M	20CS005.2	L2
9 (b)	Explain IEEE 802.11.	6M	20CS005.2	L2
10 (a)	Explain Mobile TCP.	6M	20CS005.3	L2
10 (b)	Explain Location Management.	6M	20CS005.3	L2
	OR			
11 (a)	Discuss Tunneling and Encapsulation.	6M	20CS005.3	L2
11 (b)	Explain Route Optimization.	6M	20CS005.3	L2
12 (a)	Explain indirect TCP/IP Protocols.	6M	20CS005.4	L2
12 (b)	Explain Snooping Transmission Control Protocol.	6M	20CS005.4	L2
	OR			
13 (a)	Explain Database Hoarding & Caching Techniques.	6M	20CS005.4	L2
13 (b)	Discuss the Quality-of-Service issues.	6M	20CS005.4	L2
14	Explain Push, Pull and Hybrid mechanisms.	12M	20CS005.5	L2
	OR			
15 (a)	Explain Selective Tuning.	6M	20CS005.5	L2
15 (b)	Discuss Data Synchronization Software.	6M	20CS005.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

Degree	B. Tech. (U. G.)	**Program**	CSD			**Academic Year**	2022 - 2023
Course Code	20CS005	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	V
Course	**Mobile Computing**						

1. Define GSM 2M

GSM (Global System for Mobile communication) is a digital mobile network that is widely used by mobile phone users in Europe and other parts of the world.

2. What is Medium Access Control? 2M

medium access control (MAC) is a sublayer of the data link layer of the open system interconnections (OSI) reference model for data transmission. It is responsible for flow control and multiplexing for transmission medium. It controls the transmission of data packets via remotely shared channels.

3. What is the use of DHCP? 2M

Dynamic Host Configuration Protocol (DHCP) is a client/server protocol that automatically provides an Internet Protocol (IP) host with its IP address and other related configuration information such as the subnet mask and default gateway

4. How does Client Server Computing Work? 2M

The client server computing works with a system of request and response. The client sends a request to the server and the server responds with the desired information. The client and server should follow a common communication protocol so they can easily interact with each other

5. Define Protocol 2M

The most common meaning of protocol is "**a system of rules that explain the correct conduct and procedures to be followed in formal situations,**" as in these example sentences: The soldier's actions constituted a breach of military protocol. They did not follow the proper diplomatic protocols.

6A. Explain mobile communication. 6M

Mobile communication is talking, texting or sending data or image files over a wireless network. An example of mobile communication is chatting on the cell phone with a friend. An example of mobile communication is sending email from a computer using a wireless network at your local coffee shop

Most common types of communication technology

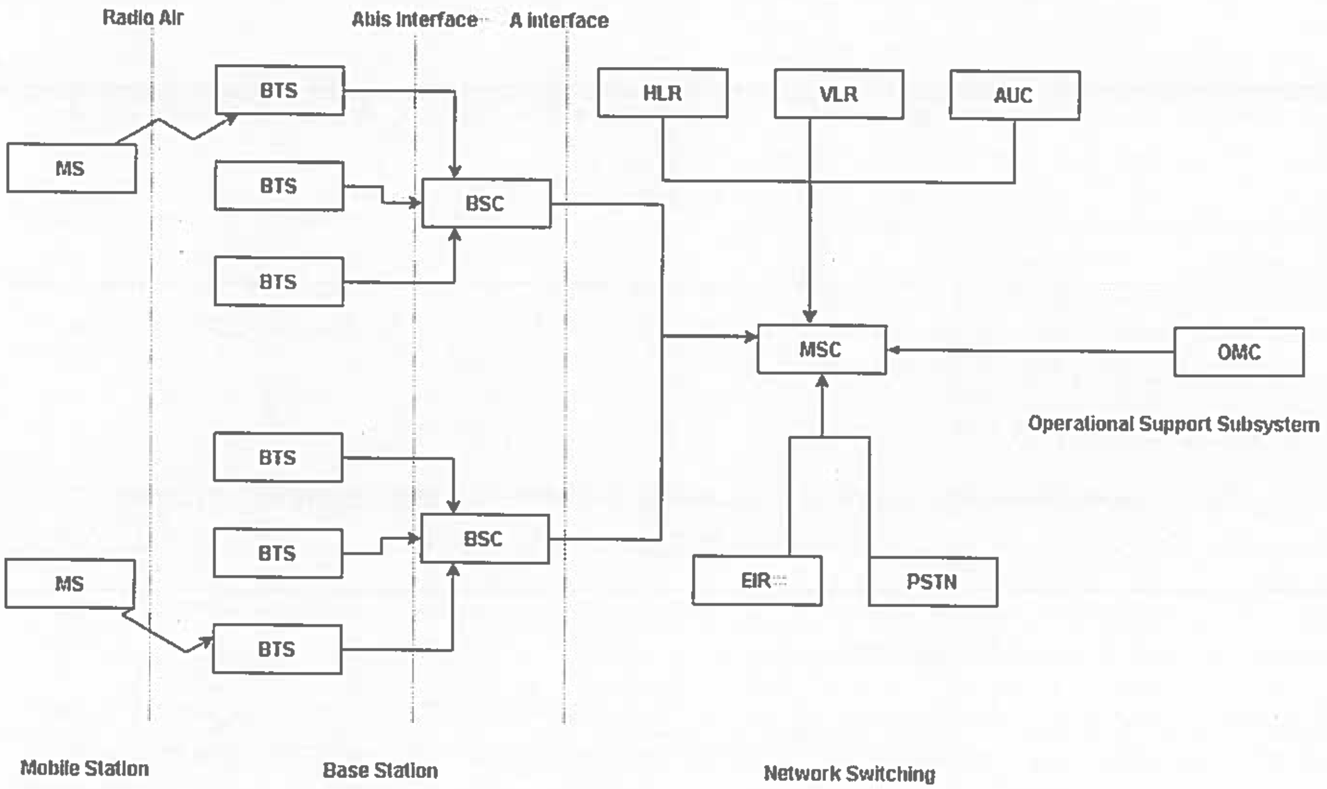
Global Systems for Mobile (GSM) Communications.
Code Division Multiple Access (CDMA)
Universal Mobile Telecommunication System (UMTS)
Long Term Evolution (LTE) using the Orthogonal Frequency Division Multiplexing (OFDM)method.
Adaptive communication.

6B. Discuss the paradigms of Mobile Computing 6M

The tasks they work on is of either high computing power and consist of large data sets. All communication between the computer systems in grid computing is done on the "data grid". The goal of grid computing is to solve more high computational problems in less time and improve productivity.

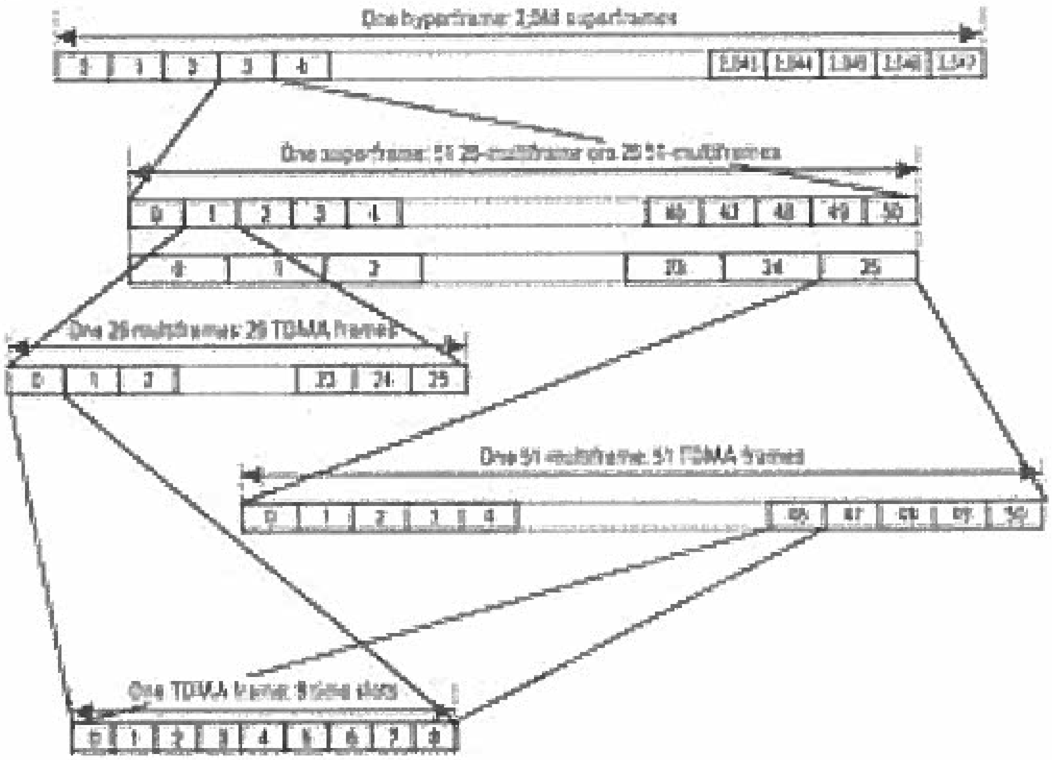
7A. Explain GSM System Architecture. 6M

GSM stands for Global System for Mobile Communication. GSM is an open and digital cellular technology used for mobile communication. **It uses 4 different frequency bands of 850 MHz, 900 MHz, 1800 MHz and 1900 MHz . It uses the combination of FDMA and TDMA.**



7B. Explain Radio Interface 6M

The radio interface is the wireless interface between two points. This term encompasses all the functionality required to maintain such interfaces. In cellular systems the interface between the terminal and the base station is referred to as the radio interface or air interface. Also called a "radio interface," the air interface **defines the frequency, channel bandwidth and modulation scheme.** For example, TDMA and CDMA modulation are used in GSM and CDMA cellular networks respectively, while OFDMA is used for LTE. OFDMA was also used for WiMAX. See air card, CDMA, TDMA

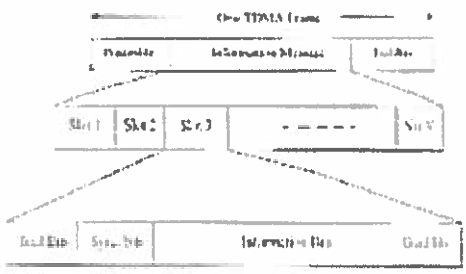


8A. Explain Time Division Multiple Access 6M

Digital modulation technique that allocates a discrete amount of frequency bandwidth to each user to permit many simultaneous conversations. Each caller is assigned a specific time slot for transmission.

Examples of TDMA include IS-136, personal digital cellular (PDC), integrated digital enhanced network (iDEN) and the second generation (2G) Global System for Mobile Communications (GSM). TDMA allows a mobile station's radio component to listen and broadcast only in its assigned time slot.

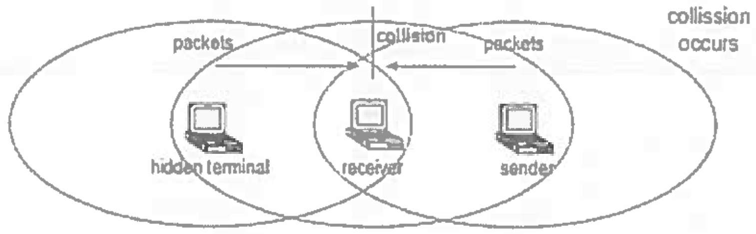
What is the efficiency of TDMA?



Efficiency of TDMA: The frame efficiency is the percentage of bits per frame which contain transmitted data. It is a measure of the percentage of transmitted data that contains information as opposed to providing overhead for the access scheme.

8B.

Hidden terminal problem – two nodes that are outside each-other's range perform simultaneous transmission to a node that is within the range of each of them, hence, there is a packet collision. Exposed terminal problem – the node is within the range of a node that is transmitting, and it cannot transmit to any node. 6M



9A. What is the difference between FDMA and CDMA?   6M

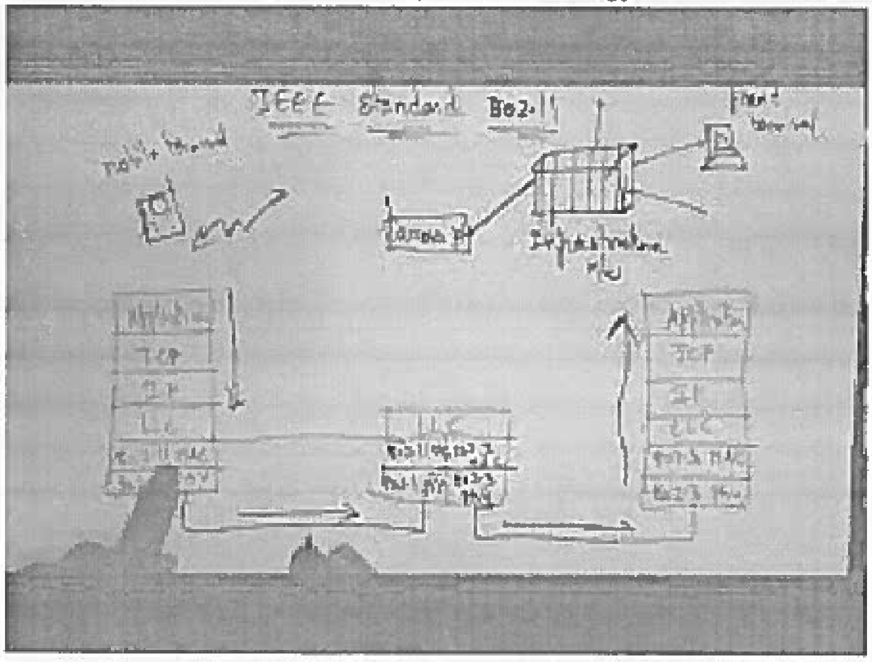
In FDMA the mode of data transfer is continuous signal. In CDMA the mode of data transfer is digital signal. The capacity of the system is low in FDMA. The capacity of the system is large in CDMA.

FDMA	CDMA
Mode of data transfer is continuous signal.	Mode of data transfer is digital signal.
It is little flexible.	It is highly flexible.

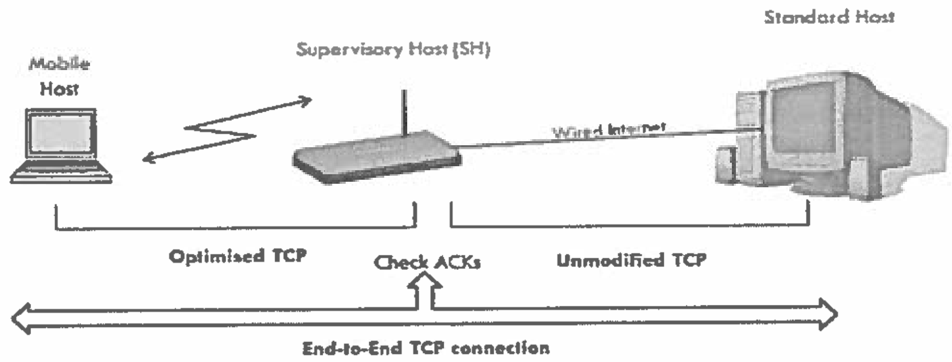
9B.   6M

The IEEE 802.11 architecture consists of several components that interact to provide a wireless LAN that supports station mobility transparently to upper layers. The basic service set (BSS) is the basic building block of an IEEE 802.11 LAN.

IEEE 802.11 refers to **the set of standards that define communication for wireless LANs (wireless local area networks, or WLANs)**. The technology behind 802.11 is branded to consumers as Wi-Fi.

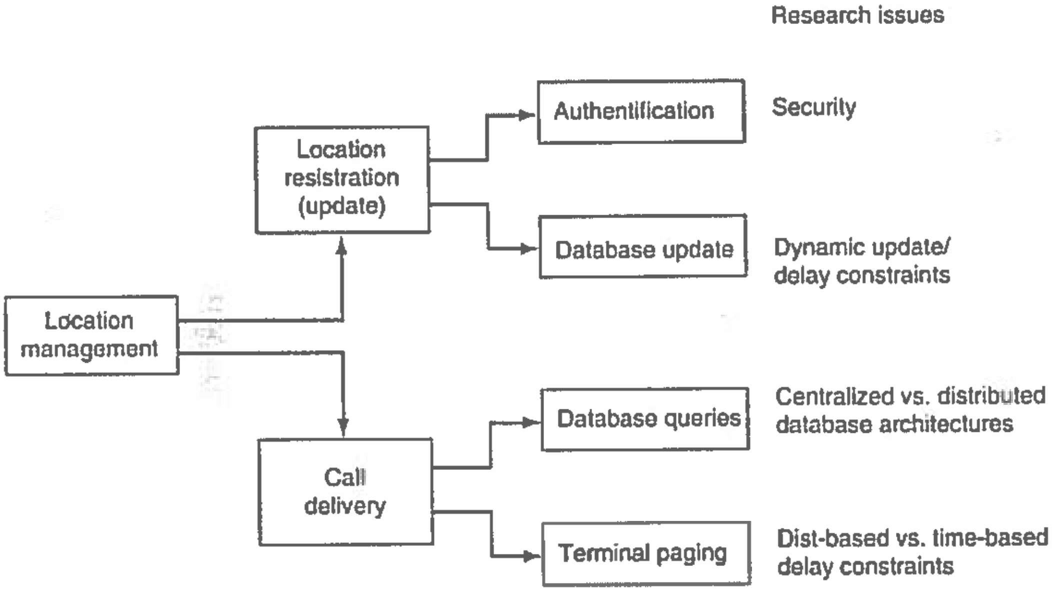


10A. M-TCP (mobile TCP) has the same goals as similar to its variants i.e. I-TCP and Snoop-TCP. It too wants to improve overall throughput, to lower the delay, to main end-to-end semantics of TCP. But, it is mainly enhanced to address problems related to lengthy or frequent disconnections.   6M



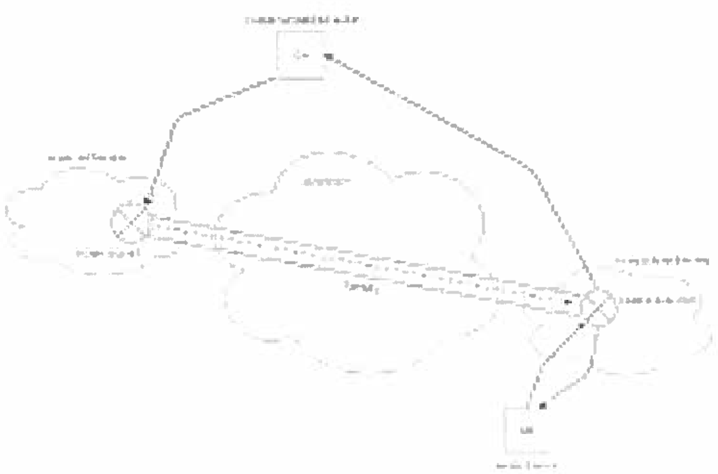
10B. Explain Location Management. 6M

Location management is the process of identifying the physical location of the user so that calls directed to that user can be routed to that location. Location management is also responsible for verifying the authenticity of users accessing the network.



11A. Tunneling and Encapsulation 6M

In computer networks, **a tunneling protocol is a communication protocol which allows for the movement of data from one network to another, by exploiting encapsulation.** It involves allowing private network communications to be sent across a public network (such as the Internet) through a process called encapsulation.



Tunneling establishes a virtual pipe for the packets available between a tunnel entry and an endpoint. It is the process of sending a packet via a tunnel and it is achieved by a mechanism called encapsulation.18-May-2022

11B. Route optimization is **the process of determining the most cost-efficient route**. It's more complex than simply finding the shortest path between two points. It needs to include all relevant factors, such as the number and location of all the required stops on the route, as well as time windows for deliveries.

Optimizing your delivery route can result in fuel savings of more than 10 percent. Route optimization means less driving time for your drivers, which results in lower fuel costs. You can also power an increase in the number of stops per day and a decrease in the amount of time taken between deliveries 6M

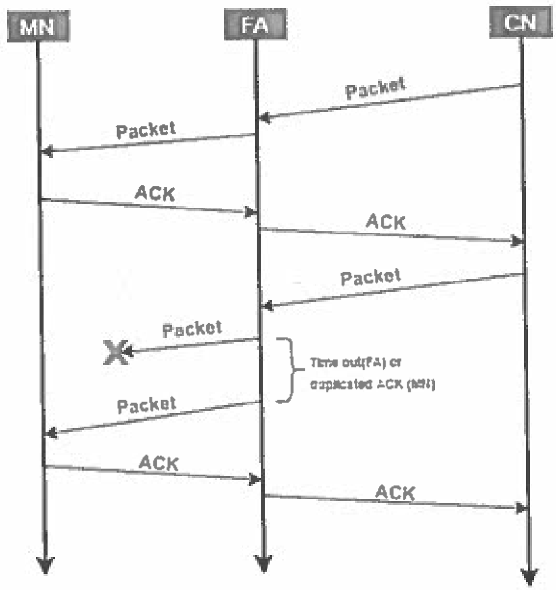
MapQuest. ...
Waze. ...
Speedy Route. ...
TruckRouter. ...
MyRouteOnline. ...
Route4Me. ...
RouteXL. ...
RAC Route Planner. RAC Route Planner is a commercial web-mapping service, an alternative to AA Route Planner to plan efficient orders of routes and receive driving directions and real-time traffic data.

12A We describe the design and implementation of I-TCP, which is an indirect transport layer protocol for mobile hosts. I-TCP **utilizes the resources of Mobility Support Routers (MSRs) to provide transport layer communication between mobile hosts and hosts on the fixed network**. Advantages of I-TCP:

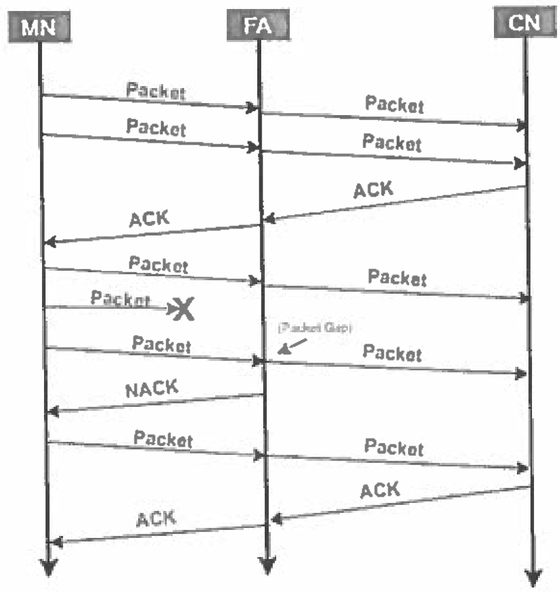
I-TCP does not require any changes in TCP protocol as used by the different hosts in network. Because of a strict partition between the two connections, transmission error on the wireless link will not propagate to the wired link. Therefore, flow will always be in a sequence. 6M

12B. Snooping TCP is one of the classical TCP improvement approaches. This approach is designed to solve the end-to-end semantics loss in I-TCP. The basic concept is to buffer packets close to the mobile node and retransmit them locally if a packet is lost. 6M

Packet Flow: CN to MN
end-to-end TCP-Connection



Packet Flow: MN to CN
end-to-end TCP-Connection



13A Hoarding is performed when the connectivity with the server is strong. In this stage the client becomes aggressive and prefetches the data based on the user access patterns. Caching on the other side have nothing to do with connectivity its an operation which gets performed locally on the frequently used data. 6M

13B What can cause QoS Issues? Problems with QoS performance generally result from one of two categories of causes. First, the network may be experiencing generalized issues, such as **saturated bandwidth, high latency or packet loss**, that are affecting all traffic.

Application QoS: Controls the bandwidth used by applications.
IP QoS: Controls the bandwidth of designated IP addresses.
Role QoS: Also called role-based QoS. It controls the bandwidth of designated roles. 6M

14A. **A pull system initiates production as a reaction to present demand, while a push system initiates production in anticipation of future demand.** In a pull system, production is triggered by actual demands for finished products, while in a push system, production is initiated independently of demands.

Push and pull distribution strategy is all about **directing your promotional route to market**. Either by the product being pushed towards customers or your customers pulling the product through the retail chain towards them 6M

Step 1: Examples of Push. Closing the door. Pushing the table. Pushing the brakes of a car. Pushing off the thumb pins. Pushing a plug inside the socket. Step 2: Examples of Pull. ...
Step 2: Examples of Pull. Opening the door. Pulling a rope. Pulling a chair out of the table. Pulling a kite. Pulling trolley luggage.

15A A model for aspects of visual attention based on the concept of selective tuning is presented. It provides for a solution to the problems of selection in an image, information routing through the visual processing hierarchy and task-specific intentional bias. 6M

Selective tuning is a **process by which client device selects only the required pushed buckets or records, tunes to them, and caches them**. Tuning means getting ready for caching at those instants and intervals when a selected record of interest broadcasts. Broadcast data has a structure and overhead.

15B Data synchronization is **the ongoing process of synchronizing data between two or more devices and updating changes automatically between them to maintain consistency within systems**. While the sheer quantity of data afforded by the cloud presents challenges, it also provides the perfect solution for big data. 6M

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Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Common to All		Academic Year	2022 - 2023
Course Code	20CEO01	Test Duration	3 Hrs.	Max. Marks 70	Semester	V
Course	Urban Environmental Services (Open Elective)					

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the impact of urban planning on health?	20CEO01.1	L1
2	What is urban sprawl and what are the impacts?	20CEO01.2	L1
3	What role does transportation planning play in urban planning?	20CEO01.3	L1
4	Recall the spatial health care access in urban planning.	20CEO01.4	L1
5	What are the factors that influence the collection of preliminary data for urban development?	20CEO01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Elucidate on Health implications of traditional urban planning.	12 M	20CEO01.1	L2
	OR			
7	List and explain the factors affecting urban planning and health development.	12M	20CEO01.1	L2
8	Explain the urban sprawl index and interpret the various factors that influence the benefits and drawbacks of urban form.	12M	20CEO01.2	L2
	OR			
9	Discuss the renewing of health-urban link in cities.	12M	20CEO01.2	L2
10	List and explain the key environmental impacts caused during the implementation of transportation development during urbanization.	12M	20CEO01.3	L2
	OR			
11	Elucidate the various process involved in the system approach concept in urban transport planning and explain its stages.	12M	20CEO01.3	L2
12	Elaborate the five dimensions of access to health care considered in urban health design.	12M	20CEO01.4	L2
	OR			
13	List and explain the various factors that influences the transport and infrastructure on the accessibility to the health services	12M	20CEO01.4	L2
14	Explain data collection for the formulation of conceptual framework of urban system.	12M	20CEO01.5	L2
	OR			
15	Elaborate the role of various governmental and non-governmental organizations/departments on successful implementation of smart cities with major emphasis on health and economy.	12M	20CEO01.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

[UES |(20CE001)] 7/12/22

Scheme of Evaluation for Urban Environmental services |(20CE001). III - I, B.Tech Regular examination:- (2022-23 AY)

PART A (All questions should attempt) [5×2 = 10m]

1. Impacts raised from the urban planning on health — 2m
2. Definition of urban sprawl — 1m
   Impacts of urban sprawl on environment — 1m
   — 2m
3. Transportation planning importance and merits / De-merits — 2m
4. Def of spatial access and health in urban planning and its importance — 2m
5. Factors that relates to Data collection in urban development — 2m

10 m

PART B (5×12 = 60 marks)

6. Brief regarding Health implications of traditional urban planning includes urban form, structure, transportation, physical, economical social and environmental behaviours. — 12m

(or)

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7.	Listing of factors affecting urban planning	2m
	Explanation of factors regarding planning & health development	10m
		12m
8.	Def of urban sprawl index	2m
	Various factors of urban sprawl & form	2m
	Benefits and drawbacks of urban form	8m
		12m
	(or)	
9.	Renewing concept	2m
	Health-urban link concept	2m
	Renewing of urban development and health	8m
		12m
10.	Transportation development intro	2m
	Listing of factors which effect environment	2m
	Explanation of factors during urbanization (includes physical, social, economical and environmental factors)	8m
		12m
	(or)	
11.	Transport planning & process	3m
	process involved in system approach and explanation	9m
		12m

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12	Access to health care concept	2m
	listing of five dimensions in access	2m
	Explanation of dimensions	8m
		12m

(or)

13.	Listing of factors influencing	2m
	transport and infrastructure	2m
	Accessibility to health services	8m
		12m

14	Data collection importance	2m
	Formulation of conceptual framework	2m
	Explanation of framework	8m
		12m

(or)

15. Role and importance of govt, non-govt, Agencies regarding urban planning, development, health development and other environmental aspects } 12m

60m

Part A = 10m
Part B - 60m
70m

—o—

8/12/22
Prepared by
(G. Chanikya)

Verified by
(Head of the Department)
(CE)

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ANSWER KEY AND SCHEME OF EVALUATION
III- I SEM / URBAN ENVIRONMENTAL SERVICES/20CEO01 /2022-23 A.Y 7/12/22

PART A

1. Non communicable diseases like heart disease, asthma, cancer and diabetes are made worse by unhealthy living and working conditions, inadequate green space, pollution such as noise, water and soil contamination, urban heat islands and a lack of space for walking, cycling and active living.

2. Urban sprawl refers to the expansion of poorly planned, low-density, auto-dependent development, which spreads out over large amounts of land, putting long distances between homes, stores, and work and creating a high segregation between residential and commercial uses with harmful impacts on the people.

   (or)

   Urban sprawl is defined as "the spreading of urban developments on undeveloped land near a city." Urban sprawl has been described as the unrestricted growth in many urban areas of housing, commercial development, and roads over large expanses of land, with little concern for urban planning.

3. Transport plays a vital role in urban development. Transport systems provide essential mobility options for people and goods and influences patterns of growth and the level of economic activity through the accessibility. The goal of urban transportation planning is to develop a plan for an efficient, balanced transportation system for an urban area one which will promote a desirable pattern of human activities.

   (or)

   Transport planning is highly essential in shaping cities, enabling economic activities, promoting community interaction, and enhancing quality of life. It is also essential for sustainable development and ensuring safe accessibility at various levels for all individuals.

4. Spatial accessibility is an important factor for planning healthcare services to maintain a quality life for the metropolitan area. These can improve our knowledge of all types of healthcare geography in all settings, including primary care in urban areas. It may provide guidance to the patients in choosing their preferred hospital and may also reduce the time cost of hospital travel .In particular, with respect to the shortage of medical resources during a public health situation, the efficiency and equity of using public medical resources are very important for residents to seek medical treatment.

5. Factors including environmental quality, geographic location, and city scale are prerequisite for conditioning urban innovation and development. Economic factors are including economic level,

industrial structure, industrial agglomeration, and technological innovation. There are four main factors that are expected to affect the future of the urban systems. These according to them, are future population growth, industrial development, improvement of transportation facilities, and communication technology revolution.

PART B

6. Urban design factors can affect public health in several ways, including physical activity, traffic accident risk, pollution exposure, access to health resources, mental health and affordability, which affects households' ability to afford other critical goods, such as healthy food and medical care. Some of the major health problems resulting from urbanization include poor nutrition, pollution-related health conditions and communicable diseases, poor sanitation and housing conditions, and related health conditions. These are areas of high human density and have limited access to potable water, sanitation and other basic services. As a consequence, many times, their health indicators are worse than those in rural areas. High human densities and lack of ventilation make them prone to communicable diseases such as tuberculosis. Among the urban problems that he listed were insufficient housing, poor sanitation, unavailability of safe drinking water, inadequate waste management, and lack of accessibility of the poor to healthcare services, all of which he acknowledged as the foundation of poor urban health.

   We can reduce the health risks by focusing on better air quality, water and sanitation and other environmental determinants; healthy urban planning; healthier and smoke-free environments; safe and healthy mobility, prevention of violence and injuries, healthy food systems and diets also includes smoke-free city ordinances and enforcement; altering the built environment and promoting alternative transport options to foster greater physical activity and reduce air pollution; new approaches to urban food environment to reduce malnutrition and obesity; affordable and healthy housing conditions .

7. There are four main factors that are expected to affect the future of the urban systems. These according to them, are future population growth, industrial development, improvement of transportation facilities, and communication technology revolution.

   Despite some differences state to state, these are the general key factors involved in urban planning.

   1. The Environment and Climate.
   2. The Residents.
   3. Areas in Need of Regeneration.
   4. Resources Available.
   5. The Future.

   Physical factors

   Urban form generally encompasses a number of physical features and non- physical characteristic including size, shape, scale, density, land uses, building types, urban block layout and distribution of green space.

   Environmental factors

   For effective management of urban areas and improve the planning should include the parameters of the natural environment as good or bad as air, drinking water quality, the presence or absence of green areas, noise, the presence or absence of unattractive objects.

Economical factors

Urban economics can help transform the way local councils approach planning proposals, enabling them to take into consideration economic assessments and leading to better decision making, an expert says. A central goal of urban and economic development planning is producing policies and programs to promote economic growth. Urban planners and economic planners always struggle to define economic development policies to improve the growth in way that enhance the quality of life in the community people live and work.

Socio-economical factors

Socio-economic factors include occupation, education, income, wealth and where someone lives.First, socio-economic factors shape the quantity and quality of green spaces and their ability to supply services by influencing management and planning decisions. Second, variation in socio-economic factors across a city alters people's desires and needs and thus demands for different ecosystem services.

There are many different factors that can affect your health. These include things like housing, financial security, community safety, employment, education and the environment. These are known as the wider determinants of health.
Health is influenced by many factors, which may generally be organized into five broad categories known as determinants of health: genetics, behavior, environmental and physical influences, medical care and social factors. These five categories are interconnected.

Environmental factors
These issues include chemical pollution, air pollution, climate change, disease-causing microbes, lack of access to health care, poor infrastructure, and poor water quality.

Physical factors
Physical environment – safe water and clean air, healthy workplaces, safe houses, communities and roads all contribute to good health. Employment and working conditions – people in employment are healthier, particularly those who have more control over their working conditions.

Social factors
In addition to genetics and lifestyle choices, many social factors interact to determine the health of an individual and community. Together, social factors that influence our health (income, education, social connections, and housing) account for up to 40% of what keeps us healthy.

8. Urban sprawl index is a measurement of urban activities including area, population, transportation, green spaces and other requirements for human settlements and benefits. Some of the main factors that have led to grow of cities are: (i) Surplus Resources (ii) Industrialization and Commercialization (iii) Development of Transport and Communication (iv) Economic Pull of the City (v) Educational and Recreational Facilities.

Advantages of urban form:

- The problem of unemployment will be solved.
- There are often roads of a better quality and well-built houses in urban areas.
- Transport facilities are highly developed and often receive regular funding for updates. It can be faster to get from place to place in a city or town.
- Due to better public transport, you can save money on a car
- Most amenities and entertainments are easy to reach. Clubs, restaurants and cinemas are more prolific in these busier areas and you often find new attractions will open in a city before anywhere else.
- Hospitals and clinics are close by for easy access to healthcare or aid in an emergency.
- Cities and towns tend to have a greater mix of cultures and ethnicities which can help when making new friends and meeting people.
- There are a greater number of jobs available in urban areas. Starting a new career could be far easier if you move to a town or city.

Drawbacks

Poor air and water quality, insufficient water availability, waste-disposal problems, and high energy consumption are exacerbated by the increasing population density and demands of urban environments.

- Busy towns or cities can feel crowded and may mean you feel more stress or pressure. You may also not be able to form such tight knit communities in urban areas.
- Urban areas tend to be more expensive to live in. Property prices are higher and so are goods and services.
- Houses are more compact in urban areas. To maximize space, flats and smaller apartments are built instead of houses with larger gardens.
- There are often fewer green spaces in a town or city. You may not always be able to enjoy natural spaces.
- Public transport might not always be as reliable as you'd like, and many towns or cities are restrictive with parking. If you prefer to drive, you may find it harder to keep a car close to where you live unless it is at great expense.
- Because of larger populations, cities can have higher levels of pollution, including noise pollution This could be damaging to your health in the long-term.
- If you have pets you may find it harder to find a place to live that allows them. It may also be harder to find a place to walk a dog or enjoy outdoor space with them.

9. Urban Renewal is an economic development tool used by local governments across the country. More specifically, it is a method of economically revitalizing areas of "blight" through public investments that stimulate private development. The purpose of urban renewal is to improve specific areas of a city that are poorly developed or underdeveloped. Economic Urban Regeneration aims to increase the number of businesses and employment in an area. This will include commercial development enhanced transport links, and high-quality housing to attract people to live and work in a region These areas can have old deteriorated buildings and bad streets and utilities or the areas can lac streets and utilities altogether. Development projects. The rising no communicable disease burden the persistent threat of infectious disease outbreaks and an increased risk of violence and injuries ar

key public health concerns in urban areas. However, urban renewal programs can also have negative effects on social and physical environments by contributing to unsustainable increases in property values and lifestyle costs, leading to social exclusion, gentrification and displacement of long-term residents of lower socio-economic levels.

Other aspects of urban renewal involve the reuse of the land for new purposes, rehabilitation of structurally sound buildings that have deteriorated or lost their original functions, and conservation a protective process designed to maintain the function and quality of an area, for instance, by requiring or assisting. Renewal Projects can facilitate urban life and improve the quality of the environment by constant renovation, increasing of well-equipped public spaces and reduction of environmental pollutions.If the city areas are improved, residents experience a better quality of living. More jobs are created, and the economy can grow. As a result, fewer people move away from the cities to rural areas where resources are often limited. Urban regeneration happens when an urban area is upgraded. The aim is to improve both the economic and social spaces within a city. This usually takes place when areas of dereliction, pollution or brown-field spaces are restored or the area is used for new purposes.

10. The growth of speedy transportation is man's greatest achievement in minimizing distances but at the same time it has also become a cause of environmental degradation. Concern over the environmental consequences of transport development is long-standing. The environmental implications of transport development have become very widely recognized. Through the emissions from combustion of fossil-derived fuels, transportation systems contribute to degraded air quality, as well as a changing climate. Transportation also leads to noise pollution, water pollution, and affects ecosystems through multiple direct and indirect interactions. The environmental effects of transport are significant because transport is a major user of energy, and burns most of the world's petroleum. This creates air pollution, including nitrous oxides and particulates, and is a significant contributor to global warming through emission of carbon dioxide. The potential negative impacts of transportation on environment can be listed as degradation of air quality, greenhouse gas emissions, increased threat of global climate change, degradation of water resources, noise and habitat loss and fragmentation.

    Energy Consumption in Transport and Environmental Pollution:

    Transport requires energy mainly for vehicle operation and to some extent also for manufacturing of the vehicle. The energy consumption in transport sector is the main cause of pollution. There are significant differences in fuel efficiencies between various modes of transport, for example, consumption of energy in cars is more among urban transport modes. Although there has been a significant improvement in the fuel efficiency in cars and other automobiles.

    Air Pollution: Transport is a major source of air pollution not only in developed but in developing countries also. Ecologists believe that the rapid increase in the number of vehicles on our roads, which has taken place without any real restriction, is fast developing into an environmental crisis. Exhaust fumes are the major source of atmospheric pollution by the motor vehicle.

    Noise Pollution: The sources of noise from road vehicles are many and varied, including break squeal, door slam, loose loads, horns, over-amplified music systems, etc. Rail noise depends on the form of propulsion, the nature and load, the speed of train and the type of track. The noise pollution problems around airports are well known.

    Land Consumption and Landscape Damage: The provision of land-based transport requires the direct utilization of land. Long strips of land are consumed, and large areas effectively divided into smaller ones (severance). Previous land uses, such as forestry, agriculture, housing and nature reserves, may be displaced, and zones adjacent to the new development rendered unsuitable for wide range of activities.

Ecological Degradation: The degradation of terrestrial and aquatic ecosystems, as measured by indicators such as reduced habitat/species diversity, primary productivity or the areal extent of ecologically valuable plant and animal communities, provides one of the most emotive aspects of the tension between transport development and environmental quality.

11. Transportation planning is an integral part of overall urban planning and needs systematic approach. The goal of urban transportation planning is to develop a plan for an efficient, balanced transportation system for an urban area one which will promote a desirable pattern of human activities. Transportation planning is also commonly referred to as transport planning internationally, and is involved with the evaluation, assessment, design, and siting of transport facilities (generally streets, highways, bike lanes, and public transport lines).
Urban transportation planning is the process that leads to decisions on transportation policies and programs. In this process, planners develop information about the impacts of implementing alternative courses of action involving transportation services, such as new highways, bus route changes, or parking restrictions. This information is used to help decision-makers in their selection of transportation policies and programs. These are: trip generation, trip distribution, modal split and traffic assignments.

Trip generation is the first step in the conventional four-step transportation forecasting process used for forecasting travel demands. It predicts the number of trips originating in or destined for a particular traffic analysis zone.

Trip distribution is the second component in the traditional four-step transportation forecasting model. This step matches trip makers' origins and destinations to develop a "trip table", a matrix that displays the number of trips going from each origin to each destination.

The modal split, also known as modal share or mode choice, is a common and widespread indicator in transportation engineering to evaluate transportation behavior. In brief, the modal split shows the percentage of travelers using a particular mode of transport compared to the ratio of all trips made.

Route assignment, route choice, or traffic assignment concerns the selection of routes between origins and destinations in transportation networks. It is the fourth step in the conventional transportation forecasting model, following trip generation, trip distribution, and mode choice. Traffic assignment models are used to estimate the traffic flows on a network.

12.

Health care or healthcare is the improvement of health via the prevention, diagnosis, treatment, amelioration or cure of disease, illness, injury, and other physical and mental impairments in people. The maintaining and restoration of health by the treatment and prevention of disease especially by trained and licensed professionals (as in medicine, dentistry, clinical psychology, and public health) They grouped these characteristics into five as of access to care: affordability, availability, accessibility, adequacy, and acceptability.

Health care affordability describes whether a person or organization has sufficient income to pay for or provide for health care costs. These costs could be insurance premiums or direct health care service costs. Affordability is one of the most important challenges influencing Americans' ability to access health care. However, no single, agreed-upon definition of health care affordability exists because it is influenced by many complex factors. Meanwhile, employers are anxious about their increasing contributions to employee health plans. In addition, federal and state governments are troubled by rising per capita spending on health care, growing budget deficits, and which priorities to fund, such as Medicare and Medicaid, education or infrastructure.

Availability: Availability means that there is an adequate supply of health care professionals who have the training and skills to meet the needs of the patients. Access to health care may vary across countries, communities, and individuals, influenced by social and economic conditions as well as health policies. Providing health care services means "the timely use of personal health services to achieve the best possible health outcomes". Lack of health insurance coverage may negatively affect health. Uninsured adults are less likely to receive preventive services for chronic conditions such as diabetes, cancer, and cardiovascular disease. So that availability is important in health care access. Reducing the income gap between primary and specialized care providers. Improving work life of primary care physicians. Increasing funding for primary care training. Expanding training program for general practitioners.

Accessibility can be defined as the ease with which health services are reached in terms of physical access, costs, time, and availability of qualified personnel. Accessibility is a prerequisite for a high-quality and efficient health system. When patients have good access to healthcare, they are more likely to receive the care they need to live a long, healthy life. Good access can improve the quality of life for patients by providing preventive care, early treatment of health problems, and management of chronic conditions. Health care access is the ability to obtain healthcare services such as prevention, diagnosis, treatment, and management of diseases, illness, disorders, and other health-impacting conditions. For healthcare to be accessible it must be affordable and convenient. Many people do not have access to adequate healthcare.

Adequacy is sufficient in quantity, quality, or amount to achieve a desired therapeutic effect. For patients, it is essential that health care be adequate. This means that services and treatments patients need, including those required by individuals with unique and complex health needs, are covered. Adequacy is the state of being sufficient for the purpose concerned. The meaning doesn't suggest abundance or excellence, or even more than what is needed. Adequacy is the quality of being good enough or great enough in amount to be acceptable. Several studies point to a real cause for concern over the adequacy

Acceptability is a multi-faceted construct that reflects the extent to which people delivering or receiving a healthcare intervention consider it to be appropriate, based on anticipated or experienced cognitive and emotional responses to the intervention. If an intervention is considered acceptable, patients are more likely to adhere to treatment recommendations and to benefit from improved clinical outcomes. Acceptability is the characteristic of a thing being subject to acceptance for some purpose. A thing is acceptable if it is sufficient to serve the purpose for which it is provided.

13. Travel behavior is the study of what people do over geography, and how people use transport. Cities around the world differ in the extent of their provision of public transport, infrastructure for walking and cycling, and car dominance. Similarly, the existence and contents of policies related to speed and blood alcohol concentration limits, for example, and their enforcement differ markedly. These examples can have profound impacts on the health of citizens and on inequalities between different groups and demonstrate the implications of transport policy decisions.
Countering the potential for transport to positively impact on health and wellbeing are the numerous negative impacts that continue to be observed globally. Transport systems that prioritize private car use and other motorized vehicles impose a greater burden of related harms such as air pollution noise pollution, road traffic casualties, community severance, poor mental health and reduced social interaction and cohesion. Through the emissions from combustion of fossil-derived fuels transportation systems contribute to degraded air quality, as well as a changing climate Transportation also leads to noise pollution, water pollution, and affects ecosystems through multiple direct and indirect interactions. Transport policies centered on private vehicle use also limit the opportunities for, and likelihood of, daily physical activity and the associated health benefits. I western countries, the health benefits of active travel greatly outweigh the harms from injuries an air pollution. Unfortunately, transportation can negatively impact human health in two broa ways, accidents and environmental pollution. The most direct impact of transportation on healt is vehicle-related injuries and deaths. Motor vehicle accidents are not limited to cars and the occupants but include accidents involving trucks, vans, busses, pedestrians, and bicyclers. It can als facilitate the spread of infectious diseases. Common transportation barriers include long trav distances, lack of vehicle, transportation cost, inadequate infrastructure, and adverse policie affecting travel. Each of these obstacles can keep a patient from accessing her providers, which in tur could impair overall health.

Transport affects health in many ways. Benefits include access to education, employment, goods, services and leisure, and opportunities for incorporating physical activity into daily living.

- Physical activity.
- Lack of exercise.
- Communicable diseases.
- Employment.
- Safety.
- Travel.
- Noise.
- Pollution.

Barriers to transportation access impact all parts of an individual's health care, including missed or delayed doctor and clinic appointments, limited pharmacy access and decreased prescription refills. Transportation barriers also cause an economic burden for patients and the health care system.

14. Urban systems design arises from disparate current planning approaches (urban design, Planning Support Systems, and community engagement), compounded by the reemergence of rational planning methods from new technology (Internet of Things (IoT), metric based analysis, and big data). The proposed methods join social considerations (Human Well-Being), environmental needs (Sustainability), climate change and disaster mitigation (Resilience), and prosperity (Economics) as the four foundational pillars. Urban systems design integrates planning methodologies to systematically tackle urban challenges, using IoT and rational methods, while human beings form the core of all analysis and objectives.

    In urban data collection, this could mean getting citizens to actively gather the data that is needed or to decide what types of data should be collected. In urban design, it could mean soliciting people's views on a design scheme or plan, or designing and building together with residents.

    Urban researchers develop and test survey instruments to capture the data needed for effectively addressing a study's research questions. Urban surveys are often multimode ventures, collecting data face to face, over the phone, through the mail, online, or any combination of the above. Procedure may involves the following

    - Step 1: Identify issues and/or opportunities for collecting data.
    - Step 2: Select issue(s) and/or opportunities) and set goals.
    - Step 3: Plan an approach and methods.
    - Step 4: Collect data.
    - Step 5: Analyze and interpret data.
    - Step 6: Act on results.

15. The main tasks of GOs, NGOs in the health system are providing services and raising health advocacy. These services include medical, social, and psychological services.They help consumers by providing assistance and help them in seeking redressal. They encourage consumers to protest against exploitation In short the NGO's make consumers to become aware of their rights raise voice against exploitation and helps them to seek redressal of grievances. The urban NGO may forge links between beneficiaries and levels of government, donors and local financial institutions. A central purpose of

the urban NGO may be to provide services directly to other organizations that support the poor perform co-coordinating or networking services. NGOs experience in health care lias not only contributed to health development but also has influenced the health care delivery system of the government. However, the major contribution of NGOs is the involvement of people in lineally prevention and promotion.

Their main aim is to make the earth a better place for every human being who is suffering. The function of NGO is to focus on all the issues concerning human rights, social, environmental and advocacy. They work to promote and improve the social and political conditions of the society on a broad scale. Areas of intervention of these NGOs are environmental protection are awareness generation, resource development and documentation, introduction of alternative livelihood, coordination and assistance with various governmental departments, habitat monitoring management and restoration etc.

These are the some of the activities and programs should avail .

1. Education
2. Reducing poverty
3. Implementing govt and non govt schemes for public beneficiaries
4. Improving green spaces and health access facilities
5. Economical development
6. Financial assistance
7. Infrastructure development
8. Employment
9. Environmental protection
10. Awareness programs
11. Health care and medical supply
12. Pollution control strategies
13. Development of infrastructure and Transportation etc

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Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Common to All	Academic Year	2022- 2023
Course Code	20EEO01	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Introduction to Renewable Energy Sources (Open Elective)				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Draw the I-V characteristics of PV cell	20EEO01.1	L1
2	State the wind power equation	20EEO01.2	L1
3	Write any 4 merits and limitations of wave energy	20EEO01.3	L1
4	List out any 4 demerits of Bio-gas plant	20EEO01.4	L1
5	What is the importance of the geothermal energy over the conventional energy sources?	20EEO01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain about the solar radiation spectra.	6M	20EEO01.1	L2
6 (b)	Explain about the working principle of a flat plate collector with a neat sketch.	6M	20EEO01.1	L2
	OR			
7 (a)	Explain about the principle of the solar thermal power generation.	6M	20EEO01.1	L2
7 (b)	Explain about any two solar PV applications.	6M	20EEO01.1	L2
8 (a)	Explain about the 2 types of wind energy systems.	6M	20EEO01.2	L2
8 (b)	Explain about any 6 factors affecting the site selection of the wind power plants.	6M	20EEO01.2	L2
	OR			
9 (a)	Derive the expression for power in the wind turbine systems.	7M	20EEO01.2	L3
9 (b)	Write a short on wind turbine generators.	5M	20EEO01.2	L1
10 (a)	Explain about the working principle of OTEC with a neat sketch. And list out its demerits.	6M	20EEO01.3	L2
10 (b)	List out any 5 merits and demerits of tidal energy.	6M	20EEO01.3	L1
	OR			
11 (a)	Explain about the OTEC closed cycle system with a neat sketch.	6M	20EEO01.3	L2
11 (b)	Write a short note on single basin system.	6M	20EEO01.3	L1
12 (a)	Explain about the principle of bio-conversion systems with a neat sketch.	5M	20EEO01.4	L2
12 (b)	Explain about single stage bio gas digesters along with any 3 advantages and disadvantages.	7M	20EEO01.4	L2
	OR			
13 (a)	Explain about the combustion characteristics of bio-gas.	6M	20EEO01.4	L1
13 (b)	Explain about the working principle of I.C engine with a neat sketch.	6M	20EEO01.4	L1
14 (a)	Explain about the basic principle of Geo-thermal energy generation with a neat sketch.	6M	20EEO01.5	L2
14 (b)	Write any 6 merits and demerits of geothermal energy.	6M	20EEO01.5	L2
	OR			
15 (a)	Compare between Geothermal power plant and conventional thermal power plant.	6M	20EEO01.5	L2
15 (b)	What is the need of an improved cooking stove (ICS) over conventional cooking stoves in the world? Explain.	6M	20EEO01.5	L1

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ANSWER KEY AND SCHEME OF EVALUATION

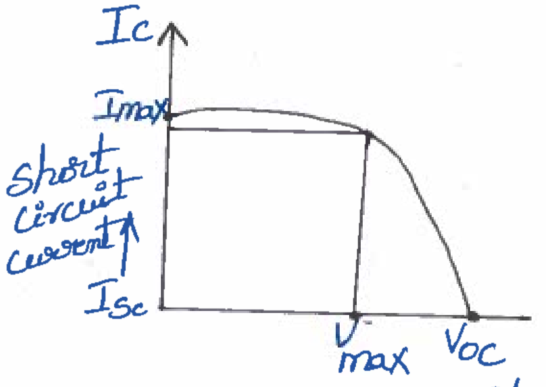
Course Code - 20EE001

Course Name - Introduction to Renewable Energy sources

Short Answer Questions (5×2 = 10 Marks)

1. Draw I-V characteristics of PV cell

A. Solar cell is a device to Convert light energy to Electrical Energy. Rated voltage = 0.5V



$$P_{max} = V_{max} \cdot I_{max}$$

$$P_{max} = I_L \times V_{OC} \times \text{fill factor}$$

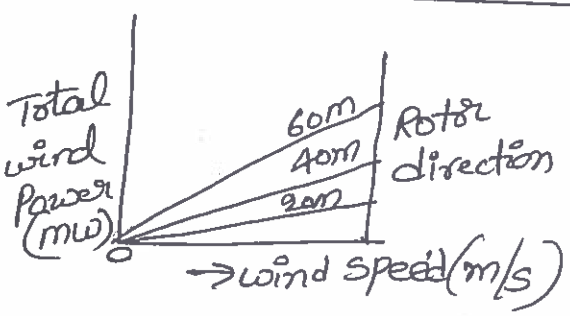
$$\left[\text{fill factor} = \frac{I_{max} \cdot V_{max}}{I_{sc} \cdot V_{oc}}\right.$$

2. State the wind Power Equation?

A. wind mill works on the principle of

$$K.E = \frac{1}{2} m v^2 \qquad \left[m = \rho . A . V \right.$$

$$= \frac{1}{2} (\rho . A . V) v^2 = \frac{1}{2} \rho . A . V^3$$



Area $A = \pi/4 . D^2$

∴ Available Power $P_a = \frac{1}{2} \rho \left(\frac{\pi}{4} D^2 \right) v^3$

$$= \frac{1}{8} . \rho . \pi . D^2 . v^3 .$$

3. write any 4 merits & limitations of wave Energy?

A.

Merits	limitations
→ Reliable	→ High Expensive
→ Environment - Eco-friendly	→ High maintance Cost
→ Enormous energy potential	→ Enviroment effect
→ No damage to land	→ Hard to scale.
→ Renewable Source of Energy	

4. List out any 4 demerits of Bio-gas Plant?

A. → Bio-gas Contains impurities
→ Effect of Temperature
→ Requires large areas.
→ Not Economically viable
→ Little technology advancements.

5. what is the importance of Geo-Thermal Energy over the Conventiona energy sources?

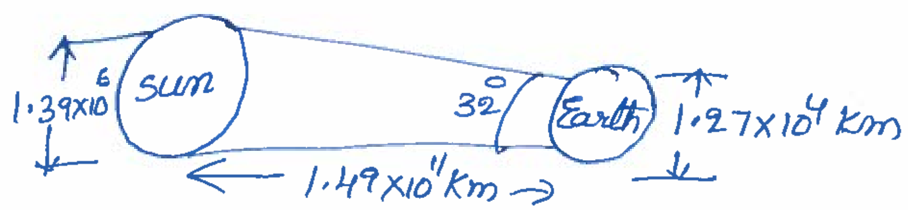
A. Energy that Can be obtained by Man through the use of heat inside the Earth., due to Several factory, including the geotherma gradient heat Energy generated.

→ Geothermal areas depent on a reservior of hot water, the volume taken out Can be reinjected making it a sustainable Energy source. An increase in mass flow rate is shown to increase both Power output & Efficient.

Long Answers (5x12 = 60 marks)

6(a) Explain about the Solar radiation spectra.

(A) Solar radiation is radiant energy (Electromagnetic) from the sun.



→ Sun is the basic source of energy for earth.

→ sun is a large sphere of very hot gases, heat being generated by the various fusion reactions in it.

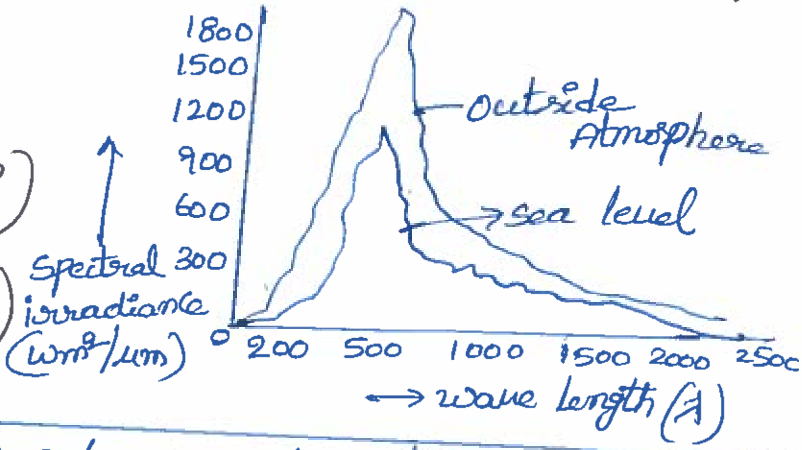
→ Energy is radiated by the sun as Electromagnetic wave of which 99% of wave length (λ) in the range of 0.2 μm to 4 μm [$1 \mu m = 10^{-6} m$]

→ Solar energy reaching the top of the earth atmosphere which consists of

8% UV rays (< 0.39 μm)

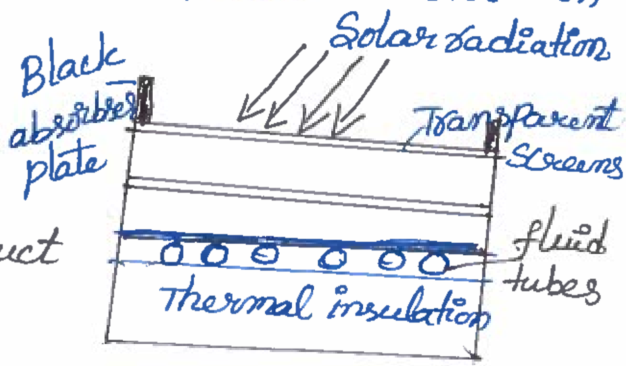
46% visible light (0.39 μm to 0.78 μm)

46% Infrared rays (0.39 μm to 0.78 μm)



5(b) Explain about the working principle of a flat plate collector with a neat sketch?

(A) → They are made in rectangular panels about 1.72 m, 2.9 m^2 in area under relatively simple to construct and erect.



→ Flat-plates can collector and absorber both direct & diffuse radiation.

→ Black surface – absorbent of the incident solar energy

→ tubes containing heating fluid to transfer the heat from the collector.

→ support structure to protect the components and hold them in place.

→ Insulation covering sides and bottom of the collector to reduce heat losses.

→ The thermal insulation prevents heat loss during fluid transfer, the screens reduce the heat loss due to convection & radiation to the atmosphere.

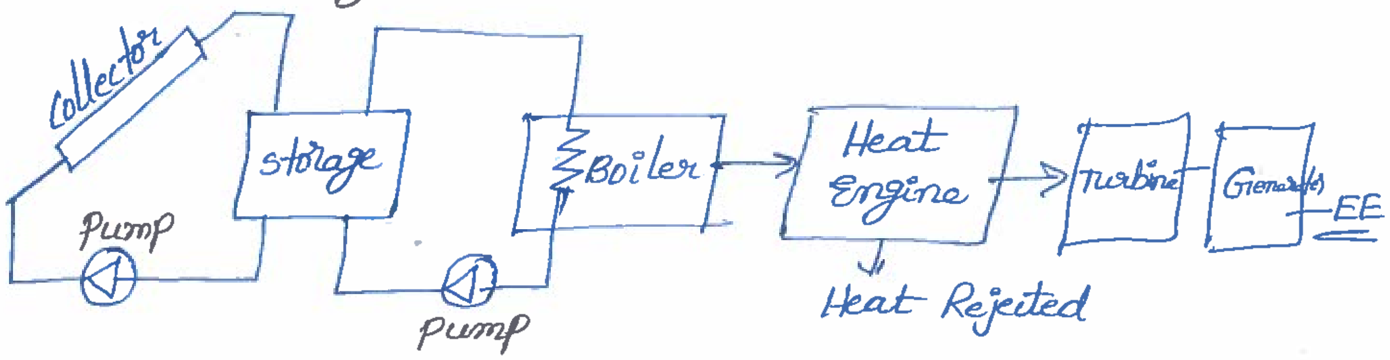
7a) Explain about the principle of Solar thermal Power generation?

(A) → In solar thermal power plants, solar radiation is used to generate electricity.

→ mirrors Concentrate the sunlight on a radiation Collector & heat up a heat-bearing medium, generally thermal oil.

→ A turbine transforms this energy into Electricity.

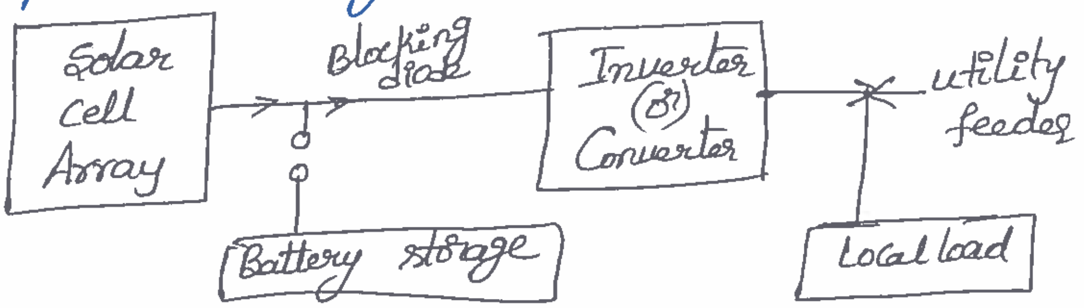
→ Solar thermal plants are an efficient way of transforming Solar energy into Electricity.



Solar Collector is a device to collect the heat energy (steam Energy) that energy Can be fed to storage pump through a Condenser through a boiler to steam energy Can be Converted to Kinetic Energy. Kinetic energy Can be passed through the turbine gets mechanical Energy through Generator gives Electrical Energy.

7b) Explain about any 2 solar PV applications?

(A)



Solar cell array is large (or) small system which Convert the insolation to useful DC power.

Blocking diode :- Array generated power flow only towards the battery (or) grid (without a blocking diode the battery would disch. back through the Solar array)

Battery storage :- In which solarly generated electrical energy may be stored.

Inverter / Converter :- Inverter DC-AC Conversion.
Converter AC-DC Conversion.

usually solid state energy which Converts the battery bus voltage to AC of frequency & phase to match that needed to integrate with the utility feeder (Transmission line).

Solar photovoltaic's (Solar PV)

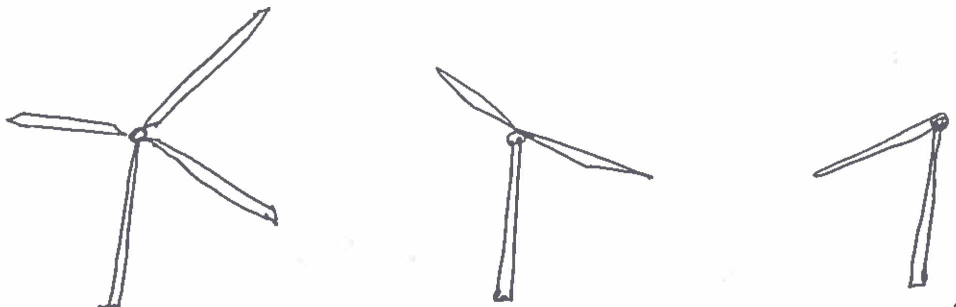
- A solar PV system generally Consists of photovoltaic modules ("Solar panels") installed as an array (Series of panels) on a roof-top to generate electricity to be used by home (or) business.

→ Solar PV array Captures the sun's rays & delivers energy to a home.

→ sun light energy, absorb solar energy, Transforms the energy into electricity.

8a) Explain about the 2 types of wind energy systems?

(A)



Single blade :- A long blade is mounted on rigid hub induction generator and gear box is also Connected, if extremely long blades mounted on rigid hub, large blade root bending moment may be rotated through it, sudden shift in wind direction which gives kinetic Energy.

2,3 blade system

In this type of design rotor drives a generator through a step-up gear box. The blade rotor is usually designed to be oriented down wing of the tower.

wind turbines :- No. of Blades.

- most Common design is the 3-blade turbine the most important reason is the stability of the turbine. A rotor with an odd no. of rotor blades (at least '3' blades) Can be Considered.
- A rotor with an even no. of blades will give stability problems for a machine with a stiff structure, the tower most blade passes into the wind shade infront of the tower.

(8b) Explain about any 6 factors affecting the site selection of the wind power plants?

(A) → favourable land cost

→ The speed generated by the wind mill depends on cubic value of velocity of wind (v^3) the small velocity increases which effect the power.

→ It is obviously desirable to select a site for windy energy power plant with high wind velocity.

→ Distance to rail, road

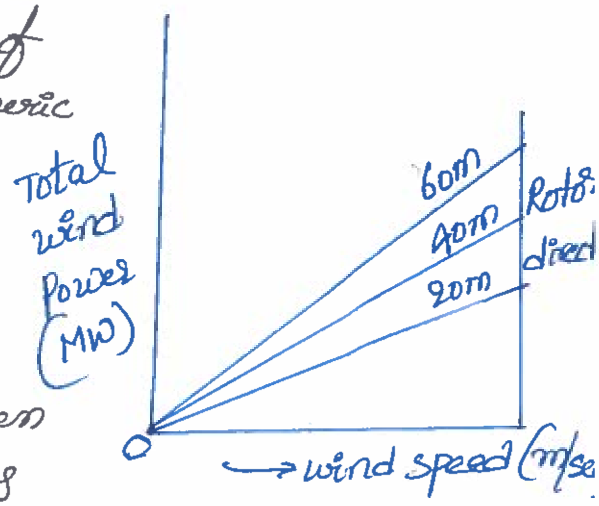
→ Nature of ground

→ Nearness of site to local centre (users)

→ wind structure at the proposed site (altitude)

(9a) Derive the Expression for power in the wind turbine systems?

(A) wind is the stabilization movement of air between areas of high & low atmospheric pressure, created by the uneven heating of the earth's surfaces, land, water and air.



→ The greater the pressure difference between these areas the harder the wind blows

→ wind also exists as the circulation of air around a high or low pressure area.

wind Energy

wind Energy is Converting of wind Power to Electrical Power through the use of wind mills (or) turbines.

— Electricity produced is sent to transformers where voltage is increased and sent to the power grid via transmission lines.

wind mill works on the principle of kinetic Energy (K.E)

$$K.E = \frac{1}{2} m v^2$$ [v - velocity, m - mass, area - A]

$$= \frac{1}{2}(\rho \cdot A \cdot V) v^2 \qquad m = \rho \cdot A \cdot V$$

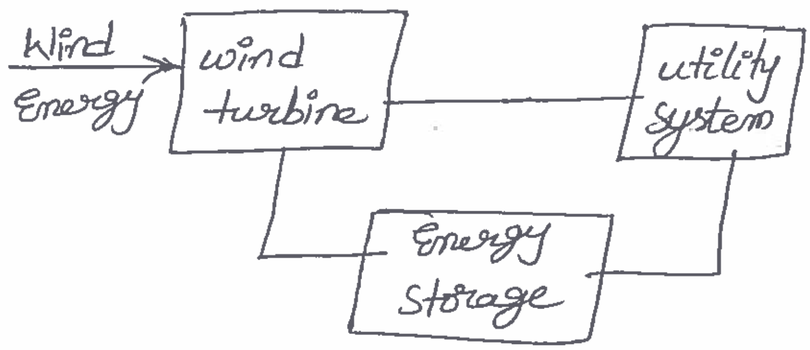
$$= \frac{1}{2} \rho \cdot A \cdot V^3$$ [air density - 1.25 kg/m^3]

$\therefore$ Available power — aero-turbine $A = \frac{\pi}{4} D^2$

$$P_a = \frac{1}{2} \rho \left(\frac{\pi}{4} D^2\right) \cdot V^3$$

$$\boxed{\therefore P_a = \frac{1}{8} \rho \cdot \pi \cdot D^2 \cdot V^3}$$

9b) write a short notes on wind turbine generators? [5M]

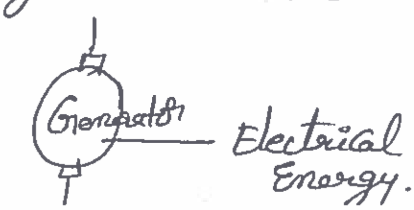


→ wind power generators Convert wind energy (mechanical energy) to electrical energy.

→ The generator is attached at one end to the wind turbine, which provides the mechanical Energy

→ At the other end, the generator is Connected to the electrical grid.

→ The generator needs to have a Cooling system to make sure there is no overheating.



wind energy is Converting of wind Power to electrical Power through the use of wind mills (or) turbines.

→ A wind mill Captures wind Energy and then uses to generator to Convert into a Electrical Energy.

(10a) Explain about the working principle of OTEC with a neat sketch & list out its demerits?

(A) OTEC (Ocean thermal Energy Conversion) is a process that Can produce electricity by using the temperature difference between deep cold ocean water & warm tropical surface water.

→ OTEC is an energy technology that Converts Solar radiation to electric Power OTEC utilizes the world's largest Solar radiation Collector, The ocean Contains enough energy Power all of the world's electrical needs.

→ Tapping ocean currents, power is Converted to high voltage DC & is Cabled to shore for Conversion to AC & integration into the Local Power distribution network.

Demerits

- OTEC plants are extremely expensive
- current plant only acheive 1% to 3% η
- limited geographical availability
- Electricity must also be transported to land.

(10b) list out any 5 merits & demerits of tidal Energy?

(A) merits

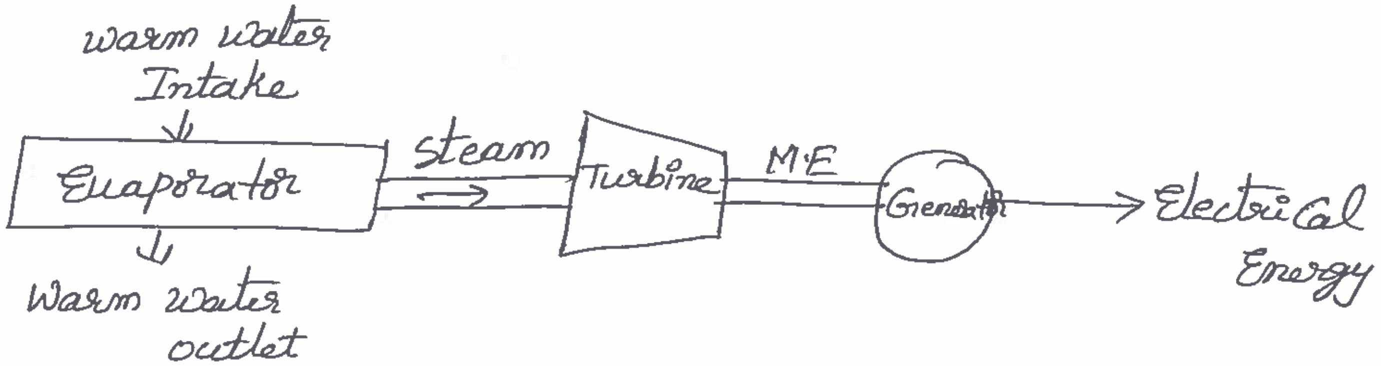
- Tidal energy is an inexhaustable Source of energy since water is available in most parts of the earth
- Environment Eco-friendly
- Built to use it effectively
- It doesn't produce any harmful gases to Cause pollution
- Tidal Energy doesn't require any kind of fossil fuel to thereby Contributing to fuel Consumption

Disadvantages

- There is high-cost involvement in the construction of tidal power plants due to corrosion-free machinery.
- The intensity of sea waves may vary so there can be an interruption in the power generation process.
- may influence aquatic life to adversely die to control of water flow & large driving forces involved.
- This technology used for utilization of tidal energy is still not very cost effective.
- They require setup in coastal regions only.

(11a) Explain about the OTEC closed cycle system with a neat sketch.

(A)



closed-cycle system (Ranking-cycle) use fluid with a low-boiling point, such as, NH_3 to rotate a turbine to generate electricity.

- when warm sea-water is pumped through a heat exchanger where the low-boiling point fluid is vapourized.

→ the expanding vapour turns the turbo-generator

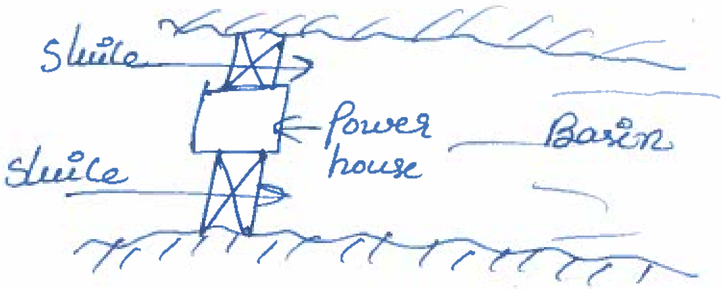
→ Then cold, deep sea-water-pumped through a second heat exchanger.

→ Then condenses the vapour back into a liquid which is then recycled through the system.

→ In this Ranking cycle turbine-generator mainly to generate electricity.

(11b) Write a short note on single basin system?

(A) → The tidal range of 5m & above available in particular locations can be utilized to operate a hydraulic turbine



→ The mechanical power of the turbine can be used to run a generator to produce Electrical Power.

→ In case of power generation by tides, the water during high tides is 1st trapped in an artificial basin & then it is allowed to escape during the period of low tides.

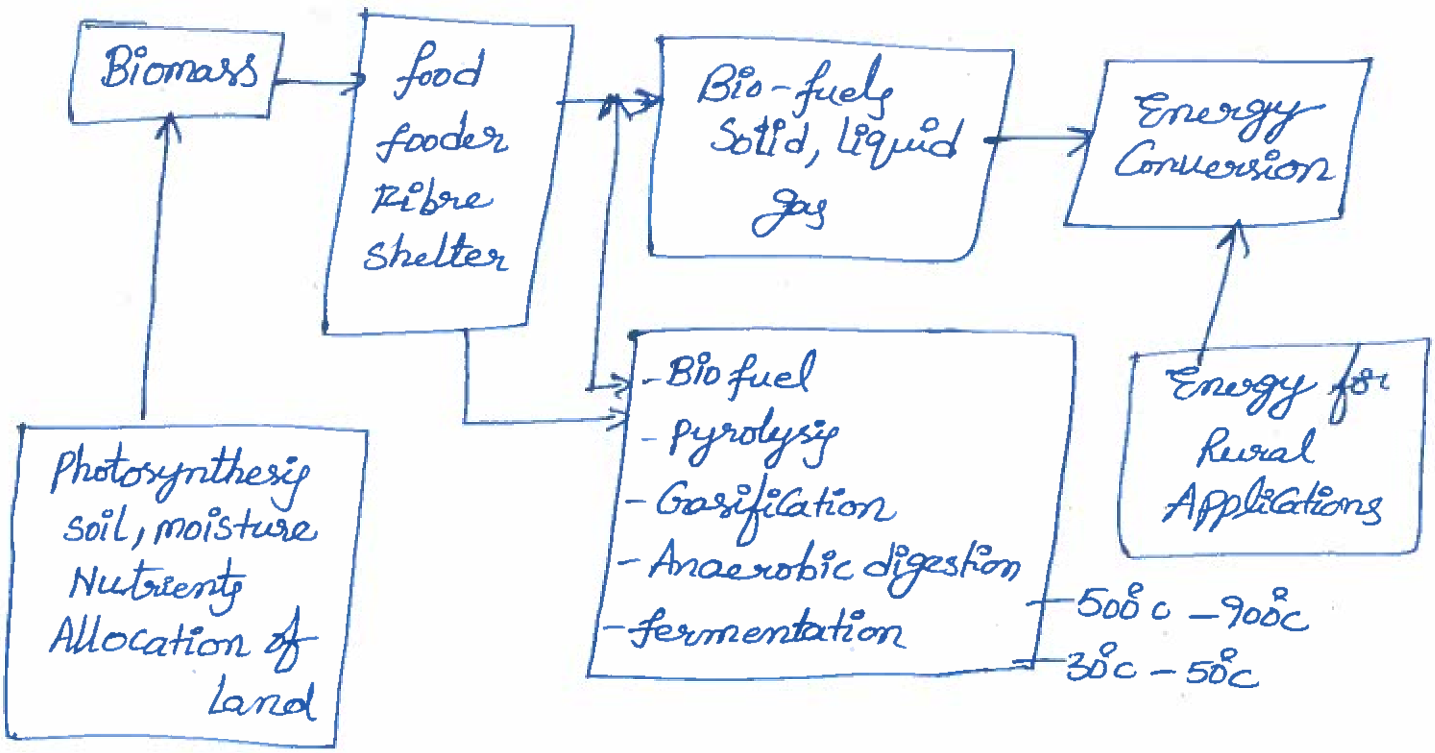
→ The water while escaping is utilised to run a hydraulic turbine coupled to a generator.

[3 Components — Power house
— Dam
— sluice gates

→ Dam is to form a barrier b/w the sea & basin
→ Basin during high tide

(12a) Explain about the principle of Bio-Conversion systems with a neat sketch?

(A)

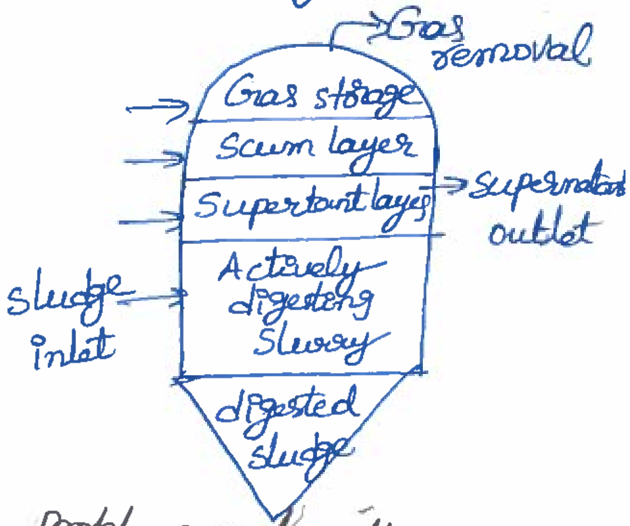


- The total mass of living matter within, a given unit of environment area plant material agricultural waste used as a fuel (or) Energy Source
- Biomass is a renewable energy source it is delivered from living (or) recently living organisms.
- Energy delivered from biomass is mostly to generate electricity (or) to produce heat.
- Thermal Energy is extracted by means of Combustion, Pyrolysis & gasification.
- Biomass Can be chemically & bio-chemically treated to Convert to a Energy-rich fuel.

12b) Explain about single stage bio-gas digesters along with 3 advantages & disadvantages?

(A) - The entire Conversion from organic Compounds into Biogas is Carried out in a single chamber

- The raw material is fed regularly in the chamber while the spent waste moves from the outlet.



- The agriculture residues produces the problem, when they are fermented in a single stage process.

Advantages

- mature for heat for updraft
- small Scale applications for Downdraft
- fluidised bed for large Scale applications
- feed characterstics Can produce gas

Disadvantages

- feed size limits
- High tar yeilds → medium tar yeid for downdraft.
- Slagging Potentia

(13a) Explain about the Combustion characteristics of bio-gas?

(A) Energy plantation can be treated as the growth of land plants for their fuel value by capturing solar radiation.

→ Energy plantation offers a renewable energy source for liquid fuel & organic chemicals; alternative to fossil & nuclear energy source.

– The forest product industry are mature trees for high yeild desirable for energy plantation.

Based on Combustion of fuel → External Combustion / Internal Combustion

External Combustion System Ex:- S.I Engine

The working medium, the steam is generated in a boiler, located outside the engine & allowed into the cylinder to operate the piston to do mechanical work.

Internal Combustion system Ex:- C.I Engine

In Internal Combustion Engine, the Combustion of fuel takes place inside the cylinder & the pressure of the air is increased tremendously.

(13b) Explain about the working principle of C.I Engine with a neat sketch?

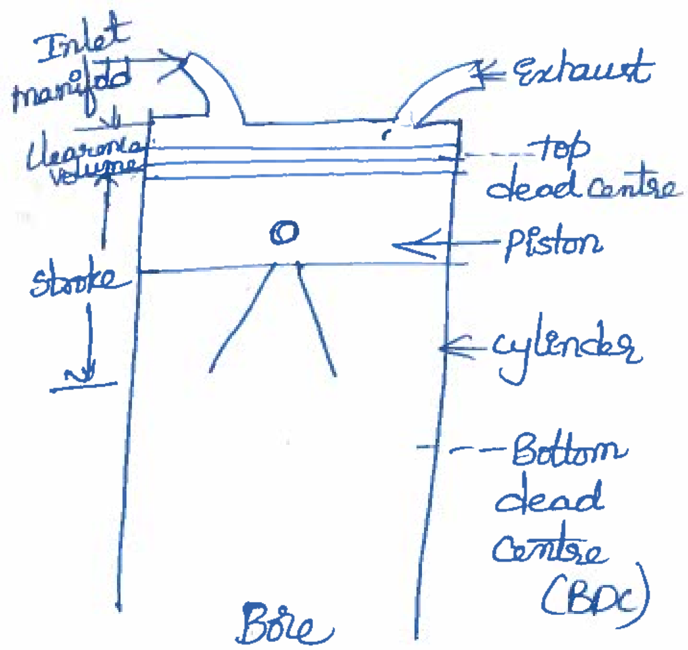
(A) Diesel engine can be made to operate on dual fuel & modifications are

– provides for entry of bio-gas

– advancing the injection timing i.e fuel improves

– provision of a system to reduce the supply of diesel

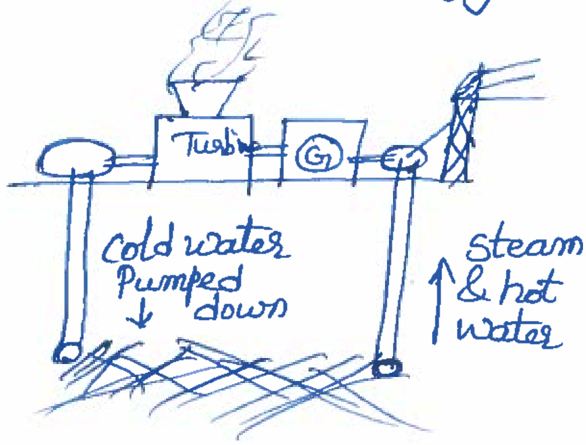
– The bio gas, diesel & intake air is mixed properly in a mixing chamber



- once Engine starts running, provision should be made for reducing the supply of diesel in order to make the speed of Engine optimum & as required for various load conditions. This is done by actuating a control rack.

14a) Explain about the basic principle of Geo-Thermal Energy generation With a neat sketch?

(A) Deep underground, the earth's rocks are naturally very hot.
- We can turn their heat Energy into Electrical Energy to use in our homes - we call this geothermal Energy.



- Cold water is pumped below the ground
- Hot rocks heat the water, turning into steam Energy
- The steam is used to generate electricity.
- Geo-thermal energy is one of the Renewable Energy Sources.
- Rain water filters in the Earth's interior where in contact with intrusion (or) active magma fans it heats upto considerable temperature. As a result it travels to the surface of the Earth as hot water (or) water steam.

14b) Write any 6 merits & demerits of geothermal Energy?

(A)

Merits	Demerits
- Good for Environment	- The Extraction of G.E Causes greenhouse Emissions
- High Efficient	- High Cost investment need
- Reliable Source of Renewable Energy	- Depletion in geothermal Sources
- There is no too little geothermal system maintenance needed	- Geothermal reservoirs cannot Easily be found
- No green house gases	- It is hard to implement G.E in big cities
- There is an unlimited supply of Geothermal Energy	- Not many places to build Power stations

(15a) Compare between Geothermal plant & Conventional Thermal Power plant?

(A)

	Geothermal	Thermal
① formation	Naturally occuring in the sea	industrial setup
② Requirement/dependance	Temperature difference of water in sea	grade of coal and petroleum
③ type	non-Conventional Source	Conventional Source
④	In exhaustable	Exhaustable
⑤	Non-polluting	Polluting
⑥	Renewable Source	Non-Renewable Source.

(15b) what is the need of an (Ics) over Conventional Cooking stoves in the world? Explain?

(A)

Issues	Traditional Ics	Smart Ics
→ Installation type	most are fixed, site selection is Important	Portable
→ Ready to go	After installation required for certain time for ensuring dry stove is throughly rubbed	Ready to use immediately
→ Body fracture & crack	Required Smoothen the Ics body with mud	No such problem
→ Corrossion problem	Yes	NO
→ Heat Trapping	NO	Yes
→ chimney clogging problem	Yes, required	No, cleaning required

Scheme of Evaluation

Short Answer Questions 5×2=10 Marks

1) I-V characterstic – [1M] } – [2M]
   formulae – [1M]

2) K.E Equation – [1M] } – [2M]
   Wind power Equation – [1M]

3) Wave Energy merits – [1M] } – [2M]
   Demerits – [1M]

4) Demerits of Bio gas – [2M]

5) Geothermal over Conventional Energy
   Explanation – [2M]

Long Answer Questions –

6a) Solar radiation diagram – [2M] } – [6M]
    Explanation – [4M]

6b) flat plate Collector diagram – [2M] } – [6M]
    Explanation – [4M]

7a) Solar Thermal Power generation – [2M] } – [6M]
    Explanation – [4M]

7b) Solar PV applications diagram – [2M] } – [6M]
    Explanation – [4M]

8a) 2 types of wind Energy system diagram – [2M] } – [6M]
    Explanation – [4M]

8b) 6 factors of selection of site – [6M]

9a) Wind power derivation & Explanation – [7M]

9b) Wind turbine diagram – [2M] } – [6M]
    Explanation – [4M]

(10a) OTEC system Explanation – [4M]
Demerits – [2M] } – [6M]

(10b) 5 merits – [3M]
5 Demerits of Tidal Energy – [3M] } – [6M]

(11a) OTEC closed cycle diagram – [2M]
Explanation – [4M] } – [6M]

(11b) Single basin diagram – [2M]
Explanation – [4M] } – [6M]

(12a) Bio-Conversion system diagram – [2M]
Explanation – [3M] } – [5M]

(12b) single stage digester diagram – [2M]
Explanation – [2M], Advantages disadvantages } – 3M } – 7M

(13a) Combustion characterstics Explanation – [6M]

(13b) C.I Engine diagram – [3M]
Explanation – [3M] } – [6M]

(14a) Geothermal Energy diagram – [3M]
Explanation – [3M] } – [6M]

(14b) 6 merits – [3M]
6 Demerits of Geothermal Energy – [3M]

(15a) Geothermal over Conventional plant Explantion – [6M]

(15b) (ICS) over Conventional type Explanation – [6M]

HOD-EEE
8/12/22

08/12/2022

NSRIT

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	Mechanical Engineering			**Academic Year**	2022 - 2023
Course Code	20MEO01	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	V
Course	**Nano Technology (Open Elective)**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define the following terms (i) Nano Science (ii) Nano Technology.	20MEO01.1	L1
2	Define Top-down and bottom-up approach.	20MEO01.2	L1
3	Define Nanoforms of Carbon.	20MEO01.3	L1
4	What do you mean by surface analysis technique?	20MEO01.4	L1
5	Define Nano Biotechnology.	20MEO01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the classification of Nanoscale particles.	6M	20MEO01.1	L1
6 (b)	What is Nanometer scale? Explain.	6M	20MEO01.1	L2
	OR			
7 (a)	What are the effects of Nano materials on Magnetic and Mechanical properties?	6M	20MEO01.1	L1
7 (b)	What are the implications of Nano science for Physics and Chemistry?	6M	20MEO01.1	L2
8	Explain Ball milling method under Top-down approach with the help of neat sketch.	12M	20MEO01.2	L2
	OR			
9 (a)	With the help of neat sketch, explain the Physical Vapour Deposition (PVD) method.	8M	20MEO01.2	L2
9 (b)	Define Lithography.	4M	20MEO01.2	L1
10 (a)	Explain Quantum wires	4M	20MEO01.3	L2
10 (b)	Explain the following (i) Single wall Carbon Nano Tubes (ii) Double wall Carbon Nano Tubes	8M	20MEO01.3	L2
	OR			
11	Describe the Buckminster fullerence with the help of neat sketch.	12M	20MEO01.3	L2
12 (a)	With the help of neat sketch, explain Scanning Electron Microscopy (SEM).	6M	20MEO01.4	L2
12 (b)	Explain X-ray diffraction technique.	6M	20MEO01.4	L2
	OR			
13	Explain the following (i) Atomic force microscopy (ii) Scanning Tunneling Microscope	12M	20MEO01.4	L2
14	Write a short note on (i) Nano computer (ii) Super chip (iii) Nano crystal	12M	20MEO01.5	L2
	OR			
15 (a)	Describe about Micro Electro Mechanical Systems (MEMS).	6M	20MEO01.5	L2
15 (b)	Explain the applications of nanotechnology in medicine.	6M	20MEO01.5	L2

NSRIT

SemesterEndRegularExamination,Nov./Dec.,2022

Degree	B.Tech.	**Program**	MechanicalEngineering			**AcademicYear**	2022-2023
CourseCode	20MEO01	**TestDuration**	3Hrs.	Max.Marks	70	**Semester**	V
Course	NanoTechnology(OpenElective)						

ANSWER KEY

PartA(ShortAnswerQuestions5x2=10Marks)

No.	Questions(1through5)	LearningOutcome(s)	DoK
1	Definethefollowingterms(i)NanoScience (ii)NanoTechnology. Nanoscience is a convergence of physics, materials science and biology, which deal with manipulation of materials at atomic and molecular scales; while nanotechnology is the ability to observe measure, manipulate, assemble, control, and manufacture matter at the nanometer scale. 2.The branch of technology that deals with dimensions and tolerances of less than 100 nanometres, especially the manipulation of individual atoms and molecules.	20MEO01.1	L1
2	DefineTop-downandbottom-upapproach. Each approach can be quite simple—the top-down approach goes from the general to the specific, and the bottom-up approach begins at the specific and moves to the general. These methods are possible approaches for a wide range of endeavors, such as goal setting, budgeting, and forecasting.	20MEO01.2	L1
3	DefineNanoformsofCarbon. Carbon nanostructures include various low-dimensional allotropes of carbon including carbon black (CB), carbon fiber, carbon nanotubes (CNTs), fullerene, and graphene. CNTs, and graphene have very unique properties. CNTs can be categorized into semiconducting or metallic according to their atomic structure.	20MEO01.3	L1
4	Whatdoyoumeanbysurfaceanalysistechnique? A surface analysis method is a technique for discovering the chemical structure of an extremely shallow and thin area called the surface number atomic layer of the solid matter. Surface analysis measurements provide a means to correlate performance with surface composition and structure. This knowledge can be used to accelerate the development of new materials or improve existing materials' performance.	20MEO01.4	L1
5	DefineNanoBiotechnology. Nanotechnology refers to the branch of science and engineering devoted to designing, producing, and using structures, devices, and systems by manipulating atoms and molecules at nanoscale, i.e. having one or more dimensions of the order of 100 nanometres (100 millionth of a millimetre) or less.	20MEO01.5	L1

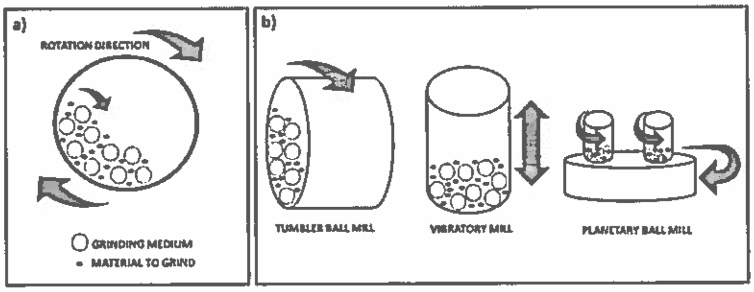
PartB(LongAnswerQuestions5x12=60Marks)

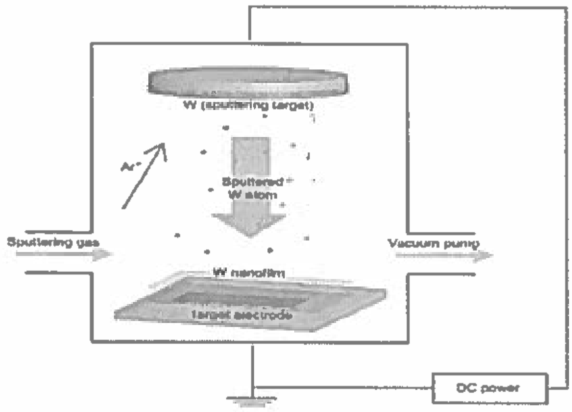
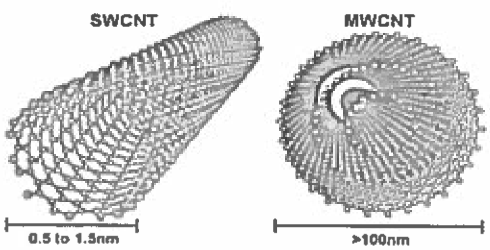
No.	Questions(6through15)	Marks	LearningOutcome(s)	DoK
6(a)	ExplaintheclassificationofNanoscaleparticles. Types of Nanomaterials For the purpose of this article, most current nanomaterials could be organized into four types: • Carbon Based Materials • Metal Based Materials • Dendrimers • Composites Carbon Based Materials	6M	20MEO01.1	L1

	These nanomaterials are composed mostly of carbon, most commonly taking the form of a hollow spheres, ellipsoids, or tubes. Spherical and ellipsoidal carbon nanomaterials are referred to as fullerenes, while cylindrical ones are called nanotubes. These particles have many potential applications, including improved films and coatings, stronger and lighter materials, and applications in electronics. Below we outline some examples of nanomaterials that are aimed at understanding their properties. As we will see, the behavior of some nanomaterials is well understood, whereas others present greater challenges. Nanoscale in One Dimension **Thin films, layers and surfaces** One-dimensional nanomaterials, such as thin films and engineered surfaces, have been developed and used for decades in fields such as electronic device manufacture, chemistry and engineering. In the silicon integrated-circuit industry, for example, many devices rely on thin films for their operation, and control of film thicknesses approaching the atomic level is routine. Monolayers (layers that are one atom or molecule deep) are also routinely made and used in chemistry. The most important example of this new class of materials is graphene. The formation and properties of these layers are reasonably well understood from the atomic level upwards, even in quite complex layers (such as lubricants) and nanocoatings. Advances are being made in the control of the composition and smoothness of surfaces, and the growth of films. Engineered surfaces with tailored properties such as large surface area or specific reactivity are used routinely in a range of applications such as in fuel cells and catalysts. The large surface area provided by nanoparticles, together with their ability to self assemble on a support surface, could be of use in all of these applications. Although they represent incremental developments, surfaces with enhanced properties should find applications throughout the chemicals and energy sectors. The benefits could surpass the obvious economic and resource savings achieved by higher activity and greater selectivity in reactors and separation processes, to enabling small-scale distributed processing (making chemicals as close as possible to the point of use). There is already a move in the chemical industry towards this. Another use could be the small-scale, on-site production of high value chemicals such as pharmaceuticals.			
6(b)	WhatisNanometerscale?Explain. A nanometre (nm) is 10^{-9}metres, which is one-thousandth of a micrometre, or one-billionth of a metre. This is the scale at which we measure atoms and the molecules they make Macro – anything that can be seen with the naked eye or anything greater than ~100 micrometer. Micro – 100 micrometers to 100 nanometers. Nano – 100 nanometers to 1 nanometer. Electrical and mechanical devices, components and systems are being manufactured in a variety of. sizes from macro to nano To put the size of a nanoparticle into perspective, compare it to a human hair. One strand of human hair is about 50 to 100 micrometers thick. One nanometer is 1/1000 of a micrometer. A nanoparticle is **100 nanometers thick** Examples of these include carbon nanotubes and carbon nanofibers. Nanomaterials with all three dimensions in the nanoscale are called nanoparticles. ... • Nanoparticles. • 2D materials. • Quantum dots. • Nanowires, nanofibres, nanorods. • Carbon nanotubes. • Biological nanomaterials. • Suspensions and dispersions. • Composites	6M	20MEO01.1	L2

OR

7(a)	WhataretheeffectsofNanomaterialsonMagneticandMechanicalproperties? Magnetic nanoparticles are nanomaterials consist of magnetic elements, such as iron, nickel, cobalt, chromium, manganese, gadolinium, and their chemical compounds. Magnetic nanoparticles are superparamagnetic because of their nanoscale size, offering great potentials in a variety of applications in their bare form or coated with a surface coating and functional groups chosen for specific uses. Especially, ferrite nanoparticles are the most explored magnetic nanoparticles, which can be greatly increased by clustering of a number of individual superparamagnetic nanoparticles into clusters to form magnetic beads. Magnetic nanoparticles can be selective attached to a functional molecules and allow transportation to a targeted location under an external magnetic field from an electromagnet or permanent magnet. In order to prevent aggregation and minimize the interaction of the particles with the system environment, surface coating may be required. The surface of ferrite nanoparticles is often modified by surfactants, silica, silicones, or phosphoric acid derivatives to increase their stability in solution. In general, coated magnetic nanoparticles have been widely used in several medical applications, such as cell isolation, immunoassay, diagnostic testing and drug delivery. **Magnetic Property** The properties of magnetic nanoparticles depend on the synthesis method and chemical structure. In most cases, the magnetic nanoparticles range from 1 to 100 nm in size and can display superparamagnetism. Superparamagnetism is caused by thermal effects that the thermal fluctuations are strong enough to spontaneously demagnetize a previously saturated assembly; therefore, these particles have zero coercivity and have no hysteresis. In this state, an external magnetic field is able to magnetize the nanoparticles with much larger magnetic susceptibility. When the field is removed, magnetic nanoparticles exhibit no magnetization. This property can be useful for controlled therapy and targeted drug delivery. **2. Magnetocaloric Effect** Some magnetic materials heat up when they are placed in a magnetic field and cool down when they are removed from a magnetic field, which is defined as the magnetocaloric effect (MCE). Magnetic nanoparticles provide a promising alternative to conventional bulk materials because of their particle size-dependent superparamagnetic features. In addition, the large surface area in magnetic nanoparticles has the potential to provide better heat exchange with the surrounding environment. By careful design of core-shell structures, it would be possible to control the heat exchange between the magnetic nanoparticles and the surrounding matrix, which provide a possible way for improving therapy technologies, such as hypertherm	6M	20MEO01.1	L1
7(b)	WhataretheimplicationsofNanoscienceforPhysicsandChemistry? Nanoscience is the study of small scale matter, the minuscule building blocks of the material and biological worlds. Typically nanoscientists study materials of less than 100 nanometres. 1 nanometre is one billionth of a metre. A human hair is about 50,000 - 100,000nm wide. Nanotechnologists are concerned with the behaviour of materials at these small dimensions and how they can be manipulated to do useful things. Nanotechnology is being used to develop smaller and more powerful electronic devices, lasers, medical diagnostics and materials with completely new properties. Nanoscience is contributing to product innovation in virtually every field of manufactured goods, enabling nearly $250 billion in products in 2008, on track to exceed $3 trillion globally by 2015. In Ireland we are well positioned to play a lead role in this worldwide social and economic revolution and are ranked 6th globally for the quality of the nanoscience research carried out in our universities, especially in TCD. Nanoscience does not belong fully to either Physics or Chemistry. Therefore, a new approach is required. An interdisciplinary degree programme marries part	6M	20MEO01.1	L2

	of both subjects so that students gain a deep and lasting understanding of the science of advanced materials that underpins the nano revolution. This is the key for the development of our Nanoscience, The ability to create new technologies or devices would not be possible without the use of **advanced materials**. Energy is an important issue for any new device, and making devices smaller approaching the nano-scale can reduce the energy cost, while increasing speed. These nanostructures or nanodevices can behave in surprising ways which are not like miniaturised versions of the macroscopic devices. Ultimately this behaviour is explicable by quantum mechanics, a branch of modern physics, but new methods of fabricating or interacting with such nanostructures is what nanoscience is all about, ideally to the benefit of technology and to people. Nanoscience incorporates applications in photonics, medical diagnostics, ultra-fast electronics and many other areas which in addition use advanced materials. Advanced materials include superconductors, polymers, lasers and optoelectronics and they can be found in applications			
8	ExplainBallmillingmethodunderTop-downapproachwiththehelpofneat sketch. Ball milling is an economic and facile technique to produce nanosized materials. It is a top-down approach of nanoparticle synthesis which includes mechanical breakdown of large substances into smaller one. It is used in producing metallic as well as ceramic nanomaterials Ball milling is a grinding method that grinds nanotubes into extremely fine powders. During the ball milling process, the collision between the tiny rigid balls in a concealed container will generate localized high pressure. Usually, ceramic, flint pebbles and stainless steel are used The top-down approach to management is a strategy in which the decision-making process occurs at the highest level and is then communicated to the rest of the team. This style can be applied at the project, team, or even the company level, and can be adjusted according to the particular group's needs. Ball milling is a mechanical technique widely used to grind powders into fine particles and blend materials. Being an environmentally-friendly, cost-effective technique, it has found wide application in industry all over the world 	12M	20MEO01.2	L2
	OR			
9(a)	Withthehelpofneatsketch,explainthePhysicalVapourDeposition(PVD)method. Physical vapour deposition (PVD) is a process used to produce a metal vapour that can be deposited on electrically conductive materials as a thin, highly adhered pure metal or alloy coating. The process is carried out in a vacuum chamber at high vacuum (10–6 torr) using a cathodic arc source. The two most common PVD processes are based on sputtering and evaporation. The former is one of the most commonly used PVD techniques in the textile industry. Different sputter deposition exist, such as DC, RF, magnetron	8M	20MEO01.2	L2

	 			
9(b)	DefineLithography. the process of printing from a plane surface (such as a smooth stone or metal plate) on which the image to be printed is ink-receptive and the blank area ink-repellent.. The definition of lithography is a method of printing from a flat surface where unnecessary ink is turned away from the surface, generally by grease. An example of lithography is printing a message on a stone using grease to repel unwanted ink.	4M	20MEO01.2	L1
10(a)	ExplainQuantumwires Quantum wires are extremely narrow structures where electron transport is possible only in a very few transverse modes (with energies less than the Fermi energy). Quantum wires can be used as electron waveguides. Semiconductor quantum wires have been used to make switchable high-speed lasers. Quantum structures such as quantum wells, quantum wires or quantum dots are characterized with very small concentration of electrons. A higher doping level is required to have any significant electron concentration if size of the nanostructure is reduced below the electron de Broglie wavelength. Nanowires are extensively used in nanoelectronic devices as connectors for the transportation of electrons. Cobalt, copper, silicon, and gold have been utilized to make nanowires. Chemical vapor deposition is used for the production of nanowires	4M	20MEO01.3	L2
10(b)	Explainthefollowing (i) SinglewallCarbonNanoTubes Single walled carbon nanotubes are an allotrope of sp2 ;hybridized carbon, similar to fullerenes. The structure can be thought of as a cylindrical tube comprised of 6-membered carbon rings, as in graphite. The cylindrical tubes may have one or both ends capped with a hemisphere of the buckyball or fullerene structure. Single-walled carbon nanotubes theoretically possess ultimate intrinsic tensile strengths in the **100–200 GPa range**, among the highest in existing material  	8M	20MEO01.3	L2

(ii) DoublewallCarbonNanoTubes

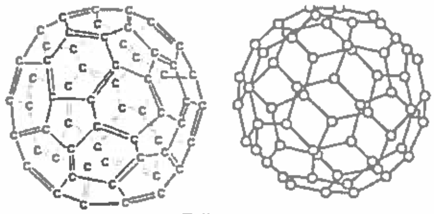
Double-walled carbon nanotubes are coaxial nanostructures composed of exactly two single-walled carbon nanotubes , one nested in another. This unique structure offers advantages and opportunities for extending our knowledge and application of the carbon nanomaterials family.

CNTs have two major structural forms: single-walled CNTs and multiwalled CNTs. Graphene is another two-dimensional carbon nanostructure, a one-atom-thick sheet comprising carbon atoms characterized by sp2 hybridization

OR

11	DescribetheBuckminsterfullerencewiththehelpofneatsketch.	12M	20MEO01.3	L2

The 1996 Nobel Prize in Chemistry was awarded to Richard Smalley, Robert Curl, and Harold Kroto for their discovery of a new allotrope of carbon, C60 called buckminsterfullerene.
Buckminsterfullerene (C60) is a spherical carbon allotrope where 60 atoms are assembled in pentagons and hexagons, in a geometry similar to a soccer ball. All the carbon atoms are connected by single and double bonds, these are often called BuckyBalls. Their cage structure and poly aromaticity cause the formation of a displaced electron cloud that allows these molecules to act as charge-transfer complexes.



Fullerene

The fullerene, C_{60}, consists of fused **five** and **six**-membered carbon rings. Each six membered rings is surrounded, alternately, by hexagons and pentagons of carbons; each pentagon is fused to five hexagons. The consequence of this structural motif is that each hexagon is like the base of a bowl; the three pentagons fused to this ring, linked by hexagons, force the structure to curve resulting in a dome-like structure that eventually curves around itself to give a structure resembling a sphere. The shape of fullerene,C_{60} resembles a **soccer ball**. All the 60 carbon atoms are equivalent and give rise to a single **13C NMR** resonance.

When an electric spark is struck between graphite electrodes, soot is produced. This soot is mainly carbon black but contains a significant amount of C_{60} carbon cluster compound, in this process smaller amounts of other fullerenes C_{32}, C_{50}, C_{70}, C_{76} and C_{84} may also be produced.
The following are easily extracted from the soot by dissolving them in benzene or hydrocarbon solvents, giving a red solution and finally mustard colour crystals. The different compounds are separated by chromatographically

12(a)	Withthehelpofneatsketch,explainScanningElectronMicroscopy(SEM).	6M	20MEO01.4	L2

The SEM is **an instrument that produces a largely magnified image by using electrons instead of light to form an image**. A beam of electrons is produced at the top of the microscope by an electron gun. The electron beam follows a vertical path through the microscope, which is held within a vacuum

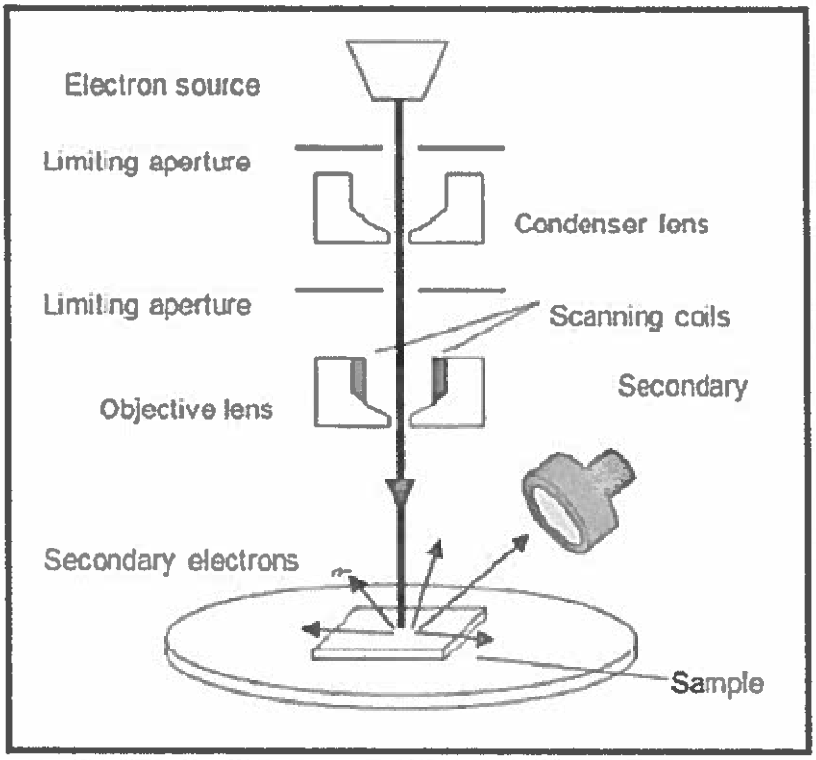
Scanning electron microscopy works by **scanning a sample with electron beams.** An electron gun fires these beams, which then accelerate down the column of the scanning electron microscope. During this action, the electron beams pass through a series of lenses and apertures, which act to focus it.

SEM imaging **occurs by scanning the sample with a high-energy beam of electrons**. When these electrons interact with the sample they create

secondary electrons, characteristic x-rays, and backscattered electrons. One or more detectors collect these signals and form images that can be seen on a computer screen.
Scanning electron microscope (SEM) is one of the most widely used instrumental methods for the **examination and analysis of micro- and nanoparticle imaging characterization of solid objects.** One of the reasons that SEM is preferred for particle size analysis is due to its resolution of 10 nm, that is, 100 Å.
A scanning electron microscope (SEM) is a **type of microscope which uses a focused beam of electrons to scan a surface of a sample to create a high resolution image.** SEM produces images that can show information on a material's surface composition and topography.

Components in a SEM

- Tungsten (W) electron filament.
- Lanthanum hexaboride (LaB_6) or Cerium hexaboride (CeB_6)
- Field Emission Gun (FEG)



12(b)	Explain X-ray diffraction technique.	6M	20MEO01.4	L2

X-ray diffraction analysis (XRD) is a technique used in materials science to determine the crystallographic structure of a material. XRD works by irradiating a material with incident X-rays and then measuring the intensities and scattering angles of the X-rays that leave the material

X-Ray diffraction analysis (XRD) is a nondestructive technique that provides detailed information about the crystallographic structure, chemical composition, and physical properties of a material [48]. It is based on the constructive interference of monochromatic X-rays and a crystalline sample

X-ray diffraction (XRD) helps to find the geometry or shape of a molecule using X-rays. The elastic scattering phenomenon of X-rays from the atoms of material has a long range order.

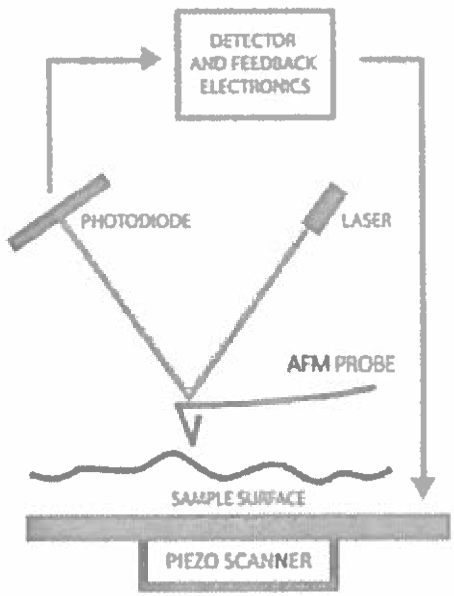
Diffraction is the bending of light around obstacles and can be explained using a concept known as Huygen's Principle. Huygen's Principle states that every wave crest consists of source points that produce wavelets, which propagate outward to form new wave fronts.

OR

13	Explainthefollowing (i) Atomicforcemicroscopy	12M	20MEO01.4	L2

Atomic Force Microscopy (AFM) is a high-resolution non-optical imaging technique first demonstrated by Binnig, Quate and Gerber in 1985 Since then it has developed into a powerful measurement tool for surface analysis. AFM allows accurate and non-destructive measurements of the topographical, electrical, magnetic, chemical, optical, mechanical, etc. properties of a sample surface with very high resolution [2] in air, liquids or ultrahigh vacuum. This unique combination of capabilities makes AFM indispensable in most advanced science and technology labs around the world

The basic operation principle of a standard AFM system with optical involves scanning an AFM probe with a sharp AFM tip over a sample surface in a raster pattern. The AFM tip is usually made of silicon or silicon nitride and is integrated near the free end of a flexible AFM cantilever. A piezoelectric ceramic scanner controls the lateral and the vertical position of the AFM probe relative to the surface. As the AFM tip moves over features of different height the deflection of the AFM cantilever changes. This deflection is tracked by a laser beam reflected from the back side of the AFM cantilever and directed into a position sensitive photodetector. A feedback loop controls the vertical extension of the scanner in order to maintain near-constant AFM cantilever deflection and hence a constant interaction force. The coordinates that the AFM tip tracks during the scan are combined to generate a three-dimensional topographic image of the surface.



(ii) ScanningTunnelingMicroscope

The scanning tunneling microscope (STM) is widely used in both industrial and fundamental research **to obtain atomic-scale images of metal surfaces**

Unlike SEM and TEM, **STM does not send electrons to a sample**. Rather, the tunneling microscope picks up electrons that escape from the surface of a sample by the process known as the quantum tunneling effect. Also, unlike SEM and TEM, the resolution of an STM does not depend on the wavelength of the electron.

The STM has ultra-high resolution and **can image single atoms**. This instrument allowed scientists to view a world that they could not view before: the world of the nanoscale. Many people believe that the invention of the STM was the birth of nanoscience.

	scanning tunneling microscope (STM), type of microscope whose principle of operation is based on the quantum mechanical phenomenon known as tunneling, in which the wavelike properties of electrons permit them to "tunnel" beyond the surface of a solid into regions of space that are forbidden to them under the rules of			
14	Writeashortnoteon (i) Nanocomputer Nanocomputer refers to a computer smaller than the microcomputer, which is smaller than the minicomputer. Microelectronic components that are at the core of all modern electronic devices employ semiconductor transistors Nanocomputing refers to the representation and manipulation of data by computers that are significantly smaller. A nanocomputer serves all the purposes of a modern personal computer, but the only difference is that it is very tiny in size. . (ii) Superchip advances in molecular medicine and cell biology also require new electrochemical systems to detect disease biomarkers and therapeutic compounds. Microelectronic technology offers powerful circuits and systems to develop innovative and miniaturized biochips for sensing at the molecular level. However, microelectronic biochips proposed in the literature often do not show the right specificity, sensitivity, and reliability required by biomedical applications. Nanotechnology offers new materials and solutions to improve the surface properties of sensing probes. The aim of the present paper is to review the most recent progress in Nano-Bio-Technology in the area of the development of new electrochemical systems for molecular detection in personalized therapy and cell culture monitoring (iii) Nanocrystal Nanocrystals are aggregates of molecules that can be combined into a crystalline form of the drug surrounded by a thin coating of surfactant. They have extensive uses in materials research, chemical engineering, and as quantum dots for biological imaging [26,27], but less so in nanomedicine for drug deliver (iv) Nanocrystals (v) Chitin. (vi) Chitosan. (vii)Nanofiber. (viii) Quantum Dot. (ix) Polymers. (x) Nanomaterials. (xi) Polysaccharides. (xii)Nanoparticles.	12M	20MEO01.5	L2
	OR			
15(a)	DescribeaboutMicroElectroMechanicalSystems(MEMS). Micro-electromechanical systems (MEMS) is a process technology used to create tiny integrated devices or systems that combine mechanical and electrical components. They are fabricated using integrated circuit (IC) batch processing techniques and can range in size from a few micrometers to millimetres. There are two types of MEMS accelerometers: variable capacitive and piezoresistive. Variable capacitives are highly sensitive and piezoresistive are low range devices used for acceleration measurement. The attractive features of MEMS, which have led to its commercial success,	6M	20MEO01.5	L2

include small size, light weight, cost effective batch fabrication, and precision control of physical dimensions.

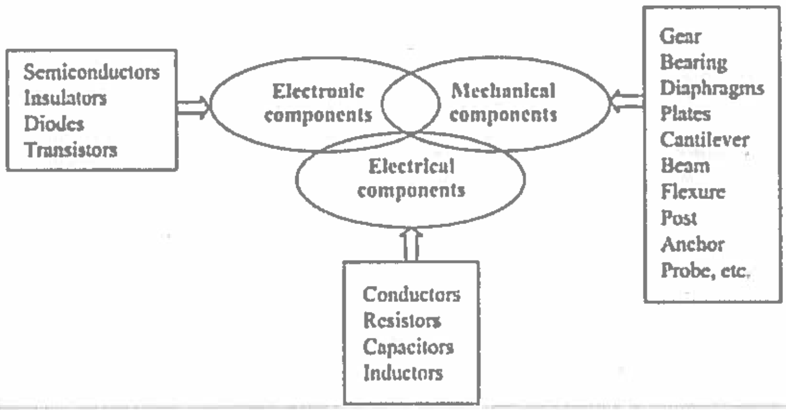
Current examples of MEMS devices include accelerometers for airbag sensors, microphones, projection display chips, blood and tire pressure sensors, optical switches, and analytical components such as lab-on-chip, biosensors and many other products.
.&
MEMS is a process technology used to create tiny integrated devices or systems that combine mechanical and electrical components. They are fabricated using integrated circuit (IC) batch processing techniques and can range in size from a few micrometers to millimetres.

A MEMS (micro-electromechanical system) is a miniature machine that has both mechanical and electronic components. The physical dimension of a MEMS can range from several millimeters to less than one micrometer, a dimension many times smaller than the width of a human hair

MEMS provides an integrated environment for engineers who design MEMS devices and integrate MEMS with systems and circuits. It is ideal for designing and optimizing MEMS devices that depend on electrostatics for sensing and actuation. It also supports piezo-electric sensors and actuators.

Several types of pressure sensor can be built using MEMS techniques. Here we will discuss two of the most common: piezoresistive and capacitive. In both of these, a flexible layer is created which acts as a diaphragm that deflects under pressure but different methods are used to measure the displacement

A MEMS (micro-electromechanical system) is a miniature machine that has both mechanical and electronic components. The physical dimension of a MEMS can range from several millimeters to less than one micrometer, a dimension many times smaller than the width of a human hair.



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15(b)	Explaintheapplicationsofnanotechnologyinmedicine.	6M	20MEO01.5	L2

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Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	**Program**	Common to All	**Academic Year**	2022 - 2023
Course Code	20ECO01	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	V
Course	**Architectures and Algorithms of IoT (Open Elective)**				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are things in IoT terminology?	20ECO01.1	L1
2	What is a protocol?	20ECO01.1	L1
3	Define a Duty Cycle	20ECO01.3	L1
4	What are events in IoT?	20ECO01.2	L1
5	List any two examples of IIoT	20ECO01.2	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 11)	Marks	Learning Outcome (s)	DoK
6	Explain four applications of IoT	12M	20ECO01.1	L2
	OR			
7	Explain the Architecture of IoT	12M	20ECO01.1	L2
8	Explain the protocol concept used in IoT design	12M	20ECO01.2	L2
	OR			
9	Classify and explain the different IoT oriented protocols	12M	20ECO01.2	L2
10	Write short notes with reference to IoT device spacing (a) Data Bases (b) Cost of Ownership (c) Power Consumption	12M	20ECO01.3	L2
	OR			
11	Explain the Cost per Transistor and Chip Size in IoT	12M	20ECO01.3	L2
12	Describe the IoT Event Analysis with an example	12M	20ECO01.4	L2
	OR			
13	Describe the IoT network model with devices, networks and hubs	12M	20ECO01.4	L2
14	Draw and explain Architecture of IIoT	12M	20ECO01.5	L2
	OR			
15	Explain the challenges and applications of IIoT	12M	20ECO01.5	L2

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(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

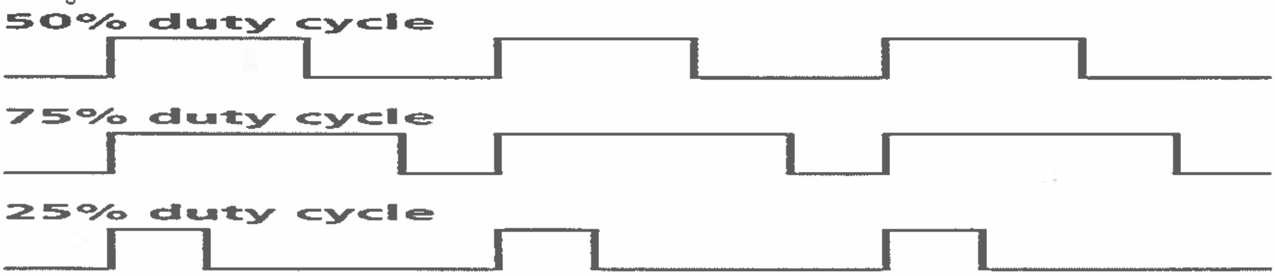
Part A

1. Things in IoT terminology are physical object, idea or action and situation or activity
   Ex: umbrella
   Which is not coming from him
   System
2. A protocol is **a set of rules and guidelines for communicating data** or a protocol is **a standardized set of rules for formatting and processing data.** Rules are defined for each step and process during communication between two or more computers.

3. Duty cycle is **the ratio of time a load or circuit is ON compared to the time the load or circuit is OFF**. Duty cycle, sometimes called "duty factor,".

Or

The duty cycle is **a measure of the peak-to-average ratio of the input signal over time.** The duty cycle can be expressed as a ratio or as a percentage.



4. **IoT** Events are event producer, event router and event consumer. **IoT Events are** device fleets for failures or changes in operation, and trigger actions when such **events** occur.
5. MAN. MAN is a Truck & Bus Company. ...
   Siemens. Siemens is a German multinational conglomerate company. ...
   Caterpillar (CAT). It is an American machinery and equipment firm.

Part B

6. Any Four applications of IoT
   1.Home automation: Maybe the most famous application of IoT is in Smart Homes. After all, who hasn't heard about connecting all the home applications like lighting, air conditioners, locks, thermostat, etc. into a single system that can be controlled from your smartphone. These IoT devices are becoming more and more popular these days because they allow you complete freedom to personalize your home as you want.
   2. Smart Cities: Cities can be made more efficient so that they require fewer resources and are more energy-efficient. This can be done with a combination of sensors in different capacities all over the city that can be used for various tasks ranging from managing the traffic, controlling handling waste management, creating smart buildings, optimizing streetlights, etc. There are many cities in the world that are working on incorporating IoT and becoming smarter
   3. Smart Agriculture: Food is an integral part of life without which we cannot survive. One way to feed everyone is better agricultural practices which can be enhanced using IoT. This can be done by first collecting data for a farm such as soil quality, sunlight levels, seed type, rainfall density from various sources like farm sensors, satellites, local weather stations, etc. and then

using this data with Machine Learning and IoT to create custom recommendations for each farm that will optimize the planting procedure, irrigation levels required, fertilizer amount, etc

4.Environment

5.Energy

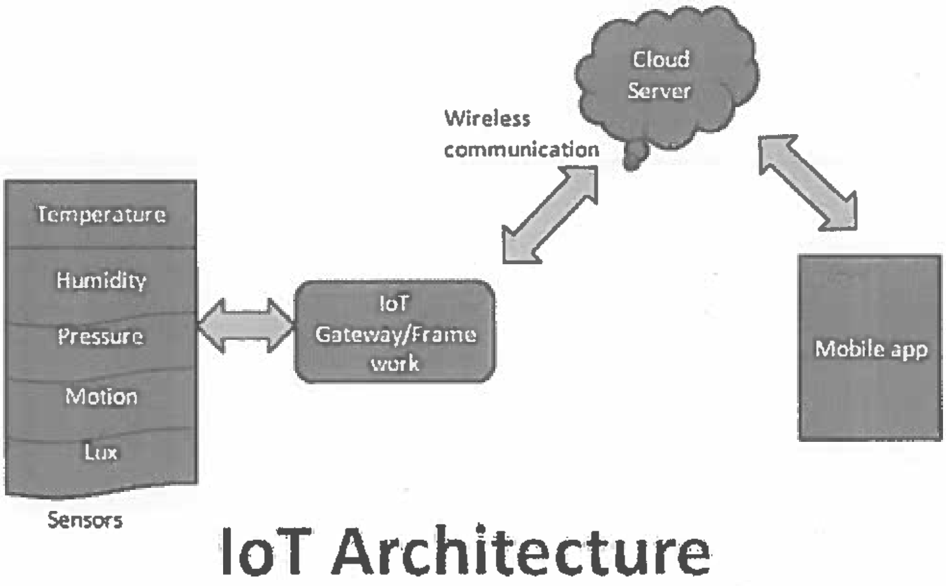
6. Retail: Retail stores can make use of IoT in a wide range of operations to make shopping a much smoother experience for customers and also easier for the employees. IoT can be used to handle the inventory, improve store operations, reduce shoplifting and theft, and prevent long queues at the cashiers.

7.Logistics

8.Industry

9.Health

7.



IoT architecture

The above figure is self-explanatory about IoT Architecture. Internet of things is based on four simple building blocks also called IoT architecture layers 1) Sensors 2) Internet of Things (IoT) framework & gateway 3) cloud server 4) mobile app

Sensors:

Sensors are everywhere, sensors sense data from the atmosphere or place. the eg. temperature sensor senses temperature from the room and shares it through IoT gateway.Sensors will sniff a wide variety of information ranging from Location, Weather/Environment conditions, running machine, from the human body, engine maintenance data to health essentials of a vehicle.

IoT Gateways & frameworks:

It is a gateway to the internet for all the things/devices that we want to interact with. Gateways act as a carrier between the internal network of sensor nodes with the external Internet or World Wide Web.They do this by collecting the data from sensor nodes and transmitting them to the internet infrastructure.

Cloud server:

The data transmitted through the gateway is stored & processed securely within the cloud server i.e. in data centres. This processed data is then used to perform intelligent actions that make all our devices Smart Devices. In the cloud, all analytics and decision making happen considering user comfort.

Mobile app:

The intuitive mobile apps will help end-users to control & monitor their devices (ranging from the room thermostat to vehicle engines) from remote locations.These apps push important information from the cloud on your smartphones, tablets.After analytics, Information is in the form of graphs, bars and in pi-diagram and display to the user in a user-friendly manner.From mobile application, we can send a command to sensors to change default values, like changing default temperature of air conditioner and many more.

8. A protocol is a set of rules and guidelines for communicating data. Rules are defined for each step and process during communication between two or more computers. Networks have to follow these rules to successfully transmit data.

Similar to programming languages, protocols are based on specific rules and regulations for computing and are designed for efficiency. Each rule is defined in different terms and is assigned a unique name. Protocols specify the standards for communication and provide detailed information on processes involved in data transmission. Such processes include:

- Type of task
- Process nature
- Data flow rate
- Data type
- Device management

A single process can be handled by more than one protocol simultaneously. This coordination of protocols creates a protocol family.

9. IoT protocols are a crucial part of the IoT technology stack — without them, hardware would be rendered useless as the IoT protocols enable it to exchange data in a structured and meaningful way. Out of these transferred pieces of data, useful information can be extracted for the end user and thanks to it, the whole deployment becomes economically profitable, especially in terms of IoT device management.

Constrained Application Protocol (CoAP)

While the existing Internet infrastructure is freely available and usable for any IoT device, it often proves too heavy and power-consuming for most IoT use cases. Created by the IETF Constrained RESTful Environments working group and launched in 2013, Constrained Application Protocol (CoAP) was designed to translate the HTTP model so that it could be used in restrictive device and network environments.

Message Queuing Telemetry Transport (MQTT)

Probably the most widely adopted standard in the Industrial Internet of Things to date, Message Queuing Telemetry Transport is a lightweight publication/subscription type (pub/sub) messaging protocol. Designed for battery-powered devices, MQTT's architecture is simple and lightweight, providing low power consumption for devices. Working on top of TCP/IP protocol, it has been especially designed for unreliable communication networks in order to respond to the problem of the growing number of small-sized cheap low-power objects that have made their appearance in the network in the recent years.

Wi-Fi

Creating a Wi-Fi network requires devices that can send wireless signals which means devices such as telephones, computers or routers, to name a few. At home, a router is used to transfer the internet connection from a public network to a private home or office network. Wi-Fi provides an Internet connection to nearby devices that are within a certain range. Another way to use Wi-Fi is to create a Wi-Fi hotspot, i.e. telephones or computers may share a wireless or wired internet connection with other devices by broadcasting a signal.

ZigBee

ZigBee-based networks are characterized by low power consumption, low throughputs (up to 250 kbps) and connectivity range of 100 meters between nodes. Typical applications include sensor networks, personal networks (WPAN), home automation, alarm systems and monitoring systems.

Bluetooth

Bluetooth is a technology that allows wireless connection of various electronic devices, such as a telephone, keyboard, computer, laptop, mouse, palmtop, printer, headset or speakerphone, and more. If you're down for a more wiki-like definition, this is an open standard described in the IEEE 802.15.1 specification and its technical specification includes three classes of ERP 1-3 transmission power with a range of, respectively, 100, 10 and 1 meter in open space. The most common class is the second one (10m) which allows you to connect devices that are in different rooms and even on different floors.

Extensible Messaging and Presence Protocol (XMPP)

Developed in 1999 by the Jabber open source community and originally meant for real-time messaging, this communication IoT protocol for message-oriented middleware is based on the XML language. It allows for real-time exchange of structured but extensible data between two or more network clients.
Since its inception, XMPP has been widely applied as a communications protocol. Over time and with the emergence of a lightweight XMPP specification: XMPP-IoT, it has gone on to be used in the context of the Internet of Things. Being an open community supported standard, XMPP IoT's strengths are addressing and scalability capabilities, which makes it perfect for consumer-oriented IoT deployments.

10. Database

Data base systems store the data transmitted from different IoT devices. Database systems types are 1.

Hot data base 2.cold data base

1.Hot databases

These are typically used for data that is frequently being queried or updated. They are often a good choice for storing data as they provide read and write capabilities with little latency at the lowest cost.

2.Cold databases

They store information in their original state with little to no changes made thereafter. In contrast with real-time data collection, storing huge volumes of historical data can be a difficult task on cold databases

Cost of ownership: Cost of ownership is also called as Total cost of ownership. Total cost of ownership is initial purchase cost of product plus cost of the purchase the product life cycle. (TCO) is an estimation of the expenses associated with purchasing, deploying, using and retiring a product or piece of equipment.TCO, or actual cost, quantifies the cost of the purchase across the product's entire lifecycle. Therefore, it offers a more accurate basis for determining the value -- cost vs. return on investment (ROI) -- of an investment than the purchase price alone.

Power consumption: There three ways through which energy is being consumption by IoT devices.

1 Data centers Data collection IoT through the sensors, surveillance, machine and appliances. All the data collected by these mediums is stored and analyzed in data centers due that consumes mega amount of energy.

2.Emboied Energy: Data need much more power in manufacturing process. New technology consists of billions of sensor nodes, microchips and count less semiconductor chips. For this inordinate amount of power requirement for IoT products

11. One of the foremost things about new technology is the hype created around it. In most cases, the initial hype regarding a new technology does not match with the content in it. The same has happened in the case of IoT due to the concentrated focus on new devices rather. Therefore, there has been no attention to the interaction between devices, applications, and data or the possibilities for integration of all the devices and systems. At the same time, devices will also play an important role in expanding the future growth of IoT. This is where you would find Moore's law IoT relationship. The increasing number of IoT devices and the introduction of new device form factors could work as a trigger for organizations to work on the management of data flow between IoT devices and applications. The new IoT landscape should not focus on presenting information to users from a back-end or cloud system.

Moore's Law Formula

The next important thing to develop a basic understanding of the law refers to Moore's law formula. Since you are looking for a law, it is reasonable to seek a quantified representation of the law. It is important to know that Moore did not introduce a specific equation or formula for his concept. Therefore, you can end up with confusion while trying to find out 'what is Moore's law formula?' to understand it. The simplest formula for explaining the law is as follows.

Processing Power in Future = Existing Processing Power. 2n

Here 'n' represents the number of years required for developing a new microprocessor, divided by two. Let us take an example to understand the law better. The Intel 8008 had around 3500 transistors in 1972. After 10 years, how many transistors can you expect in the microprocessor? In this case, the time difference is 10 years, thereby implying that 'n' will be 5. Therefore, you can use the formula as follows,

Processing Power in Future = 3,500. 25

12. An event-driven architecture uses events to trigger and communicate between decoupled services and is common in modern applications built with micro services. An event is a change in state, or an update, like an item being placed in a shopping cart on an e-commerce website. Events can either carry the state (the item purchased, its price, and a delivery address) or events can be identifiers (a notification that an order was shipped).

Event-driven architectures have three key components: event producers, event routers, and event consumers. A producer publishes an event to the router, which filters and pushes the events to consumers. Producer services and consumer services are decoupled, which allows them to be scaled, updated, and deployed independently.

Benefits of an event-driven architecture

Scale and fail independently

By decoupling your services, they are only aware of the event router, not each other. This means that your services are interoperable, but if one service has a failure, the rest will keep running. The event router acts as an elastic buffer that will accommodate surges in workloads.

Develop with agility

You no longer need to write custom code to poll, filter, and route events; the event router will automatically filter and push events to consumers. The router also removes the need for heavy coordination between producer and consumer services, speeding up your development process.

Audit with ease

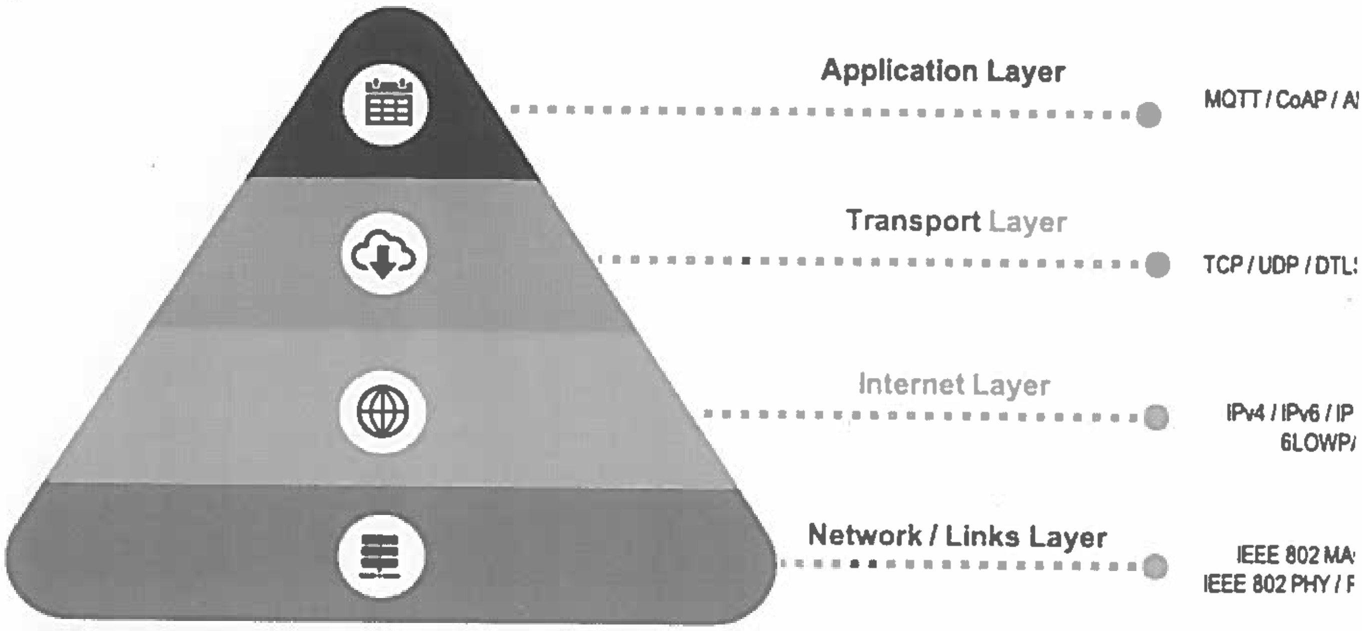
An event router acts as a centralized location to audit your application and define policies. These policies can restrict who can publish and subscribe to a router and control which users and resources have permission to access your data. You can also encrypt your events both in transit and at rest.

Cut costs

Event-driven architectures are push-based, so everything happens on-demand as the event presents itself in the router. This way, you're not paying for continuous polling to check for an event. This means less network bandwidth consumption, less CPU utilization, less idle fleet capacity, and less SSL/TLS handshakes.

13.IoT Network Model

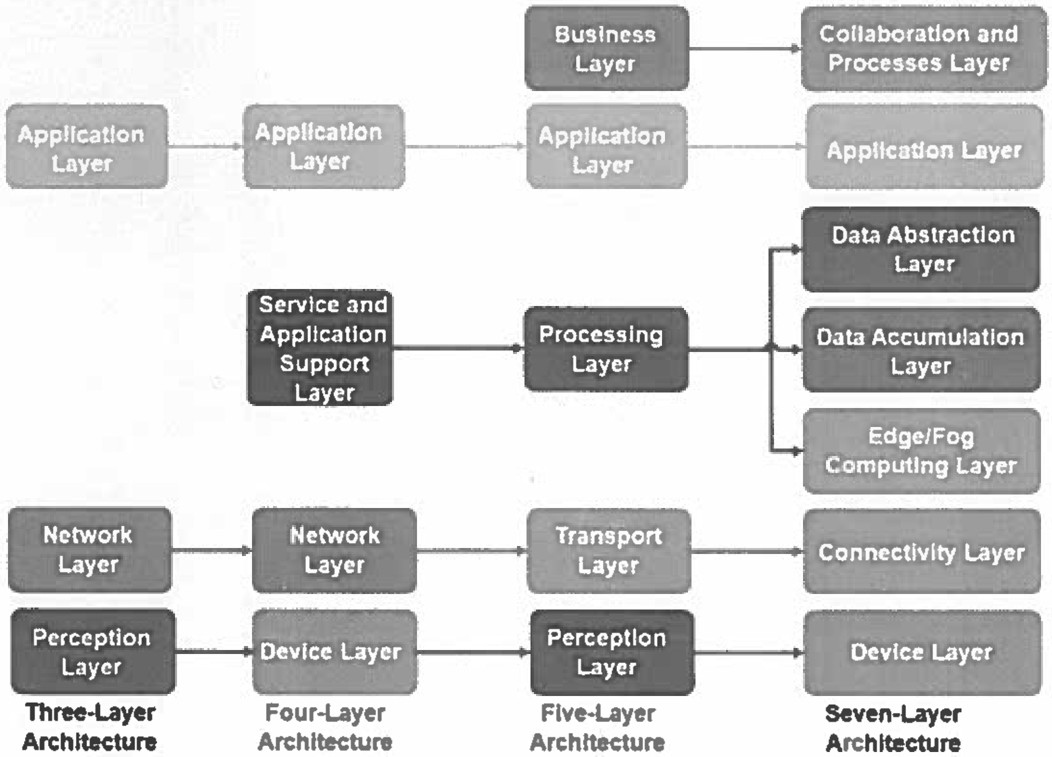
IoT Network refers to the communication technologies used by Internet of Things (IoT) devices to share or spread the data to other device or interfaces available within reachable distance. There are various types of IoT networks available for IoT devices / IoT sensors to communicate. It is critical to choose proper networking protocol for given requirements in order to be able to collect data at real-time and access insights through IoT applications.



The first links layer is in line with the industry standards like that of IEEE 802 MAC and IEEE 802 PHY dealing with local and metropolitan area networks. It is restricted to short data transmission of uniformly sized cells. The next layer is the internet layer (IPv4/IPv6/IP Routing) that is internet-ready connected devices / systems that communicate within internet connected domains with the help of a device unique identification. Following the internet layer is the transport layer consisting of TCP/UDP/DTLS/HTTP over wire helps communicate between systems as part of transportation principles and protocols. At the apex is the application layer which accommodates industry standard approach like MQTT, CoAP, API for application communication between devices / systems.

Types of IoT Networks: RFID,LPWAN,Wi-Fi,Cellular and BLE

14. IIoT architecture is **defined by layers of interaction, including device, communications and semantic**. In the device layer, architects design how devices interact with communications systems to connect and interconnect in a structure. In the communications layer, systems use protocols to exchange actionable information.



the initial widely accepted IIoT architecture is constructed based on the three-layer architecture , namely the perception, network, and application layers. The perception layer consists of the "things" identification and sensing technologies to collect and exchange the data. The network layer enables the communication and data transmission between the perception layer and the application layer. In most cases, it also involves data aggregation and curation process. Lastly, the application layer confluxes the data aggregated and virtualises the analysed result based on society, business and government demands. Different business interests reflect various IIoT applications for this layer, such as smart cities, intelligent health and smart transport.

As three-layer architecture confronted the interoperability and scalability problem to well-suit into existing Internet and telecommunication networks, Wu et al. extended the three-layer architecture into five-layer architecture by proposing a new business layer that resides on the top of the application layer and further dividing the previous network layer into processing layer and transport layer. The transport layer is responsible for transmitting the data generated from the perception layer into the processing layer.

15. IIoT applications

Remote Monitoring

Radar-level sensors provide local displays so that operators can easily manage levels through a singular dashboard. These systems make for easy measuring points on moving and rotating machinery, so operators are constantly fed real-time data regarding the equipment's functionality. This, in turn, gives insights into overall equipment life-cycles and repair needs, allowing for predictive maintenance.

Predictive Maintenance

In the power industry, drones with equipment monitors and sensors are being used to monitor power line networks and evaluate risks. These drones can anticipate scenarios such as estimating when a tree is likely to fall on a line, resulting in costly maintenance and repair. That way, companies will be informed before the damage is done. In this fashion, predictive maintenance enables cost-effective repairs and intervention before the damage is even done.

Automation

One vital use of automation via IIoT is smart irrigation in industrial farming. Water is a precious resource, but farmers typically have to keep a consistent watering schedule to ensure proper plant care. Smart irrigation systems, however, are automating this process while conserving water. The IIoT device reads moisture levels in the soil and reports to the sprinkler system when water is needed. This way, water, money, and time are all saved.

These three categories of IIoT implementation give industries unprecedented precision and efficiency with the right application.

The Challenges of IIoT Implementation

Only now are many realizing some of the broader challenges of managing an industrial IoT network complete with a repertoire of useful devices. Like with any networked device, IIoT components are open to cyber security risks. Meanwhile, the application of these devices to fill their potential requires pre-planning and assessment.

When implementing your IIoT system, consider these common challenges to wide-spread IoT success:

- Failure to align KPIs with clear business objectives
- Improper organizational alignment
- Lack of IoT experience
- IoT security threats

The dangers of not taking these challenges seriously can have more than just monetary risks. An autonomous machine can compromise employee safety like a vehicle that has been hacked or infected with malware. Overcoming many of these challenges requires running a cyber security risk assessment at consistent intervals during the IIoT device's lifecycle and training staff around the proper implementation.

9/12/22

9/12/22

Semester End Regular Examination, Nov./Dec., 2022

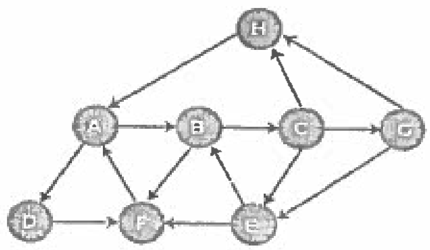
Degree	B. Tech.	**Program**	Common to All		**Academic Year**	2022 - 2023
Course Code	20CSO01	**Test Duration**	3 Hrs. Max. Marks	70	**Semester**	V
Course	**Data Structures and Algorithms (Open Elective)**					

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the different representations of graphs?	20CSO01.2	L1
2	What are the various characteristics of an algorithm?	20CSO01.2	L1
3	What are the applications of an array?	20CSO02.5	L1
4	Define divide and conquer technique with example.	20CSO02.1	L1
5	Define performance analysis of an algorithm.	20CSO02.1	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

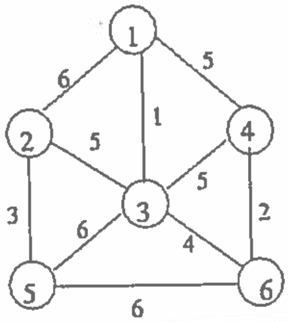
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What is an algorithm? Write an algorithm for calculating the average of 5 numbers.	8M	20CSO01.1	L2
6 (b)	Explain the Big Oh, Omega and Theta notations.	4M	20CSO01.1	L2
	OR			
7 (a)	Explain about types of arrays and its representations.	6M	20CSO01.1	L2
7 (b)	Explain various operations on data structures.	6M	20CSO01.1	L2
8 (a)	Define stack and write an algorithm to implement stack using array.	8M	20CSO01.2	L2
8 (b)	Explain the differences between stack and a queue.	4M	20CSO01.2	L2
	OR			
9 (a)	Explain about Singly linked list operations and write its algorithms.	8M	20CSO01.2	L2
9 (b)	List out any 4 applications of linked lists.	4M	20CSO01.2	L1
10 (a)	What is a binary tree? Construct a binary tree for given array of elements and write in-order, pre order and post order traversals along with algorithm.	8M	20CSO01.3	L2
10 (b)	Explain Binary tree, Full Binary tree, complete binary tree and balanced binary tree with examples.	4M	20CSO01.3	L2
	OR			
11 (a)	Write an algorithm for construction of DFS.	4M	20CSO01.3	L2
11 (b)	Construct Depth First Search Path from the following. (see graph and adjacency lists below)	8M	20CSO01.3	L2



Adjacency Lists

A : B, D
B : C, F
C : E, G, H
G : E, H
E : B, F
F : A
D : F
H : A

No.	Questions	Marks	Learning Outcome (s)	DoK
12 (a)	Write an algorithm for Bubble sort.	6M	20CSO01.4	L2
12 (b)	Sort the following elements using Bubble sort. 14, 26, 15, 55, 75, 50, 60, 30.	6M	20CSO01.4	L2
	OR			
13 (a)	Write an algorithm for Merge Sort.	6M	20CSO01.4	L2
13 (b)	Search the element 10 from the given list using linear search technique 32, 14, 65, 67, 23, 36, 6, 100.	6M	20CSO01.4	L2
14 (a)	Write prims algorithm for finding minimum spanning tree.	4M	20CSO01.5	L2

14 (b)	Construct minimum spanning tree using Prim's algorithm from the following graph.  	8M	20CSO01.5	L2
	OR			
15 (a)	Explain All pairs shortest path algorithm with example.	6M	20CS502.5	L2
15 (b)	Find the optimal solution for the fractional knapsack problem making use of greedy approach. Consider- n = 5, w = 60 kg (w1, w2, w3, w4, w5) = (5, 10, 15, 22, 25) (b1, b2, b3, b4, b5) = (30, 40, 45, 77, 90)	6M	20CS502.5	L2

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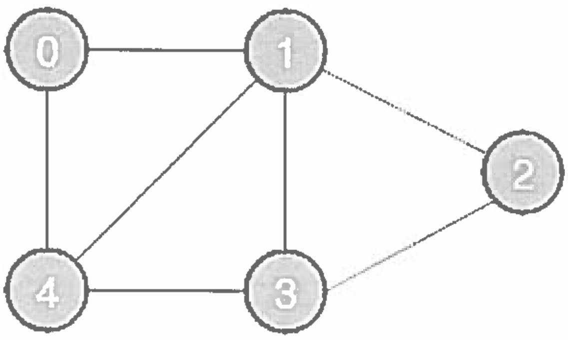
Key for Data Structures and Algorithms(Open Elective)

Short Answers

1.Different representations of graphs.

A graph is a Non-linear data structure that consists of the following two components:
1. A finite set of vertices also called as nodes.
2. A finite set of ordered pair of the form (u, v) called as edge. The pair is ordered because (u, v) is not the same as (v, u) in case of a directed graph(di-graph). The pair of the form (u, v) indicates that there is an edge from vertex u to vertex v. The edges may contain weight/value/cost.
Graphs are used to represent many real-life applications: Graphs are used to represent networks. The networks may include paths in a city or telephone network or circuit network. Graphs are also used in social networks like linkedIn, Facebook. For example, in Facebook, each person is represented with a vertex(or node). Each node is a structure and contains information like person id, name, gender, and locale. See this for more applications of graph.
Following is an example of an undirected graph with 5 vertices.



- The following two are the most commonly used representations of a graph.
  1. Adjacency Matrix
  2. Adjacency List
  There are other representations also like, Incidence Matrix and Incidence List. The choice of graph representation is situation-specific. It totally depends on the type of operations to be performed and ease of use.
  Adjacency Matrix:
  Adjacency Matrix is a 2D array of size V x V where V is the number of vertices in a graph. Let the 2D array be adj[][], a slot adj[i][j] = 1 indicates that there is an edge from vertex i to vertex j. Adjacency matrix for undirected graph is always symmetric.

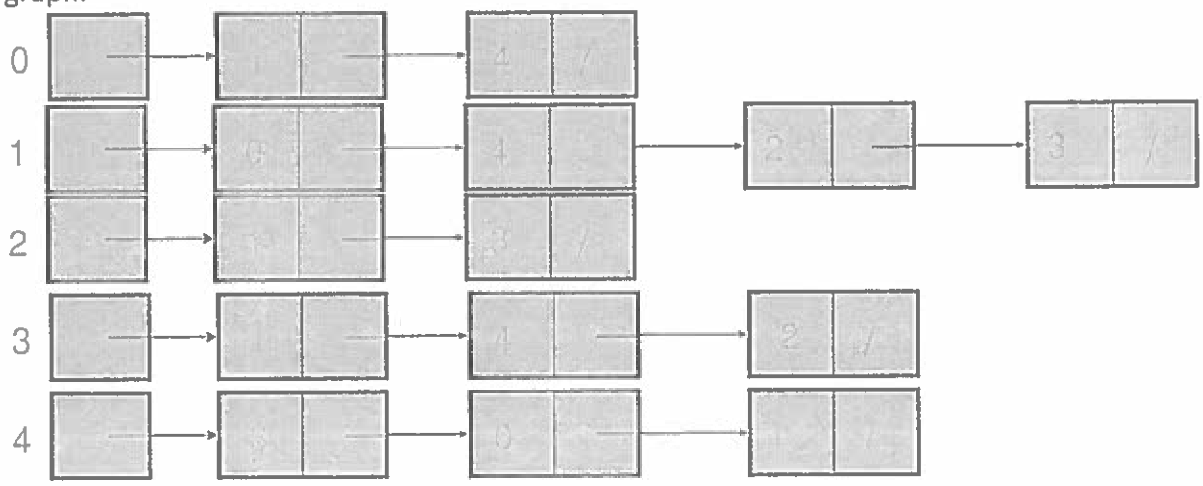
Adjacency Matrix is also used to represent weighted graphs. If adj[i][j] = w, then there is an edge from vertex i to vertex j with weight w.

- The adjacency matrix for the above example graph is:

$$
\begin{array}{c|ccccc|}
 & 0 & 1 & 2 & 3 & 4 \\
\hline
0 & 0 & 1 & 0 & 0 & 1 \\
1 & 1 & 0 & 1 & 1 & 1 \\
2 & 0 & 1 & 0 & 1 & 0 \\
3 & 0 & 1 & 1 & 0 & 1 \\
4 & 1 & 1 & 0 & 1 & 0 \\
\hline
\end{array}
$$

Adjacency List:

An array of lists is used. The size of the array is equal to the number of vertices. Let the array be an array[]. An entry array[i] represents the list of vertices adjacent to the *i*th vertex. This representation can also be used to represent a weighted graph. The weights of edges can be represented as lists of pairs. Following is the adjacency list representation of the above graph.



2. Incidence Matrix:

otal number of vertices by total number of edges.

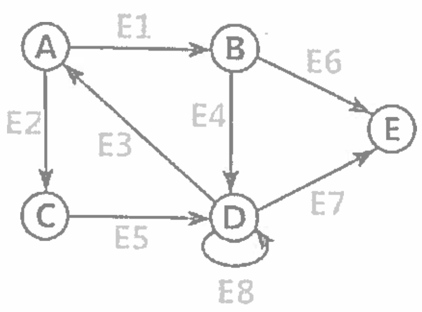
It means if a graph has 4 vertices and 6 edges, then it can be represented using a matrix of 4X6 class. In this matrix, columns represent edges and rows represent vertices.

This matrix is filled with either **0 or 1** or -1. Where,

- 0 is used to represent row edge which is not connected to column vertex.
- 1 is used to represent row edge which is connected as outgoing edge to column vertex.
- -1 is used to represent row edge which is connected as incoming edge to column vertex.

Example

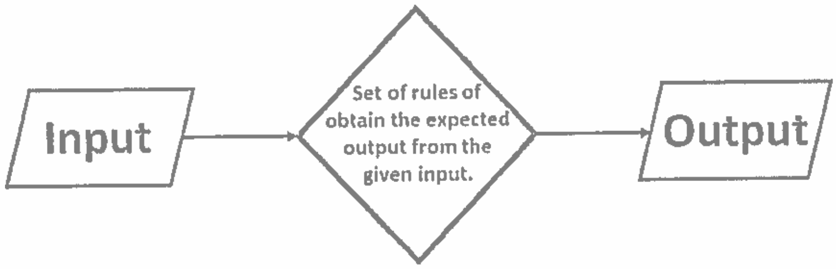
Consider the following directed graph representation.



$$
\begin{array}{c}
\\ A \\ B \\ C \\ D \\ E
\end{array}
\begin{array}{c}
\begin{array}{cccccccc} E1 & E2 & E3 & E4 & E5 & E6 & E7 & E8 \end{array} \\
\begin{pmatrix}
1 & 1 & -1 & 0 & 0 & 0 & 0 & 0 \\
-1 & 0 & 0 & 1 & 0 & 1 & 0 & 0 \\
0 & -1 & 0 & 0 & 1 & 0 & 0 & 0 \\
0 & 0 & 1 & -1 & -1 & 0 & 1 & 1 \\
0 & 0 & 0 & 0 & 0 & -1 & -1 & 0
\end{pmatrix}
\end{array}
$$

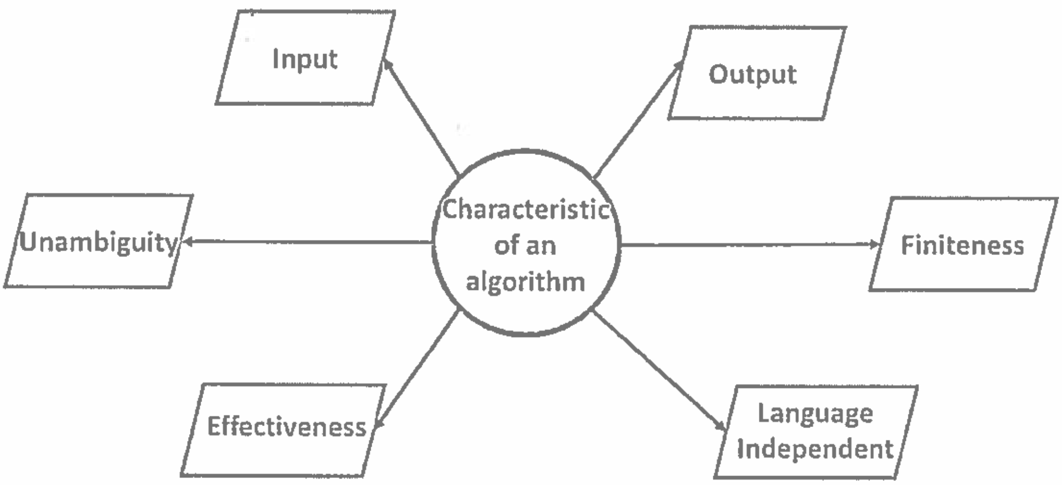
2.Charactaristics of an algorithm:

- An algorithm is a set of commands that must be followed for a computer to perform calculations or other problem-solving operations.
- According to its formal definition, an algorithm is a finite set of instructions carried out in a specific order to perform a particular task.
- It is not the entire program or code; it is simple logic to a problem represented as an informal description in the form of a flowchart or pseudocode.



Algorithm

- Problem: A problem can be defined as a real-world problem or real-world instance problem for which you need to develop a program or set of instructions. An algorithm is a set of instructions.
- Algorithm: An algorithm is defined as a step-by-step process that will be designed for a problem.
- Input: After designing an algorithm, the algorithm is given the necessary and desired inputs.
- Processing unit: The input will be passed to the processing unit, producing the desired output.
- Output: The outcome or result of the program is referred to as the output.
  After defining what an algorithm is, you will now look at algorithm characteristics.
  Characteristics of an Algorithm
  An algorithm has the following characteristics:



- Input: An algorithm requires some input values. An algorithm can be given a value other than 0 as input.
- Output: At the end of an algorithm, you will have one or more outcomes.
- Unambiguity: A perfect algorithm is defined as unambiguous, which means that its instructions should be clear and straightforward.
- Finiteness: An algorithm must be finite. Finiteness in this context means that the algorithm should have a limited number of instructions, i.e., the instructions should be countable.
- Effectiveness: Because each instruction in an algorithm affects the overall process, it should be adequate.
- Language independence: An algorithm must be language-independent, which means that its instructions can be implemented in any language and produce the same results.

3.Applications of Arrays

- Arrays are used to implement vectors and lists which are an important part of C++ STL.
- Arrays are also used to implement stack and queues.
- Trees also use array implementation whenever possible as arrays are easy to handle compared to pointers. Tress, in turn, are used to implement various other types of data structures.
- Matrices which are an important part of the mathematical library in any programming languages is implemented using arrays.
- Adjacency list implementation of graph uses vectors which are again implemented using arrays.
- Data structures like a heap, map, and set use binary search tree and balanced binary trees which uses can be implemented using arrays.

- Arrays are used to maintain multiple variables with the same name..
- CPU scheduling algorithms use implemented using arrays.
- All sorting algorithms use arrays at its core.

4.Divide and conquer technique with an example

Divide and Conquer is an algorithmic paradigm in which the problem is solved using the Divide, Conquer, and Combine strategy.

A typical Divide and Conquer algorithm solves a problem using following three steps:

1. **Divide:** This involves dividing the problem into smaller sub-problems.
2. **Conquer:** Solve sub-problems by calling recursively until solved.
3. **Combine:** Combine the sub-problems to get the final solution of the whole problem.

Standard algorithms that follow Divide and Conquer algorithm

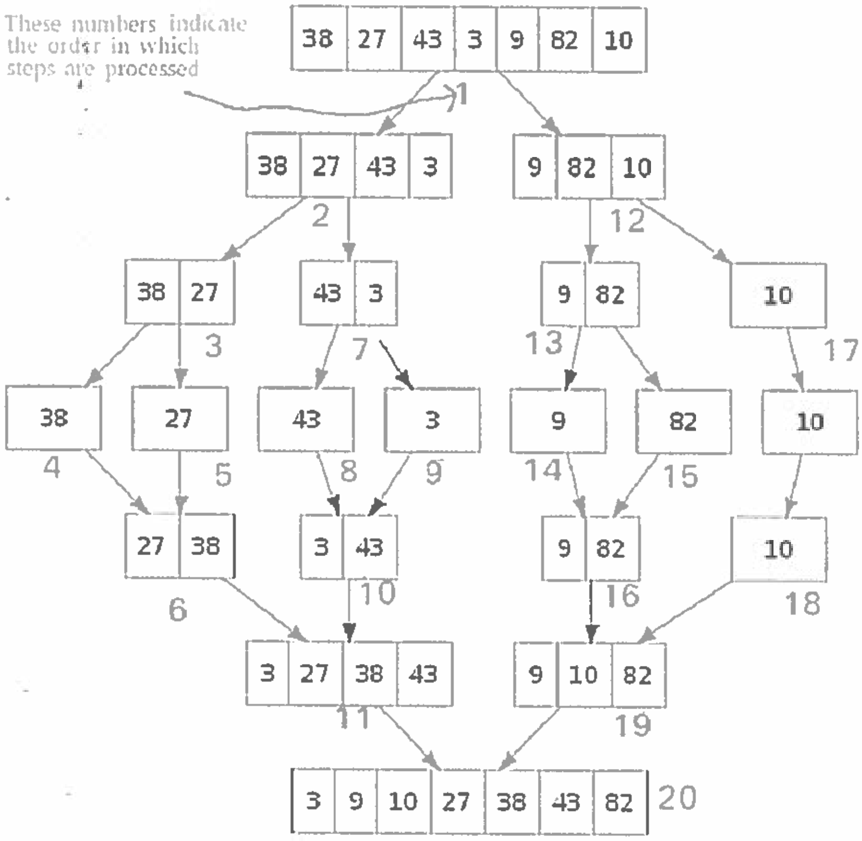
The following are some standard algorithms that follow Divide and Conquer algorithm.

1. **Quicksort** is a sorting algorithm. The algorithm picks a pivot element and rearranges the array elements so that all elements smaller than the picked pivot element move to the left side of the pivot, and all greater elements move to the right side. Finally, the algorithm recursively sorts the subarrays on the left and right of the pivot element.
2. **Merge Sort** is also a sorting algorithm. The algorithm divides the array into two halves, recursively sorts them, and finally merges the two sorted halves.
3. **Closest Pair of Points** The problem is to find the closest pair of points in a set of points in the x-y plane. The problem can be solved in O(n^2) time by calculating the distances of every pair of points and comparing the distances to find the minimum. The Divide and Conquer algorithm solves the problem in O(N log N) time.
4. **Strassen's Algorithm** is an efficient algorithm to multiply two matrices. A simple method to multiply two matrices needs 3 nested loops and is O(n^3). Strassen's algorithm multiplies two matrices in O(n^2.8974) time.
5. **Cooley–Tukey Fast Fourier Transform (FFT) algorithm** is the most common algorithm for FFT. It is a divide and conquer algorithm which works in O(N log N) time.
6. **Karatsuba algorithm for fast multiplication** does the multiplication of two *n*-digit numbers in at most single-digit multiplications in general (and exactly when *n* is a power of 2). It is, therefore, faster than the classical algorithm, which requires n^2 single-digit products. If $n = 2^{10} = 1024$, in particular, the exact counts are 3^{10} = 59, 049 and $(2^{10})^2$ = 1, 048, 576, respectively.

Example of Divide and Conquer algorithm

A classic example of Divide and Conquer is Merge Sort demonstrated below. In Merge Sort, we divide array into two halves, sort the two halves recursively, and then merge the sorted halves.

7.



5.Perform analysis of algoriths.

Analyzing the performance of an algorithm is an important part of its design. One of the ways to estimate the performance of an algorithm is to analyze its complexity.

Complexity theory is the study of how complicated algorithms are. To be useful, any algorithm should have three key features:

- It should be correct. An algorithm won't do you much good if it doesn't give you the right answers.
- A good algorithm should be understandable. The best algorithm in the world won't do you any good if it's too complicated for you to implement on a computer.
- A good algorithm should be efficient. Even if an algorithm produces a correct result, it won't help you much if it takes a thousand years or if it requires 1 billion terabytes of memory.

There are two possible types of analysis to quantify the complexity of an algorithm:

- Space complexity analysis: Estimates the runtime memory requirements needed to execute the algorithm.

- Time complexity analysis: Estimates the time the algorithm will take to run.

Space complexity analysis

Space complexity analysis estimates the amount of memory required by the algorithm to process input data. While processing the input data, the algorithm needs to store the transient temporary data structures in memory. The way the algorithm is designed affects the number, type, and size of these data structures. In an age of distributed computing and with increasingly large amounts of data that needs to be processed, space complexity analysis is becoming more and more important. The size, type, and number of these data structures will dictate the memory requirements for the underlying hardware. Modern in-memory data structures used in distributed computing—such as **Resilient Distributed Datasets (RDDs)**—need to have efficient resource allocation mechanisms that are aware of the memory requirements at different execution phases of the algorithm.

Space complexity analysis is a must for the efficient design of algorithms. If proper space complexity analysis is not conducted while designing a particular algorithm, insufficient memory availability for the transient temporary data structures may trigger unnecessary disk spillovers, which could potentially considerably affect the performance and efficiency of the algorithm.

Time complexity analysis

Time complexity analysis estimates how long it will take for an algorithm to complete its assigned job based on its structure. In contrast to space complexity, time complexity is not dependent on any hardware that the algorithm will run on. Time complexity analysis solely depends on the structure of the algorithm itself. The overall goal of time complexity analysis is to try to answer these important questions—will this algorithm scale? How well will this algorithm handle larger datasets?

To answer these questions, we need to determine the effect on the performance of an algorithm as the size of the data is increased and make sure that the algorithm is designed in a way that not only makes it accurate but also scales well. The performance of an algorithm is becoming more and more important for larger datasets in today's world of "big data."

In many cases, we may have more than one approach available to design the algorithm. The goal of conducting time complexity analysis, in this case, will be as follows:

"Given a certain problem and more than one algorithm, which one is the most efficient to use in terms of time efficiency?"

PART-B

6 (a) An algorithm is a set of **step-by-step procedures**, or a set of rules to follow, for completing a specific task or solving a particular problem. The word algorithm was first coined in the 9th century. Algorithms are all around us. Common examples include: the recipe for baking a cake, the method we use to solve a long division problem, the process of doing laundry, and the functionality of a search engine are all examples of an algorithm. Here's what baking a cake might look like, written out as a list of instructions, just like an algorithm:

1. Preheat the oven
2. Gather the ingredients
3. Measure out the ingredients
4. Mix together the ingredients to make the batter
5. Grease a pan
6. Pour the batter into the pan
7. Put the pan in the oven
8. Set a timer
9. When the timer goes off, take the pan out of the oven
10. Enjoy!

Algorithmic programming is all about writing a **set of rules** with a finite number of steps that instruct the computer how to perform a task. A computer program is essentially an algorithm that tells the computer what specific steps to execute, in what specific order, in order to carry out a specific task. Algorithms are written using particular syntax, depending on the programming language being used.

Types of Algorithms

Algorithms are classified based on the concepts that they use to accomplish a task. While there are many types of algorithms, the most fundamental types of computer science algorithms are:

1. **Divide and conquer algorithms** – divide the problem into smaller subproblems of the same type; solve those smaller problems, and combine those solutions to solve the original problem.
2. **Brute force algorithms** – try all possible solutions until a satisfactory solution is found.
3. **Randomized algorithms** – use a random number at least once during the computation to find a solution to the problem.
4. **Greedy algorithms** – find an optimal solution at the local level with the intent of finding an optimal solution for the whole problem.
5. **Recursive algorithms** – solve the lowest and simplest version of a problem to then solve increasingly larger versions of the problem until the solution to the original problem is found.
6. **Backtracking algorithms** – divide the problem into subproblems, each which can be attempted to be solved; however, if the desired solution is not reached, move backwards in the problem until a path is found that moves it forward.
7. **Dynamic programming algorithms** – break a complex problem into a collection of simpler subproblems, then solve each of those subproblems only once, storing their solution for future use instead of re-computing their solutions.

Algorithm for Calculating average of 5 numbers.

Step 1: Start

Step 2: Read a , b, c , d , e;

Step 3: avg = (a+b+c+d+e)/5

Step 4: Print avg

Step 5: Stop

6(b) Asymptotic Notations

Asymptotic notations are the mathematical notations used to describe the running time of an
algorithm when the input tends towards a particular value or a limiting value.
For example: In bubble sort, when the input array is already sorted, the time taken by the algorithm is linear i.e. the best case.
But, when the input array is in reverse condition, the algorithm takes the maximum time (quadratic) to sort the elements i.e. the worst case.
When the input array is neither sorted nor in reverse order, then it takes average time. These
durations are denoted using asymptotic notations.

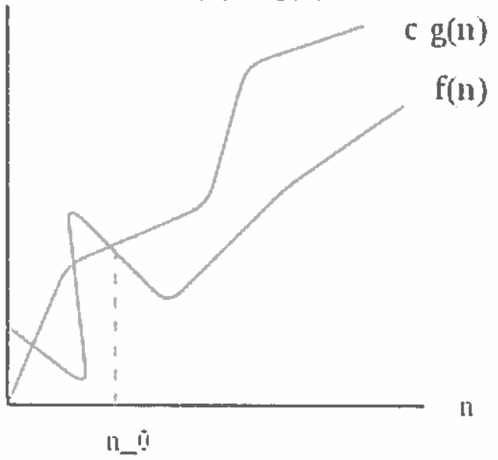
There are mainly three asymptotic notations:

- **Big-O notation**
- **Omega notation**
- **Theta notation**

Big-O Notation (O-notation)

Big-O notation represents the upper bound of the running time of an algorithm. Thus, it gives the worst-case complexity of an algorithm.

" O(g(n)) = { f(n): there exist positive constants c and n0 such that 0 ≤ f(n) ≤ cg(n) for all n ≥ n0 } "



f(n) = O(g(n))

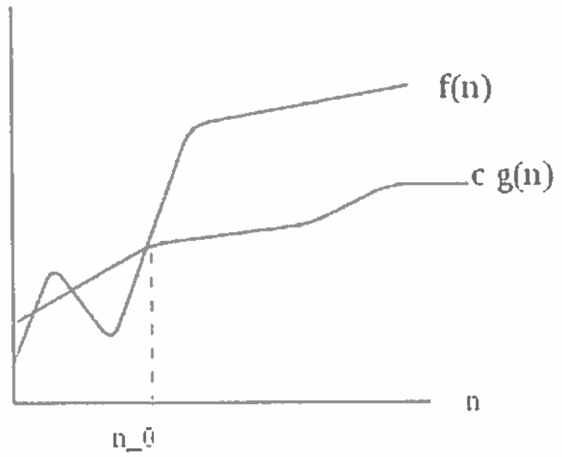
The above expression can be described as a function f(n) belongs to the set O(g(n)) if there exists a positive constant c such that it lies between 0 and cg(n), for sufficiently large n.
For any value of n, the running time of an algorithm does not cross the time provided by O(g(n)).

Since it gives the worst-case running time of an algorithm, it is widely used to analyze an algorithm as we are always interested in the worst-case scenario.

Omega Notation (Ω-notation)

Omega notation represents the lower bound of the running time of an algorithm. Thus, it provides the best case complexity of an algorithm.

" $\Omega(g(n))$ = { f(n): there exist positive constants c and n0 such that $0 \le cg(n) \le f(n)$ for all $n \ge n0$ } "



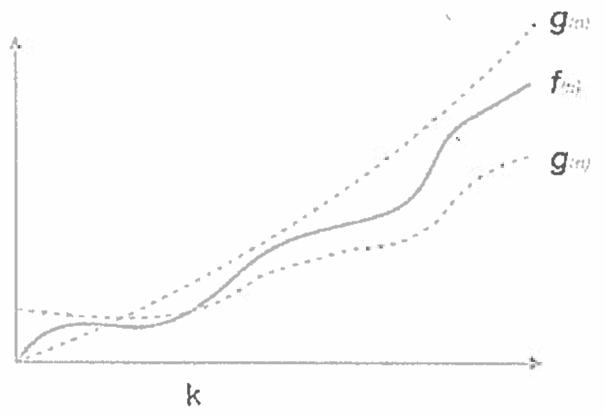
f(n) = Omega(g(n))

The above expression can be described as a function f(n) belongs to the set Ω(g(n)) if there exists a positive constant c such that it lies above cg(n), for sufficiently large n.
For any value of n, the minimum time required by the algorithm is given by Omega Ω(g(n)).

Theta Notation (Θ-notation)

Theta notation encloses the function from above and below. Since it represents the upper and
the lower bound of the running time of an algorithm, it is used for analyzing the average-case complexity of an algorithm.
For a function g(n), Θ(g(n)) is given by the relation:

"$\Theta(g(n))$ = { f(n): there exist positive constants c1, c2 and n0 such that $0 \le c1g(n) \le f(n) \le c2g(n)$ for all $n \ge n0$ }"



The above expression can be described as a function f(n) belongs to the set Θ(g(n)) if there exist
positive constants c1 and c2 such that it can be sandwiched between c1g(n) and c2g(n), for sufficiently large n.
If a function f(n) lies anywhere in between c1g(n) and c2g(n) for all n ≥ n0, then f(n) is said to be

7(a) The various types of arrays are as follows.

- One dimensional array
- Multi-dimensional array

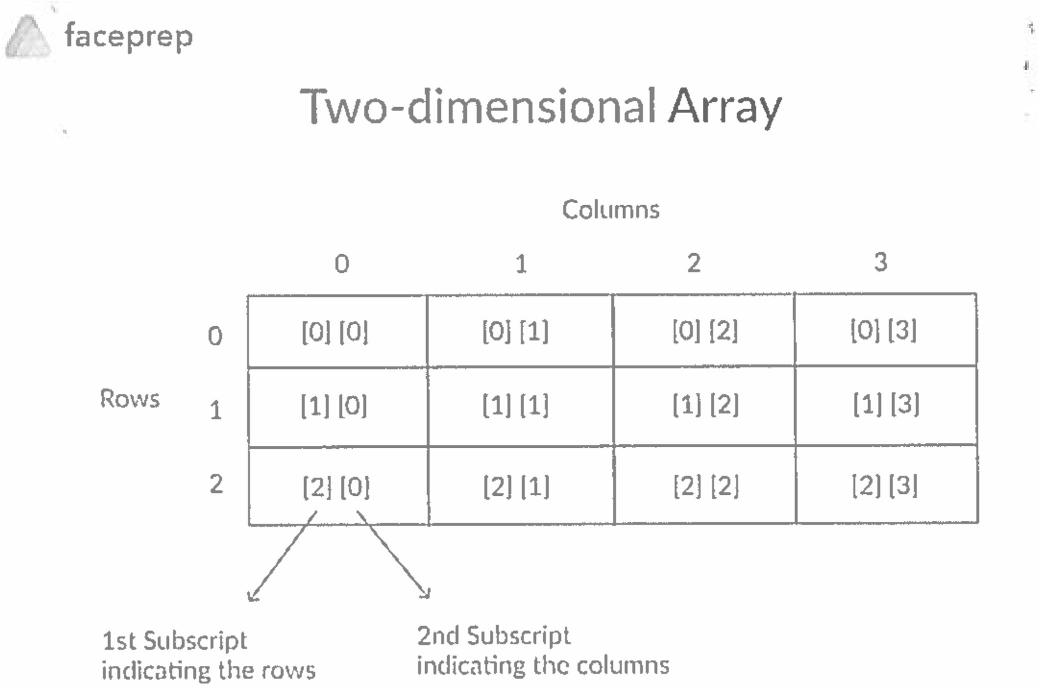
One-Dimensional Array

A one-dimensional array is also called a single dimensional array where the elements will be accessed in sequential order. This type of array will be accessed by the subscript of either a column or row index.

one-dimensional array, sometimes known as a single-dimensional array, is one in which the elements are accessed in sequence. The subscript of a column or row index will be used to access this type of array. A single subscript, in this case, represents each element. The items are saved in memory in sequential order. For example, A [1], A [2],..., A [N].

Multi-Dimensional Array

When the number of dimensions specified is more than one, then it is called as a multi-dimensional array. Multidimensional arrays include 2D arrays and 3D arrays.



A two-dimensional array will be accessed by using the subscript of row and column index. For traversing the two-dimensional array, the value of the rows and columns will be considered. In the two-dimensional array **face [3] [4]**, the first index specifies the number of rows and the second index specifies the number of columns and the array can hold 12 elements (3 * 4).

Similarly, in a three-dimensional array, there will be three dimensions. The array **face [5] [10] [15]** can hold 750 elements (5 * 10 * 15).

7(b) Various operations on Data Structures

Data Structure is the way of storing data in computer's memory so that it can be used easily and efficiently. There are different data-structures used for the storage of data. It can also be define as a mathematical or logical model of a particular organization of data items. The representation of particular data structure in the main memory of a computer is called as storage structure. **For Examples:** Array, Stack, Queue, Tree, Graph, etc.

Operations on different Data Structure:

There are different types of operations that can be performed for the manipulation of data in every data structure. Some operations are explained and illustrated below:

1.Traversing: Traversing a Data Structure means to visit the element stored in it. It visits data in a systematic manner.This can be done with any type of DS.

2.Searching :Searching means to find a particular element in the given data-structure. It is considered as successful when the required element is found. Searching is the operation which we can performed on data-structures like array, linked-list, tree, graph, etc.

3. Insertion: It is the operation which we apply on all the data-structures. Insertion means to add an element in the given data structure. The operation of insertion is successful when the required element is added to the required data-structure. It is unsuccessful in some cases when the size of the data structure is full and when there is no space in the data-structure to add any additional element. The insertion has the same name as an insertion in the data-structure as an array, linked-list, graph, tree. In stack, this operation is called Push. In the queue, this operation is called Enqueue.

4. **Deletion:** It is the operation which we apply on all the data-structures. Deletion means to delete an element in the given data structure. The operation of deletion is successful when the required element is deleted from the data structure. The deletion has the same name as a deletion in the data-structure as an array, linked-list, graph, tree, etc. In stack, this operation is called Pop. In Queue this operation is called Dequeue.

8(a) Define stack and write an algorithm to implement stack using array

Stack is a linear data structure which follows a particular order in which the operations are performed. The order may be LIFO(Last In First Out) or FILO(First In Last Out).

Basic Operations

Stack operations may involve initializing the stack, using it and then de-initializing it. Apart from these basic stuffs, a stack is used for the following two primary operations –

- **push()** – Pushing (storing) an element on the stack.
- **pop()** – Removing (accessing) an element from the stack.

To use a stack efficiently, we need to check the status of stack as well. For the same purpose, the following functionality is added to stacks –

- **peek()** – get the top data element of the stack, without removing it.
- **isFull()** – check if stack is full.
- **isEmpty()** – check if stack is empty.

step 1: Start

Step 2: Declare Stack[MAX]; //Maximum size of Stack

Step 3: Check if the stack is full or not by comparing top with (MAX-1)
If the stack is full, Then print "Stack Overflow" i.e, stack is full and cannot be pushed with another element

Step 4: Else, the stack is not full

Increment top by 1 and Set, a[top] = x
which pushes the element x into the address pointed by top.
// The element x is stored in a[top]

Step 5: Stop

Step 1: Start

Step 2: Declare Stack[MAX]

Step 3: Push the elements into the stack

Step 4: Check if the stack is empty or not by comparing top with base of array i.e 0

If top is less than 0, then stack is empty, print "Stack Underflow"

Step 5: Else, If top is greater than zero the stack is not empty, then store the value pointed by top in a variable x=a[top] and decrement top by 1. The popped element is x.

Step 1: Start

Step 2: Declare Stack[MAX]

Step 3: Push the elements into the stack

Step 4: Print the value stored in the stack pointed by top.

Step 6: Stop

8(b) Explain the differences between stack and queue.

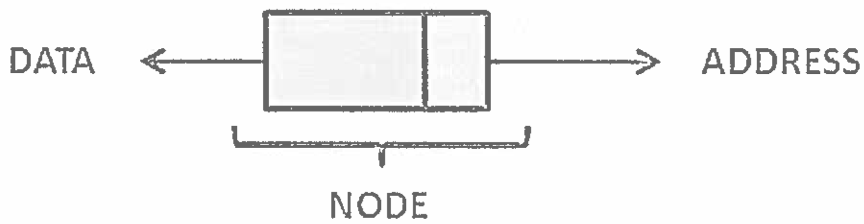
Stacks	Queues
Stacks are based on the LIFO principle, i.e., the element inserted at the last, is the first element to come out of the list.	Queues are based on the FIFO principle, i.e., the element inserted at the first, is the first element to come out of the list.
Insertion and deletion in stacks takes place only from one end of the list called the top.	Insertion and deletion in queues takes place from the opposite ends of the list. The insertion takes place at the rear of the list and the deletion takes place from the front of the list.
Insert operation is called push	Insert operation is called enqueue operation.

Stacks	Queues
operation.	
Delete operation is called pop operation.	Delete operation is called dequeue operation.
In stacks we maintain only one pointer to access the list, called the top, which always points to the last element present in the list.	In queues we maintain two pointers to access the list. The front pointer always points to the first element inserted in the list and is still present, and the rear pointer always points to the last inserted element.
Stack is used in solving problems works on recursion.	Queue is used in solving problems having sequential processing.
Stack does not have any types.	Queue is of three types – 1. Circular Queue 2. Priority queue 3. double-ended queue.
Can be considered as a vertical collection visual.	Can be considered as a horizontal collection visual.

9(a) Explain about singly linked list operations and write it's algorithms.

A singly linked list defined as all nodes are linked together in a few sequential manners, hence, it also knows as a linear linked list.

therefore, clearly it has the beginning and the end. the main problem which comes with this list is that we cannot access the predecessor of the node from the current node.



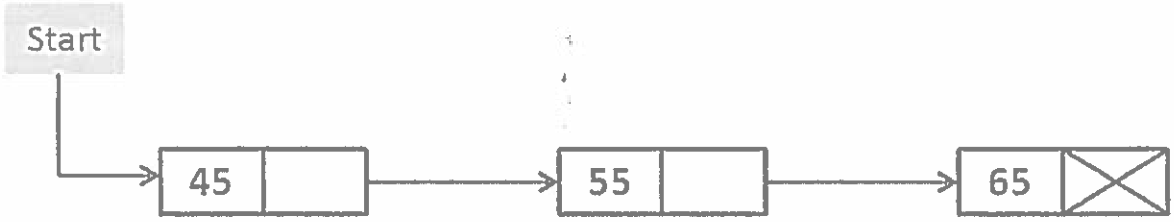


Fig: Singly Linked List

we can say that a singly linked list is a dynamic data structure because it may shrink or grow. hence, the shrinking and growing depending on the operation made.

Operations on Single Linked List

The following operations are performed on a Single Linked List

- **Insertion**
- **Deletion**
- **Display**

Insertion

In a single linked list, the insertion operation can be performed in three ways. They are as follows...

1. Inserting At Beginning of the list
2. Inserting At End of the list
3. Inserting At Specific location in the list

Inserting At Beginning of the list

We can use the following steps to insert a new node at beginning of the single linked list...

- **Step 1** - Create a **newNode** with given value.
- **Step 2** - Check whether list is **Empty** (**head == NULL**)
- **Step 3** - If it is **Empty** then, set **newNode→next = NULL** and **head = newNode**.
- **Step 4** - If it is **Not Empty** then, set **newNode→next = head** and **head = newNode**.

Inserting At End of the list

We can use the following steps to insert a new node at end of the single linked list...

- **Step 1** - Create a **newNode** with given value and **newNode → next** as **NULL**.
- **Step 2** - Check whether list is **Empty** (**head == NULL**).
- **Step 3** - If it is **Empty** then, set **head = newNode**.
- **Step 4** - If it is **Not Empty** then, define a node pointer **temp** and initialize with **head**.

- **Step 5** - Keep moving the **temp** to its next node until it reaches to the last node in the list (until **temp → next** is equal to **NULL**).
- **Step 6** - Set **temp → next = newNode.**

Inserting At Specific location in the list (After a Node)

We can use the following steps to insert a new node after a node in the single linked list...

- **Step 1** - Create a **newNode** with given value.
- **Step 2** - Check whether list is **Empty** (**head == NULL**)
- **Step 3** - If it is **Empty** then, set **newNode → next = NULL** and **head = newNode.**
- **Step 4** - If it is **Not Empty** then, define a node pointer **temp** and initialize with **head.**
- **Step 5** - Keep moving the **temp** to its next node until it reaches to the node after which we want to insert the newNode (until **temp1 → data** is equal to **location**, here location is the node value after which we want to insert the newNode).
- **Step 6** - Every time check whether **temp** is reached to last node or not. If it is reached to last node then display **'Given node is not found in the list!!! Insertion not possible!!!'** and terminate the function. Otherwise move the temp to next node.
- **Step 7** - Finally, Set **'newNode → next = temp → next'** and **'temp → next = newNode'**

Deletion

In a single linked list, the deletion operation can be performed in three ways. They are as follows...

1. Deleting from Beginning of the list
2. Deleting from End of the list
3. Deleting a Specific Node

Deleting from Beginning of the list

We can use the following steps to delete a node from beginning of the single linked list...

- **Step 1** - Check whether list is **Empty** (**head == NULL**)
- **Step 2** - If it is **Empty** then, display **'List is Empty!!! Deletion is not possible'** and terminate the function.
- **Step 3** - If it is **Not Empty** then, define a Node pointer **'temp'** and initialize with **head.**
- **Step 4** - Check whether list is having only one node (**temp → next == NULL**)
- **Step 5** - If it is **TRUE** then set **head = NULL** and delete **temp** (Setting **Empty** list conditions)
- **Step 6** - If it is **FALSE** then set **head = temp → next**, and delete **temp.**

Deleting from End of the list

We can use the following steps to delete a node from end of the single linked list...

- **Step 1** - Check whether list is **Empty** (**head == NULL**)

- **Step 2** - If it is **Empty** then, display **'List is Empty!!! Deletion is not possible'** and terminate the function.
- **Step 3** - If it is **Not Empty** then, define two Node pointers **'temp1'** and **'temp2'** and initialize **'temp1'** with **head**.
- **Step 4** - Check whether list has only one Node (**temp1 → next == NULL**)
- **Step 5** - If it is **TRUE**. Then, set **head = NULL** and delete **temp1**. And terminate the function. (Setting **Empty** list condition)
- **Step 6** - If it is **FALSE**. Then, set **'temp2 = temp1 '** and move **temp1** to its next node. Repeat the same until it reaches to the last node in the list. (until **temp1 → next == NULL**)
- **Step 7** - Finally, Set **temp2 → next = NULL** and delete **temp1**.

Deleting a Specific Node from the list

We can use the following steps to delete a specific node from the single linked list...

- **Step 1** - Check whether list is **Empty** (**head == NULL**)
- **Step 2** - If it is **Empty** then, display **'List is Empty!!! Deletion is not possible'** and terminate the function.
- **Step 3** - If it is **Not Empty** then, define two Node pointers **'temp1'** and **'temp2'** and initialize **'temp1'** with **head**.
- **Step 4** - Keep moving the **temp1** until it reaches to the exact node to be deleted or to the last node. And every time set **'temp2 = temp1'** before moving the **'temp1'** to its next node.
- **Step 5** - If it is reached to the last node then display **'Given node not found in the list! Deletion not possible!!!'**. And terminate the function.

- **Step 6** - If it is reached to the exact node which we want to delete, then check whether list is having only one node or not
- **Step 7** - If list has only one node and that is the node to be deleted, then set **head = NULL** and delete **temp1** (**free(temp1)**).
- **Step 8** - If list contains multiple nodes, then check whether **temp1** is the first node in the list (**temp1 == head**).
- **Step 9** - If **temp1** is the first node then move the **head** to the next node (**head = head → next**) and delete **temp1**.
- **Step 10** - If **temp1** is not first node then check whether it is last node in the list (**temp1 → next == NULL**).
- **Step 11** - If **temp1** is last node then set **temp2 → next = NULL** and delete **temp1** (**free(temp1)**).
- **Step 12** - If **temp1** is not first node and not last node then set **temp2 → next = temp1 → next** and delete **temp1** (**free(temp1)**).

Displaying a Single Linked List

We can use the following steps to display the elements of a single linked list...

- **Step 1** - Check whether list is **Empty** (**head == NULL**)
- **Step 2** - If it is **Empty** then, display **'List is Empty!!!'** and terminate the function.

- **Step 3 -** If it is **Not Empty** then, define a Node pointer **'temp'** and initialize with **head.**
- **Step 4 -** Keep displaying **temp → data** with an arrow (--->) until **temp** reaches to the last node
- **Step 5 -** Finally display **temp → data** with arrow pointing to **NULL (temp → data ---> NULL).**

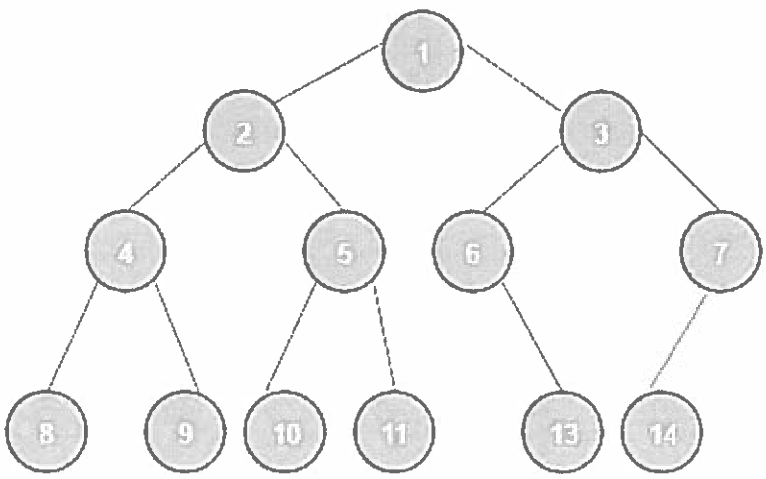
9(b) List out any 4 applications of Linked lists.

Applications of linked list in computer science:

1. Implementation of stacks and queues
2. Implementation of graphs: Adjacency list representation of graphs is the most popular which uses a linked list to store adjacent vertices.
3. Dynamic memory allocation: We use a linked list of free blocks.
4. Maintaining a directory of names
5. Performing arithmetic operations on long integers
6. Manipulation of polynomials by storing constants in the node of the linked list
7. representing sparse matrices

10(a). what is a binary tree? Construct a binary tree for given array of elements and write in-order, pre order and post order traversals along with algorithms.

Binary Tree is defined as a tree data structure where each node has at most 2 children. Since each element in a binary tree can have only 2 children, we typically name them the left and right child.



Traversal is a process to visit all the nodes of a tree and may print their values too. Because, all nodes are connected via edges (links) we always start from the root (head) node. That is, we cannot randomly access a node in a tree. There are three ways which we use to traverse a tree –

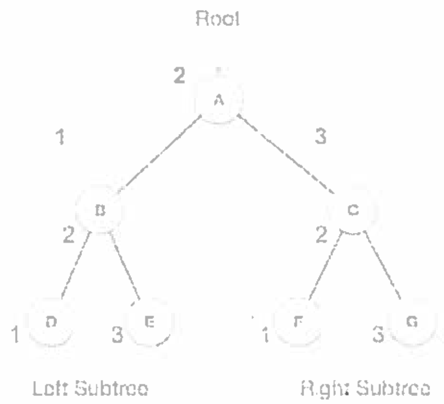
- In-order Traversal
- Pre-order Traversal
- Post-order Traversal

Generally, we traverse a tree to search or locate a given item or key in the tree or to print all the values it contains.

In-order Traversal

In this traversal method, the left subtree is visited first, then the root and later the right sub-tree. We should always remember that every node may represent a subtree itself.

If a binary tree is traversed **in-order**, the output will produce sorted key values in an ascending order.



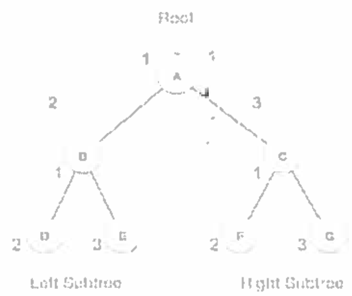
We start from **A**, and following in-order traversal, we move to its left subtree **B**. **B** is also traversed in-order. The process goes on until all the nodes are visited. The output of inorder traversal of this tree will be –

D → B → E → A → F → C → G

Algorithm
Until all nodes are traversed –
Step 1 – Recursively traverse left subtree.
Step 2 – Visit root node.
Step 3 – Recursively traverse right subtree.

Pre-order Traversal

In this traversal method, the root node is visited first, then the left subtree and finally the right subtree.



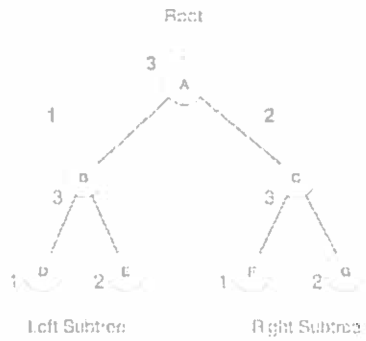
We start from **A**, and following pre-order traversal, we first visit **A** itself and then move to its left subtree **B**. **B** is also traversed pre-order. The process goes on until all the nodes are visited. The output of pre-order traversal of this tree will be –

A → B → D → E → C → F → G

Algorithm
Until all nodes are traversed –
Step 1 – Visit root node.
Step 2 – Recursively traverse left subtree.
Step 3 – Recursively traverse right subtree.

Post-order Traversal

In this traversal method, the root node is visited last, hence the name. First we traverse the left subtree, then the right subtree and finally the root node.



We start from **A**, and following Post-order traversal, we first visit the left subtree **B**. **B** is also traversed post-order. The process goes on until all the nodes are visited. The output of post-order traversal of this tree will be –

D → E → B → F → G → C → A

Algorithm
Until all nodes are traversed –
Step 1 – Recursively traverse left subtree.
Step 2 – Recursively traverse right subtree.
Step 3 – Visit root node.

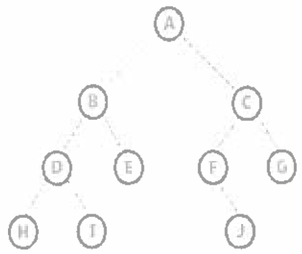
10(b). Explain Binary tree, Full Binary tree, complete Binary tree and balanced Binary tree with examples.

In a normal tree, every node can have any number of children. A binary tree is a special type of tree data structure in which every node can have a **maximum of 2 children**. One is known as a left child and the other is known as right child.

A tree in which every node can have a maximum of two children is called Binary Tree.

In a binary tree, every node can have either 0 children or 1 child or 2 children but not more than 2 children.

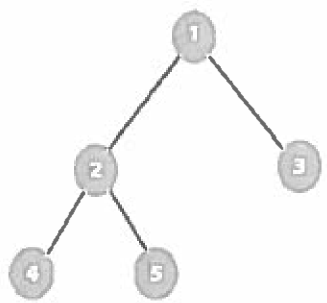
Example



Full Binary Tree

A full binary tree can be defined as a binary tree in which all the nodes have 0 or two children. In other words, the full binary tree can be defined as a binary tree in which all the nodes have two children except the leaf nodes.

The below tree is a full binary tree:

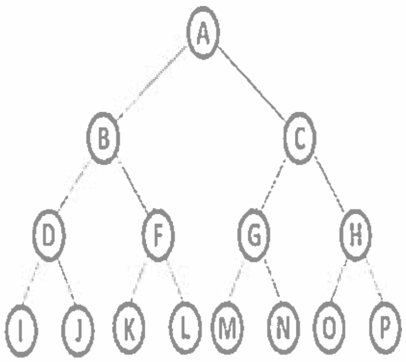


Complete Binary Tree

In a binary tree, every node can have a maximum of two children. But in strictly binary tree, every node should have exactly two children or none and in complete binary tree all the nodes must have exactly two children and at every level of complete binary tree there must be 2^{level} number of nodes. For example at level 2 there must be $2^2 = 4$ nodes and at level 3 there must be $2^3 = 8$ nodes.

A binary tree in which every internal node has exactly two children and all leaf nodes are at same level is called Complete Binary Tree.

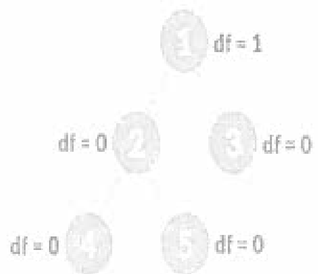
Complete binary tree is also called as **Perfect Binary Tree.**



A balanced binary tree, also referred to as a height-balanced binary tree, is defined as a binary tree in which the height of the left and right subtree of any node differ by not more than 1.

Following are the conditions for a height-balanced binary tree:

1. difference between the left and the right subtree for any node is not more than one
2. the left subtree is balanced
3. the right subtree is balanced



Balanced Binary Tree with depth at each level

11(a). Write an algorithm for construction of DFS.

DFS (Depth First Search) algorithm

The step by step process to implement the DFS traversal is given as follows –

- **Step 1** - Define a Stack of size total number of vertices in the graph.
- **Step 2** - Select any vertex as **starting point** for traversal. Visit that vertex and push it on to the Stack.

- **Step 3** - Visit any one of the non-visited **adjacent** vertices of a vertex which is at the top of stack and push it on to the stack.
- **Step 4** - Repeat step 3 until there is no new vertex to be visited from the vertex which is at the top of the stack.
- **Step 5** - When there is no new vertex to visit then use **back tracking** and pop one vertex from the stack.
- **Step 6** - Repeat steps 3, 4 and 5 until stack becomes Empty.
- **Step 7** - When stack becomes Empty, then produce final spanning tree by removing unused edges from the graph

11(b). Construct Depth First Search path from the following DFS traversal of a graph produces a **spanning tree** as final result. **Spanning Tree** is a graph without loops. We use **Stack data structure** with maximum size of total number of vertices in the graph to implement DFS traversal.

We use the following steps to implement DFS traversal...

- **Step 1** - Define a Stack of size total number of vertices in the graph.

- **Step 2** - Select any vertex as **starting point** for traversal. Visit that vertex and push it on to the Stack.

- **Step 3** - Visit any one of the non-visited **adjacent** vertices of a vertex which is at the top of stack and push it on to the stack.

- **Step 4** - Repeat step 3 until there is no new vertex to be visited from the vertex which is at the top of the stack.

- **Step 5** - When there is no new vertex to visit then use **back tracking** and pop one vertex from the stack.

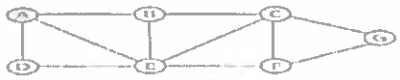
- **Step 6** - Repeat steps 3, 4 and 5 until stack becomes Empty.

- **Step 7** - When stack becomes Empty, then produce final spanning tree by removing unused edges from the graph

Back tracking is coming back to the vertex from which we reached the current vertex.**Example**

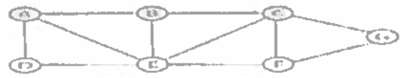
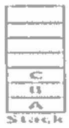
Consider the following example graph to perform DFS traversal

Step 1:
- Select the vertex A as starting point (visit A).
- Push A on to the Stack.

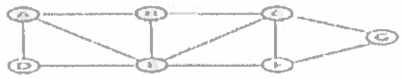



Step 2:
- Visit any adjacent vertex of A which is not visited (B).
- Push newly visited vertex B on to the Stack.


Step 3:
- Visit any adjacent vertext of B which is not visited (C).
- Push C on to the Stack.

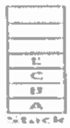



Step 4:
- Visit any adjacent vertext of C which is not visited (E).
- Push E on to the Stack

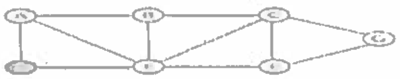
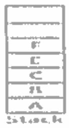



Step 5:
- Visit any adjacent vertext of E which is not visited (D).
- Push D on to the Stack

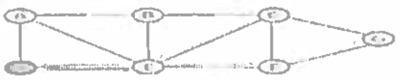
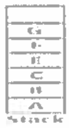

Step 6:
- There is no new vertiex to be visited from D. So use back track.
- Pop D from the Stack.

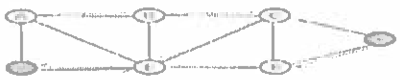
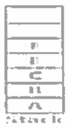
Step 7:
- Visit any adjacent vertex of E which is not visited (F).
- Push F on to the Stack.

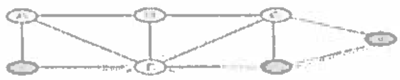
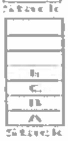
Step 8:
- Visit any adjacent vertex of F which is not visited (G).
- Push G on to the Stack.

Step 9:
- There is no new vertiex to be visited from G. So use back track.
- Pop G from the Stack.

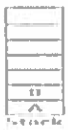



Step 10:
- There is no new vertiex to be visited from F. So use back track.
- Pop F from the Stack.

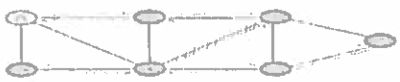



Step 11:
- There is no new vertiex to be visited from E. So use back track.
- Pop E from the Stack.


Step 12:
- There is no new vertiex to be visited from C. So use back track.
- Pop C from the Stack.

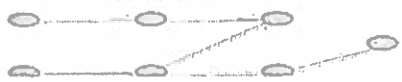



Step 13:
- There is no new vertiex to be visited from B. So use back track.
- Pop B from the Stack.




Step 14:
- There is no new vertiex to be visited from A. So use back track.
- Pop A from the Stack.


- Stack became Empty. So stop DFS Traversal.
- Final result of DFS traversal is following spanning tree.



12(a). Write an algorithm for Bubble Sort.

Bubble Algorithm

```
begin BubbleSort(list)

  for all elements of list
    if list[i] > list[i+1]
      swap(list[i], list[i+1])
    end if
  end for

  return list

end BubbleSort
```

12(b). Sort the following elements using bubble sort.

10,5,15,0,12

The above illustration can be summarized in a tabular form as shown below:

Pass	Unsorted list	comparison	Sorted list
1	{10,5,15,0,12}	{10,5}	{5,10,15,0,12}
	{5,10,15,0,12}	{10,15}	{5,10,15,0,12}
	{5,10,15,0,12}	{15,0}	{5,10,0,15,12}
	{5,10,0,15,12}	{15,12}	{5,10,0,12,15}
2	{5,10,0,12,15}	{5,10}	{5,10,0,12,15}
	{5,10,0,12,15}	{10,0}	{5,0,10,12,15}
	{5,0,10,12,15}	{10,12}	{5,0,10,12,15}
3	{5,0,10,12,15}	{5,0}	{0,5,10,12,15}
	{5,0,10,12,15}	{5,10}	{5,0,10,12,15}
	{5,0,10,12,15}	SORTED	

As shown in the illustration, with every pass, the largest element bubbles up to the last thereby sorting the list with every pass. As mentioned in the introduction, each element is compared to its adjacent element and swapped with one another if they are not in order.

13(a). Write an algorithm for merge sort.

Merge sort is the first algorithm we are going to study in Divide and Conquer. According to Divide and Conquer, it first divides an array into smaller subarrays and then merges them together to get a sorted array.

Algorithm for merge sort :

```
MERGE-SORT(A, start, end)
 if start < right
   middle = floor((start+end)/2)
   MERGE-SORT(A, start, middle)
   MERGE-SORT(A, middle+1, end)
```

So, our first task is to make two temporary arrays.

```
temp1 = A[start, middle]
temp2 = A[middle+1, end]
```

Now, we have to iterate over these arrays and compare the elements and then put the smaller elements into the bigger array.

```
i = 1
j = 1
k = start
while i <= temp1.length and j <= temp2.length
  if temp1[i] < temp2[j]
    A[k] = temp1[i]
    i=i+1
  else
    A[k] = temp2[j]
    j=j+1
  k=k+1
```

13 (b)Search element 10 from the given list 32,14,65,67,23,36,6,100

(13).b, given elements :- 32, 14, 65, 67, 23, 36, 6, 100.

Search element = 10.

Step 1 :-

32	14	65	67	23	36	6	100

$32 \neq 10$

Step 2 :-

32	14	65	67	23	36	6	100

$14 \neq 10$

Step 3 :-

32	14	65	67	23	36	6	100

$65 \neq 10$

Step 4 :-

32	14	65	67	23	36	6	100

$67 \neq 10$

Step 5 :-

32	14	65	67	23	36	6	100.

$23 \neq 10$

Step 6 :-

32	14	65	67	23	36	6	100.

$36 \neq 0$

Step 7 :-

32	14	65	67	23	36	6	100

$6 \neq 10$

Step 8 :-

32	14	65	67	23	36	6	100

$100 \neq 10$

∴ Since the given Search element is not present in the given list of elements.

∴ The Linear Search is Unsuccessful.

14(a). Write prims algorithm for finding minimum spanning tree.

Spanning tree - A spanning tree is the subgraph of an undirected connected graph.

Minimum Spanning tree - Minimum spanning tree can be defined as the spanning tree in which the sum of the weights of the edge is minimum. The weight of the spanning tree is the sum of the weights given to the edges of the spanning tree.

Now, let's start the main topic.

Prim's Algorithm is a greedy algorithm that is used to find the minimum spanning tree from a graph. Prim's algorithm finds the subset of edges that includes every vertex of the graph such that the sum of the weights of the edges can be minimized.

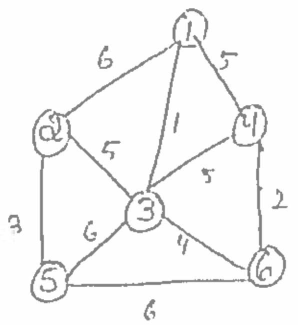
Prim's algorithm starts with the single node and explores all the adjacent nodes with all the connecting edges at every step. The edges with the minimal weights causing no cycles in the graph got selected.

Algorithm

1. Step 1: Select a starting vertex
2. Step 2: Repeat Steps 3 and 4 until there are fringe vertices
3. Step 3: Select an edge 'e' connecting the tree vertex and fringe vertex that has minimum wei ght
4. Step 4: Add the selected edge and the vertex to the minimum spanning tree T
5. [END OF LOOP]
6. Step 5: EXIT

14(b). Construct minimum spanning tree using Prim's algorithm from the following graph.

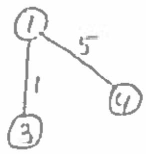
(14) b)



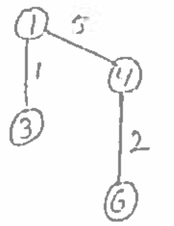
Sol:- Using Prim's algorithm:-

Step 1:-

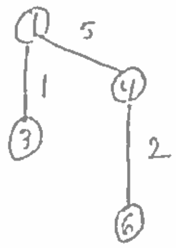
Step 2:-



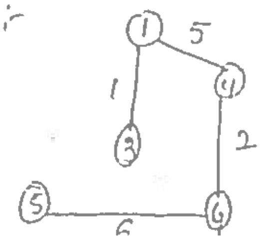
Step 3:-



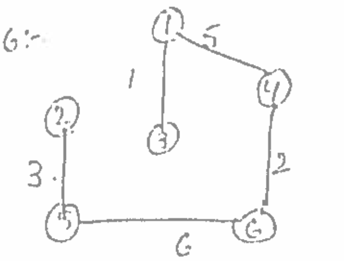
Step 4:-



Step 5:-



Step 6:-



$\therefore$ The cost of the minimum spanning tree =

$1+5+2+6+3$

= 17 Units

15(a). Explain All pairs shortest path algorithm with example.

The all pair shortest path algorithm is also known as Floyd-Warshall algorithm is used to find all pair shortest path problem from a given weighted graph. As a result of this algorithm, it will generate a matrix, which will represent the minimum distance from any node to all other nodes in the graph.

1. In all pair shortest path, when a weighted graph is represented by its weight matrix W then objective is to find the distance between every pair of nodes.
2. We will apply dynamic programming to solve the all pairs shortest path.
3. In all pair shortest path algorithm, we first decomposed the given problem into sub problems.
4. In this principle of optimally is used for solving the problem.
5. It means any sub path of shortest path is a shortest path between the end nodes.

Steps:

i. Let Aki,jAi,jk be the length of shortest path from node i to node j such that the label for every intermediate node will be ≤ k.
ii. Now, divide the path from i node to j node for every intermediate node, say 'k' then there arises two case.

a. Path going from i to j via k. b. Path which is not going via k.

iii. Select only shortest path from two cases.

iv. Using recursive method we compute shortest path.

v. Initially: A0=W[i,j]A0=W[i,j]
vi. Next computations: Aki,j=minAk−1i,j,Ak−1i,k,Ak−1k,jAi,jk=minAi,jk−1,Ai,kk−1,Ak,jk−1.

```
Algorithm All_pair(W, A)
{
For i = 1 to n do
For j = 1 to n do
A [i , j] = W [i , j]
For k = 1 to n do
        {
        For i = 1 to n do
                {
                For j = 1 to n do
                {
                A [i , j] = min(A [i, j], A [i, k] + A [k, j])
                }
                }}}
```

Algorithm:

Analysis of Algorithm:

i. The first double for loop takes O (n2) time.

ii. The nested three for loop takes O (n3) time.

iii. Thus, the whole algorithm takes O (n3) time.

Example: Compute all pair shortest *path for following figure 7.*

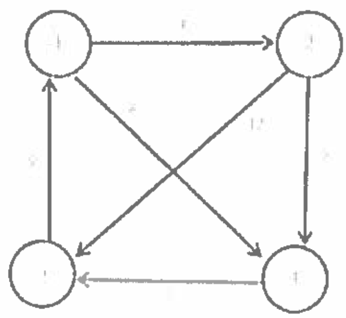


Figure 7

Solution:

$$A^1 = \begin{bmatrix} 0 & 4 & 8 & \infty \\ \infty & 0 & 5 & 12 \\ \infty & \infty & 0 & 7 \\ 5 & \infty & \infty & 0 \end{bmatrix} \quad A^2 = \begin{bmatrix} 0 & 4 & 8 & \infty \\ \infty & 0 & 5 & 12 \\ 12 & \infty & 0 & 7 \\ 5 & 9 & 13 & 0 \end{bmatrix} \quad A^3 = \begin{bmatrix} 0 & 4 & 8 & 16 \\ 17 & 0 & 5 & 12 \\ 12 & \infty & 0 & 7 \\ 5 & 9 & 13 & 0 \end{bmatrix} \quad A^4 = \begin{bmatrix} 0 & 4 & 8 & 16 \\ 17 & 0 & 5 & 12 \\ 12 & 16 & 0 & 7 \\ 5 & 9 & 13 & 0 \end{bmatrix}$$

Thus the shortest distances between all pair are obtained.

15(b). Find the optimal solution for the fractional knapsack problem making use of greedy approach. Consider n=5, w=60kg

(w1,w2,w3,w4,w5) = (5,10,15,22,25)

(b1,b2,b3,b4,b5) = (30,40,45,77,90)

Step-01:
Compute the value / weight ratio for each item->

Items: (1 , 2 , 3 , 4 , 5)
Value: (30 , 40 , 45 , 77 , 90)
Weight: (5 , 10 , 15 , 22 , 25)
Ratio: (6 , 4 , 3 , 3.5 , 3.6)
Step-02:
Sort all the items in the decreasing order of their value / weight ratios->
I1 : 6
I2: 4
I5: 3.6
I4: 3.5
I3: 3
Step-03:
Start filling the knapsack by putting the items in it one by one.
Knapsack Weight : 60 , Items in the Knapsack: 0 , Profit : 0
Knapsack Weight : 60–5=55 , Items in the Knapsack: I1, Profit: 30
Knapsack Weight : 55–10=45 , Items in the Knapsack (I1,I2) , Profit :30+40=70
Knapsack Weight : 45–25=20 , Items in the Knapsack: (I1,I2,I5) , Profit :70+90=160
Now,

Knapsack weight left to be filled is 20 kg but item-4 has a weight of 22 kg.

-Had the problem been a 0/1 knapsack problem, we would have stopped and reported that

the knapsack has items < I1 , I2 , I5 > and the knapsack's total cost is 160.

-But because in fractional knapsack problem we can even take the fraction of any item.

-So, our knapsack will contain the items->

I1 , I2 , I5 , (20/22) I4

Now,

Total cost of the knapsack

= 160 + (20/22) * 77

= 230

So,final profit of knapsack problem, which we can get by fractional greedy method is, 230.

Thus,we can get optimal solution of knapsack problem using fractional greedy method...which overcomes the disadvantage of the 0/1 knapsack problem,that doesn't give optimal solution.

15 (b) **Greedy Technique: General method**

The greedy method is one of the strategies like Divide and conquer used to solve the problems. This method is used for solving optimization problems. An optimization problem is a problem that demands either maximum or minimum results. Let's understand through some terms.

The Greedy method is the simplest and straightforward approach. It is not an algorithm, but it is a technique. The main function of this approach is that the decision is taken on the basis of the currently available information. Whatever the current information is present, the decision is made without worrying about the effect of the current decision in future.

This technique is basically used to determine the feasible solution that may or may not be optimal. The feasible solution is a subset that satisfies the given criteria. The optimal solution is the solution which is the best and the most favorable solution in the subset. In the case of feasible, if more than one solution satisfies the given criteria then those solutions will be considered as the feasible, whereas the optimal solution is the best solution among all the solutions.

Characteristics of Greedy method

- To construct the solution in an optimal way, this algorithm creates two sets where one set contains all the chosen items, and another set contains the rejected items.
- A Greedy algorithm makes good local choices in the hope that the solution should be either feasible or optimal.

Applications of Greedy Algorithm

- It is used in finding the shortest path.
- It is used to find the minimum spanning tree using the prim's algorithm or the Kruskal's algorithm.
- It is used in a job sequencing with a deadline.

- This algorithm is also used to solve the fractional knapsack problem.

Knapsack problem:

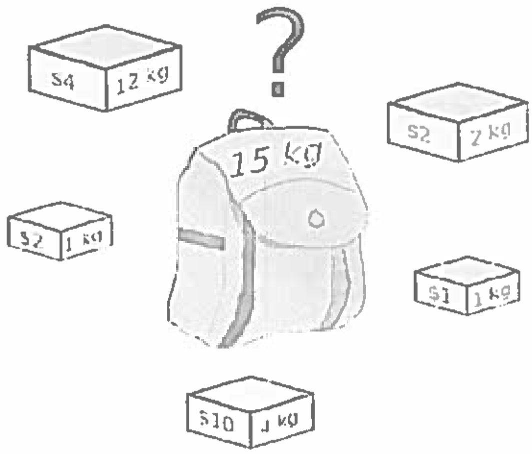
Knapsack Problem-

You are given the following-

- A knapsack (kind of shoulder bag) with limited weight capacity.
- Few items each having some weight and value.
- **The problem states-**

Which items should be placed into the knapsack such that

- The value or profit obtained by putting the items into the knapsack is maximum.
- And the weight limit of the knapsack does not exceed.



Knapsack Problem

Knapsack Problem Variants-

Knapsack problem has the following two variants-

1. Fractional Knapsack Problem
2. 0/1 Knapsack Problem

Fractional Knapsack Problem-

In Fractional Knapsack Problem,

- As the name suggests, items are divisible here.
- We can even put the fraction of any item into the knapsack if taking the complete item is not possible.
- It is solved using Greedy Method.

Fractional Knapsack Problem Using Greedy Method-

Fractional knapsack problem is solved using greedy method in the following steps-

Step-01:

For each item, compute its value / weight ratio.

Step-02:

Arrange all the items in decreasing order of their value / weight ratio.

Step-03:

Start putting the items into the knapsack beginning from the item with the highest ratio.

Put as many items as you can into the knapsack.

Time Complexity-

- The main time taking step is the sorting of all items in decreasing order of their value / weight ratio.
- If the items are already arranged in the required order, then while loop takes O(n) time.
- The average time complexity of **Quick Sort** is O(nlogn).
- Therefore, total time taken including the sort is O(nlogn).

FRACTIONAL KNAPSACK PROBLEM-

Problem-

For the given set of items and knapsack capacity = 60 kg, find the optimal solution for the fractional knapsack problem making use of greedy approach.

Item	Weight	Value

1	5	30
2	10	40
3	15	45
4	22	77
5	25	90

Find the optimal solution for the fractional knapsack problem making use of greedy approach. Consider-

$$n = 5$$

$$w = 60 \text{ kg}$$

$$(w1, w2, w3, w4, w5) = (5, 10, 15, 22, 25)$$

$$(b1, b2, b3, b4, b5) = (30, 40, 45, 77, 90)$$

A thief enters a house for robbing it. He can carry a maximal weight of 60 kg into his bag. There are 5 items in the house with the following weights and values. What items should thief take if he can even take the fraction of any item with him?

Item	Weight	Value
1	5	30
2	10	40
3	15	45
4	22	77
5	25	90

Solution-

Step-01:

Compute the value / weight ratio for each item-

Items	Weight	Value	Ratio
1	5	30	6
2	10	40	4
3	15	45	3
4	22	77	3.5
5	25	90	3.6

Step-02:

Sort all the items in decreasing order of their value / weight ratio-

I1 I2 I5 I4 I3

(6) (4) (3.6) (3.5) (3)

Step-03:

Start filling the knapsack by putting the items into it one by one.

Knapsack Weight	Items in Knapsack	Cost

60	Ø	0
55	I1	30
45	I1, I2	70
20	I1, I2, I5	160

Now,

- Knapsack weight left to be filled is 20 kg but item-4 has a weight of 22 kg.
- Since in fractional knapsack problem, even the fraction of any item can be taken.
- So, knapsack will contain the following items-

< I1 , I2 , I5 , (20/22) I4 >

Total cost of the knapsack

= 160 + (20/27) x 77

= 160 + 70

= 230 units

Important Note-

Had the problem been a 0/1 knapsack problem, knapsack would contain the following items-

< I1 , I2 , I5 >

Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech. (U. G.)	Program	Common to All			Academic Year	2022 – 2023
Course Code	20AIO01	Test Duration	3 Hrs.	Max. Marks	70	Semester	V
Course	Machine Learning for Engineers (Open Elective)						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define machine learning.	20AIO01.1	L1
2	List out any two applications of Machine Learning.	20AIO01.2	L1
3	Define classification and clustering.	20AIO01.3	L1
4	List any two genetic operators.	20AIO01.4	L1
5	Write Baye's theorem.	20AIO01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain different types of machine learning.	6M	20AIO01.1	L2
6 (b)	Explain about linear regression.	6M	20AIO01.1	L2
	OR			
7	Describe about vectors and candidate elimination algorithm.	12M	20AIO01.1	L2
8	Explain multi-layered perceptron model with suitable example.	12M	20AIO01.2	L2
	OR			
9	Describe support vector machine algorithm with a suitable example.	12M	20AIO01.2	L2
10	Explain decision tree algorithm with a suitable application.	12M	20AIO01.3	L2
	OR			
11	Describe nearest neighbor algorithm with example.	12M	20AIO01.3	L2
12	Explain about principal component analysis and independent component analysis.	12M	20AIO01.4	L2
	OR			
13	Explain genetic algorithm along with operators and applications.	12M	20AIO01.4	L2
14	Explain about Monte Carlo methods.	12M	20AIO01.5	L2
	OR			
15	Describe Bayesian network.	12M	20AIO01.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

1	**Define machine learning** Machine learning is an application of AI that enables systems to learn and improve from experience without being explicitly programmed. Machine learning focuses on developing computer programmed that can access data and use it to learn for themselves.
2	**List out any two applications of machine learning** Virtual personal assistants Self-driving cars Google translates Traffic alerts Product recommendations Speech Recognition
3	**Define classification and clustering** Classification is used for supervised learning. The process of classifying the input instances based on their corresponding class labels is known as classification, as classification have labels so there is need of training and testing, examples are logistic regression naive bayes classifier Clustering is used for unsupervised learning, grouping the instances based on their similarity without the help of class labels is know as clustering.in clustering there is no need for training and testing, example k-means clustering, fuzzy c-means
4	**List any two genetic operator** 1) Selection Operator: The idea is to give preference to the individuals with good fitness scores and allow them to pass their genes to successive generations. 2) Crossover Operator: This represents mating between individuals. Two individuals are selected using selection operator and crossover sites are chosen randomly. Then the genes at these crossover sites are exchanged thus creating a completely new individual (offspring)
5	**List any ~~two genetic operator~~** Write Bayes theorem Bayes' theorem describes the probability of occurrence of an event related to any condition. To prove the Bayes' theorem, use the concept of conditional probability formula, which is P(Ei|A)=P(Ei∩A)P(A). Bayes' Theorem describes the probability of occurrence of an event related to any condition.
6 (a)	**Explain the difference types of machine learning** Based on the methods and way of learning, machine learning is divided into mainly two types, which are: 1. Supervised Machine Learning 2. Unsupervised Machine Learning **Supervised Machine Learning** As its name suggests, Supervised machine learning is based on supervision. It means in the supervised learning technique, we train the machines using the "labelled" dataset, and based on the training, the machine predicts the output. Here, the labelled data specifies that some of the inputs are already mapped to the output. More preciously, we can say; first, we train the machine with the input and corresponding output, and then we ask the machine to predict the output using the test dataset. Let's understand supervised learning with an example. Suppose we have an input dataset of cats and dog images. So, first, we will provide the training to the machine to understand the images, such as the shape & size of the tail of cat and dog, Shape of eyes, colour, height (dogs are taller, cats are smaller), etc. After completion of training, we input the picture of a cat and ask the machine to identify the object and predict the output. Now, the machine is well trained, so it will check all the features of the object, such as height, shape, colour, eyes, ears, tail, etc., and find that it's a cat. So, it will put it in the Cat category. This is the process of how the machine identifies the objects in Supervised Learning. The main goal of the supervised learning technique is to map the input variable(x) with the output variable(y). Some real-world applications of supervised learning are Risk Assessment, Fraud Detection, Spam filtering, etc. **Unsupervised Machine Learning**

Unsupervised learning is different from the Supervised learning technique; as its name suggests, there is no need for supervision. It means, in unsupervised machine learning, the machine is trained using the unlabeled dataset, and the machine predicts the output without any supervision.

In unsupervised learning, the models are trained with the data that is neither classified nor labelled, and the model acts on that data without any supervision.

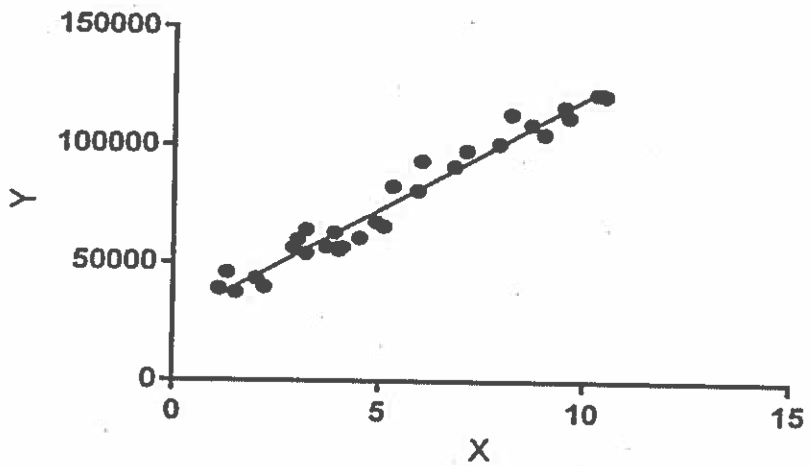
The main aim of the unsupervised learning algorithm is to group or categories the unsorted dataset according to the similarities, patterns, and differences. Machines are instructed to find the hidden patterns from the input dataset.

Let's take an example to understand it more preciously; suppose there is a basket of fruit images, and we input it into the machine learning model. The images are totally unknown to the model, and the task of the machine is to find the patterns and categories of the objects.

So, now the machine will discover its patterns and differences, such as colour difference, shape difference, and predict the output when it is tested with the test dataset.

6 (b) **Explain about linear regression**

Linear Regression is a machine learning algorithm based on supervised learning. It performs a regression task. Regression models a target prediction value based on independent variables. It is mostly used for finding out the relationship between variables and forecasting. Different regression models differ based on – the kind of relationship between dependent and independent variables they are considering, and the number of independent variables getting used. There are many names for a regression's dependent variable. It may be called an outcome variable, criterion variable, endogenous variable, or regressand. The independent variables can be called exogenous variables, predictor variables, or regressors.



Linear regression performs the task to predict a dependent variable value (y) based on a given independent variable (x). So, this regression technique finds out a linear relationship between x (input) and y(output). Hence, the name is Linear Regression. In the figure above, X (input) is the work experience and Y (output) is the salary of a person. The regression line is the best fit line for our model .Hypothesis function for Linear Regression :

$$y = \theta_1 + \theta_2 . x$$

While training the model we are given : x: input training data (univariate – one input variable(parameter)) y: labels to data (Supervised learning) When training the model – it fits the best line to predict the value of y for a given value of x. The model gets the best regression fit line by finding the best θ1 and θ2 values. θ1: intercept θ2: coefficient of x Once we find the best θ1 and θ2 values, we get the best fit line. So when we are finally using our model for prediction, it will predict the value of y for the input value of x. How to update θ1 and θ2 values to get the best fit line ?

Cost Function (J):

By achieving the best-fit regression line, the model aims to predict y value such that the error difference between predicted value and true value is minimum. So, it is very important to update the θ1 and θ2 values, to reach the best value that minimize the error between predicted y value (pred) and true y value (y).

$$minimize \frac{1}{n} \sum_{i=1}^{n} (pred_i - y_i)^2$$

$$J = \frac{1}{n} \sum_{i=1}^{n} (pred_i - y_i)^2$$

Cost function(J) of Linear Regression is the Root Mean Squared Error (RMSE) between predicted y value (pred) and true y value (y).

Gradient Descent: To update θ1 and θ2 values in order to reduce Cost function (minimizing RMSE value) and achieving the best fit line the model uses Gradient Descent. The idea is to start with random $\theta 1$ and $\theta 2$ values and then iteratively updating the values, reaching minimum cost.

7 **Describe about vectors and candidate elimination algorithm**

Version spaces:
Consider the following examples:
From these examples, can you come up with a description of days on which you enjoy playing your favorite water sport?
Is the answer that you enjoy playing water sports on sunny days?
Is it that you enjoy playing water sports on warm days?
Or is it sunny and warm days?
Or perhaps it's sunny and warm days with strong wind?
In Version Spaces, we look for all answers that are consistent with the examples.

- an incremental learning method
- makes use of positive and negative examples
- will keep track of all possible definitions that are consistent with the examples seen (but not explicitly)

Rather than literally storing all possible answers, we store two sets of information:
G: the most general classification rules consistent with the examples.
S: the most specific classification rules consistent with the examples.

❖ Candidate Elimination Learning Algorithm

Candidate Elimination Learning Algorithm is a method for learning concepts from data that is supervised. In this blog, we'll explain the candidate elimination learning algorithm with examples.
Given a hypothesis space H and a collection E of instances, the candidate elimination procedure develops the version space progressively.
The examples are introduced one by one, with each one potentially shrinking the version space by deleting assumptions that contradict the example. For each new case, the candidate elimination method updates the general and particular boundaries.
To understand the algorithm better, let us have a look at some terminologies and what it means.

What is Version Space?
It's a cross between a generic and a specific theory. It didn't simply write one hypothesis; it wrote a list of all feasible hypotheses based on the training data.
With regard to hypothesis space H and training examples D, the version space, denoted as VSH,D, is the subset of hypotheses from H that are consistent with the training instances in D.

For example, consider the following dataset. The classic example of EnjoySport.
In the above example, We have two hypotheses from H in the case above, both of which are consistent with the training dataset.
h1=< Sunny, Warm, ?, Strong, ?, ?> and
h2=< ?, Warm, ?, Strong, ?, ?>
As a result, the collection of hypotheses h1, h2 is referred to as a Version Space.
Specific Hypothesis:
If a hypothesis, h, covers none of the negative cases and there is no other hypothesis, h', that covers none of the negative examples, then h is strictly more general than h', then h is said to be the most specific hypothesis.
The specific hypothesis fills in important details about all the variables given in the hypothesis.
S = < 'ϕ', 'ϕ', 'ϕ', ..., 'ϕ' >
G
eneral Hypothesis:
In general, a hypothesis is an explanation for anything. The general hypothesis explains the relationship between the key variables in general. I want to watch Avengers, for example, is a general hypothesis for selecting a movie.
G = < '?', '?', '?', ...'?'>
Representations:

o The most specific hypothesis is represented using ϕ.
o The most general hypothesis is represented using ?.

Why Candidate Elimination Algorithm?
Candidate Elimination Learning Algorithm addresses several of the limitations of

FIND-S.
Although the FIND-S algorithm outputs a hypothesis from H, that is consistent with the training examples, this is just one of many hypotheses from H that might fit the training data equally well.
The key idea in the CANDIDATE-ELIMINATION Algo is to output a description of the set of all hypotheses consistent with the training examples.
Candidate Elimination:
Unlike Find-S(#Link to Find-S) algorithm, the Candidate Elimination algorithm considers not just positive but negative samples as well. It relies on the concept of version space.
At the end of the algorithm, we get both specific and general hypotheses as our final solution.
For a positive example, we move from the most specific hypothesis to the most general hypothesis.
For a negative example, we move from the most general hypothesis to the most specific hypothesis.
Candidate Elimination Algorithm:
1. Initialize both specific and general hypotheses.
S = < 'ϕ', 'ϕ', 'ϕ', ..., 'ϕ' >
G = < '?', '?', '?', ..., '?'>
Depending on the number of attributes.
2. Take the next example, if the taken example is positive make a specific hypothesis to general.
3. If the taken example is negative make the general hypothesis to a more specific hypothesis.
Let's have a look at an example to see how the Candidate Elimination Algorithm works.
1. Initializing both specific and general hypotheses.
G0 = < <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?>,
<?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ? >>
S0 = < ϕ, ϕ, ϕ, ϕ, ϕ, ϕ>
2. When the first training example is supplied (in this case, a positive example), the Candidate Elimination method evaluates the S boundary and determines that it is too specific, failing to cover the positive example.
As a result, the border is shifted to the least general hypothesis that covers this new case. S1 indicates the updated border.
No update for G1 is needed in this example as G0 accurately represents the training instance.
G1 = G0 = < <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?>,
<?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ? >>
S1 = < 'Sunny', 'warm', 'normal', 'strong', 'warm ', 'same'>
3. When the second (also positive) training example is observed, it has a similar effect of generalizing S to S2, while leaving G intact (i.e., G2 = G1 = G0).
G2 = G0 = < <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?>,
<?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ?> , <?, ?, ?, ?, ?, ? >>
S2 = < 'Sunny', 'warm', '?', 'strong', 'warm ', 'same'>
4. Similarly, considering the training instance 3, This negative example demonstrates that the version space's G border is extremely general;
As a result, the hypothesis in the G border must be specialized until it appropriately categorizes this new negative case.
G3 = < <'Sunny', ?, ?, ?, ?, ?>, <?, 'warm', ?, ?, ?, ?>, <?, ?, ?, ?, ?, ?>, <?, ?,
?, ?, ?, ?>, <?, ?, ?, ?, ?, ?>, <?, ?, ?, ?, ?, 'same'> >
S3 = S2 = < 'Sunny', 'warm', '?', 'strong', 'warm ', 'same'>
5. The fourth training example, generalizes the version space's S boundary. It also results in the removal of one G border member, as this one fails to cover the new positive example.

	G4 = < <'Sunny', ?, ?, ?, ?, ?>, <?, 'warm', ?, ?, ?, ?> > S4 = <'Sunny', 'warm', ?, 'strong', ?, ?> Finally, the result is produced by synchronizing the G4 and S4 algorithms. The above diagram depicts the whole version space, including the hypotheses bounded by S4 and G4. The order in which the training examples are given has no impact on the learned version space. The final hypothesis is, G = <['Sunny', ?, ?, ?, ?, ?>, <?, 'warm', ?, ?, ?, ?>> S = <'Sunny', 'warm', ?, 'strong', ?, ?>

8 (a) **Explain multi-layered perceptron model with suitable example.**

Perceptron

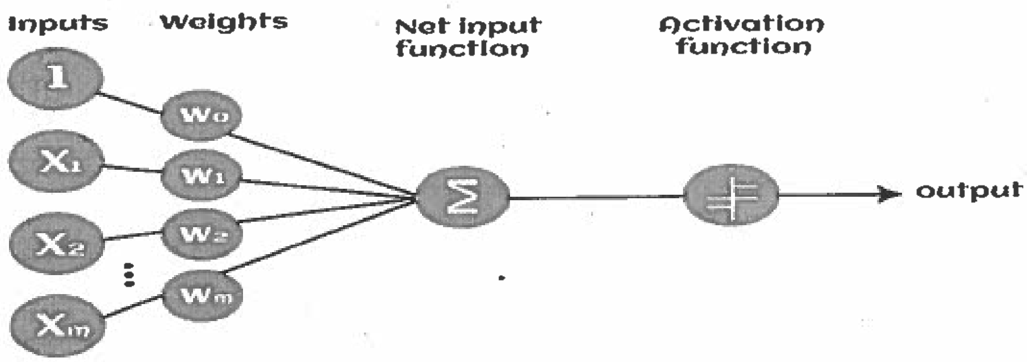
In Machine Learning and Artificial Intelligence, Perceptron is the most commonly used term for all folks. It is the primary step to learn Machine Learning and Deep Learning technologies, which consists of a set of weights, input values or scores, and a threshold. Perceptron is a building block of an Artificial Neural Network. Initially, in the mid of 19th century, Mr. Frank Rosenblatt invented the Perceptron for performing certain calculations to detect input data capabilities or business intelligence. Perceptron is a linear Machine Learning algorithm used for supervised learning for various binary classifiers. This algorithm enables neurons to learn elements and processes them one by one during preparation. In this tutorial, "Perceptron in Machine Learning," we will discuss in-depth knowledge of Perceptron and its basic functions in brief. Let's start with the basic introduction of Perceptron.

What is the Perceptron model in Machine Learning?

Perceptron is Machine Learning algorithm for supervised learning of various binary classification tasks. Further, Perceptron is also understood as an Artificial Neuron or neural network unit that helps to detect certain input data computations in business intelligence. Perceptron model is also treated as one of the best and simplest types of Artificial Neural networks. However, it is a supervised learning algorithm of binary classifiers. Hence, we can consider it as a single-layer neural network with four main parameters, i.e., input values, weights and Bias, net sum, and an activation function.

Basic Components of Perceptron

Mr. Frank Rosenblatt invented the perceptron model as a binary classifier which contains three main components. These are as follows:



- **Input Nodes or Input Layer:**

This is the primary component of Perceptron which accepts the initial data into the system for further processing. Each input node contains a real numerical value.

- **Wight and Bias:**

Weight parameter represents the strength of the connection between units. This is another most important parameter of Perceptron components. Weight is directly proportional to the strength of the associated input neuron in deciding the output. Further, Bias can be considered as the line of intercept in a linear equation.

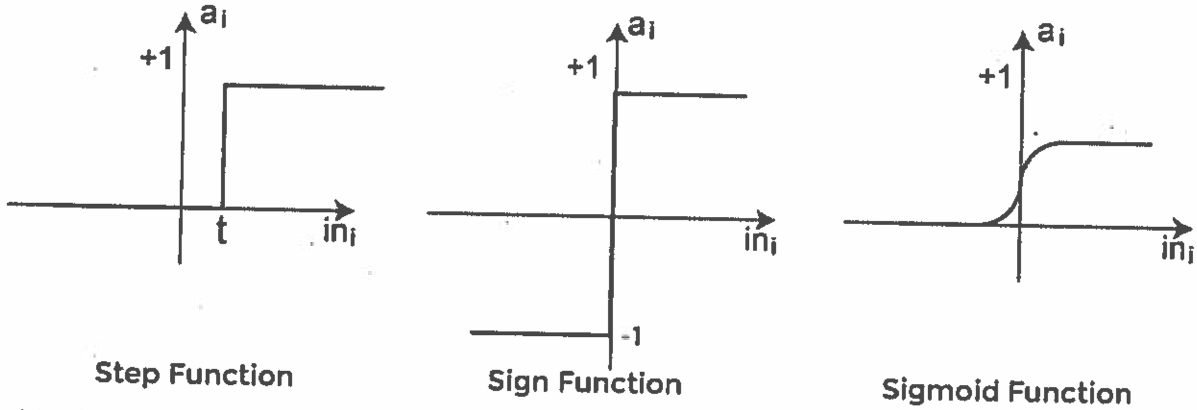
- **Activation Function:**

These are the final and important components that help to determine whether the neuron will fire or not. Activation Function can be considered primarily as a step function.

Types of Activation functions:

- Sign function

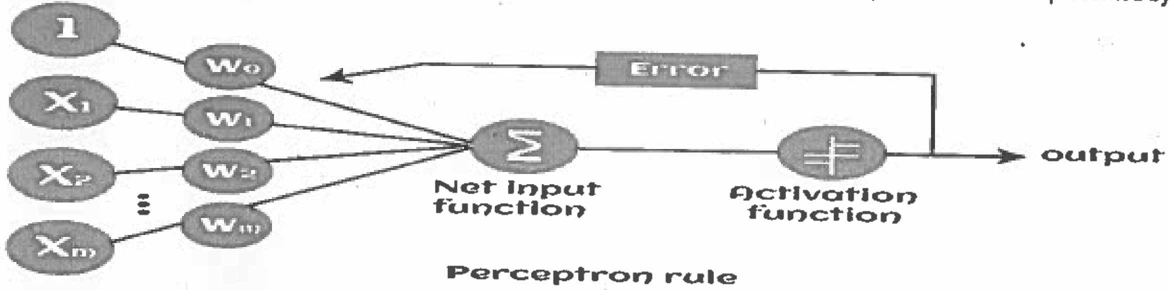
- Step function, and
- Sigmoid function



The data scientist uses the activation function to take a subjective decision based on various problem statements and forms the desired outputs. Activation function may differ (e.g., Sign, Step, and Sigmoid) in perceptron models by checking whether the learning process is slow or has vanishing or exploding gradients

How does Perceptron work?

In Machine Learning, Perceptron is considered as a single-layer neural network that consists of four main parameters named input values (Input nodes), weights and Bias, net sum, and an activation function. The perceptron model begins with the multiplication of all input values and their weights, then adds these values together to create the weighted sum. Then this weighted sum is applied to the activation function 'f' to obtain the desired output. This activation function is also known as the step function and is represented by 'f'.



Perceptron rule

This step function or Activation function plays a vital role in ensuring that output is mapped between required values (0,1) or (-1,1). It is important to note that the weight of input is indicative of the strength of a node. Similarly, an input's bias value gives the ability to shift the activation function curve up or down.

Perceptron model works in two important steps as follows:

Step-1

In the first step first, multiply all input values with corresponding weight values and then add them to determine the weighted sum. Mathematically, we can calculate the weighted sum as follows:

$\sum wi*xi = x1*w1 + x2*w2 + \ldots wn*xn$

Add a special term called bias 'b' to this weighted sum to improve the model's performance.

$\sum wi*xi + b$

Step-2

In the second step, an activation function is applied with the above-mentioned weighted sum, which gives us output either in binary form or a continuous value as follows:

$Y = f(\sum wi*xi + b)$

Types of Perceptron Models

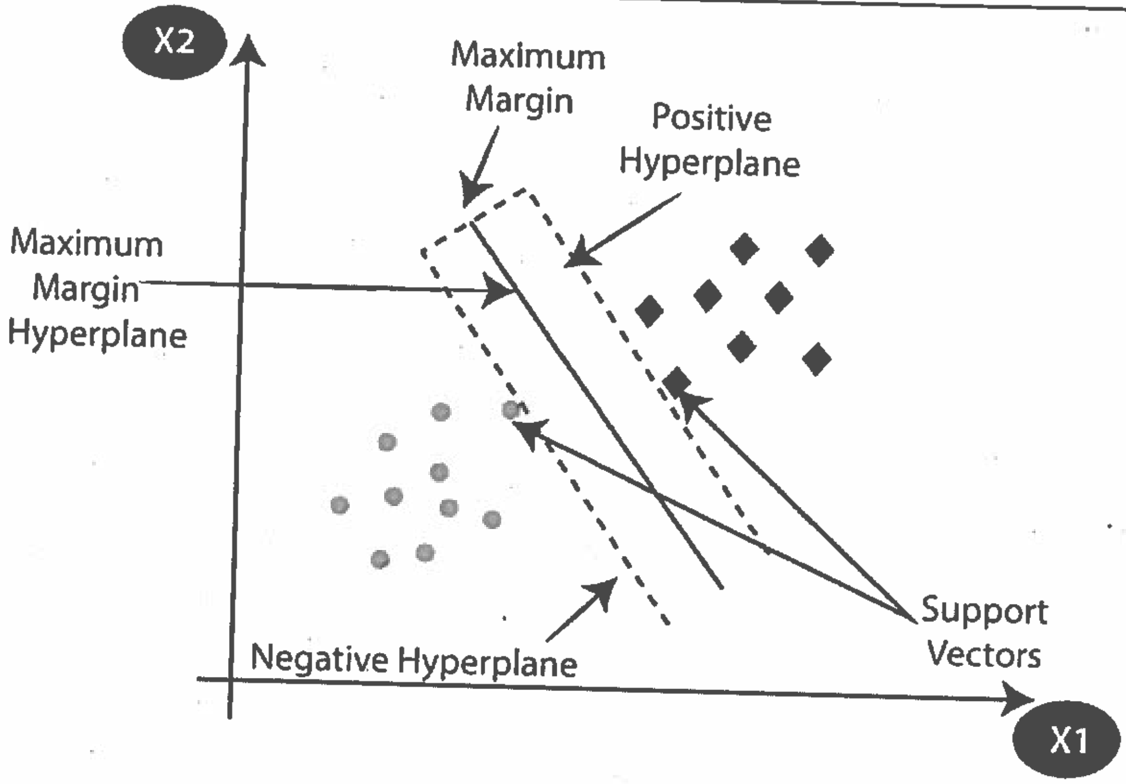
Based on the layers, Perceptron models are divided into two types. These are as follows:

1. Single-layer Perceptron Model
2. Multi-layer Perceptron model

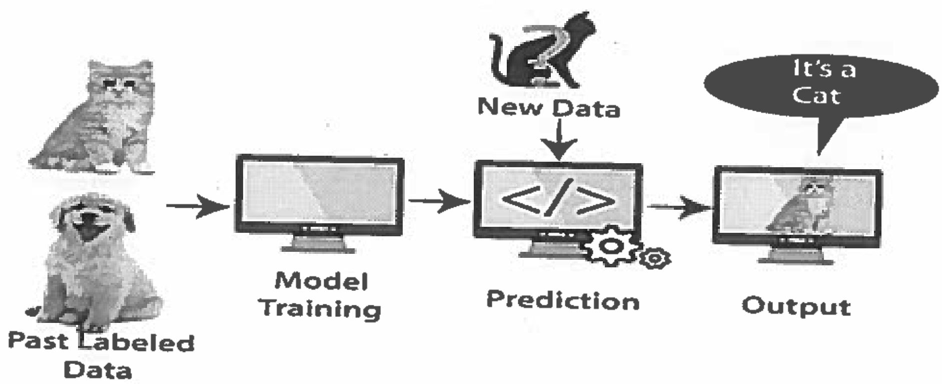
Single Layer Perceptron Model:

This is one of the easiest Artificial neural networks (ANN) types. A single-layered perceptron model consists feed-forward network and also includes a threshold transfer function inside the model. The main objective of the single-layer perceptron model is to analyze the linearly separable objects with binary outcomes.

	In a single layer perceptron model, its algorithms do not contain recorded data, so it begins with inconstantly allocated input for weight parameters. Further, it sums up all inputs (weight). After adding all Inputs, if the total sum of all inputs is more than a pre-determined value, the model gets activated and shows the output value as +1. If the outcome is same as pre-determined or threshold value, then the performance of this model is stated as satisfied, and weight demand does not change. However, this model consists of a few discrepancies triggered when multiple weight inputs values are fed into the model. Hence, to find desired output and minimize errors, some changes should be necessary for the weights input. "Single-layer perceptron can learn only linearly separable patterns." **Multi-Layered Perceptron Model:** Like a single-layer perceptron model, a multi-layer perceptron model also has the same model structure but has a greater number of hidden layers. The multi-layer perceptron model is also known as the Backpropagation algorithm, which executes in two stages as follows: o Forward Stage: Activation functions start from the input layer in the forward stage and terminate on the output layer. o Backward Stage: In the backward stage, weight and bias values are modified as per the model's requirement. In this stage, the error between actual output and demanded originated backward on the output layer and ended on the input layer. Hence, a multi-layered perceptron model has considered as multiple artificial neural networks having various layers in which activation function does not remain linear, similar to a single layer perceptron model. Instead of linear, activation function can be executed as sigmoid, TanH, ReLU, etc., for deployment. A multi-layer perceptron model has greater processing power and can process linear and non-linear patterns. Further, it can also implement logic gates such as AND, OR, XOR, NAND, NOT, XNOR, NOR. Advantages of Multi-Layer Perceptron: o A multi-layered perceptron model can be used to solve complex non-linear problems. o It works well with both small and large input data. o It helps us to obtain quick predictions after the training. o It helps to obtain the same accuracy ratio with large as well as small data. Disadvantages of Multi-Layer Perceptron: o In Multi-layer perceptron, computations are difficult and time-consuming. o In multi-layer Perceptron, it is difficult to predict how much the dependent variable affects each independent variable. o The model functioning depends on the quality of the training. Perceptron Function Perceptron function "f(x)" can be achieved as output by multiplying the input 'x' with the learned weight coefficient 'w'. Mathematically, we can express it as follows: f(x)=1; if w.x+b>0 otherwise, f(x)=0 o 'w' represents real-valued weights vector o 'b' represents the bias o 'x' represents a vector of input x values.
9	**Describe support vector machine algorithm with a suitable example.** Support Vector Machine or SVM is one of the most popular Supervised Learning algorithms, which is used for Classification as well as Regression problems. However, primarily, it is used for Classification problems in Machine Learning. The goal of the SVM algorithm is to create the best line or decision boundary that can segregate n-dimensional space into classes so that we can easily put the new data point in the correct category in the future. This best decision boundary is called a hyperplane. SVM chooses the extreme points/vectors that help in creating the hyperplane. These extreme cases are called as support vectors, and hence algorithm is termed as Support Vector Machine. Consider the below diagram in which there are two different categories that are classified using a decision boundary or hyperplane:



Example: SVM can be understood with the example that we have used in the KNN classifier. Suppose we see a strange cat that also has some features of dogs, so if we want a model that can accurately identify whether it is a cat or dog, so such a model can be created by using the SVM algorithm. We will first train our model with lots of images of cats and dogs so that it can learn about different features of cats and dogs, and then we test it with this strange creature. So as support vector creates a decision boundary between these two data (cat and dog) and choose extreme cases (support vectors), it will see the extreme case of cat and dog. On the basis of the support vectors, it will classify it as a cat. Consider the below diagram:



SVM algorithm can be used for Face detection, image classification, text categorization, etc.

Types of SVM
SVM can be of two types:

Linear SVM: Linear SVM is used for linearly separable data, which means if a dataset can be classified into two classes by using a single straight line, then such data is termed as linearly separable data, and classifier is used called as Linear SVM classifier.
Non-linear SVM: Non-Linear SVM is used for non-linearly separated data, which means if a dataset cannot be classified by using a straight line, then such data is termed as non-linear data and classifier used is called as Non-linear SVM classifier.
Hyperplane and Support Vectors in the SVM algorithm:

Hyperplane: There can be multiple lines/decision boundaries to segregate the classes in n-dimensional space, but we need to find out the best decision boundary that helps to classify the data points. This best boundary is known as the hyperplane of SVM.

The dimensions of the hyperplane depend on the features present in the dataset, which means if there are 2 features (as shown in image), then hyperplane will be a straight line. And if there are 3 features, then hyperplane will be a 2-dimension plane.

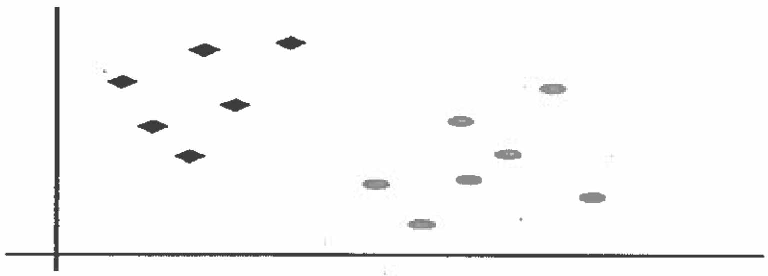
We always create a hyperplane that has a maximum margin, which means the maximum distance between the data points.

Support Vectors:

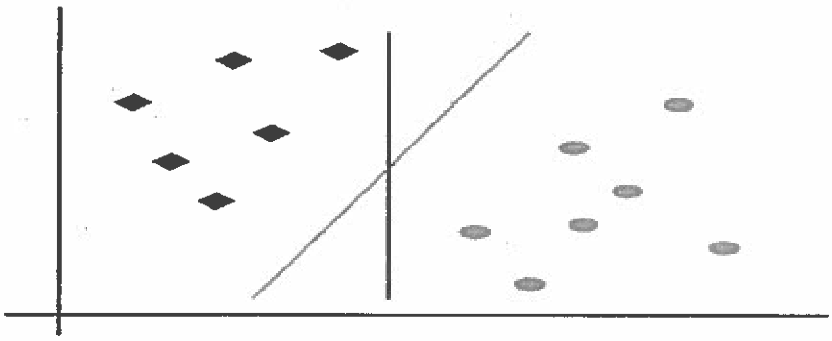
The data points or vectors that are the closest to the hyperplane and which affect the position of the hyperplane are termed as Support Vector. Since these vectors support the hyperplane, hence called a Support vector.

How does SVM works?
Linear SVM:

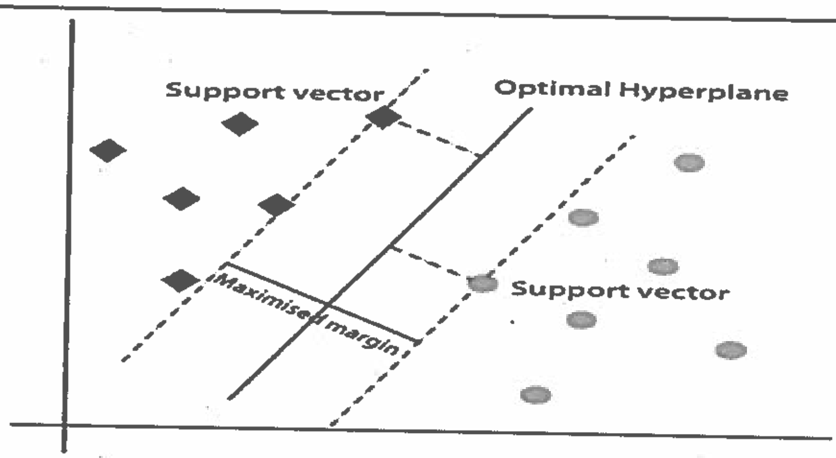
The working of the SVM algorithm can be understood by using an example. Suppose we have a dataset that has two tags (green and blue), and the dataset has two features x1 and x2. We want a classifier that can classify the pair(x1, x2) of coordinates in either green or blue. Consider the below image:



So as it is 2-d space so by just using a straight line, we can easily separate these two classes. But there can be multiple lines that can separate these classes. Consider the below image:

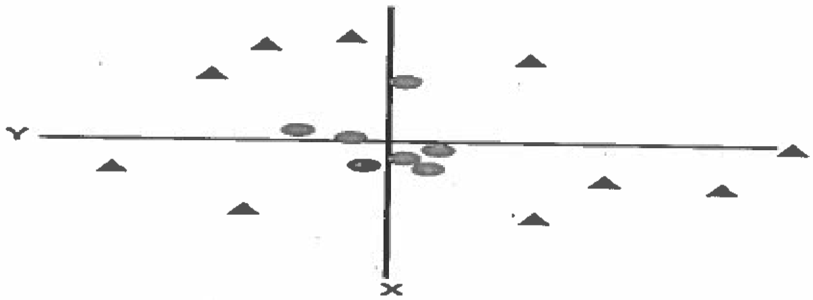


Hence, the SVM algorithm helps to find the best line or decision boundary; this best boundary or region is called as a **hyperplane**. SVM algorithm finds the closest point of the lines from both the classes. These points are called support vectors. The distance between the vectors and the hyperplane is called as **margin**. And the goal of SVM is to maximize this margin. The **hyperplane** with maximum margin is called the **optimal hyperplane**.



Non-Linear SVM:

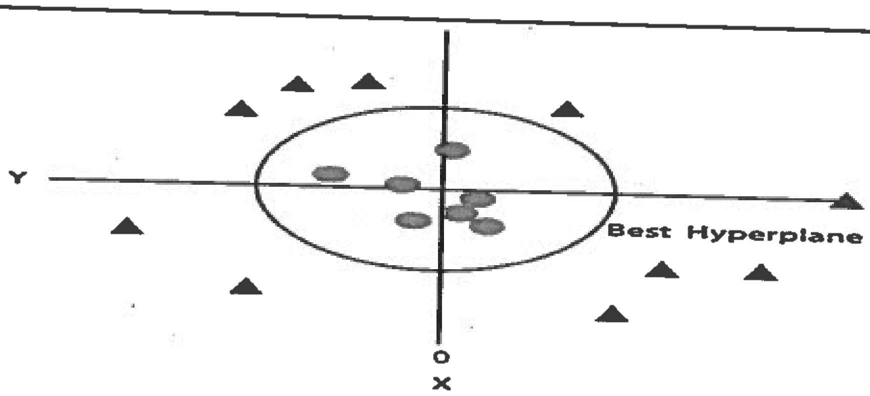
If data is linearly arranged, then we can separate it by using a straight line, but for non-linear data, we cannot draw a single straight line. Consider the below image:



So to separate these data points, we need to add one more dimension. For linear data, we have used two dimensions x and y, so for non-linear data, we will add a third dimension z. It can be calculated as:

$z=x^2+y^2$

By adding the third dimension, the sample space will become as below image:

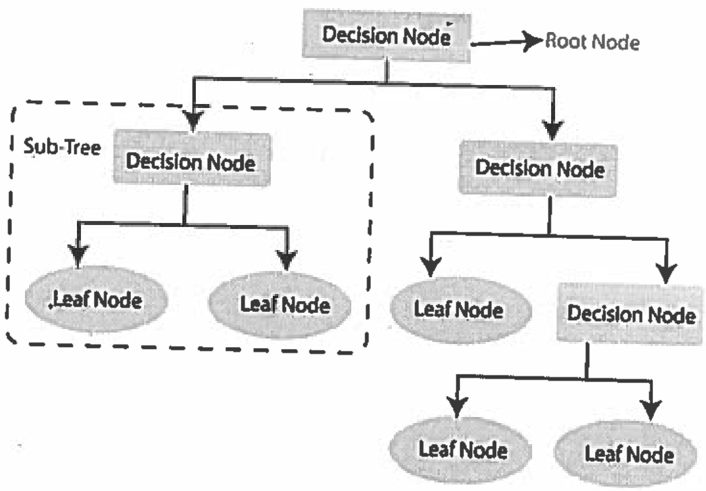


Hence, we get a circumference of radius 1 in case of non-linear data.

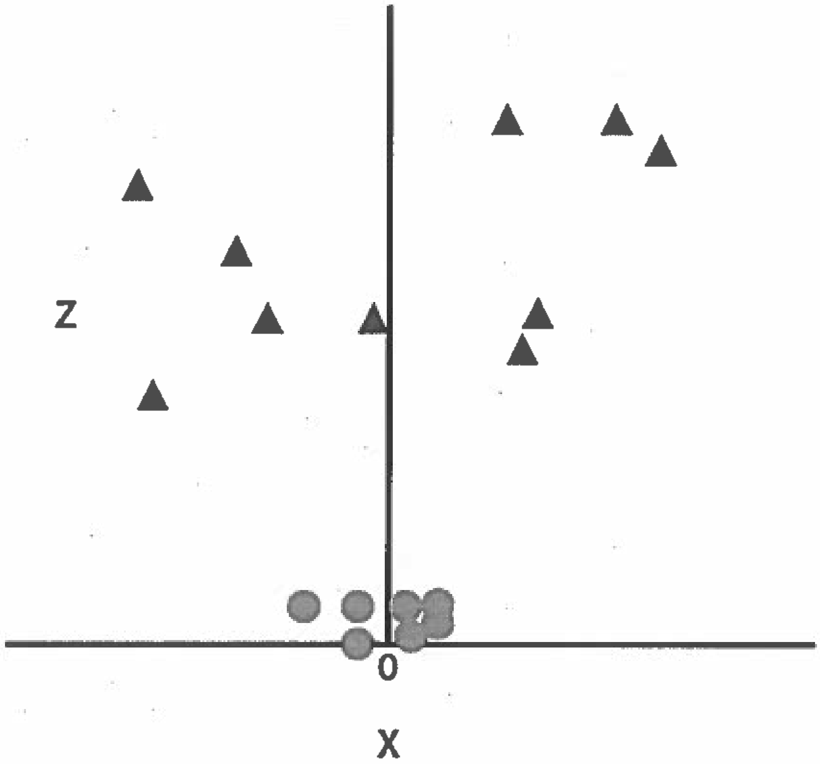
10 **Explain decision tree algorithm with a suitable application**

Decision Tree Classification Algorithm

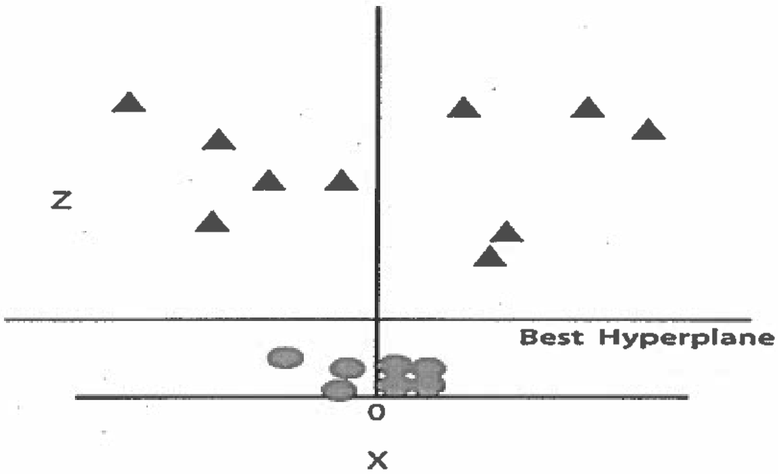
- o Decision Tree is a Supervised learning technique that can be used for both classification and Regression problems, but mostly it is preferred for solving Classification problems. It is a tree-structured classifier, where internal nodes represent the features of a dataset, branches represent the decision rules and each leaf node represents the outcome.
- o In a Decision tree, there are two nodes, which are the Decision Node and Leaf Node. Decision nodes are used to make any decision and have multiple branches, whereas Leaf nodes are the output of those decisions and do not contain any further branches.
- o The decisions or the test are performed on the basis of features of the given dataset.
- o *It is a graphical representation for getting all the possible solutions to a problem/decision based on given conditions.*
- o It is called a decision tree because, similar to a tree, it starts with the root node, which expands on further branches and constructs a tree-like structure.
- o In order to build a tree, we use the CART algorithm, which stands for Classification and Regression Tree algorithm.
- o A decision tree simply asks a question, and based on the answer (Yes/No), it further split the tree into subtrees.
- o Below diagram explains the general structure of a decision tree:



o



So now, SVM will divide the datasets into classes in the following way. Consider the below image:



Since we are in 3-d Space, hence it is looking like a plane parallel to the x-axis. If we convert it in 2d space with z=1, then it will become as:

	o It calculates how much information a feature provides us about a class. o According to the value of information gain, we split the node and build the decision tree. o A decision tree algorithm always tries to maximize the value of information gain, and a node/attribute having the highest information gain is split first. It can be calculated using the below formula: o Information Gain= Entropy(S)- [(Weighted Avg) *Entropy(each feature) **Entropy:** Entropy is a metric to measure the impurity in a given attribute. It specifies randomness in data. Entropy can be calculated as: Entropy(s)= -P(yes)log2 P(yes)- P(no) log2 P(no) **Where,** o **S= Total number of samples** o **P(yes)= probability of yes** o **P(no)= probability of no** 2. Gini Index: o Gini index is a measure of impurity or purity used while creating a decision tree in the CART(Classification and Regression Tree) algorithm. o An attribute with the low Gini index should be preferred as compared to the high Gini index. o It only creates binary splits, and the CART algorithm uses the Gini index to create binary splits. o Gini index can be calculated using the below formula: Gini Index= $1- \sum_j P_j^2$
11	**Describe Nearest neighbor algorithm with example.** **K-NEAREST NEIGHBOR METHODS:** K-Nearest Neighbour is one of the simplest Machine Learning algorithms based on Unsupervised Learning technique. • K-NN algorithm assumes the similarity between the new case/data and available cases and put the new case into the category that is most similar to the available categories. • K-NN algorithm stores all the available data and classifies a new data point based on the similarity. This means when new data appears then it can be easily classified into a well suite category by using K- NN algorithm. • K-NN algorithm can be used for Regression as well as for Classification but mostly it is used for the Classification problems. • K-NN is a non-parametric algorithm, which means it does not make any assumption on underlying data. • In ML, we have two different learners' algorithms as early learner and late learner and this method comes under lazy learner. **• Lazy Learner Algorithm:** It is also called a lazy learner algorithm, which it will means it will not learn function from the training data it will just memorize the training data. • It simply stores training data and wait until it gets as a test tuple. • It works only when it gets a new example. • Memorizing will be vanishing data immediately. • Learning and understanding will not vanish. • It is very easy to process the data/information.

Decision Tree Terminologies

- **Root Node:** Root node is from where the decision tree starts. It represents the entire dataset, which further gets divided into two or more homogeneous sets.

- **Leaf Node:** Leaf nodes are the final output node, and the tree cannot be segregated further after getting a leaf node.

- **Splitting:** Splitting is the process of dividing the decision node/root node into sub-nodes according to the given conditions.

- **Branch/Sub Tree:** A tree formed by splitting the tree.

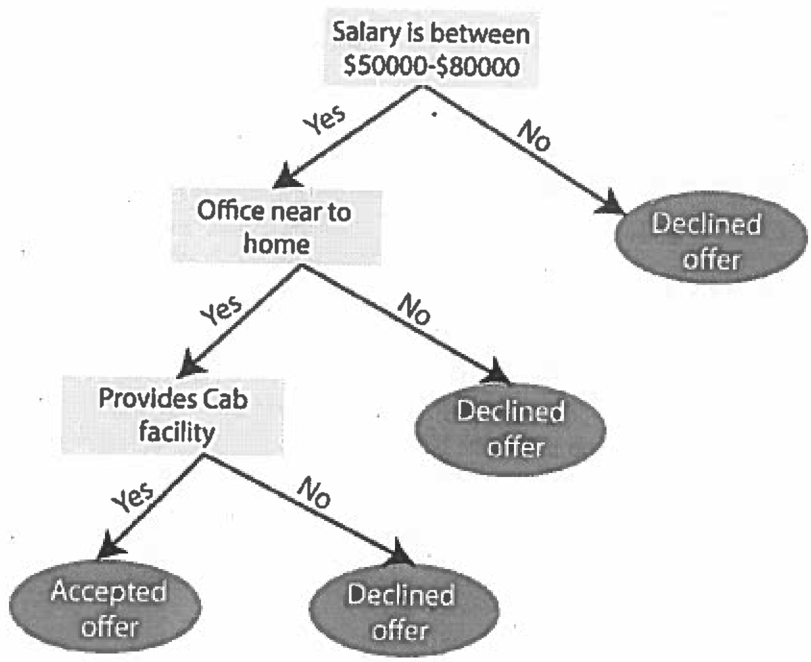
- **Pruning:** Pruning is the process of removing the unwanted branches from the tree.

- **Parent/Child node:** The root node of the tree is called the parent node, and other nodes are called the child nodes.

Algorithm:

- **Step-1:** Begin the tree with the root node, says.S, which contains the complete dataset.
- **Step-2:** Find the best attribute in the dataset using Attribute Selection Measure (ASM).
- **Step-3:** Divide the S into subsets that contains possible values for the best attributes.
- **Step-4:** Generate the decision tree node, which contains the best attribute.
- **Step-5:** Recursively make new decision trees using the subsets of the dataset created in step -3. Continue this process until a stage is reached where you cannot further classify the nodes and called the final node as a leaf node.

Example: Suppose there is a candidate who has a job offer and wants to decide whether he should accept the offer or Not. So, to solve this problem, the decision tree starts with the root node (Salary attribute by ASM). The root node splits further into the next decision node (distance from the office) and one leaf node based on the corresponding labels. The next decision node further gets split into one decision node (Cab facility) and one leaf node. Finally, the decision node splits into two leaf nodes (Accepted offers and Declined offer). Consider the below diagram:



1. Information Gain:

- Information gain is the measurement of changes in entropy after the segmentation of a dataset based on an attribute.

Ex- Given New data:-

$X = (\text{Maths} = 6, \ CP = 8)$

$K = 3$ (We can take)

↓

(It gives information about Nearest Neighbours)

Classification :- Fail / Pass

Given - Available Data

Roll no	Maths	CP	Result
1	4	3	F
2	6	7	P
3	7	8	P
4	5	5	F
5	8	8	P

$$d = \sqrt{(X_{01} - X_{A1})^2 + (X_{02} - X_{A2})^2}$$

Here O_1, O_2 values are fixed as they are taken by new data. $O_1 = 6, \ O_2 = 8$.

Calculating Distance values :-

$d_1 = \sqrt{(6-4)^2 + (8-3)^2} = \sqrt{(2)^2 + (5)^2} = \sqrt{4+25} = \sqrt{29} = 5.38$

$A_1 = 4, A_2 = 3$

$d_2 = \sqrt{(6-6)^2 + (8-7)^2} = \sqrt{(0)^2 + (1)^2} = \sqrt{0+1} = \sqrt{1} = 1$

$A_1 = 6, A_2 = 7$

$d_3 = \sqrt{(6-7)^2 + (8-8)^2} = \sqrt{(-1)^2 + (0)^2} = \sqrt{1+0} = \sqrt{1} = 1$

$A_1 = 7, A_2 = 8$

$d_4 = \sqrt{(6-5)^2 + (8-5)^2} = \sqrt{(1)^2 + (3)^2} = \sqrt{1+9} = \sqrt{10} = 3.16$

$A_1 = 5, A_2 = 5$

$d_5 = \sqrt{(6-8)^2 + (8-8)^2} = \sqrt{(-2)^2 + 0} = \sqrt{4+0} = \sqrt{4} = 2$

$A_1 = 8, A_2 = 8$

Algorithm:

Steps for implementing:

The K-NN working can be explained on the basis of the below algorithm:

- Step-1: Select the number K of the neighbors
- Step-2: Calculate the Euclidean distance of K number of neighbors
- Step-3: Take the K nearest neighbors as per the calculated Euclidean distance.
- Step-4: Among these k neighbors, count the number of the data points in each NEAREST NEIGHBOR METHODS IN ML 10 October 2022 02:39 PM New Section 1 Page 1 category.
- Step-5: Assign the new data points to that category for which the number of the neighbor is maximum.
- Step-6: Our model is ready.

The above formula will be used in this algorithm.

In which,

O= Observed Value

A= Actual Value Now, will take an example of a problem in which we will use the above algorithm and classify the new data by using the available data.

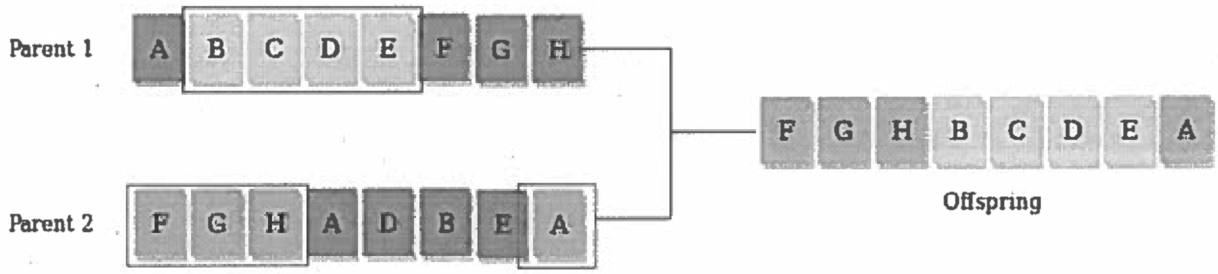
	Firstly, we need to take the input dataset and divide it into two subparts X and Y, where X is the training set, and Y is the validation set. **2. Representing data into a structure** Now we will represent our dataset into a structure. Such as we will represent the two-dimensional matrix of independent variable X. Here each row corresponds to the data items, and the column corresponds to the Features. The number of columns is the dimensions of the dataset. **3. Standardizing the data** In this step, we will standardize our dataset. Such as in a particular column, the features with high variance are more important compared to the features with lower variance. If the importance of features is independent of the variance of the feature, then we will divide each data item in a column with the standard deviation of the column. Here we will name the matrix as Z. **4. Calculating the Covariance of Z** To calculate the covariance of Z, we will take the matrix Z, and will transpose it. After transpose, we will multiply it by Z. The output matrix will be the Covariance matrix of Z. **5. Calculating the Eigen Values and Eigen Vectors** Now we need to calculate the eigenvalues and eigenvectors for the resultant covariance matrix Z. Eigenvectors or the covariance matrix are the directions of the axes with high information. And the coefficients of these eigenvectors are defined as the eigenvalues. **6. Sorting the Eigen Vectors** In this step, we will take all the eigenvalues and will sort them in decreasing order, which means from largest to smallest. And simultaneously sort the eigenvectors accordingly in matrix P of eigenvalues. The resultant matrix will be named as P*. **7. Calculating the new features Or Principal Components** Here we will calculate the new features. To do this, we will multiply the P* matrix to the Z. In the resultant matrix Z*, each observation is the linear combination of original features. Each column of the Z* matrix is independent of each other. **8. Remove less or unimportant features from the new dataset.** The new feature set has occurred, so we will decide here what to keep and what to remove. It means, we will only keep the relevant or important features in the new dataset, and unimportant features will be removed out. **Independent component analysis** Independent Component Analysis (ICA) is a machine learning technique to separate independent sources from a mixed signal. Unlike principal component analysis which focuses on maximizing the variance of the data points, the independent component analysis focuses on independence, i.e. independent components. X,S are column vector of random variables X=AS Where A is full constant matrix. PCA X be column vector of random variables. Q be orthogonal, S is uncorrelated. X=QS Works similar but independent works to find source from group where as PCA works to combine into group ./
13	**Explain genetic algorithm along with operators and applications.** A genetic algorithm is an adaptive heuristic search algorithm inspired by "Darwin's theory of evolution in Nature." It is used to solve optimization problems in machine learning. It is one of the important algorithms as it helps solve complex problems that would take a long time to solve. Genetic Algorithms are being widely used in different real-world applications, for example, Designing electronic circuits, code-breaking, image processing, and artificial creativity. In this topic, we will explain Genetic algorithm in detail, including basic terminologies used in Genetic algorithm, how it works, advantages and limitations of genetic algorithm, etc. Population: Population is the subset of all possible or probable solutions, which can solve the given problem. Chromosomes: A chromosome is one of the solutions in the population for the given problem, and the collection of gene generate a chromosome.

	As we need to consider, (k=3) three nearest neighbour values from distance values, So, we consider, d_2 first as its has least value as $d_2 = 1$, then $d_3 = 1$ as they are equal to each other and near to each. Finally we take $d_5 = 2-1 = 1$ the difference is as compared to d_1 and d_4 so we take d_5 d_2, d_3, d_5: from available Data, d_2, d_3, d_5 all three are pass. So, the new data of 'x' student pass. ∴ We classified 'x' student as pass (P) by using K- Nearest Neighbour Alg Method.
12	**Explain about principal component analysis and independent component analysis.** **Principal Component Analysis** Principal Component Analysis is an unsupervised learning algorithm that is used for the dimensionality reduction in machine learning. It is a statistical process that converts the observations of correlated features into a set of linearly uncorrelated features with the help of orthogonal transformation. These new transformed features are called the Principal Components. **Steps for PCA algorithm** 1. **Getting the dataset**

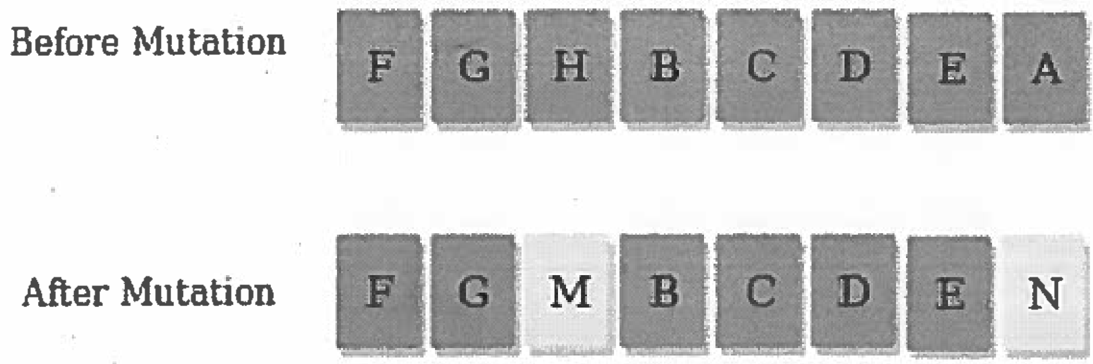
	Characters A-Z, a-z, 0-9, and other special symbols are considered as genes A string generated by these characters is considered as chromosome/solution/Individual Fitness score is the number of characters which differ from characters in target string at a particular index. So individual having lower fitness value are given more preference.
14)	**Explain about Monte carlo methods.** Monte Carlo techniques involves three basic steps: 1. Set up the predictive model, identifying both the dependent variable to be predicted and the independent variables (also known as the input, risk or predictor variables) that will drive the prediction. 2. Specify probability distributions of the independent variables. Use historical data and/or the analyst's subjective judgment to define a range of likely values and assign probability weights for each. 3. Run simulations repeatedly, generating random values of the independent variables. Do this until enough results are gathered to make up a representative sample of the near infinite number of possible combinations. You can run as many Monte Carlo Simulations as you wish by modifying the underlying parameters you use to simulate the data. However, you'll also want to compute the range of variation within a sample by calculating the variance and standard deviation, which are commonly used measures of spread. Variance of given variable is the expected value of the squared difference between the variable and its expected value. Standard deviation is the square root of variance. Typically, smaller variances are considered better. The technique breaks down into five simple steps: 1. Setting up a probability distribution for important variables. 2. Building a cumulative probability distribution for each variable. 3. Establishing an interval of random numbers for each variable. 4. Generating random numbers. 5. Actually simulating a series of trials. How does Monte Carlo Simulation work? Unlike a normal forecasting model, Monte Carlo Simulation predicts a set of outcomes based on an estimated range of values versus a set of fixed input values. In other words, a Monte Carlo Simulation builds a model of possible results by leveraging a probability distribution, such as a uniform or normal distribution, for any variable that has inherent uncertainty. It, then, recalculates the results over and over, each time using a different set of random numbers between the minimum and maximum values. In a typical Monte Carlo experiment, this exercise can be repeated thousands of times to produce a large number of likely outcomes. Monte Carlo Simulations are also utilized for long-term predictions due to their accuracy. As the number of inputs increase, the number of forecasts also grows, allowing you to project outcomes farther out in time with more accuracy. When a Monte Carlo Simulation is complete, it yields a range of possible outcomes with the probability of each result occurring.

Gene: A chromosome is divided into a different gene, or it is an element of the chromosome.
Allele: Allele is the value provided to the gene within a particular chromosome.
Fitness Function: The fitness function is used to determine the individual's fitness level in the population. It means the ability of an individual to compete with other individuals. In every iteration, individuals are evaluated based on their fitness function.
Genetic Operators: In a genetic algorithm, the best individual mate to regenerate offspring better than parents. Here genetic operators play a role in changing the genetic composition of the next generation.
Operators of Genetic Algorithms

Once the initial generation is created, the algorithm evolves the generation using following operators –
1) Selection Operator: The idea is to give preference to the individuals with good fitness scores and allow them to pass their genes to successive generations.
2) Crossover Operator: This represents mating between individuals. Two individuals are selected using selection operator and crossover sites are chosen randomly. Then the genes at these crossover sites are exchanged thus creating a completely new individual (offspring). For example –



3) Mutation Operator: The key idea is to insert random genes in offspring to maintain the diversity in the population to avoid premature convergence. For example –



The whole algorithm can be summarized as –

1) Randomly initialize populations p
2) Determine fitness of population
3) Until convergence repeat:
    a) Select parents from population
    b) Crossover and generate new population
    c) Perform mutation on new population
    d) Calculate fitness for new population

Example problem and solution using Genetic Algorithms

Given a target string, the goal is to produce target string starting from a random string of the same length. In the following implementation, following analogies are made –

	One simple example of a Monte Carlo Simulation is to consider calculating the probability of rolling two standard dice. There are 36 combinations of dice rolls. Based on this, you can manually compute the probability of a particular outcome. Using a Monte Carlo Simulation, you can simulate rolling the dice 10,000 times (or more) to achieve more accurate predictions.
15	**Describe Bayesian network**

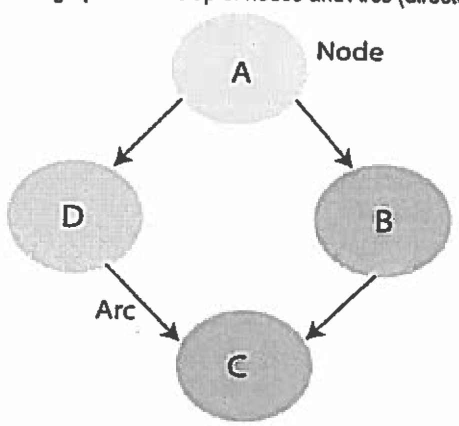
"A Bayesian network is a probabilistic graphical model which represents a set of variables and their conditional dependencies using a directed acyclic graph."

It is also called a Bayes network, belief network, decision network, or Bayesian model.

Bayesian networks are probabilistic, because these networks are built from a probability distribution, and also use probability theory for prediction and anomaly detection.
Bayesian Network can be used for building models from data and experts opinions, and it consists of two parts:

Directed Acyclic Graph
Table of conditional probabilities.
The generalized form of Bayesian network that represents and solve decision problems under uncertain knowledge is known as an Influence diagram.

A Bayesian network graph is made up of nodes and Arcs (directed links), where:



Each node corresponds to the random variables, and a variable can be continuous or discrete.
Arc or directed arrows represent the causal relationship or conditional probabilities between random variables. These directed links or arrows connect the pair of nodes in the graph.
These links represent that one node directly influence the other node, and if there is no directed link that means that nodes are independent with each other
In the above diagram, A, B, C, and D are random variables represented by the nodes of the network graph.
If we are considering node B, which is connected with node A by a directed arrow, then node A is called the parent of Node B.
Node C is independent of node A.
Joint probability distribution:
If we have variables x1, x2, x3,....., xn, then the probabilities of a different combination of x1, x2, x3.. xn, are known as Joint probability distribution.
P[x1, x2, x3,....., xn], it can be written as the following way in terms of the joint probability distribution.

= P[x1| x2, x3,....., xn]P[x2, x3,....., xn]

= P[x1| x2, x3,....., xn]P[x2|x3,....., xn]....P[xn-1|xn]P[xn].

In general for each variable Xi, we can write the equation as:

P(Xi|Xi-1,........., X1) = P(Xi |Parents(Xi))

SCHEME OF EVALUATION

1	Definition of machine learning.	2M
2	List the applications of machine learning.	2M
3	Definitions of classification and clustering.	2M
4	Listing two genetic operators.	2M
5	Definition of baye's theorem.	2M
6(a)	Supervised Learning - 2 M Examples-1 M Unsupervised Learning -2 M Examples- 1 M	6M
6(b)	Definition Of Linear Regression-2M Diagram-2M Derivation and Example-2M	6M
7	Describe about vectors and candidate elimination algorithm. vectors-2M Candidate Elimination Definition and Algorithm-10M	12M
8	Explain multi-layered perceptron model with suitable example. Perceptron-2M Types Of Preceptrons - 3M Multilayered perceptron with diagram and example-7M	12M
9	Describe support vector machine algorithm with a suitable example. Support Vector Machine Definition and diagram-8M Exam of SVM-4M	12M
10	Explain decision tree algorithm with a suitable application. Decision Tree Definition-2M Algorithm/Steps -3M Example and final representation of Decision Tree-7M	12M
11	Describe Nearest neighbor algorithm with example. Definition of K-NN-3M Algorithm-3M Example and final solution-7M	12M
12	Explain about principal component analysis and independent component analysis. PCA Definition-2M Steps for PCA-5M ICA Definition-2M Formula/Derivation of ICA-3M	12M
13	Explain genetic algorithms along with operators and applications. Definition-3M Types of operators with examples-9M	12M
14	Explain about Monte carlo methods. Definition-3M Basic Steps - 9M	12M
15	Describe Bayesian network. Definition-3M Diagrams-2M Description of diagrams-3M Description-4M	12M

NBhargavi

Semester End Regular Examination, Nov/Dec., 2022

Degree	B. Tech.	Program	Common to All	Academic Year	2021 - 2022
Course Code	20DSO01	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Introduction to Database Management Systems (Open Elective)				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any 4 database applications	20DSO01.1	L1
2	What is an Entity?	20DSO01.2	L1
3	Define the term Data Dictionary.	20DSO01.4	L1
4	How to create table with syntax in SQL?	20DSO01.3	L1
5	Define DROP command with syntax.	20DSO01.5	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain Database and DBMS along with the importance of database design.	6M	20DSO01.1	L2
6 (b)	What are the problems in file system data management? Explain with relevant example.	6M	20DSO01.1	L2
	OR			
7	Explain different data models.	12M	20DSO01.1	L2
8 (a)	Discuss correlated nested queries along with a query to find the names of sailors who have reserved a red boat?	6M	20DSO01.2	L2
8 (b)	Explain different types of joins with examples	6M	20DSO01.2	L2
	OR			
9	Explain about ER model.	12M	20DSO01.2	L2
10 (a)	Explain various DML, DDL commands in SQL.	6M	20DSO01.3	L2
10 (b)	Explain Set Operators	6M	20DSO01.3	L2
	OR			
11 (a)	Explain Order by, Group by and Having Clauses with example	6M	20DSO01.3	L2
11 (b)	Explain Aggregate functions with examples	6M	20DSO01.3	L2
12 (a)	Explain the components of PL/SQL block.	6M	20AI603.4	L2
12 (b)	Explain about control statements in PL/SQL block	6M	20AI603.4	L2
	OR			
13 (a)	Explain about triggers.	6M	20AI603.4	L2
13 (b)	Explain about cursors.	6M	20AI603.4	L2
14	Explain about Normalization and need for normalization along with the problems caused by Redundancy	12M	20AI603.5	L2
	OR			
15	Explain Third NF and BCNF with relevant table structure.	12M	20AI603.5	L2

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SCHEME OF VALUATION
&
ANSWER KEY

Degree	B. Tech. (U. G.)	**Program**	CSE	**Test**	**END EXAM**	**Academic Year**	2022 - 2023
Course Code	20DSO01	**Test Duration**	180 Min.	**Max. Marks**	70	**Semester**	I
Course	**INTRODUCTION TO DATABASE MANAGEMENT SYSTEM**						

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	**List any 4 database applications**	20DSO01.1	**L1**

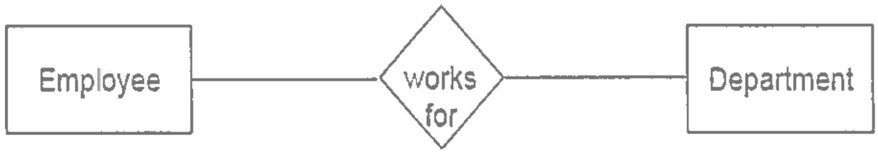
For this we can give two marks for any four applications of database

Railway reservation system
Library management system
Banking
Education sector
Credit card exchanges
Social media sites
Online shopping
Broadcast communication
Airline reservation system

2	**What is an Entity?**	20DSO01.2	L1

Entities – It is a real-world thing which can be a person, place, or even a concept. For Example: Department, Admin, Courses, Teachers, Students, Building, etc are some of the entities of a School Management System.

- Consider an organization as an example- manager, product, employee, department etc. can be taken as an entity.



3 **Define the term Data Dictionary.** 20AI603.3 **L1**

A Data Dictionary is a collection of names, definitions, and attributes about data elements that are being used or captured in a database, information system, or part of a research project. It describes the meanings and purposes of data elements within the context of a project, and provides guidance on interpretation, accepted meanings and representation. A Data Dictionary also provides metadata about data elements.

4 **How to create table with syntax in SQL?** 20DSO01.4 L1

The CREATE TABLE statement is used to create a new table in a database.

Syntax

CREATE TABLE *table_name* (
   column1 datatype,
   column2 datatype,
   column3 datatype,
  ....
);

The column parameters specify the names of the columns of the table.

The datatype parameter specifies the type of data the column can hold (e.g. varchar, integer, date, etc.).

5 **Define DROP command with syntax.** 20DSO01.5 L1

DROP is used to delete a whole database or just a table.The DROP statement destroys the objects like an existing database, table, index, or view.

A DROP statement in SQL removes a component from a relational database management system (RDBMS).

Syntax:

DROP object object_name

Examples:

DROP TABLE table_name;

table_name: Name of the table to be deleted.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Learning Outcome (s)	DoK
6 (a)	Explain Database and DBMS along with the importance of database	6M	20DSO01.1

		Outcome (s)	
6 (a)	**Explain Database and DBMS along with the importance of database design.** **For writing about databse and DBMS – 3 marks** **For writing about importance of database design. – 3 marks**	**6M**	20DSO01.1

A Database is a collection of related data organized in a way that data can be easily accessed, managed and updated. Any piece of information can be a data, for example name of your school. Database is actually a place where related piece of information is stored and various operations can be performed on it.

A **DBMS** is a software that allows creation, definition and manipulation of database. Dbms is actualy a tool used to perform any kind of operation on data in database. Dbms also provides protection and security to database. It maintains data consistency in case of multiple users. Here are some examples of popular dbms, MySql, Oracle, Sybase, Microsoft Access and IBM DB2 etc.

Functions of DBMS

- Provides data Independence
- Concurrency Control
- Provides Recovery services
- Provides Utility services
- Provides a clear and logical view of the process that manipulates data.

Advantages of DBMS

- Segregation of applicaion program.
- Minimal data duplicacy.
- Easy retrieval of data.
- Reduced development time and maintainance need.

6 (b)	**What are the problems in file system data management? Explain with relevant example.** **For explaning disadvantages of file systems – 6marks**	**6M**	20DSO01.1

DISADVANTAGES OF FILE SYSTEM DATA MANAGEMENT

1. Data redundancy and inconsistency.
2. Integrity Problems.
3. Security Problems
4. Difficulty in accessing data.
5. Data isolation

6. Limited Data Sharing
7. Atomicity Problems

1) Data redundancy and inconsistency:
Data redundancy means duplication of data and inconsistency means that the duplicated values are different.

2) Integrity problems:
Data integrity means that the data values in the data base should be accurate in the sense that the value must satisfy some rules.

3) Security Problem:
Data security means prevention of data accession by unauthorized users.

4) Difficulty in accessing data:
Difficulty in accessing data arises whenever there is no application program for a specific task.

5) Data isolation:
This problem arises due to the scattering of data in various files with various formats. Due to the above disadvantages of the earlier data processing system, the necessity for an effective data processing system arises. Only at that time the concept of DBMS emerges for the rescue of a large number of organizations.

OR

7	**Explain different data models.** **For explaining all the three models – 12 marks**	**12M**	20DSO01.1

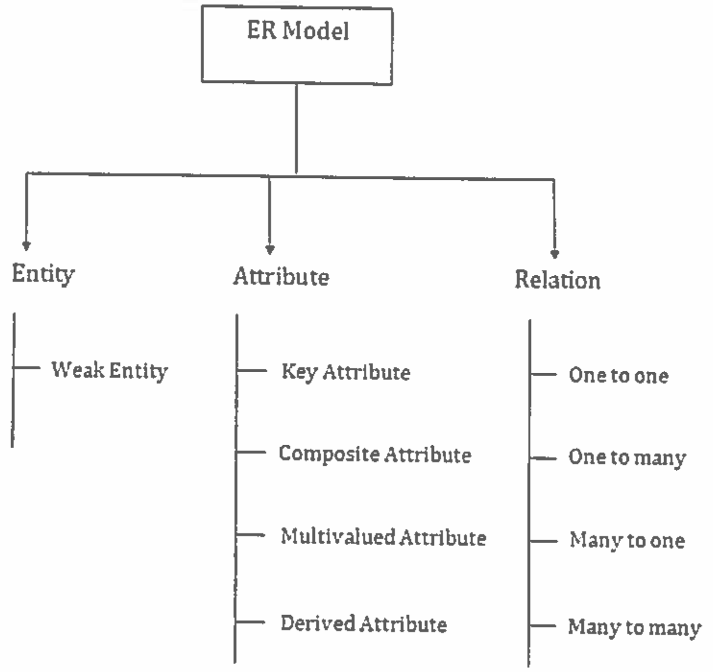
<u>DATA MODELS:</u>

- Data Model is the modeling of the data description, data semantics, and consistency constraints of the data.
- It provides the conceptual tools for describing the design of a database at each level of data abstraction.
- The various data models that have been proposed fall into three different groups:
    1) Object – based logical models
    2) Record – based logical models
    3) Physical data models

Object – Based Logical Models :
Object –based logical models are used in describing data at the conceptual and view levels.
They are characterized by the fact that they provide fairly flexible structuring capabilities and allow data constraints to be specified explicitly.
There are many different models:

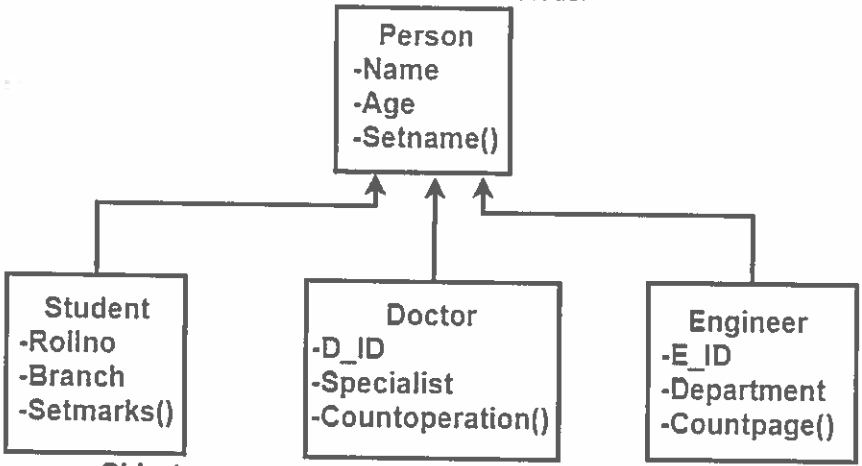
a) The entity- relationship model
b) The object- oriented model
c) The binary model
d) The semantic data model
e) The infological model
f) The functional data model

The Entity- Relationship Model :
Entity relationship (ER) models are based on the real-world entities and their relationships.



B) Object-Oriented Model:

- Like ER model, the object-oriented model is based on a collection of objects
- An object contains values stored in instance variables within objects.
- Unlike the record – oriented models, these values are themselves objects.
- Thus, objects contain objects objects to an arbitrarily deep level of nesting.
- An object also contains bodies of code that operate on the object.
- These bodies of code are called methods.



- **Objects** –
  An object is an abstraction of a real world entity or we can say it is an instance of class. Objects encapsulates data and code into a single unit which provide data abstraction by hiding the implementation details from the user. For example: Instances of student, doctor, engineer in above figure.

- **Attribute** –
  An attribute describes the properties of object. For example: Object is STUDENT and its attribute are Roll no, Branch, Setmarks() in the Student class.

- **Methods** –
  Method represents the behavior of an object. Basically, it represents the real-world action. For example: Finding a STUDENT marks in above figure as Setmarks().

- **Class**
  A class is a collection of similar objects with shared structure i.e. attributes and behavior i.e. methods. An object is an instance of class. For example: Person, Student, Doctor, Engineer in above figure.

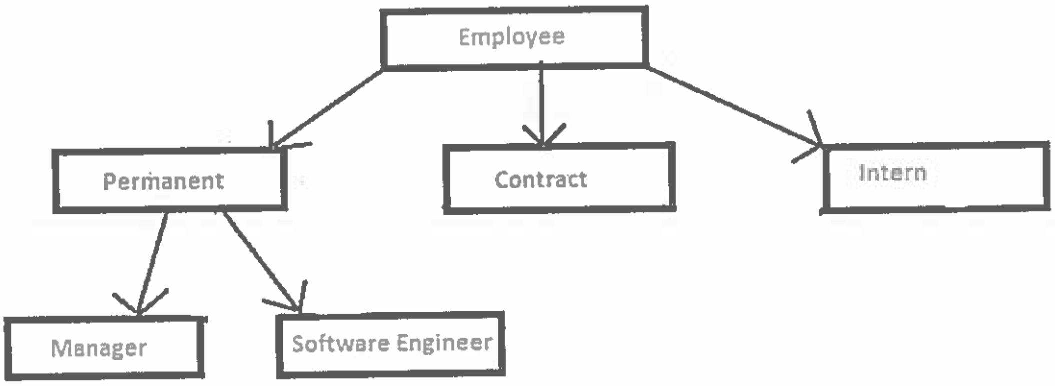
```
class student
{
char Name[20];
int roll_no;
–
–
public:
void search();
void update();
}
```

RECORD-BASED DATA MODEL :

When the database is organized in some fixed format of records of several than the model is known as Record-Based Data Model.
It has a fixed number of fields or attributes in each record type and each field is usually of a fixed length.
Further, it is classified into three types-

1. **Hierarchical Data Model :**
   In hierarchical type, the model data are represented by collection of records. In this, relationships among the data are represented by links. In this model, tree data structure is used.

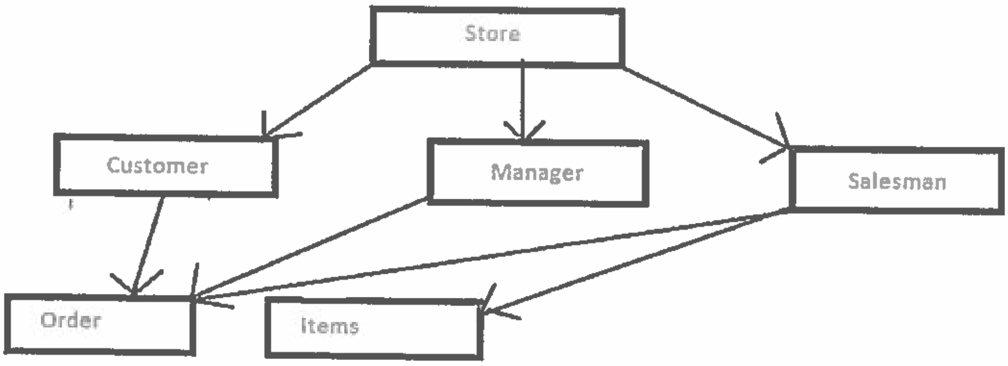
It was developed in 1960s by IBM, to manage large amount of data for complex manufacturing projects. The basic logic structure of hierarchical data model is upside-down "tree".



Advantages
Simplicity, Data Integrity, Data security, Efficiency, Easy availability of expertise.
Disadvantages
Complexity, Inflexibility, Lack of Data Independence, Lack of querying facility, Data Manipulation Language, Lack Of standards.

2. **Network Data Model :**
   In network type, the model data are represented by collection of records. In this, relationships among the data are represented by links. Graph data structures are used in this model. It permits a record to have more than one parent.

For Example- Social Media sites like Facebook, Instagram etc.



Advantages
Simplicity, Data Integrity, Data Independence, Database standards.
Disadvantages
System Complexity, Lack of structural Independence.

3. **Relational Data Model :**
   Relational Data Model uses tables to represent the data and the relationship among these data. Each table has multiple columns and each column is identified by a unique name. It is a low-level model.

Activity Code	Activity Name
23	Patching
24	Overlay
25	Crack Sealing

Key=24

Activity Code	Date	Route No.
24	01/12/01	I 95
24	02/08/01	I 66

Date	Activity Code	Route No.
01/12/01	24	I-95
01/15/01	23	I-495
02/08/01	24	I-66

Advantages
Structural Independence, Simplicity, Ease of designing, Implementation, Ad-Hoc query capability.
Disadvantages
Hardware Overheads, Ease of design can result in bad design.

PHYSICAL DATA MODELS:
Physical data models are used to describe data at the lowest level.
In contrast to logical data models, there are very few physical data models in use. Two of the widely known ones are :

a) Unifying model
b) Frame memory

8 (a)	Discuss correlated nested queries along with a query to find the names of sailors who have reserved a red boat? For explaining about nested queries – 4 marks For writing query – 2marks	20DSO01.2	L2

Nested queries :
In nested queries, a query is written inside a query. The result of inner query is used in execution of outer query. We will use STUDENT, COURSE, STUDENT_COURSE tables for understanding nested queries.

There are mainly two types of nested queries:

1. Independent Nested Queries: In independent nested queries, query execution starts from innermost query to outermost queries. The execution of inner query is independent of outer query, but the result of inner query is used in execution of outer query. Various operators like IN, NOT IN, ANY, ALL etc are used in writing independent nested queries.
2. Co-related Nested Queries: In co-related nested queries, the output of inner query depends on the row which is being currently executed in outer query. e.g.; If we want to find out S_NAME of STUDENTs who are enrolled in C_ID 'C1', it can be done with the help of co-related nested query as:

Select S_NAME from STUDENT S where EXISTS (select * from STUDENT_COURSE SC where S.S_ID=SC.S_ID and SC.C_ID='C1');

Find the names of sailors who have reserved a red boat, and list in the order of age.

SELECT S.sname, S.age FROM Sailors S, Reserves R, Boats B WHERE S.sid = R.sid AND R.bid = B.bid AND B.color = 'red'
ORDER BY S.age ORDER BY S.age [ASC] (default)
ORDER BY S.age DESC

Tables used in this note:

Sailors(sid: integer, sname: string, rating: integer, age: real);

Boats(bid: integer, bname: string, color: string);

Reserves(sid: integer, bid: integer, day: date).

Sailors

Sid	Sname	Rating	Age
22	Dustin	7	45
29	Brutus	1	33
31	Lubber	8	55.5
32	Andy	8	25.5
58	Rusty	10	35
64	Horatio	7	35
71	Zorba	10	16
74	Horatio	9	40
85	Art	3	25.5
95	Bob	3	63.5

Boats

bid	bname	color
101	Interlake	blue
102	Interlake	red
103	Clipper	green
104	Marine	red

Reserves

sid	bid	day
22	101	1998-10-10
22	102	1998-10-10
22	103	1998-10-8
22	104	1998-10-7
31	102	1998-11-10
31	103	1998-11-6
31	104	1998-11-12
64	101	1998-9-5
64	102	1998-9-8
74	103	1998-9-8

Figure 1: Instances of Sailors, Boats and Reserves

Create the Tables:

```
CREATE TABLE sailors ( sid integer not null,
                       sname varchar(32),
                       rating integer,
                       age real,
                       CONSTRAINT PK_sailors
                       PRIMARY KEY (sid) );
CREATE TABLE reserves ( sid integer not null,
                        bid integer not null,
                        day datetime not null,
                        CONSTRAINT PK_reserves
                        PRIMARY KEY (sid, bid, day),
                        FOREIGN KEY (sid) REFERENCES sailors(sid),
                        FOREIGN KEY (bid) REFERENCES boats(bid) );
```

8 (b)	**Explain different types of joins with examples** **For explaining about four types of joins - 6 Marks**	**6M**	20DSO01.2

SQL JOIN

As the name shows, JOIN means to combine something. In case of SQL, JOIN means "to combine two or more tables".

In SQL, JOIN clause is used to combine the records from two or more tables in a database.

Types of SQL JOIN

1. INNER JOIN
2. LEFT JOIN
3. RIGHT JOIN
4. FULL JOIN

1. INNER JOIN

In SQL, INNER JOIN selects records that have matching values in both tables as long as the condition is satisfied. It returns the combination of all rows from both the tables where the condition satisfies.

Syntax

```
SELECT table1.column1, table1.column2, table2.column1,....
        FROM table1
        INNER JOIN table2
        ON table1.matching_column = table2.matching_column;
```

Query

```
SELECT EMPLOYEE.EMP_NAME, PROJECT.DEPARTMENT
        FROM EMPLOYEE
        INNER JOIN PROJECT
        ON PROJECT.EMP_ID = EMPLOYEE.EMP_ID;
```

2. LEFT JOIN

The SQL left join returns all the values from left table and the matching values from the right table. If there is no matching join value, it will return NULL.

Syntax

```
SELECT table1.column1, table1.column2, table2.column1,....
FROM table1
LEFT JOIN table2
ON table1.matching_column = table2.matching_column;
```

Query

```
SELECT EMPLOYEE.EMP_NAME, PROJECT.DEPARTMENT
FROM EMPLOYEE
```

LEFT JOIN PROJECT
ON PROJECT.EMP_ID = EMPLOYEE.EMP_ID;

3. RIGHT JOIN

In SQL, RIGHT JOIN returns all the values from the values from the rows of right table and the matched values from the left table. If there is no matching in both tables, it will return NULL.

Syntax

```
SELECT table1.column1, table1.column2, table2.column1,....
  FROM table1
   RIGHT JOIN table2
                    ON table1.matching_column = table2.matching_column;
```

Query

```
SELECT EMPLOYEE.EMP_NAME, PROJECT.DEPARTMENT
FROM EMPLOYEE
RIGHT JOIN PROJECT
ON PROJECT.EMP_ID = EMPLOYEE.EMP_ID;
```

4. FULL JOIN

In SQL, FULL JOIN is the result of a combination of both left and right outer join. Join tables have all the records from both tables. It puts NULL on the place of matches not found.

Syntax

```
 SELECT table1.column1, table1.column2, table2.column1,....
          FROM table1
          FULL JOIN table2
          ON table1.matching_column = table2.matching_column;
```

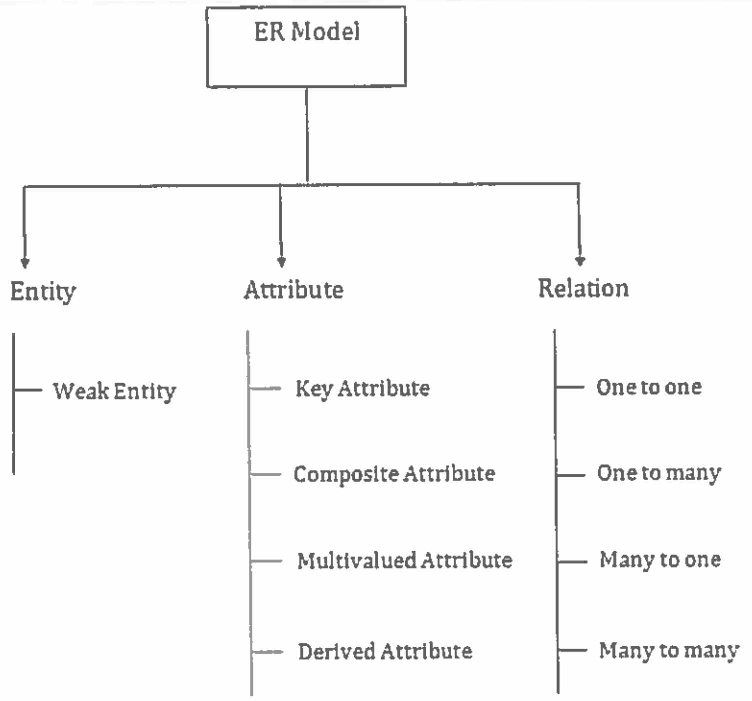
Query

```
SELECT EMPLOYEE.EMP_NAME, PROJECT.DEPARTMENT
         FROM EMPLOYEE
FULL JOIN PROJECT
ON PROJECT.EMP_ID = EMPLOYEE.EMP_ID;
```

9	**Explain about ER model.** **For explaining about entity - 4 Marks** **For explaining about Attribute - 4 Marks** **For explaining about Relation - 4 Marks**	**12M**	**20AI603.2**

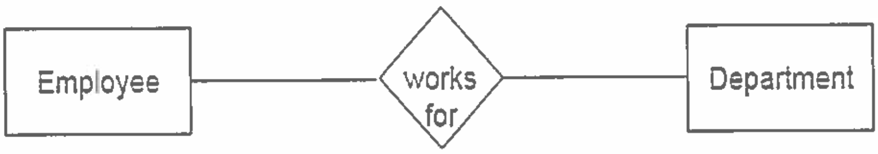
The Entity- Relationship Model :

Entity relationship (ER) models are based on the real-world entities and their relationships.



Components
ER diagram basically having three components:

- **Entities** – It is a real-world thing which can be a person, place, or even a concept. For Example: Department, Admin, Courses, Teachers, Students, Building, etc are some of the entities of a School Management System.
- Consider an organization as an example- manager, product, employee, department etc. can be taken as an entity.

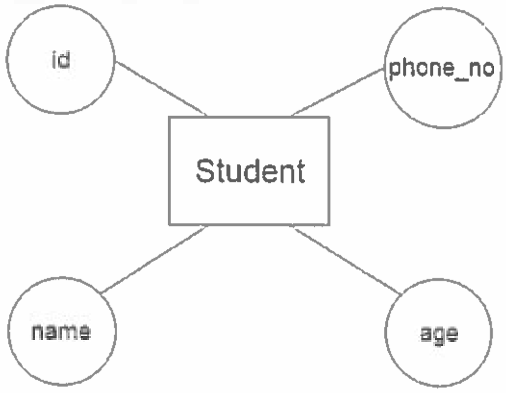


Weak EntityJ: An entity that depends on another entity called a weak entity. The weak entity doesn't contain any key attribute of its own. The weak entity is represented by a double rectangle.

- **Attributes –**

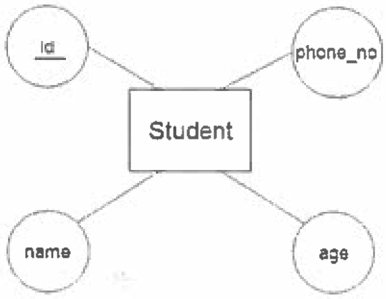
➢ An entity which contains a real-world property called an attribute.
➢ The attribute is used to describe the property of an entity. Eclipse is used to represent an attribute
➢ For Example: The entity employee has the property like employee id, salary, age, etc.

For example, id, age, contact number, name, etc. can be attributes of a student.



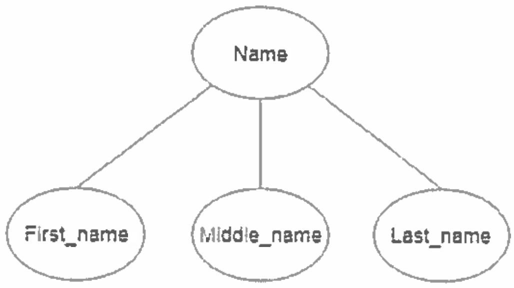
Key Attribute

The key attribute is used to represent the main characteristics of an entity.
It represents a primary key.
The key attribute is represented by an ellipse with the text underlined.



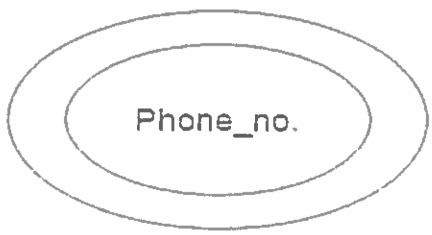
b. Composite Attribute

An attribute that composed of many other attributes is known as a composite attribute. The composite attribute is represented by an ellipse, and those ellipses are connected with an ellipse.



c. Multivalued Attribute

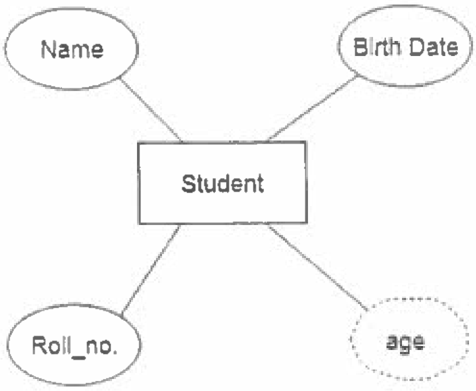
An attribute can have more than one value. These attributes are known as a multivalued attribute. The double oval is used to represent multivalued attribute.
For example, a student can have more than one phone number.



d. Derived Attribute

An attribute that can be derived from other attribute is known as a derived attribute. It can be represented by a dashed ellipse.

For example, A person's age changes over time and can be derived from another attribute like Date of birth.



Relationship – Relationship tells how two attributes are related. A relationship is used to describe the relation between entities. Diamond or rhombus is used to represent the relationship.

Types of relationship are as follows:

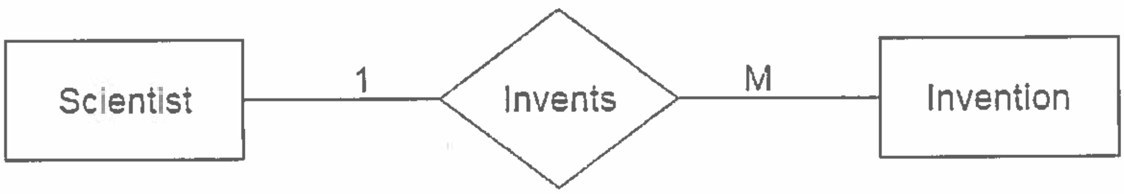
a. One-to-One Relationship

- For Example: Employee works for a department.
- When only one instance of an entity is associated with the relationship, then it is known as one to one relationship.
- **For example,** A female can marry to one male, and a male can marry to one female.
-

- **b. One-to-many relationship**

When only one instance of the entity on the left, and more than one instance of an entity on the right associates with the relationship then this is known as a one-to-many relationship.

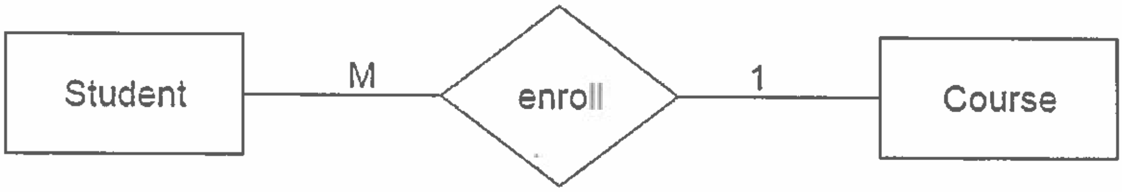
For example, Scientist can invent many inventions, but the invention is done by the only specific scientist.



- **c. Many-to-one relationship**

When more than one instance of the entity on the left, and only one instance of an entity on the right associates with the relationship then it is known as a many-to-one relationship.

For example, Student enrolls for only one course, but a course can have many students.



- **d. Many-to-many relationship**

When more than one instance of the entity on the left, and more than one instance of an entity on the right associates with the relationship then it is known as a many-to-many relationship.
For example, Employee can assign by many projects and project can have many employees.

In a university,
- A student is an entity,
- University is the database,
- Name and age and sex are the attributes.
- The relationships among entities define the logical association between entities.

10 (a)	**Explain various DML, DDL commands in SQL.** **For explaining DML commands – 3 Marks** **For explaining DDL commands – 3 Marks**	**6M**	20DSO01.3

SQL Commands
- SQL commands are instructions. It is used to communicate with the database. It is also used to perform specific tasks, functions, and queries of data.
- SQL can perform various tasks like create a table, add data to tables, drop the table, modify the table, set permission for users.

Types of SQL Commands
SQL is a structured query language, which is used to deal with structured data. Structured data is data that is generally stored in the form of relations or tables.
Whenever we store the data in tables or relations, we need SQL commands. Moreover, these commands are also required to retrieve the data which is stored in tables.
DDL Commands in SQL
DDL is an abbreviation of **Data Definition Language.**
The DDL Commands in Structured Query Language are used to create and modify the schema of the database and its objects. The syntax of DDL commands is predefined for describing the data. The commands of Data Definition Language deal with how the data should exist in the database.
Following are the five DDL commands in SQL:
1. CREATE Command
2. DROP Command
3. ALTER Command

4. TRUNCATE Command
5. RENAME Command

CREATE Command

CREATE is a DDL command used to create databases, tables, triggers and other database objects.

Examples of CREATE Command in SQL

Example 1: This example describes how to create a new database using the CREATE DDL command.

Syntax to Create a Database:

CREATE Database Database_Name;

Suppose, you want to create a Books database in the SQL database. To do this, you have to write the following DDL Command:

Create Database Books;

Example 2: This example describes how to create a new table using the CREATE DDL command.

Syntax to create a new table:

```
CREATE TABLE table_name
(
  column_Name1 data_type ( size of the column ) ,
  column_Name2 data_type ( size of the column) ,
  column_Name3 data_type ( size of the column) ,
  ...
  column_NameN data_type ( size of the column )
) ;
```

Suppose, you want to create a **Student** table with five columns in the SQL database. To do this, you have to write the following DDL command:

```
CREATE TABLE Student
(
  Roll_No. Int ,
  First_Name Varchar (20) ,
  Last_Name Varchar (20) ,
  Age Int ,
  Marks Int ,
) ;
```

Example 3: This example describes how to create a new index using the CREATE DDL command.

Syntax to Create a new index:

CREATE INDEX Name_of_Index **ON** Name_of_Table (column_name_1 , column_name_2 , … . , column_name_N);

Let's take the Student table:

Stu_Id	Name	Marks	City	State
100	Abhay	80	Noida	U.P
101	Sushil	75	Jaipur	Rajasthan
102	Ankit	90	Gurgaon	Haryana
103	Yogesh	93	Lucknow	U.P

Suppose, you want to create an index on the combination of the **City** and **State** field of the **Student** table. For this, we have to use the following DDL command:

Suppose, you want to create an index on the combination of the **City** and **State** field of the **Student** table. For this, we have to use the following DDL command:

CREATE INDEX index_city_State **ON** Employee (Emp_City, Emp_State);
Example 4: This example describes how to create a trigger in the SQL database using the DDL CREATE command.
Syntax to create a trigger:
CREATE TRIGGER [trigger_name]
[BEFORE | **AFTER**]
{ **INSERT** | **UPDATE** | **DELETE** }
ON [table_name] ;

DROP Command
DROP is a DDL command used to delete/remove the database objects from the SQL database.
We can easily remove the entire table, view, or index from the database using this DDL command.
Examples of DROP Command in SQL
Example 1: This example describes how to remove a database from the SQL database.
Syntax to remove a database:
DROP DATABASE Database_Name;
Suppose, you want to delete the Books database from the SQL database.
To do this, you have to write the following DDL command:
DROP DATABASE Books;

Example 2: This example describes how to remove the existing table from the SQL database.
Syntax to remove a table:
DROP TABLE Table_Name;
Suppose, you want to delete the Student table from the SQL database. To do this, you have to write the following DDL command:
DROP TABLE Student;
Example 3: This example describes how to remove the existing index from the SQL database.
Syntax to remove an index:
DROP INDEX Index_Name;
Suppose, you want to delete the index_city from the SQL database. To do this, you have to write the following DDL command:
DROP INDEX Index_city;
ALTER Command :
ALTER is a DDL command which changes or modifies the existing structure of the database, and it also changes the schema of database objects.
We can also add and drop constraints of the table using the ALTER command.
Examples of ALTER Command in SQL
Example 1: This example shows how to add a new field to the existing table.
Syntax to add a newfield in the table:
ALTER TABLE name_of_table **ADD** column_name column_definition;
Suppose, you want to add the 'Father's_Name' column in the existing Student table. To do this, you have to write the following DDL command:
ALTER TABLE Student **ADD** Father's_Name **Varchar**(60);
Example 2: This example describes how to remove the existing column from the table.
Syntax to remove a column from the table:
ALTER TABLE name_of_table **DROP** Column_Name_1 , column_Name_2 , ..., column_Name_N;
Suppose, you want to remove the Age and Marks column from the existing Student table. To do this, you have to write the following DDL command:
ALTER TABLE StudentDROP Age, Marks;
Example 3: This example describes how to modify the existing column of the existing table.

Syntax to modify the column of the table:
ALTER TABLE table_name **MODIFY** (column_name column_datatype(**size**));
Suppose, you want to change the character size of the Last_Namefield of the Student table. To do this, you have to write the following DDL command:
ALTER TABLE table_name **MODIFY** (Last_Name **varchar**(25));

<u>TRUNCATE Command</u>

TRUNCATE is another DDL command which deletes or removes all the records from the table.
This command also removes the space allocated for storing the table records.

Syntax of TRUNCATE command
TRUNCATE TABLE Table_Name;

Example

Suppose, you want to delete the record of the Student table. To do this, you have to write the following TRUNCATE DDL command:
TRUNCATE TABLE Student;
The above query successfully removed all the records from the student table. Let's verify it by using the following SELECT statement:
SELECT * FROM Student;

<u>RENAME Command</u>

RENAME is a DDL command which is used to change the name of the database table.

Syntax of RENAME command
RENAME **TABLE** Old_Table_Name **TO** New_Table_Name;

Example

RENAME **TABLE** Student **TO** Student_Details ;
This query changes the name of the table from Student to Student_Details.

DML Commands in SQL

DML is an abbreviation of **Data Manipulation Language.**
The DML commands in Structured Query Language change the data present in the SQL database. We can easily access, store, modify, update and delete the existing records from the database using DML commands.

Following are the four main DML commands in SQL:

1. SELECT Command
2. INSERT Command
3. UPDATE Command
4. DELETE Command

SELECT DML Command

SELECT is the most important data manipulation command in Structured Query Language. The SELECT command shows the records of the specified table. It also shows the particular record of a particular column by using the WHERE clause.

Syntax of SELECT DML command
SELECT column_Name_1, column_Name_2, ….., column_Name_N
FROM Name_of_table;
Here, **column_Name_1, column_Name_2, ….., column_Name_N** are the names of those columns whose data we want to retrieve from the table.
If we want to retrieve the data from all the columns of the table, we have to use the following SELECT command:
SELECT * FROM table_name;

Examples of SELECT Command

Example 1: This example shows all the values of every column from the table.

SELECT * FROM Student;
This SQL statement displays the following values of the student table:

Student_ID	**Student_Name**	**Student_Marks**
BCA1001	Abhay	85
BCA1002	Anuj	75
BCA1003	Bheem	60
BCA1004	Ram	79
BCA1005	Sumit	80

Example : This example describes how to use the WHERE clause with the SELECT DML command.
Let's take the following Student table:

Student_ID	**Student_Name**	**Student_Marks**
BCA1001	Abhay	80
BCA1002	Ankit	75
BCA1003	Bheem	80
BCA1004	Ram	79
BCA1005	Sumit	80

If you want to access all the records of those students whose marks is 80 from the above table, then you have to write the following DML command in SQL:
SELECT * FROM Student **WHERE** Stu_Marks = 80;
The above SQL query shows the following table in result:

Student_ID	**Student_Name**	**Student_Marks**

BCA1001	Abhay	80
BCA1003	Bheem	80
BCA1005	Sumit	80

INSERT DML Command

INSERT is another most important data manipulation command in Structured Query Language, which allows users to insert data in database tables.

Syntax of INSERT Command

INSERT INTO TABLE_NAME (column_Name1 , column_Name2 , column_Name3 , ... column_NameN)
VALUES (value_1, value_2, value_3, ... value_N) ;

Examples of INSERT Command

Example 1: This example describes how to insert the record in the database table.

Let's take the following student table, which consists of only 2 records of the student.

Stu_Id	Stu_Name	Stu_Marks	Stu_Age
101	Ramesh	92	20
201	Jatin	83	19

Suppose, you want to insert a new record into the student table. For this, you have to write the following DML INSERT command:

INSERT INTO Student (Stu_id, Stu_Name, Stu_Marks, Stu_Age)
VALUES (104, Anmol, 89, 19);

UPDATE DML Command

UPDATE is another most important data manipulation command in Structured Query Language, which allows users to update or modify the existing data in database tables.

Syntax of UPDATE Command

UPDATE Table_name **SET** [column_name1= value_1, ….., column_nameN = value_N]
WHERE CONDITION;

Here, 'UPDATE', 'SET', and 'WHERE' are the SQL keywords, and 'Table_name' is the name of the table whose values you want to update.

Examples of the UPDATE command

Example 1: This example describes how to update the value of a single field.

Let's take a Product table consisting of the following records:

Product_Id	Product_Name	Product_Price	Product_Quantity
P101	Chips	20	20
P102	Chocolates	60	40
P103	Maggi	75	5
P201	Biscuits	80	20
P203	Namkeen	40	50

Suppose, you want to update the Product_Price of the product whose Product_Id is P102. To do this, you have to write the following DML UPDATE command:
UPDATE Product **SET** Product_Price = 80
WHERE Product_Id = 'P102' ;

DELETE DML Command
DELETE is a DML command which allows SQL users to remove single or multiple existing records from the database tables.
This command of Data Manipulation Language does not delete the stored data permanently from the database. We use the WHERE clause with the DELETE command to select specific rows from the table.
Syntax of DELETE Command
DELETE FROM Table_Name **WHERE** condition;

10 (b)	**Explain Set Operators .** **For explaining four set operators – 6 Marks**	20DSO01.3	**L2**

- SET Operators in SQL
- SET operators are special type of operators which are used to *combine the result of two queries.*
- Operators covered under SET operators are:

1. **UNION**
2. **UNION ALL**
3. **INTERSECT**
4. **MINUS**

- There are certain rules which must be followed to perform operations using SET operators in SQL. Rules are as follows:

1. **The number and order of columns must be the same.**
2. **Data types must be compatible.**

- Let us see each of the SET operators in more detail with the help of examples.

- All the examples will be written using the MySQL database.
- Consider we have the following tables with the given data.
- **Table 1: t_employees**

	Name	Department	Salary	Year_of_Experience
	Prashant Wagh	R&D	49000	1
	Abhishek Pawar	Production	45000	1
	Gautam Jain	Development	56000	4
	Shubham Mahale	Accounts	57000	2
	Rahul Thakur	Production	76000	4
	Bhushan Wagh	R&D	75000	2
	Anand Singh	Marketing	28000	1

- **Table 3: t_students**

	Name	Hometown	Percentage	Favourite_Subject
	Soniya Jain	Udaipur	89	Physics
	Harshada Sharma	Kanpur	92	Chemistry
	Anuja Rajput	Jaipur	78	History
	Pranali Singh	Nashik	88	Geography

	Renuka Deshmukh	Panipat	90	Biology
	Swati Kumari	Faridabad	93	English
	Prachi Jaiswal	Gurugram	96	Hindi

- **Table 4: t2_students**

	Name	Hometown	Percentage	Favourite_Subject
1	Soniya Jain	Udaipur	89	Physics
2	Ishwari Dixit	Delhi	86	Hindi
3	Anuja Rajput	Jaipur	78	History
4	Pakhi Arora	Surat	70	Sanskrit
5	Renuka Deshmukh	Panipat	90	Biology
6	Jayshree Patel	Pune	91	Maths
7	Prachi Jaiswal	Gurugram	96	Hindi

- 1. UNION:

  - UNION will be used to combine the result of two select statements.
  - Duplicate rows will be eliminated from the results obtained after performing the UNION operation.
- **Example 1:**
- Write a query to perform union between the table t_employees and the table t2_employees.
- **Query:**

- **mysql> SELECT *FROM t_employees UNION SELECT *FROM t2_employees;**
- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_employees table and perform a UNION operation with the records fetched by the second SELECT query from the t2_employees table.
- You will get the following output:

	Name	Departme nt	S al ar	Year_of_Experience
	Aakash Singh	Developme	720	2
	Abhishek Pawar	Production	450	1
	Pranav Deshmukh	HR	599	3
	Shubham Mahale	Accounts	570	2
	Sunil Kulkarni	Developme	870	3
	Bhushan Wagh	R&D	750	2
	Paras Jaiswal	Marketing	320	1
	Prashant Wagh	R&D	490	1
	Gautam Jain	Developme	560	4
	Rahul Thakur	Production	760	4
	Anand Singh	Marketing	280	1

- Since we have performed union operation between both the tables, so only the records from the first and second table are displayed except for the duplicate records.
- **Example 2:**
- Write a query to perform union between the table t_students and the table t2_students.
- **Query:**

mysql> SELECT *FROM t_students UNION SELECT *FROM t2_students;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_students table and perform a UNION operation with the records fetched by the second SELECT query from the t2_students table.
- 2. UNION ALL

  - This operator combines all the records from both the queries.
  - Duplicate rows will be not be eliminated from the results obtained after performing the UNION ALL operation.

- **Example 1:**
- Write a query to perform union all operation between the table t_employees and the table t2_employees.
- **Query:**

1. mysql> **SELECT *FROM** t_employees **UNION** ALL **SELECT *FROM** t2_employees;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_employees table and perform UNION ALL operation with the records fetched by the second SELECT query from the t2_employees table.
- You will get the following output:

ID	Name	Department	Salary	Ye
1	Aakash Singh	Development	72000	2
2	Abhishek Pawar	Production	45000	1
3	Pranav Deshmukh	HR	59900	3
4	Shubham Mahale	Accounts	57000	2
5	Sunil Kulkarni	Development	87000	3
6	Bhushan Wagh	R&D	75000	2
7	Paras Jaiswal	Marketing	32000	1
1	Prashant Wagh	R&D	49000	1

2	Abhishek Pawar	Production	45000	1
3	Gautam Jain	Development	56000	4
4	Shubham Mahale	Accounts	57000	2
5	Rahul Thakur	Production	76000	4
6	Bhushan Wagh	R&D	75000	2
7	Anand Singh	Marketing	28000	1

- Since we have performed union all operation between both the tables, so all the records from the first and second table are displayed, including the duplicate records.
- **Example 2:**
- Write a query to perform union all operation between the table t_students and the table t2_students.
- **Query:**

1. mysql> **SELECT *FROM** t_students **UNION** ALL **SELECT *FROM** t2_students;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_students table and perform UNION ALL operation with the records fetched by the second SELECT query from the t2_students table.
- 3. INTERSECT:

  - It is used to combine two SELECT statements, but it only returns the records which are common from both SELECT statements.

- **Example 1:**
- Write a query to perform intersect operation between the table t_employees and the table t2_employees.
- **Query:**

1. mysql> **SELECT *FROM** t_employees **INTERSECT SELECT *FROM** t2_employees;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_employees table and perform INTERSECT operation with the records fetched by the second SELECT query from the t2_employees table.
- You will get the following output:

ID	Name	Hometown	Percentage	Favourite_Subject
2	Abhishek Pawar	Production	45000	1
4	Shubham Mahale	Accounts	57000	2
6	Bhushan Wagh	R&D	75000	2

- Since we have performed intersect operation between both the tables, so only the common records from both the tables are displayed.
- **Example 2:**
- Write a query to perform intersect operation between the table t_students and the table t2_students.
- **Query:**

1. mysql> **SELECT *FROM** t_students **INTERSECT SELECT *FROM** t2_students;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_students table and perform a UNION operation with the records fetched by the second SELECT query from the t2_students table.
- You will get the following output:

ID	Name	Hometown	Percentage
1	Soniya Jain	Udaipur	89
3	Anuja Rajput	Jaipur	78

5	Renuka Deshmukh	Panipat	90	
7	Prachi Jaiswal	Gurugram	96	

- Since we have performed intersect operation between both the tables, so only the common records from both the tables are displayed.

4. **MINUS**

- It displays the rows which are present in the first query but absent in the second query with no duplicates.
  - **Example 1:**
  - Write a query to perform a minus operation between the table t_employees and the table t2_employees.
  - **Query:**

1. mysql> **SELECT *FROM** t_employees MINUS **SELECT *FROM** t2_employees;

- Here, in a single query, we have written two SELECT queries. The first SELECT query will fetch the records from the t_employees table and perform MINUS operation with the records fetched by the second SELECT query from the t2_employees table.
- You will get the following output:
- Since we have performed Minus operation between both the tables, so only the unmatched records from both the tables are displayed.

11 (a)	**Explain Order by, Group by and Having Clauses with example** **For explaining Order by clause – 2 Marks** **For explaining Group by clause – 2 Marks** **For explaining Having clause – 2 Marks**	**6M**	20DSO01.3

SQL gives you options for retrieving, analyzing, and displaying the information you need with the GROUP BY, HAVING, and ORDER BY clauses.

GROUP BY clauses

Sometimes, rather than retrieving individual records, you want to know something about a group of records. The GROUP BY clause is the tool you need.

Suppose you're the sales manager of another location, and you want to look at the performance of your sales force. If you do a simple SELECT, such as the following query:

Example:

SELECT InvoiceNo, SaleDate, Salesperson, TotalSale
FROM SALES;

To do the real analysis, you can combine the GROUP BY clause with one of the *aggregate* functions (also called *set functions*) to get a quantitative picture of sales performance. For example, you can see which salesperson is selling more of the profitable high-ticket items by using the average (AVG) function as follows:

Example :

```
SELECT Salesperson, AVG(TotalSale)
 FROM SALES
 GROUP BY Salesperson;
```

ORDER BY clauses

Use the ORDER BY clause to display the output table of a query in either ascending or descending alphabetical order. Whereas the GROUP BY clause gathers rows into groups and sorts the groups into alphabetical order, ORDER BY sorts individual rows. The ORDER BY clause must be the last clause that you specify in a query.

If the query also contains a GROUP BY clause, the clause first arranges the output rows into groups. The ORDER BY clause then sorts the rows within each group. If you have no GROUP BY clause, then the statement considers the entire table as a group, and the ORDER BY clause sorts all its rows according to the column (or columns) that the ORDER BY clause specifies.

To illustrate this point, consider the data in the SALES table. The SALES table contains columns for InvoiceNo, SaleDate, Salesperson, and TotalSale.

If you use the following example, you see all the data in the SALES table — but in an arbitrary order:

```
SELECT * FROM SALES ;
```

In one implementation, this may be the order in which you inserted the rows in the table; in another implementation, the order may be that of the most recent updates. The order can also change unexpectedly if anyone physically reorganizes the database. That's one reason it's usually a good idea to specify the order in which you want the rows.

You may, for example, want to see the rows in order by the SaleDate like this:

```
SELECT * FROM SALES ORDER BY SaleDate ;
```

This example returns all the rows in the SALES table in order by SaleDate. For rows with the same SaleDate, the default order depends on the implementation. You can, however, specify how to sort the rows that share the same SaleDate.

You may want to see the sales for each SaleDate in order by InvoiceNo, as follows:

```
SELECT * FROM SALES ORDER BY SaleDate, InvoiceNo ;
```

This example first orders the sales by SaleDate; then for each SaleDate, it orders the sales by InvoiceNo. But don't confuse that example with the following query:

```
SELECT * FROM SALES ORDER BY InvoiceNo, SaleDate ;
```

This query first orders the sales by INVOICE_NO. Then for each different InvoiceNo, the query orders the

sales by SaleDate. This probably won't yield the result you want, because it's unlikely that multiple sale dates will exist for a single invoice number.
The following query is another example of how SQL can return data:

```
SELECT * FROM SALES ORDER BY Salesperson, SaleDate ;
```

This example first orders by Salesperson and then by SaleDate. After you look at the data in that order, you may want to invert it, as follows:

```
SELECT * FROM SALES ORDER BY SaleDate, Salesperson ;
```

This example orders the rows first by SaleDate and then by Salesperson.

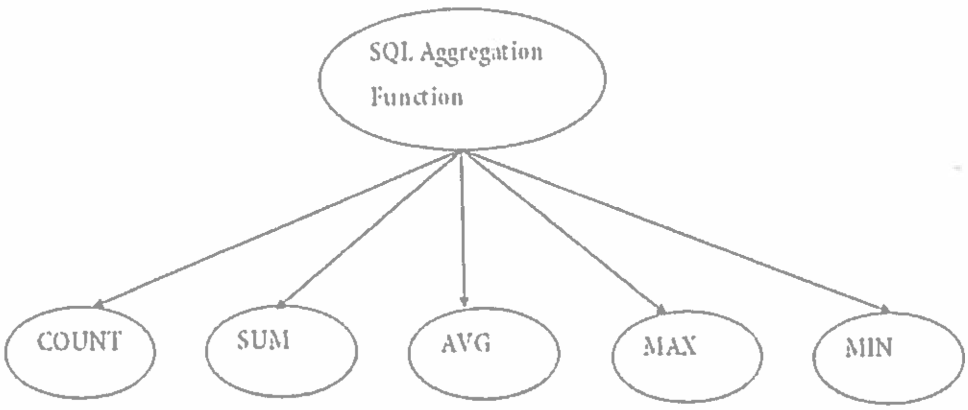
HAVING clause :
The HAVING clause was added to SQL because the WHERE keyword cannot be used with aggregate functions.
HAVING Syntax
SELECT *column_name(s)*
FROM *table_name*
WHERE *condition*
GROUP BY *column_name(s)*
HAVING *condition*
ORDER BY *column_name(s);*

Example
```
SELECT COUNT(CustomerID), Country
FROM Customers
GROUP BY Country
HAVING COUNT(CustomerID) > 5;
```

11 (b)	**Explain Aggregate functions with examples** **For writing about aggregate function – 2 Marks** **For writing about the types aggregate functions – 4 Marks**	**6M**	**20DSO01.3**

SQL Aggregate Functions
- SQL aggregation function is used to perform the calculations on multiple rows of a single column of a table. It returns a single value.
- It is also used to summarize the data.

Types of SQL Aggregation Function



COUNT FUNCTION

- COUNT function is used to Count the number of rows in a database table. It can work on both numeric and non-numeric data types.
- COUNT function uses the COUNT(*) that returns the count of all the rows in a specified table. COUNT(*) considers duplicate and Null.

Syntax

COUNT(*)
or
COUNT([ALL|DISTINCT] expression)

2. SUM Function

Sum function is used to calculate the sum of all selected columns. It works on numeric fields only.

Syntax

SUM()
or
SUM([ALL|DISTINCT] expression)

Example: SUM()

```
SELECT SUM(COST)
FROM PRODUCT_MAST;
```

Output:

670

Example: SUM() with WHERE

```
SELECT SUM(COST)
FROM PRODUCT_MAST
WHERE QTY>3;
```

3. AVG function

The AVG function is used to calculate the average value of the numeric type. AVG function returns the average of all non-Null values.

Syntax

AVG()
or
AVG([ALL|DISTINCT] expression)

Example:

```
SELECT AVG(COST)
FROM PRODUCT_MAST;
```

4. MAX Function
MAX function is used to find the maximum value of a certain column. This function determines the largest value of all selected values of a column.

Syntax

```
MAX()
or
MAX( [ALL|DISTINCT] expression )
```

Example:

```
SELECT MAX(RATE)
FROM PRODUCT_MAST;
```

5. MIN Function
MIN function is used to find the minimum value of a certain column. This function determines the smallest value of all selected values of a column.

Syntax

```
MIN()
or
MIN( [ALL|DISTINCT] expression )
```

Example:

```
SELECT MIN(RATE)
FROM PRODUCT_MAST;
```

12 a) **Explain the components of PL/SQL block. -- 6 Marks**
For writing about three components -- 6 Marks

__**Components of PL/SQL Block**__

PL/SQL blocks have a pre-defined structure in which the code is to be grouped. Below are different sections of PL/SQL blocks.

1. Declaration section
2. Execution section
3. Exception-Handling section

Declaration Section

This is the first section of the PL/SQL blocks. This section is an optional part. This is the section in which the declaration of variables, cursors, exceptions, subprograms, pragma instructions and collections that are needed in the block will be declared.

Below are few more characteristics

- This particular section is optional and can be skipped if no declarations are needed.
- This should be the first section in a PL/SQL block, if present.

- This section starts with the keyword 'DECLARE' for triggers and anonymous block. For other subprograms, this keyword will not be present. Instead, the part after the subprogram name definition marks the declaration section.
- This section should always be followed by execution section

Execution Section

Execution part is the main and mandatory part which actually executes the code that is written inside it. Since the PL/SQL expects the executable statements from this block this cannot be an empty block, i.e., it should have at least one valid executable code line in it. Below are few more characteristics of this part.

- This can contain both PL/SQL code and SQL code.
- This can contain one or many blocks inside it as a nested block.
- This section starts with the keyword 'BEGIN'.
- This section should be followed either by 'END' or Exception-Handling section (if present)

Exception-Handling Section:

The exception is unavoidable in the program which occurs at run-time and to handle this Oracle has provided an Exception-handling section in blocks. This section can also contain PL/SQL statements. This is an optional section of the PL/SQL blocks.

- This is the section where the exception raised in the execution block is handled.
- This section is the last part of the PL/SQL block.
- Control from this section can never return to the execution block.
- This section starts with the keyword 'EXCEPTION'.
- This section should always be followed by the keyword 'END'.

The Keyword 'END' marks the end of PL/SQL block.

PL/SQL Block Syntax

Below is the syntax of the PL/SQL block structure.

Syntax of PL/SQL Block Structure:

```
DECLARE          --optional
        <declarations>
        .
        .
BEGIN              --mandatory
        <executable statements. At least one executable statement is mandatory>
        .
        .
        .
EXCEPTION     --optional
        <exception handler>
        .
        .
        .
END;               --mandatory
/
```

```
DECLARE --optional
    <declarations>

BEGIN   --mandatory
    <executable statements. At least one executable statement is mandatory>

EXCEPTION --optional
    <exception handles>

END;   --mandatory
/
```

12(b) **Explain about control statements in PL/SQL block .**
For explaining the three control statements – 6 Marks

PL/SQL Control Statements
PL/SQL has three categories of control statements: conditional selection statements, loop statements and sequential control statements.
PL/SQL categories of control statements are:

- Conditional selection statements, which run different statements for different data values.

The conditional selection statements are IF and CASE.

- Loop statements, which run the same statements with a series of different data values.

The loop statements are the basic LOOP, FOR LOOP, and WHILE LOOP.
The EXIT statement transfers control to the end of a loop. The CONTINUE statement exits the current iteration of a loop and transfers control to the next iteration. Both EXIT and CONTINUE have an optional WHEN clause, where you can specify a condition.

- Sequential control statements, which are not crucial to PL/SQL programming.

The sequential control statements are GOTO, which goes to a specified statement, and NULL, which does nothing

<u>**Conditional Selection Statements**</u>

The **conditional selection statements,** IF and CASE, run different statements for different data values.
The IF statement either runs or skips a sequence of one or more statements, depending on a condition.
The IF statement has these forms:

- IF THEN

- IF THEN ELSE
- IF THEN ELSIF

LOOP Statements

Loop statements run the same statements with a series of different values. The loop statements are:

- Basic LOOP
- FOR LOOP
- Cursor FOR LOOP
- WHILE LOOP

The statements that exit a loop are:

- EXIT
- EXIT WHEN

The statements that exit the current iteration of a loop are:

- CONTINUE
- CONTINUE WHEN

Sequential Control Statements

- Unlike the IF and LOOP statements, the **sequential control statements** GOTO and NULL are not crucial to PL/SQL programming.
- The GOTO statement, which goes to a specified statement, is seldom needed. Occasionally, it simplifies logic enough to warrant its use.
- The NULL statement, which does nothing, can improve readability by making the meaning and action of conditional statements clear.

Explain about triggers – 6 Marks

13 a) **For writing about triggers – 2 marks**

For writing the syntax - 2Marks

For writing the example – 2 Marks

PL/SQL Trigger

Trigger is invoked by Oracle engine automatically whenever a specified event occurs.Trigger is stored into database and invoked repeatedly, when specific condition match.

Triggers are stored programs, which are automatically executed or fired when some event occurs.

Triggers are written to be executed in response to any of the following events.

- A database manipulation (DML) statement (DELETE, INSERT, or UPDATE).
- A database definition (DDL) statement (CREATE, ALTER, or DROP).
- A database operation (SERVERERROR, LOGON, LOGOFF, STARTUP, or SHUTDOWN).

Triggers could be defined on the table, view, schema, or database with which the event is associated.

Creating a trigger:

Syntax for creating trigger:

1. **CREATE** [OR REPLACE] **TRIGGER** trigger_name
2. {BEFORE | **AFTER** | **INSTEAD OF** }
3. {**INSERT** [OR] | **UPDATE** [OR] | **DELETE**}

4. [OF col_name]
5. **ON** table_name
6. [REFERENCING OLD **AS** o NEW **AS** n]
7. [FOR EACH ROW]
8. **WHEN** (condition)
9. **DECLARE**
10. Declaration-statements
11. **BEGIN**
12. Executable-statements
13. EXCEPTION
14. Exception-handling-statements
15. **END**;

Example:

```
CREATE OR REPLACE TRIGGER display_salary_changes
BEFORE DELETE OR INSERT OR UPDATE ON customers
FOR EACH ROW
WHEN (NEW.ID > 0)
DECLARE
  sal_diff number;
BEGIN
  sal_diff := :NEW.salary  - :OLD.salary;
  dbms_output.put_line('Old salary: ' || :OLD.salary);
  dbms_output.put_line('New salary: ' || :NEW.salary);
  dbms_output.put_line('Salary difference: ' || sal_diff);
END;
/
```

13(b)

Explain about cursors.
For writing about cursors – 3 Marks
For writing the types of cursors – 3 Marks

PL/SQL Cursor :

When an SQL statement is processed, Oracle creates a memory area known as context area. A cursor is a pointer to this context area. It contains all information needed for processing the statement. In PL/SQL, the context area is controlled by Cursor. A cursor contains information on a select statement and the rows of data accessed by it.

A cursor is used to referred to a program to fetch and process the rows returned by the SQL statement, one at a time. There are two types of cursors:

- Implicit Cursors
- Explicit Cursors

1) PL/SQL Implicit Cursors

The implicit cursors are automatically generated by Oracle while an SQL statement is executed, if you don't use an explicit cursor for the statement.

These are created by default to process the statements when DML statements like INSERT,

UPDATE, DELETE etc. are executed.
he following table soecifies the status of the cursor with each of its attribute.

Attribute	Description
%FOUND	Its return value is TRUE if DML statements like INSERT, DELETE and UPDATE affect at least one row or more rows or a SELECT INTO statement returned one or more rows. Otherwise it returns FALSE.
%NOTFOUND	Its return value is TRUE if DML statements like INSERT, DELETE and UPDATE affect no row, or a SELECT INTO statement return no rows. Otherwise it returns FALSE. It is a just opposite of %FOUND.
%ISOPEN	It always returns FALSE for implicit cursors, because the SQL cursor is automatically closed after executing its associated SQL statements.
%ROWCOUNT	It returns the number of rows affected by DML statements like INSERT, DELETE, and UPDATE or returned by a SELECT INTO statement.

PL/SQL Implicit Cursor Example
Create customers table and have records:

ID	NAME	AGE	ADDRESS	SALARY
1	Ramesh	23	Allahabad	20000
2	Suresh	22	Kanpur	22000
3	Mahesh	24	Ghaziabad	24000
4	Chandan	25	Noida	26000
5	Alex	21	Paris	28000
6	Sunita	20	Delhi	30000

Let's execute the following program to update the table and increase salary of each customer by 5000. Here, SQL%ROWCOUNT attribute is used to determine the number of rows affected:
Create procedure:

```
DECLARE
   total_rows number(2);
BEGIN
   UPDATE  customers
   SET salary = salary + 5000;
   IF sql%notfound THEN
      dbms_output.put_line('no customers updated');
   ELSIF sql%found THEN
      total_rows := sql%rowcount;
      dbms_output.put_line( total_rows || ' customers updated ');
   END IF;
```

12. END;
13. /

14 Explain about Normalization and need for normalization along with the problems caused by Redundancy.

For explaining about the problems caused due to redundancy - 5 Marks

For explaining about the Normalization – 3 Marks

For explaining about the Need of Normalization - 4 Marks

2 types of Normal forms.

Redundancy means having multiple copies of same data in the database. This problem arises when a database is not normalized. Suppose a table of student details attributes are: student Id, student name, college name, college rank, course opted.

As it can be observed that values of attribute college name, college rank, course is being repeated which can lead to problems. Problems caused due to redundancy are: Insertion anomaly, Deletion anomaly, and Updation anomaly.

1. **Insertion Anomaly –**
   If a student detail has to be inserted whose course is not being decided yet then insertion will not be possible till the time course is decided for student.

This problem happens when the insertion of a data record is not possible without adding some additional unrelated data to the record.

2. Deletion Anomaly –
If the details of students in this table are deleted then the details of college will also get deleted which should not occur by common sense.
This anomaly happens when deletion of a data record results in losing some unrelated information that was stored as part of the record that was deleted from a table.
It is not possible to delete some information without losing some other information in the table as well.

3. Updation Anomaly –
Suppose if the rank of the college changes then changes will have to be all over the database which will be time-consuming and computationally costly.

Need of **Normalization:**

A large database defined as a single relation may result in data duplication. This repetition of data may result in:

- Making relations very large.
- It isn't easy to maintain and update data as it would involve searching many records in relation.
- Wastage and poor utilization of disk space and resources.
- The likelihood of errors and inconsistencies increases.

Normalization:

- process of organizing the data in the database.
- used to minimize the redundancy from a relation or set of relations.
- It is also used to eliminate undesirable characteristics like Insertion, Update, and Deletion Anomalies.
- divides the larger table into smaller and links them using relationships.
- The normal form is used to reduce redundancy from the database table.

Types of Normal Forms - 1NF, 2NF, 3NF, 4NF, 5NF, BCNF

(OR)

Explain Third NF and BCNF with relevant table structure.

15 a) **For explaining about third NF and BCNF – 8 Marks**

For writing the differences – 4 Marks

. Third Normal Form (3NF) :

A relation is said to be in Third Normal Form (3NF), if it is in 2NF and when no non key attribute is transitively dependent on the primary key i.e., there is no transitive dependency. Also it should satisfy one of the below given conditions. For the function dependency C->D:

- C should be a super key and,
- D should be a prime attribute i.e, D should be a part of the candidate key.

3NF is used to reduce data duplication and to attain data integrity.

Example:

For the relation R(L, M, N, O, P) with functional dependencies as {L->M, MN->P, PO->L}:

The candidate keys will be : {LNO, MNO, NOP}

as the closure of LNO = {L, M, N, O, P}

closure of MNO = {L, M, N, O, P}

closure of NOP = {L, M, N, O, P}

This relation is in 3NF as it is already in 2NF and has no transitive dependency. Also there is no non prime attribute that is deriving a non prime attribute.

2. Boyce-Codd Normal Form (BCNF) **:**

BCNF stands for Boyce-Codd normal form and was made by R.F Boyce and E.F Codd in 1974.A functional dependency is said to be in BCNF if these properties hold:

- It should already be in 3NF.
- For a functional dependency say P->Q, P should be a super key.

BCNF is an extension of 3NF and it is has more strict rules than 3NF. Also, it is considered to be more stronger than 3NF.

Example:
for the relation R(A, B, C, D) with functional dependencies as {A->B, A->C, C->D, C->A}:

The candidate keys will be : {A, C}

as the closure of A = {A, B, C, D}

closure of C = {A, B, C, D}

This relation is in BCNF as it is already in 3Nf (there is no prime attribute deriving no prime attribute) and on the left hand side of the functional dependency there is a candidate key.

Difference between 3NF and BCNF :

S.NO.	3NF	BCNF
1.	In 3NF there should be no transitive dependency that is no non prime attribute should be transitively dependent on the candidate key.	In BCNF for any relation A->B, A should be a super key of relation.
2.	It is less stronger than BCNF.	It is comparatively more stronger than 3NF.
3.	In 3NF the functional dependencies are already in 1NF and 2NF.	In BCNF the functional dependencies are already in 1NF, 2NF and 3NF.
4.	The redundancy is high in 3NF.	The redundancy is comparatively low in BCNF.
5.	In 3NF there is preservation of all functional dependencies.	In BCNF there may or may not be preservation of all functional dependencies.

6.	It is comparatively easier to achieve.	It is difficult to achieve.

prepared by
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8/12/22

8/12/2022
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Semester End Regular Examination, Nov./Dec., 2022

Degree	B. Tech.	Program	Common to All	Academic Year	2022 - 2023
Course Code	20SHO01	Test Duration	3 Hrs. Max. Marks 70	Semester	V
Course	Women and Society (Open Elective)				

Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Write the difference between gender and sex.	20SHO01.2	L1
2	What is socialist Feminism?	20SHO01.1	L1
3	What is Uniform Civil code	20SHO01.2	L1
4	Define gender stereotyping with one example.	20SHO01.4	L1
5	Define the term 'taboo' with one example.	20SHO01.1	L1

Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	What are the theories of social construction of gender? Explain with relevant examples	6M	20SHO01.1	L2
6 (b)	Write a brief note on the status of women in Independent India.	6M	20SHO01.1	L2
	OR			
7 (a)	Gender equality and women empowerment are key to reducing poverty in India. Elucidate.	6M	20SHO01.1	L2
7 (b)	Explain androcentrism.	6M	20SHO01.1	L2
8 (a)	Write a short note on the Socio-Economic Status of women in ancient India.	8M	20SHO01.2	L2
8 (b)	How do stereotypes affect women's right to equality?	4M	20SHO01.2	L2
	OR			
9 (a)	What were the issues in Shah Bano Begum case? Explain.	6M	20SHO01.2	L2
9 (b)	Explain in detail the Women's rights movement of 1848-1920 and its significance.	6M	20SHO01.2	L2
10 (a)	Describe gender norms.	6M	20SHO01.3	L2
10 (b)	Explain stereotyping. Why is it important to break gender norms and stereotypes?	6M	20SHO01.3	L2
	OR			
11 (a)	What is feminism? Explain the types of feminism with relevant examples.	6M	20SHO01.3	L2
11(b)	What do you mean by domestic violence? How does domestic violence affect us? Explain.	6M	20SHO01.3	L2
12 (a)	Environment movements often contain economic and identity issues. Discuss.	6M	20SHO01.4	L2
12 (b)	Mention about 19th century social and religious reform movements.	6M	20SHO01.4	L1
	OR			
13 (a)	Mention the measures to ensure the safety and security of women in India.	8M	20SHO01.4	L2
13 (b)	Discuss the prominent psychological theories of gender role and gender identity development.	4M	20SHO01.4	L2
14 (a)	Critically evaluate the representation of women identity in India cinema.	6M	20SHO01.5	L2
14 (b)	Explain with examples the social construct of gender.	6M	20SHO01.5	L2
	OR			
15 (a)	Explain the role of media in women's empowerment in Indian society.	6M	20SHO01.5	L2
15 (b)	Explain the role of Role of Women in Panchayat Development and also mention the challenges.	6M	20SHO01.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Women and Society 20SHO01 (Open Elective)

Part A (Short Answer Questions 5 x 2 = 10 Marks)

Content 1 mark ; Grammar 0.5 mark; Spelling correction : 0.5 mark

1. **Write the difference between gender and sex?**

   We often use 'Sex' and 'gender' interchangeably but they both have different meanings. Sex refers to a set of biological attributes in humans and animals whereas gender refers to the socially constructed roles, behaviors, expressions and identities of girls, women, boys, men, and gender diverse people

2. **What is socialist Feminism?**

   Socialist feminism is a two-pronged theory that broadens Marxist feminism 's argument for the role of capitalism in the oppression of women and radical feminism 's theory of the role of gender and the patriarchy. Socialist feminists reject radical feminism's main claim that patriarchy is the only, or primary, source of oppression of women.

3. **What is Uniform Civil code?**

   A Uniform Civil Code means that all sections of the society irrespective of their religion shall be treated equally according to a national civil code, which shall be applicable to all uniformly. They cover areas like- Marriage, divorce, maintenance, inheritance, adoption and succession of the property.

4. **Define gender stereotyping with one example.**

   Gender stereotyping is defined as an overgeneralization of characteristics, differences and attributes of a certain group based on their gender. Gender stereotypes create widely accepted biases about certain characteristics or traits and perpetuate the notion that each gender and associated behaviors are binary.

5. **Define the term 'taboo' with one example.**

   A forbidden act considered so offensive to norms, particularly mores (moral norms) as to be reviled and unthinkable.

   Examples of Taboo bestiality: Having sex with animals. cannibalism: Eating human flesh.

Part B (Long Answer Questions 5 x 12 = 60 Marks)

6M - Content mark 4 ; Grammar 1 mark; Spelling correction : 1 mark
4M = Content mark 2 ; Grammar 1 mark; Spelling correction : 1 mark

6. A. What are the theories of social construction of gender? Explain with relevant examples

If you argue with someone who thinks that gender is a social construct, and you force them to answer certain hard questions, you will find that they stop making sense after a certain point. For example, if I ask you, "Is it the case that sex has no effect whatsoever on behavior?" then you'll have to answer "No" unless you're crazy and think that people don't have bodies, and endocrine systems, and so forth. But if sex affects behavior, then that means you can't completely separate sex and gender

There is the argument about sex being a murky category, or a spectrum, but these kinds of arguments have two problems. First, they're inevitably based on a continuum fallacy. Second, the number of people who are born with ambiguous sex is vanishingly small — less than 1% of the population. When someone is pointing to a tiny minority of cases to try and make an obvious dichotomy look like a spectrum, it is likely that they have some (ideological) motivation for doing so.

6b. Write a brief note on the status of women in Independent India

'Women's standing in India has evolved as a result of education and other societal progress. They are also given the freedom to pursue their objectives, obtain an education, and make their job goals a reality. Even in marriage, women are given the liberty to express themselves. Women in India today are well aware of their rights and benefits, and they are no longer politically, socially, economically, or educationally backward. They now have the same opportunities and rights as everyone else. They are capable of achieving any position or status in life.

7a. Gender equality and women empowerment are key to reducing poverty in India. Elucidate

Gender is the term that mainly refers to the economic, social, and cultural attributes and opportunities associated with being male or female. In most communities, being a male or a female is not merely a matter of different biological or physical characteristics—males and females face different expectations about how they should dress up, behave, or work. Whether in the family, workplace, or the public sphere, the relations between men or women also reflect the understandings of the talents, characteristics, & behavior appropriate to women and men. Gender thus differs from sex, that it is social and cultural rather than biological.

Gender equality in India is the desired state of equal ease of access to ample resources & opportunities regardless of gender, including economic participation and decision-making, and valuing different behaviors, aspirations, and needs equally, regardless of gender. Gender Equality in India is all about equal footing in all walks of life. Gender equality in India is a fundamental human right and a necessary foundation for a peaceful, prosperous & sustainable world.

Gender equality in India is the goal, while gender neutrality and gender equities are the practices and ways of thinking that help achieve the goal. Gender parity is used to measure gender balance in a given situation that can help achieve gender equality but is not the aim.

As per UNICEF, gender equality means that "women and men, girls and boys, should enjoy the same rights and liberties, resources, opportunities, and protections. It is, however, not important that girls and boys, or women and men, be the same, or that they be treated exactly alike."

7 **B. Explain Androcentrism?**

Androcentrism is the evaluation of individuals and cultures based on male perspectives, standards, and values. The term refers to a male-centered worldview which does not necessarily present explicitly negative views of women and girls, but positions men and boys as representative of the human condition or experience and women and girls as diverging from the human condition. It is a complex, subtle, and often unacknowledged form of sexism, existing on a continuum which includes misogyny and patriarchal attitudes, but it is also informed by patriarchal cultures in which men are granted more power and influence, and thus the right to evaluate and interpret individuals and cultures. Androcentrism exists in all fields of study and cultural expressions, including the arts, sciences, medicine, law, fine arts, and media.

8a. Write a short note on the Socio -Economic Status of women in ancient India.

Status of Women in India Essay: Status of women in a nation specifies the nation's worth, as Dr Jawaharlal Nehru once stated. In ancient India, the culture was backward to deprive women of their fundamental rights and equality with men of the Indian society. As a result of the situation, several prominent entities like Raja Ram Mohan Roy and Swami Vivekananda began a movement against women's struggle with inequality and subjugation. Mahatma Gandhi also induced the revival of the status of women through active participation in Freedom struggles. As a result, ever since the Independence struggle, women's status in India has slightly gone up.

The article contains the Status of Women in India Essay for the students and children of different age groups and classes. It is an essential source for spreading awareness about equality and equal rights for women and letting the future generation understand the benefit of uplifting women's status.
Through recent times, self-consciousness and Education among women elevated their development and progress, empowering them and assuring them to pay the path of success in various fields. Currently, India does not have any shortage of women in the medical field, technical field, teaching field, legal sector, or any other occupation. India witnessed the growth of empowered women taking higher positions in various offices and institutions. For example, Indira Gandhi became the first woman Prime Minister of India and successfully guided India for fourteen years while contributing to domestic and financial development.

Status of women in India is higher and better than many of their counterparts, including several other developing nations of the World. Still, there is room for more improvement in raising the status of women in India. There are evident how several women of middle-class society in India are conscious of their privileges and rights and have undergone an economical and vocational transformation in recent years. Many women in Indian society are employed successfully in several sectors, and many of them also occupy senior positions in the business or administration.

Even after the marked change in women's economic and social status in India, they still undergo a lot of burdens as they have to work outside and maintain household chores. India still comes out as a patriarchal and male dominating society. Women of society still have to depend somewhat on a male to seek protection and help in various life stages. Men are still regarded in Indian society as superior to women of society. Furthermore, in many families, the birth of a son is considered as an honoured privilege. In contrast, that of a girl is regarded as a curse and liability and cannot seek compulsory Education.

In Vedic India, women used to enjoy higher social life and family status. Women were given immense respect and considered Mother, Devi, and Shakti, which is still the same among many Indian families. Women then were adored and respected more than they are now, and were considered the entity where God resides. Yet, even after all the respect given to them, many people consider them as a weaker sex, exploit them, and harass them both physically and mentally. This is significant hypocrisy about Indian women that resides in the society.

Previously, Indian women also had to go through Sati Pratha or were deprived of joy and colours after their husbands' death, which has been eradicated from the society and women are now happier than earlier after the freedom struggle.

8b. How do stereotypes affect women's right to equality?

Stereotypes, about what women can or cannot do affect women's right to equality by forcing the society to give them certain roles and not allow others. This is unequal treatment because the choice of the woman is not considered and she is not free to do what she wants.Women are considered inferior to men. There is a belief that women do not have technical mind and therefore they cannot be scientists. It is thought that women are good at only certain jobs such as teaching and nursing. These stereotypes about women's capability or incapability of doing certain jobs badly affect women's right to equality.

9a. What were the issues in Shah Bano Begum case? Explain.

Shah Bano Begum and others brought to the forefront issues of citizenship, minority identity, and national sovereignty amidst an environment of fear and tension during the mid 1980s. The case and its ensuing controversy reflect the threat religious fundamentalism can pose to liberal democracy without the intervention of a uniform civil code.Ahmed Khan v. Shah Bano Begum Case was highly controversial and drew in a great amount of political scrutiny. Shah Bano case now stands as a landmark judgement that successfully empowers Muslim women with the right to mainteln the Daniel Latiffi case, the court dealt with a prominent question of whether the Muslim Women's Act abrogated the Judgement of the Hon'ble Supreme Court in Shah Bano Case. The constitutional validity of the Muslim Women's Act was challenged as it was viewed as discriminatory.nance beyond the period of 'iddat' under Section 125 of CrPC.

9b. Explain in detail the Women's rights movement of 1848-1920 and its significance.

The Women's Rights Movement was a long and persistent battle fought by many brave female advocates that came before us such as Elizabeth Cady Stanton, Lucretia Mott and Susan B. Anthony. These women selflessly dedicated their lives to the ratification of the 19th Amendment, which forever changed the lives of womankind in America.The Women's Rights Movement, 1848–1917 The fight for women's suffrage in the United States began with the women's rights movement in the mid-nineteenth century. This reform effort encompassed a broad spectrum of goals before its leaders decided to focus first on securing the vote for women.

10 a. Describe Gender norms?

Gender norms are socially and culturally mediated principles that govern the expected behavior of women, men, girls, and boys in a society. Examples of gender norms include the idea that women should be passive, men should be leaders, girls should be good at sewing, and men should be good at physical tasks.Gender norms are social principles that govern the behavior of girls, boys, women, and men in society and restrict their gender identity into what is considered to be appropriate. Gender norms are neither static nor universal and change over time.Gender refers to the characteristics of women, men, girls and boys that are socially constructed. This includes norms, behaviours and roles associated with being a woman, man, girl or boy, as well as relationships with each other. As a social construct, gender varies from society to society and can change over time.Gender norms are learnt early in life through the process of gender socialisation, which sets common standards and expectations to which girls and boys, and, later, women and men, should conform.Often women and girls are confined to fulfilling roles as mothers, wives and caretakers. Gender norms position girls as caretakers, which leads to gender inequality in how roles are distributed at the household level. This also results in a lack of education due to the restriction of outside opportunities.

10b. Explain stereotyping. Why is it important to break gender norms and stereotypes?

In general, gender stereotyping involves how men and women are expected to act, speak, dress, and conduct themselves, based on their sex. These preconceived gender roles can limit men's and women's capacity to pursue professional careers and prevent them from making individual choices about their lives.

For example, if a girl is expected to become a housewife and care for her family, then society might lose a natural talent for physics. If a boy is valued only for his potential to make a great firefighter, some school might lose a highly talented kindergarten teacher.

Although gender norms exist for different sexes, women have been oppressed throughout history. Society still has deeply ingrained sexist attitudes toward women in general and their role in the modern world.

The most common gender stereotypes for women include:

Girls like wearing pink clothes.

Women should be polite, accommodating, and nurturing.

Women should not be too aggressive, outspoken, or smart.

Housekeeping and childcare are women's responsibility.

Women should educate their children and care for them in every way.

Women shouldn't be part of the workplace. Career and professional advancement shouldn't be important for women.

Women don't make great scientists.

Men are also commonly expected to adhere to specific gender roles:

Boys like racing cars.

Men should be strong, aggressive, and bold.

Men are providers and protectors.

Advanced professional qualifications should be important only for men.

Men don't have to participate in childcare or housekeeping.

Men always have the final say in choosing the place to live and the school for his children.

Gender stereotype examples

Every country and ethnic group has its specific gender role expectations. Traditional gender roles can be very different from culture to culture, and in some cultures, women face dangerous discrimination and violence. Stereotypes of women are more common, but society often expects men to conform to stereotypical gender roles as well.

Some research studies show that women don't like it when men show their true feelings. Men who are considered less masculine and more emotionally expressive might be judged as being poorly adjusted. Young boys who are tender, emotional, and vulnerable are often teased, humiliated, and beaten up. If they're not living up to what society considers acceptable male standards, they might face bullying and name-calling from a young age. Many cultures encourage men to be stoic soldiers, to lead in difficult situations, and never show emotion. Whether they're frightened, anxious, or unhappy, some societies don't let them show their real character.

It's hard to believe that there are places in this world where women don't have the right to vote and can't compete in sports. It's equally hard to believe how, in these modern times, it's still considered unacceptable for men to cry or show emotion. There are a million reasons why we should all fight together for gender equality and put an end to gender stereotypes.

11a. What is feminism? Explain the types of feminism with relevant examples.

Feminism is one of the oldest movements in global history. There's no single definition, but feminism boils down to ending gender discrimination and bringing about gender equality. Within this goal, there are many types of feminism. Instead of describing them in isolation from each other, feminism can be divided into "waves."

The wave metaphor is the most common explanation for feminism's movements, though it's not without flaws. It can oversimplify a complicated history of values, ideas, and people that are often in conflict with each other. With this simplification, one might think feminism's history is a straightforward arc. The reality is much messier. There are many sub-movements building on (and fighting with) each other. That being said, the wave metaphor is a useful starting point. It doesn't tell the whole story, but it helps outline it. There are four waves:

The first wave

The first wave in the late 19th-century was not the first appearance of feminist ideals, but it was the first real political movement for the Western world. In 1792, Mary Wollstonecraft published the revolutionary Vindication of the Rights of Woman. In 1848, about 200 women met in a church. They came up with 12 resolutions asking for specific rights, such as the right to vote. Reproductive rights also became an important issue for early feminists. After years of feminist activism, Congress finally passed the 19th amendment in 1920 and gave women the vote. This was almost 30 years after New Zealand became the first country where women could vote.

The second wave

Second-wave feminism took place in the 1960s and '70s. It built on first-wave feminism and challenged what women's role in society should be. Inspired by the Civil Rights movement and protests against the Vietnam War, activists focused on the institutions that held women back. This meant taking a closer look at why women were oppressed. Traditional gender and family roles were questioned. Queer theory became more established. There were major victories in this era including the Equal Pay Act of 1963, Roe v. Wade in 1973, and other Supreme Court cases.

The third wave

Thanks to the institutional victories of second-wave feminism, women enjoyed more rights and power going into the 1990s. They were able to think about other aspects of their identity, welcoming individuality and rebellion. This was an era of reclaiming. Important cultural touchstones include Eve Ensler's The Vagina Monologues, the Guerilla Girls, and punk rock riot grrls. Many women more freely expressed their sexuality in how they spoke, dressed, and acted. This sometimes bewildered 2nd-wave feminists, many of whom had resisted traditional femininity. While many ideas and mini-movements swirled around in this time, the one "rule" was that there weren't rules. A woman should choose how she lived her life.

The fourth wave

Some people think we're still in the third wave of feminism since the fourth wave isn't so much of a shift as the continued growth of the movement. However, with the MeToo movement and a resurgence of attacks on women's rights, many believe we're living in a new wave. Social media activism has propelled the movement firmly into the technological age. It builds on the third wave's emphasis on inclusivity and asks hard questions about what empowerment, equality, and freedom really mean.

11b. What do you mean by domestic violence? How does domestic violence affect us? Explain.

The term domestic violence refers to abusive behavior in any personal relationship that allows one partner to intimidate, or to gain power and control over the other. This is often thought of to occur between married spouses or in other intimate relationships, but actually refers to any family relationship, or persons living in the same home.This is often thought of to occur between married spouses or in other intimate relationships, but actually refers to any family relationship, or persons living in the same home. Domestic violence includes physical, sexual, psychological, and emotional abuse, as well as threats of violence or economic control.This is often thought of to occur between married spouses or in other intimate relationships, but actually refers to any family relationship, or persons living in the same home. Domestic violence includes physical, sexual, psychological, and emotional abuse, as well as threats of Domestic violence statistics estimate that about 4 million women each year are subjected to abuse committed by their male partners. It is the leading cause of serious injury to women between the ages of 15 and 44, and more than 30 percent of women murdered each year are killed by a former husband or boyfriend.violence or economic control.

12a. Environment movements often contain economic and identity issues ... Discuss.

Domestic violence is violence or abuse in a domestic setting, such as within cohabitation or marriage. Domestic violence is often used as a synonym for intimate partner violence, which involves a spouse or intimate partner in an intimate partner relationship.

The abuser often believes that the abuse is an entitlement, acceptable, justified, or unlikely to be reported. Victims often feel trapped by the abuser in domestic violence situations through isolation by their abuser from family and friends, lack of finances, fear, shame, cultural acceptance, and power and control. Victims can develop physical disabilities and chronic health problems as well as severe psychological disorders. Women's involvement in environmental movements of the United States can be traced back to the early 20th century when women of upper and middle-class backgrounds became active in urban organizations advocating for reform in environmental issues such as sanitation, smoke and noise abatement, civic cleanliness and purity in food and drugs.Women's access to control of natural resources, land ownership and property management is a developing issue and is the subject of continuous debate in both the environmental realm and women's rights movement. "World wide, physical violence by husbands against wives is estimated to range between 10% and 50% (p824)".We can see this work in action at the School for Environment and Sustainability (SEAS). Ivette Perfecto, a SEAS faculty member, elevates women in the environmental justice movement. As a founding member of the Alliance for Women in Agroecology, Dr. Perfecto aims to recognize the obstacles that women farmers face in an overwhelmingly male field.

12 b. Mention about 19th century social and religious reform movements?

19th Century Social and Religious Reform Movements are important from the perspective of the upcoming UPSC exam. Candidates looking forward to appearing for the exam must refer to the details discussed below.

The candidates can go through the relevant topics useful for their upcoming exams from the links provided below:
Bipin Chandra Pal   Raja Ram Mohan Roy Ishwar Chandra Vidyasagar   Swami Vivekananda
The Indian society in the first half of the 19th century was caste-ridden, decadent and rigid.
The conquest of India by the British during the eighteenth and nineteenth centuries, exposed some serious weaknesses and drawbacks of Indian social institutions.
When the British came to India, they introduced the English language as well as certain modern ideas. These ideas were those of liberty, social and economic equality, fraternity, democracy and justice which had a tremendous impact on Indian society.
As a consequence, several individuals and movements sought to bring about changes in social and religious practices with a view to reforming and revitalizing society.
These efforts, collectively known as the Renaissance, were complex social phenomena. It is important to note that this phenomenon occurred when India was under the colonial domination of the British.
There were some enlightened Indians like Raja Ram Mohan Roy, Ishwar Chand Vidyasagar, Dayanand Saraswati and many others who were willing to fight and bring reforms to society so that it could face the challenges of the West.
Types of Reform Movements
Basically, there were two kinds of reform movements in the 19th century in India. Given below are the details about the same, important from the civil services exam preparation:

Reformist
These movements responded with the time and scientific temper of the modern era.
Revivalist
These movements started reviving ancient Indian traditions and thoughts and believed that western thinking ruined Indian culture and ethos.
Reformist Movements
Some of the reformist movements of the 18th and 19th centuries are discussed below:
Brahmo Samaj; Aligarh Movement; Prarthana Samaj ; Deoband Movement ; Ramakrishna Mission; Satyashodhak Samaj;

Young Bengal Movement

13 A Mention the measures to ensure the safety and security of women in India.

Several crimes are happening against women in India every minute. It is now time for us to be aware of the women protection act that preserves us and our rights. Awareness about these laws is essential since power comes with knowledge. There are rights placed to protect women as a parent, wife, daughter, employee, and, most importantly, as a woman. Hence, be mindful of the laws provided by the Government of India that protect women and their rights.

Ten Women Protection Act and Laws that you should know
Below are the ten women protection act and law that every woman should know:

1. The Prohibition of Child Marriage Act, 2006

2. Special Marriage Act, 1954
3. Dowry Prohibition Act, 1961
4. Indian Divorce Act, 1969

5. Maternity Benefit Act, 1861

6. Medical Termination of Pregnancy Act, 1971

7. Sexual Harassment of Women at Workplace Act, 2013

8. Indecent Representation of Women(Prevention) Act,1986

9. National Commission for Women Act, 1990

10. Equal Remuneration Act, 1976

Every woman in India should be aware of the women protection act and law, as mentioned above. The Indian Government places these laws and acts in the interest of women. However, we can only fight against injustice happening with us at the home, workplace, or in society if we are aware of our rights.

13 B Discuss the prominent psychological theories of gender role and gender identity development Gender Identity Development

'Police' and 'Public Order' are State subjects under the Seventh Schedule to the Constitution of India. State Governments are thus responsible for safety and security of the citizens including women and girls.

However, safety and security of women and children in the country is utmost priority for the Government. The Ministry of Women and Child Development has been administering various special laws relating to women such as the Protection of Women from Domestic Violence Act, 2005; Dowry Prohibition Act, 1961; Indecent Representation of Women (Prohibition) Act, 1986; the Sexual Harassment of Women at Workplace (Prevention, Prohibition and Redressal) Act, 2013 and the Prohibition of Child Marriage Act, 2006. The said Ministry is also administering the Juvenile Justice (care and protection of children) Act, 2015, the Commissions for Protection of Child Rights Act, 2005 and the Protection of Children from Sexual Offences Act, 2012.

The Criminal Law (Amendments), Act 2013 was enacted for effective legal deterrence against sexual offences. Further, the Criminal Law (Amendment) Act, 2018 was enacted to prescribe even more stringent penal provisions including death penalty for rape of a girl below the age of 12 years. The Act also inter-alia mandates completion of investigation and trials within 2 months each.

14 A. Critically evaluate the representation of women identity in India cinema.

Women just like in any other industry lack representation in media as well. Most of our stories revolve around men where women are the either 'sidekicks' or 'the love interest'. Pop Culture has its fair share of encouraging gender roles where men are the breadwinners,decision makers or the man who saves the day while women are often shown as caregivers or damsel in distress.

Bollywood movies, scenes and songs are filled with sexist remarks which not only objectifies women but are promoting such unacceptable behaviour. For instance, Salman Khan who plays a cop in the movie Dabangg says "Pyaar se de rahe hai rakhlo varna thappad maar le bhi de sakte hai". He is literally threatening to beat the female character played by Sonakshi Sinha if she doesn't accept the money.

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Women just like in any other industry lack representation in media as well. Most of our stories revolve around men where women are the either 'sidekicks' or 'the love interest'. Pop Culture has its fair share of encouraging gender roles where men are the breadwinners,decision makers or the man who saves the day while women are often shown as caregivers or damsel in distress.

14.B. Explain with examples the social construct of gender.

Introduction

The process by which an adolescent develops its unique and individual identity is known as the gender identity development. In particular in particular with this process the individual consider themselves as masculine or feminine or a combination of two. In this context it is important to note that gender refers to a broad set of qualities such as personal attributes, social roles, customs, behavior, and activity which distinguish between the femininity and masculinity.

Overview of this presentation Here is an overview of this presentation which includes:

Introduction

The history of the study of gender

Back in history, the study begins with the differences in social and biological structure of muscularity and femininity. However the forerunner history is the women's history which hurt by 1960. This includes the study of gender and cultural process. The first part is how men in early generation understood themselves as men. The second part focused on women's scholarship, their upliftment, differences in the social custom, and factors including races are classes, during 1990s. Ways in which gender is studied

This area is considered as an interdisciplinary study which is attributed to categories of analysis represented by gender and the gender identity:The gender study of gender includes following areas: Women's studies which is concerned with feminism politics, social role of women cultural attributes Men's studies include the history analysis of gender dominated society, the differences in customs, and the contribution to the progress of mank. Queer studies which includes the sexual diversity focusing on gay, lesbian, transgender, intersex people and culture, and bisexual.

Psychological theories about gender development

Five important theories of gender development are:

Psychodynamic - This theory found the role of family especially the mother in shaping the individual gender identity.

Symbolic interactionism - This theory is based on communication which is the difference of how to accept and deliver speech among boys and girls.

Social learning - The outward motivational factors which approves or disapprove the behaviour among boys and girls, is known as social learning theory.

Cognitive learning - This theory is based on punishment and external rewards that helps the child to develop identity of their gender by own.

Standpoint - This theory dictates the importance of understanding the gender

Roles of biological, social and cultural influence on gender: Stereotyping and discrimination

Biological difference makes it easier to distinguish between male and female. From biological point of view, intersex individual is the one born with both male and female sex organs.

Stereotype and social stigma include: women as weak sex, they stay at home and raise family, while men go out for work. This also forms discrimination based on social feedback, way of interaction, and job preferences given to individuals.

Present day society include all gender as successful scientist, lawyers, business owners, doctors, and engineers which demonstrate the ability of his or her gender.

Scientific evidence for gender difference: School and work, relationships and sexuality, health and wellness, stress, coping, and psychopathology

Scientific evidence for difference:

Girls are comparative weaker and reserved during "period days".

Dress code at school and work for male and female makes them different. At other social places, for example in parties and beaches, dress code are difference for male and females.

Low population of women in policial decision making position reflect inequalities since ages. This is infact reflective in oldest democracies for example of France and US.Project of women world leaders – established in 1997 have worked towards engagement of women for their progress and change attitude against the existing stereotypes and discrimination. Since 2010, the UN members have increased their women share by 63%.

Future perspectives

Feminist empiricism – They have radical future. They argues on "context of discovery", "context of justification", and "goodness". Issues across culture to consider for reform – Includes: Parenting, abortion, violence against women, and marriage issues. Feminist theorizing – admit biology which is constrained with essentialism. Ideology of power have to be defined equally in society. Other factors that also needs to be considered - Poverty, health aspects, employability, and equal upbringing in family to equalize gender in society.

15 a. Explain the role of media in women's empowerment in Indian society.

The increase in the participation and access of women to self-expression and decision-making through the media and new technologies of communication is in a way empowering women. The powerful and positive role that the media can play in the empowerment of women and gender equality should be supported and further explored.

Media, print and electronic help women to communicate their needs. .Achieving mature thought requires achieving mature use of language. Unless and until the media play a crucial role in the empowerment of women by creating awareness among the people to shed the feudalistic mind-set that denigrated women, true liberation will not be a reality.

The women's liberation movement is also sometimes seen as being synonymous with radical feminism because both were concerned with freeing members of society from oppressive social structure. Both have sometimes been characterized as a threat to men, particularly when the movements use rhetoric about "struggle" and " revolution."

15. b. Explain the role of Role of Women in Panchayat Development and also mention the challenges.

Panchayats have been the backbone of grassroot democracy in the Indian villages since its beginning. Gandhi had aptly favoured the panchayati raj and his dream got translated with the passage of the Constitution (73rdAmendment) Act, 1992 (or simply the Panchayati Raj Act), which introduced the three-tier Panchayati Raj system to ensure people's participation in rural reconstruction in general and that of women in particular. It came into force with effect from April 24, 1993.

Provisions for women in the Act: The Act provides for the reservation of not less than one-third of the total number of seats for women (including the number of seats reserved for the SCs and STs).

Further not less than one-third of the total number of offices of chairpersons in the Panchayats at each level shall be reserved for women. This would be rotated among different Panchayats at each level.

Role of Women in Panchayats

Participation in election: The Act provides for the reservation of not less than one-third of the total number of seats for women. It is an attempt to ensure greater participation of women in election process directly and indirectly. It would be the nursery of creating women politicians for national politics. Even the participation of common women citizens in various activities such as attending Gram Sabha meeting, etc. has reportedly increased (68-78 percent).

Participation in rural development: Women are actively participating in rural development as per their capacity right from labourers to policy- makers.

Participation in decision-making: The participation of women as elected as well as non-elected members are rising due to reservation for women. It acts as pull factor for women to participate in meeting. They give their suggestions for various works and problems faced by them.

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