NSRIT

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ANSWER KEY & SCHEME OF EVALUATION

B. Tech. (S3 Regular & Supplementary Jan. 2023)

ACADEMIC REGULATION 2020

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Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech.	**Program**	CE, EEE& MECH			**Academic Year**	2022 - 2023
Course Code	20BSX13	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Numerical Methods & Transforms**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Prove that $(1+\Delta)(1-\nabla)=1$	20BSX13.1	L1
2	Write the iteration formula for Jacobi's method	20BSX13.2	L1
3	Write Simpson's 3/8th rule	20BSX13.3	L1
4	Find L([t])	20BSX13.4	L1
5	State the Shifting property of Fourier transforms	20BSX13.5	L1

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

<table>
<tr><th>No.</th><th>Questions (6 through 15)</th><th>Marks</th><th>Learning Outcome (s)</th><th>DoK</th></tr>
<tr><td>6 (a)</td><td>Solve the following equations by Gauss – Seidel method:
5x + 2y +z =12, x + 4y +2z =15, x+2y+5z = 20</td><td>6M</td><td>20BSX13.1</td><td>L2</td></tr>
<tr><td>6 (b)</td><td>Find a real root of the equation 3x - cosx = 1 by Newton - Raphson method, correct to three decimal places, near x= 0.6</td><td>6M</td><td>20BSX13.1</td><td>L2</td></tr>
<tr><td colspan="5" align="center">OR</td></tr>
<tr><td>7</td><td>Find the root of the equation $x^3-2x-5=0$, using bisection method, correct to three decimal places</td><td>12M</td><td>20BSX13.1</td><td>L2</td></tr>
<tr><td>8</td><td>Using Lagrange's interpolation formula, find the value of y when x = 9, if the following values of x and y are given:
<table>
<tr><td>x</td><td>5</td><td>7</td><td>11</td><td>13</td><td>17</td></tr>
<tr><td>y</td><td>150</td><td>392</td><td>1452</td><td>2366</td><td>5202</td></tr>
</table></td><td>12M</td><td>20BSX13.2</td><td>L2</td></tr>
<tr><td colspan="5" align="center">OR</td></tr>
<tr><td>9</td><td>Find the number of men getting wages below Rs. 15 from the data
<table>
<tr><td>Wages in Rs.</td><td>0-10</td><td>10-20</td><td>20-30</td><td>30-40</td></tr>
<tr><td>Frequency</td><td>9</td><td>30</td><td>35</td><td>42</td></tr>
</table></td><td>12M</td><td>20BSX13.2</td><td>L2</td></tr>
<tr><td>10 (a)</td><td>Evaluate $\int_0^6 \frac{1}{1+x^2}dx$ by using Trapezoidal rule, by dividing the interval (0, 6) into 6 parts</td><td>6M</td><td>20BSX13.3</td><td>L2</td></tr>
<tr><td>10 (b)</td><td>Solve $y' = x^2y-1, y(0)=1$ using Euler's [redacted] method and compute $y(0.1)$. Take h = 0.02</td><td>6M</td><td>20BSX13.3</td><td>L3</td></tr>
<tr><td colspan="5" align="center">OR</td></tr>
<tr><td>11 (a)</td><td>Evaluate $\int_0^1 \frac{1}{x}\, dx$ by Simpson's $\frac{1}{3}$ rd rule with 4 subintervals</td><td>6M</td><td>20BSX13.3</td><td>L2</td></tr>
<tr><td>11 (b)</td><td>Using Runge-Kutta method, evaluate $y(0.1)$ given that $y' = x + y^2, y(0) = 1$</td><td>6M</td><td>20BSX13.3</td><td>L3</td></tr>
<tr><td>12 (a)</td><td>Show that $\int_0^\infty e^{-2t}\cos 3t\, dt = \frac{2}{13}$</td><td>4M</td><td>20BSX13.4</td><td>L2</td></tr>
</table>

12 (b)	Using convolution theorem, evaluate $L^{-1}\left(\frac{s}{(s^2+a^2)^2}\right)$	8M	20BSX13.4	L3
	OR			
13 (a)	Evaluate Laplace transform of $\frac{\cos at - \cos bt}{t}$	6M	20BSX13.4	L2
13 (b)	Solve $y'' + 4\,y' + 3y = e^{-t}$, if $y(0) = y'(0) = 1$ using transform method	6M	20BSX13.4	L3
14	Find the Fourier sine and cosine transform of $f(x) = \begin{cases} x, & \text{for } 0 < x < 1 \\ 2 - x, & \text{for } 1 < x < 2 \\ 0, & \text{for } x > 2 \end{cases}$	12M	20BSX13.5	L2
	OR			
15	Find the Fourier integral of $f(x) = \begin{cases} 1 \text{ for } \lvert x\rvert < 1 \\ 0 \text{ for } \lvert x\rvert > 1 \end{cases}$	12M	20BSX13.5	L2

ANSWER KEY AND SCHEME OF EVALUATION

12 b) $L^{-1}\frac{s}{(s^2+a^2)^2}$

Sol:- $f(t) = L^{-1}\frac{s}{s^2+a^2} = \cos at$

$g(t) = L^{-1}\frac{1}{(s^2+a^2)} = \frac{1}{a}\sin at$ 2M

By convolution Theorem

If $L^{-1}\bar{f}(s) = f(t)$, $L^{-1}(\bar{g}(s) = g(t)$, then

$L^{-1}[\bar{f}(s)\cdot\bar{g}(s)] = \int_0^t f(u)\,g(t-u)\,du = F*G$ 1M

$L^{-1}\left[\frac{s}{(s^2+a^2)}\cdot\frac{1}{(s^2+a^2)}\right] = \int_0^t \cos au \,\frac{\sin a(t-u)}{a}\,du$

[∵ $f(u) = \cos au$

$g(t-u) = \frac{1}{a}\sin a(t-u)$

$= \frac{1}{2a}\int_0^t [\sin at - \sin(2au - at)]\,dt$

$= \frac{1}{2a}\left| u\sin at + \frac{1}{2a}\cos(2au-at)\right|_0^t$

$= \frac{1}{2a}\,t\sin at$ 3M

Hence $L^{-1}\left(\frac{s}{(s^2+a^2)^2}\right) = \frac{1}{2a}\,t\sin at$

13(a) $L\left[\frac{\cos at-\cos bt}{t}\right]$

Sol If $L\,f(t)=\bar{f}(s)$, then $L\left[\frac{1}{t}f(t)\right]=\int_s^\infty \bar{f}(s)\,ds$

$L(\cos at-\cos bt)=\frac{s}{s^2+a^2}-\frac{s}{s^2+b^2}$ 3M

$\therefore L\,\frac{\cos at-\cos bt}{t}=\int_s^\infty\left(\frac{s}{s^2+a^2}-\frac{s}{s^2+b^2}\right)ds$

$=\frac{1}{2}\log(s^2+a^2)-\frac{1}{2}\log(s^2+b^2)\Big|_s^\infty$ 3M

$=\frac{1}{2}\lim_{s\to\infty}\log\frac{s^2+a^2}{s^2+b^2}-\frac{1}{2}\log\frac{s^2+a^2}{s^2+b^2}$

$=\frac{1}{2}\log(1)-\frac{1}{2}\log\frac{s^2+a^2}{s^2+b^2}=\log\left(\frac{s^2+b^2}{s^2+a^2}\right)^{1/2}$

13(b) $y''+4y'+3y=e^{-t}$, if $y(0)=y'(0)=1$

$L(y'')=s^2\bar{y}-s\,y(0)-y'(0)$

$L(y')=s\bar{y}-y(0)$

$L(y)=\bar{y}$

$\therefore L[y''+4y'+3y]=Le^{-t}$ 3M

$s^2\bar{y}-s\,y(0)-y'(0)+4s\bar{y}-4y(0)+3\bar{y}=\frac{1}{s+1}$

$(s^2+4s+3)\bar{y}-s-5=\frac{1}{s+1}$

$(s^2+4s+3)\bar{y}=\frac{1}{s+1}+s+5=\frac{s^2+6s+6}{s+1}$

ANSWER KEY AND SCHEME OF EVALUATION

$$\bar{y} = \frac{s^2+6s+6}{(s+1)(s^2+4s+3)}$$

Now apply I.L.T

$$L^{-1}[\bar{y}] = L^{-1}\left[\frac{s^2+6s+6}{(s+1)(s+1)(s+3)}\right]$$

$$y = L^{-1}\left[\frac{A}{s+1} + \frac{B}{(s+1)^2} + \frac{C}{(s+3)}\right]$$

$$\frac{s^2+6s+6}{(s+1)^2(s+3)} = \frac{A}{s+1} + \frac{B}{(s+1)^2} + \frac{C}{(s+3)}$$

$$s^2+6s+6 = A(s+1)(s+3) + B(s+3) + C(s+1)^2$$

Put $s = -1$

$1-6+6 = 2B \Rightarrow B = \frac{1}{2}$

Put $s = -3$

$4C = 9-18+6 = -3$

$C = -\frac{3}{4}$

$A+C = 1 \Rightarrow A = 1+\frac{3}{4} = \frac{7}{4}$

$$\therefore y = \frac{7}{4}L^{-1}\frac{1}{s+1} + \frac{1}{2}L^{-1}\frac{1}{(s+1)^2} - \frac{3}{4}L^{-1}\frac{1}{(s+3)}$$

$$\boxed{y = \frac{7}{4}e^{-t} + \frac{1}{2}t\cdot e^{-t} - \frac{3}{4}e^{-3t}}$$

3M

14) $f(x) = \begin{cases} x & ; \quad 0 < x < 1 \\ 2-x & ; \quad 1 < x < 2 \\ 0 & ; \quad 2 < x < \infty \end{cases}$

The fourier cosine transform of $f(x)$ is given by

$$F_c(s) = \int_0^\infty f(x) \cos(sx)\, dx$$

$$= \int_0^1 x \cos sx\, dx + \int_1^2 (2-x) \cos sx\, dx + \int_2^\infty (0) \cos sx\, dx$$

1 Mark

Consider $\int_0^1 x \cos sx\, dx$

at $x = u$ $\quad \cos sx = v$

Then $u' = 1$, $u'' = 0$, $v_1 = \frac{\sin sx}{s}$, $v_2 = -\frac{\cos sx}{s^2}$

we have $\int uv\, dx = uv_1 - u'v_2 - u''v_3 + \dots$

$$\therefore \int_0^1 x \cos sx\, dx = \left[x\left(\frac{\sin sx}{s}\right) - (1)\left(\frac{-\cos sx}{s^2}\right) + 0\right]_{x=0}^{1}$$

$$= \left[\frac{x \sin sx}{s} + \frac{\cos sx}{s^2}\right]_{x=0}^{1}$$

$$= \left(\frac{\sin s}{s} + \frac{\cos s}{s^2}\right) - \left[0 + \frac{1}{s^2}\right] = \frac{\sin s}{s} + \frac{\cos s}{s^2} - \frac{}{s^2}$$

$\longrightarrow$ 2 Marks

Now consider $\int_1^2 (2-x)\cos sx\,dx$

Let $2-x=u$, & $\cos sx = v$

Then $u'=-1$, $u''=0$, $v_1=\frac{\sin sx}{s}$, $v_2=\frac{-\cos sx}{s^2}$

$$\int_1^2 (2x)\cos sx\,dx = \left[(2-x)\frac{\sin sx}{s} - (-1)\left(-\frac{\cos x}{s^2}\right) + 0\right]_{x=1}^{2}$$

$$= \left.\frac{(2-x)\sin sx}{s} - \frac{\cos sx}{s^2}\right|_{x=1}^{2}$$

$$= -\frac{\cos 2s}{s^2} - \frac{\sin s}{s} + \frac{\cos s}{s^2} \quad \text{— 2 Mark}$$

By substitutin in ① we get

$$F_c(s) = \frac{\sin s}{s} + \frac{\cos s}{s^2} - \frac{1}{s^2} - \frac{\cos 2s}{s^2} - \frac{\sin s}{s} + \frac{\cos s}{s^2}$$

$$= \frac{2\cos s - 1 - \cos 2s}{s^2} \quad \text{— 1M}$$

The fourier sin transform of $f(x)$ is

$$F_s(s) = \int_0^\infty f(x)\sin sx\,dx$$

$$\left.\begin{aligned} &= \int_0^1 x\sin sx\,dx + \int_1^2 (2-x)\sin sx\,dx + \\ &\quad \int_2^\infty (0)\sin sx\,dx \end{aligned}\right\} \text{1M}$$

Consider $\int_0^1 x \sin sx \, dx$

Let $x = u$, $\sin sx = v$ $\Big\}$

Then $u' = 1$, $v'' = 0$, $v_1 = -\frac{\cos sx}{s}$, $v_2 = -\frac{\sin sx}{s^2}$

$\therefore \int_0^1 x \sin sx \, dx = \int_0^1 uv \, dx = (uv_1 - u'v_2 + u''v_3 \cdots)\Big|_0^1$

$= \left[x\left(-\frac{\cos sx}{s}\right) - (1)\left(-\frac{\sin sx}{s^2}\right) + 0\right]_{x=0}^{1}$

$= -\frac{x\cos sx}{s} + \frac{\sin sx}{s^2}\Big|_0^1$

$= -\frac{\cos s}{s} + \frac{\sin s}{s^2}$ —— 2M

Now Consd $\int_1^2 (2-x)\sin sx \, dx$

$2-x = u$, $\sin sx = v$

Then $u' = -1$, $u'' = 0$, $v_1 = -\frac{\cos sx}{s}$, $v_2 = -\frac{\sin sx}{s^2}$

$\int_1^2 (2-x)\sin sx \, dx = \int_1^2 uv \, dx = (uv_1 - u'v_2 + u''v_3 \cdots)_1^2$

$= \left[(2-x)\left(-\frac{\cos sx}{s}\right) - (-1)\left(-\frac{\sin sx}{s^2}\right) + 0\right]_{x=1}^{2}$

$= -\frac{\sin 2s}{s^2} + \frac{\cos s}{s} + \frac{\sin s}{s^2}$ $\Big\}$ 2M

By substituti

$F_s(s) = -\frac{\cos s}{s} + \frac{\sin s}{s^2} - \frac{\sin 2s}{s^2} + \frac{\cos s}{s} + \frac{\sin s}{s^2}$

$= \frac{2\sin s - \sin 2s}{s^2}$ $\Big\}$ 1M

ANSWER KEY AND SCHEME OF EVALUATION

15) The given function is $f(x) = \begin{cases} 1 ; & |x| < 1 \\ 0 ; & |x| > 1 \end{cases}$

$$\Rightarrow f(x) = \begin{cases} 0 ; & -\infty < x < -1 \\ 1 ; & -1 < x < 1 \\ 0 ; & 1 < x < \infty \end{cases}$$

The fourier integral of $f(x)$ is

$$f(x) = \int_0^\infty [A(\lambda)\cos\lambda x \, d\lambda + B\lambda \sin\lambda x \, d\lambda] \quad - ①$$

where $A(\lambda) = \frac{1}{\pi}\int_{-\infty}^{\infty} f(x)\cos\lambda x \, d\lambda \quad - ②$

$B(\lambda) = \frac{1}{\pi}\int_{-\infty}^{\infty} f(x)\sin\lambda x \, d\lambda \quad - ③$

} 2M

$$\text{Now } A\lambda = \frac{1}{\pi}\left[\int_{\infty}^{-1}(0)\cos\lambda x + \int_{-1}^{1}(1)\cos\lambda x \, dx + \int_{1}^{\infty} 0 \cos\lambda x \, dx\right]$$

$$= \frac{1}{\pi}\left(\frac{\sin\lambda x}{\lambda}\right)_{x=-1}^{1} = \frac{1}{\pi\lambda}(\sin\lambda - \sin(-\lambda))$$

$$= \frac{2\sin\lambda}{\pi\lambda}$$

} 4M

$$B(\lambda) = \frac{1}{\pi} \int_{-\infty}^{1} (0) \sin\lambda x \, dx + \int_{-1}^{1} (1) \sin\lambda x \, dx + \int_{1}^{\infty} (0) \sin\lambda x \, d$$

$$\left.\begin{aligned} &= \frac{1}{\pi}\left(-\frac{\cos\lambda x}{\lambda}\right)_{-1}^{1} = -\frac{1}{\pi\lambda}\left(\cos\lambda x\right)_{-1}^{1} \\ &= -\frac{1}{\pi\lambda}\left[\cos\lambda - \cos(-\lambda)\right] = 0 \end{aligned}\right\} 4M$$

By substituting $A(\lambda)$ & $B(\lambda)$ values in ①, we get

$$f(x) = \int_{0}^{\infty} \frac{2\sin\lambda}{\pi\lambda} \cos\lambda x \, d\lambda$$

$$= \frac{2}{\pi} \int_{0}^{\infty} \frac{\sin\lambda \cos\lambda x}{\lambda} \, d\lambda \quad \text{—2M}$$

11 b)

Using R-K Method,

$y' = x+y^2,\quad y(0) = 1 \qquad y(0.1) = ?$

RK 4th order

$y_1 = y_0 + K$

$K = \frac{1}{6}[K_1 + 2K_2 + 2K_3 + K_4]$

$K_1 = h.f(x_0, y_0) \qquad = h[x_0 + y_0^2]$

$K_2 = h.f(x_0+\frac{h}{2}, y_0+\frac{K_1}{2}) = h[(x_0+\frac{h}{2}) + (y_0+\frac{K_1}{2})^2]$

$K_3 = h.f(x_0+\frac{h}{2}, y_0+\frac{K_2}{2}) = h[(x_0+\frac{h}{2}) + (y_0+\frac{K_2}{2})^2]$

$K_4 = h.f(x_0+h, y_0+K_3) = h[(x+h) + (y_0+K_3)^2]$ 4M

$h = 0.1,\ K_1 = 0.1\ \ K_2 = 0.1152,\ K_3 = 0.1168,\ K_4 = 0.1347$

$K = \frac{1}{6}(K_1 + 2K_2 + 2K_3 + K_4) = 0.1165$

So $y(0.1) = y_0 + K = 1 + 0.1165 = 1.1165$ 2M

If student do R.K (2nd order also give 6M

12 (a) $\int_0^\infty e^{-2t} \cos 3t\, dt$ This is in the for $\int_0^\infty e^{-st} f(t)\, dt$

$Lf(t) = L[\cos 3t] = \frac{s}{s^2+9}$ where $s = 2$

$= \frac{2}{4+9} = \frac{2}{13}$ 4M

ANSWER KEY AND SCHEME OF EVALUATION

11) (a) $\int_0^1 \frac{1}{x} dx$

Simpson's $\frac{1}{3}$rd rule $= \frac{h}{3}[(y_0+y_n)+4(y_1+y_3+y_5+y_7\ldots)$
$+2[y_2+y_4+y_6+\ldots)]$ 2M

$y = f(x) = \frac{1}{x}$

Table $h = 0.25$

x	0	0.25	0.5	0.75	1.0
y	undefined y_0	$\frac{1}{0.25}$ y_1	$\frac{1}{0.5}$ y_2	$\frac{1}{0.75}$ y_3	1 y_4

2M

Simpsons $\frac{1}{3}$ rule $= \frac{h}{3}[(y_0+y_4)+4(y_1+y_3)+2y_2]$

But we can not find y_0 becaus $\frac{1}{0}$ is undefined

∴ The solution is not possible.

2M

$h = 0.02$

$x_1 = x_0 + h = 0.02$, $x_2 = 0.04$, $x_3 = 0.06$, $x_4 = 0.08$, $x_5 = 0.1$

$y_1 = y(0.02) = y_0 + hf(x_0, y_0) = y_0 + h[x_0^2 y_0 - 1]$

$= 1 + 0.02[-1] = 0.98$

$y_2 = y(0.04) = y_1 + hf(x_0, y_1)$

$= 0.98 + 0.02[x_1^2 y_1 - 1]$

$= 0.96$

2M

$y_3 = y(0.06) = y_2 + hf(x_2, y_2)$

$= 0.96 + 0.02[x_2^2 y_2 - 1]$

$= 0.94$

$y_4 = y(0.08) = y_3 + hf(x_3, y_3)$

$= y_3 + h[x_3^2 y_3 - 1]$

$= 0.92$

$y_5 = y(0.1) = y_4 + hf(x_4, y_4)$

$= 0.92 + 0.02[(0.08)^2(0.02) - 1]$

2M

$$\boxed{y(0.1) = 0.9001}$$

ANSWER KEY AND SCHEME OF EVALUATION

10 a) $\int_0^6 \frac{1}{1+x^2}\,dx$

$y = f(x) = \frac{1}{1+x^2}$ Data – 1M

x	0	1	2	3	4	5	6
$y = \frac{1}{1+x^2}$	$\frac{1}{1}$	$\frac{1}{2}$	$\frac{1}{5}$	$\frac{1}{10}$	$\frac{1}{17}$	$\frac{1}{26}$	$\frac{1}{32}$
	y_0	y_1	y_2	y_3	y_4	y_5	y_6

2M

$h = 1$

Trapezoidal rule $= \frac{h}{2}\left[(y_0 + y_6) + 2[y_1 + y_2 + y_3 + y_4 + y_5]\right]$ } 1M

$= \frac{1}{2}\left[\left(1 + \frac{1}{37}\right) + 2\left[\frac{1}{2} + \frac{1}{5} + \frac{1}{10} + \frac{1}{17} + \frac{1}{26}\right]\right]$

$= \frac{1}{2}\left[(1 + 0.027) + 2[0.5 + 0.2 + 0.1 + 0.0588 + 0.0385]\right]$

$= 1.4108$ } 2M

10 b) $\frac{dy}{dx} = x^2 y - 1$

Euler's Method $y(x_1) = y_0 + h\, f(x_0, y_0)$

$y(0) = 1$ $x_0 = 0,\ y(0) = 1$

} 2M

Numerical Methods & Transforms

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ANSWER KEY AND SCHEME OF EVALUATION

Part A

① P.T $(1+\Delta)(1-\nabla) = 1$

Sol:- We know that $1+\Delta = E$

and $1-\nabla = E^{-1}$

$\therefore (1+\Delta)(1-\nabla) = E \cdot E^{-1} = 1$ } 2M

2. Jacobi's iteration formula

$a_1x + b_1y + c_1z = d_1, \quad a_2x+b_2y+c_2z = d_2, \quad a_3x+b_3y+c_3z = d_3$

$x_{i+1} = \frac{1}{a_1}[d_1 - b_1y_i - c_1z_i]$

$y_{i+1} = \frac{1}{b_2}[d_2 - a_2x_i - c_2z_i]$ } 2M

$z_{i+1} = \frac{1}{c_3}[d_3 - a_3x_i - b_3y_i$

3. Simpson's $\frac{3}{8}$th rule $= \frac{3h}{8}\Big[(y_0+y_n) + 3[y_1+y_2+y_4+y_5+\dots]$

$+ 2[y_3+y_6+y_9+\dots]\Big]$ } 2M

4 $L(t) = \frac{1}{s^2}$ $\left[L(t^n) = \frac{n!}{s^{n+1}}\right]$ } 2M

5. Shifting Property

If $F(s)$ is the complex fourier transforms of $f(x)$, then $F\{f(x-a)\} = e^{isax} F(s)$ $\Big\}$ 2M

6(a) $5x+2y+z=12$

$x+4y+2z=15$

$x+2y+5z=20$

$$\left.\begin{aligned} x &= \frac{1}{5}[12-2y-z] \\ y &= \frac{1}{4}[15-x-2z] \\ z &= \frac{1}{5}[20-x-2y] \end{aligned}\right\} 2M.$$

First iteration

$y_0=z_0=0$

$$x_1 = \frac{1}{5}[12-2y_0-z_0] = \frac{12}{5} = 2.4$$

$$\left.\begin{aligned} y_1 &= \frac{1}{4}[15-x_1-2z_0] = \frac{1}{4}[15-2.4] = 3.15 \\ z_1 &= \frac{1}{5}[20-x_1-2y_1] = \frac{1}{5}[20-2.4-2(3.15) = 2.26 \\ &\text{Second iteration} \\ x_2 &= \frac{1}{5}[12-2(3.15)-2.26] = 0.688 \end{aligned}\right\} 2M$$

$$y_2 = \frac{1}{4}[15-0.688-2(2.26)] = 2.448$$

$$z_2 = \frac{1}{5}[20-0.688-2(2.448)] = 2.883$$

Third iterate

$x_3 = 0.8442$

$y_3 = 2.0975$

$z_3 = 2.992$

Fourth iterate

$x_4 = 0.9626$

$y_4 = 2.0133$

$z_4 = 3.002$

Fifth iterate

$x_5 = 1$

$y_5 = 2$

$z_5 = 3$

2M

∴ Solution $x = 1,\ y = 2,\ z = 3$

6 (b) Let $f(x) = 3x - \cos x - 1$

$x_0 = 0.6$

$f'(x) = 3 + \sin x$

2 M

Newtons iteration formula

$$x_{n+1} = x_n - \frac{f(x_n)}{f'(x_n)} = x_n - \frac{3x_n - \cos x_n - 1}{3 + \sin x_n}$$

$$= \frac{\cancel{3x_n} + x_n \sin x_n - \cancel{3x_n} + \cos x_n + 1}{3 + \sin x_n}$$

$$= \frac{x_n \sin x_n + \cos x_n + 1}{3 + \sin x_n}$$

Putting $n=0$, the first approximation x_1 is given by →2M

$$x_1 = \frac{x_0 \sin x_0 + \cos x_0 + 1}{3 + \sin x_0} = \frac{(0.6)\sin(0.6) + \cos(0.6) + 1}{3 \sin(0.6)} = 0.6071$$

Putting $n=1$ in (i), the second approximation is

$$x_2 = \frac{x_1 \sin x_1 + \cos x_1 + 1}{3 + \sin x_1} = 0.6071 \quad \text{clearly } x_1 = x_2 \Big\} 2M$$

Hence the desired root is 0.6071 correct to four decimal places.

7) $f(x) = x^3 - 2x - 5 = 0$

Find the interval

$x=0 \quad f(0) = -5$

$x=1 \quad f(1) = -6$

$x=2 \quad f(2) = -1$

$x=3 \quad f(3) = 16$

2M

$a = 2, \quad b = 3$

$$x_1 = \frac{a+b}{2}$$

iterati	a	b	$x_i = \frac{a+b}{2}$	$f(x)$
1	2	3	2.5	5.675
2	2	2.5	2.25	1.8926
3	2	2.25	2.125	0.3457
4	2	2.125	2.0625	-0.3573
5	2.0625	2.125	2.0938	-0.00838
6	2.0938	2.125	2.109~~3~~9	0.3423
7	2.0938	2.1094	2.1016	0.6789
8	2.0938	2.1016	2.0977	0.635
9	2.0938	2.0977	2.0958	0.0132
10	2.0938	2.0958	2.0948	2.779×10^{-3}
11	2.0948	2.0943		

10 M

The root is 2.0943 Ans

8)

	x_0	x_1	x_2	x_3	x_4
x	5	7	11	13	17
y	150	392	1452	2366	5202
	y_0	y_1	y_2	y_3	y_4

} 1M

$$y(x) = \frac{(x-x_1)(x-x_2)(x-x_3)(x-x_4)}{(x_0-x_1)(x_0-x_2)(x_0-x_3)(x_0-x_4)}.y_0 + \frac{(x-x_0)(x-x_2)(x-x_3)(x-x_4)}{(x_1-x_0)(x_1-x_2)(x_1-x_3)(x_1-x_4)}.y_1$$

$$+\frac{(x-x_0)(x-x_1)(x-x_3)(x-x_4)}{(x_2-x_0)(x_2-x_1)(x_2-x_3)(x_2-x_4)}.y_2 + \frac{(x-x_0)(x-x_1)(x-x_2)(x-x_4)}{(x_3-x_0)(x_3-x_1)(x_3-x_2)(x_3-x_4)}.y_3$$

$$+\frac{(x-x_0)(x-x_1)(x-x_2)(x-x_3)}{(x_4-x_0)(x_4-x_1)(x_4-x_2)(x_4-x_3)}.y_4$$

} 2M

$$y(9) = \frac{(9-7)(9-11)(9-13)(9-17)}{(5-7)(5-11)(5-13)(5-17)}.150 + \frac{(9-5)(9-11)(9-13)(9-17)}{(7-5)(7-11)(7-13)(7-17)}.392$$

$$+\frac{(9-5)(9-7)(9-13)(9-17)}{(11-5)(11-7)(11-13)(11-17)}.1452 + \frac{(9-5)(9-7)(9-11)(9-13)}{(13-5)(13-7)(13-11)(13-17)}.2366$$

$$+\frac{(9-5)(9-7)(9-11)(9-13)}{(17-5)(17-7)(17-11)(17-13)}.5202$$

} 2M

$$= -16.6 + 209.07 + 1290.66 - 788.66 + 115.6$$

$$= 810$$

} 1M

9) Prepare cumulative frequency table

Wages less than (x)	10	20	30	40
no of men (y)	9	39	74	116

} 2M

Now the difference table

x	y	Δy	Δ^2	Δ^3
x_0 10	y_0 9			
		30		
20	39		5	
		35		2
30	74		7	
		42		
40	116			

} 3M

Taking $x_0 = 10$, $x = 15$, we have $p = \frac{x - x_0}{h} = \frac{15-10}{10} = 0.5$

Newtons forward interpolation

$$y_{15} = y_{10} + p\Delta y_0 + \frac{p(p-1)}{2!}\Delta^2 y_0 + \frac{p(p-1)(p-2)}{3!}\Delta^3 y_0$$

$$= 9 + 0.5 \times 30 + \frac{(0.5)(-0.5)}{2} \cdot 5 + \frac{(0.5)(-0.5)(-1.5)}{3 \times 2} \times 2$$

} 3M

$$= 24$$ − 1M

Number men getting wages below 15 are 24

CV&T (20BSX14) END SEM(2022).

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech	Program	ECE			Academic Year	2022 - 2023
Course Code	20BSX14	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	Complex Variables & Transforms						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Harmonic function	20BSX14.1	L1
2	Find the poles of $\frac{z+1}{z^2+1}$	20BSX14.2	L2
3	State Euler's formula	20BSX14.3	L2
4	Find the Laplace transform of 1-e^t	20BSX14.4	L1
5	Define inverse Fourier transform	20BSX14.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Determine the analytic function $f(z) = u + iv, if\ u - v = \frac{\cos x+\sin x-e^{-y}}{2(\cos x-\cosh y)}$ and $f\left(\frac{\pi}{2}\right) = 0$	6M	20BSX14.1	L3
6 (b)	If f(z) is an analytic function with constant modulus, show that f(z) is constant	6M	20BSX14.1	L2
	OR			
7 (a)	Find the regular function whose imaginary part is v = log (x^2 + y^2) + x - 2y	6M	20BSX14.1	L2
7 (b)	Use Cauchy's integral formula to calculate $\oint_C \frac{\sin\pi z+\cos\pi z}{(z-1)(z-2)}dz\ where\ C\ is\ \|z\| = 4$	6M	20BSX14.1	L3
8 (a)	Find the Laurents series expansion of $f(z) = \frac{7z-2}{(z+1)z(z-2)}$ in the region $1 < z + 1 < 3$	6M	20BSX14.2	L2
8 (b)	Find the residue of the function $\frac{ze^z}{(z-1)^2}$ at the each pole	6M	20BSX14.2	L3
9 (a)	Evaluate $\oint_C \frac{z^2+4}{z^3+2z^2+2z}dz$, where C is the circle given by $(i)\|z\| = 1;\ (ii)\|z + 1 - i\| = 1;\ (iii)\|z + i + 1\| = 1;\ (iv)\ \|z - 1\| = 5$	6M	20BSX14.2	L2
9 (b)	Find Laurent's series of $f(z) = \frac{e^z}{z(1-z)}$ about z=1. Find the region converges	6M	20BSX14.2	L3
10 (a)	Find a Fourier series to represent $x - x^2$ from $x = -\pi\ to\ x = \pi$	6M	20BSX14.3	L2
10 (b)	Expand $f(x) = \sqrt{(1 - \cos x)}$. when $0 < x < 2\pi$	6M	20BSX14.3	L3

	is a Fourier series. Hence evaluate $\frac{1}{1.3}+\frac{1}{3.5}+\frac{1}{5.7}+\cdots$			
	OR			
11 (a)	Find the Fourier series expansion of $f(x)=2x-x^2$ in (0,3) and hence deduce that $\frac{1}{1^2}+\frac{1}{2^2}+\cdots=\frac{\pi}{2}$	6M	20BSX14.3	L2
11 (b)	Obtain a half-range cosine series for consider $f(x)=\pi-x\ in\ 0<x<\pi.$ Hence show that $\sum_{r=0}^{\infty}\frac{1}{(2r+1)^2}=\frac{\pi^2}{8}$	6M	20BSX14.3	L3
12 (a)	Find the Laplace transform of $\frac{\cos at-\cos bt}{t}+t\sin at$	6M	20BSX14.4	L2
12 (b)	Apply convolution theorem to solve $\frac{s^2}{(s^2+a^2)(s^2+b^2)}$	6M	20BSX14.4	L3
	OR			
13 (a)	Find the inverse Laplace transform of $\frac{s}{s^4+4a^4}$	6M	20BSX14.4	L2
13 (b)	Using the Laplace transform method to solve $(D^2-3D+2)y=4e^{2t}$ with $y(0)=3; y'(0)=5$	6M	20BSX14.4	L3
14 (a)	Find the Fourier cosine transform of x	6M	20BSX14.5	L2
14 (b)	Find the Fourier transform of $f(x)\begin{cases}a^2-x^2 & for\ \lvert x\rvert\le a\\ 0 & for\ \lvert x\rvert>a\end{cases}$ Hence deduce that $\int_0^{\infty}\frac{\sin t-t\cos t}{t^3}\,dt=\frac{\pi}{4}$	6M	20BSX14.5	L3
	OR			
15 (a)	Find the Fourier Cosine transform of e^{-ax} Hence evaluate $\int_0^{\infty}\frac{\cos\tau x}{x^2+a^2}\,dx$	6M	20BSX14.5	L2
15 (b)	Show that the inverse finite Fourier sine transform of $F_n(x)=\frac{1}{\pi}\left\{1+\cos n\pi-2\cos\frac{n\pi}{2}\right\}\ is\ f(x)=$ $\begin{cases}1 & 0<x<\pi/2\\ -1 & \frac{\pi}{2}<x<\pi\end{cases}$	6M	20BSX14.5	L3

E.C.E Branch (A&B)
Key & Scheme of Valuation
(CV&T
20BSX14)

NSRIT

Marks: 70.

①

N S RAJU INSTITUTE OF TECHNOLOGY
(AUTONOMOUS)
SONTYAM, ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

COMPLEX VARIABLES & TRANSFORMS (CV&T)
END SEMESTER EXAMS: 2022 December
26.12.2022. (20BSX14): MARKS: 70:

SECTION – A (2X5=10)

① Define Harmonic function.

Ans:– Solutions of Laplace equations having continuous second order partial derivatives are called Harmonic functions

$$\frac{\delta^2\phi}{\delta x^2} + \frac{\delta^2\phi}{\delta y^2} = 0$$ are known as Harmonic function. $\longrightarrow$ 2 Marks.

② Find the poles of $\frac{z+1}{z^2+1}$

Sol:– Poles of $f(z)$ are obtained by putting the denominator equal to zero.

$$\text{Let } z^2+1=0 \Rightarrow (z^2-i^2)=0$$

$$\Rightarrow (z-i)(z+i)=0$$

$$\Rightarrow z=-i, i$$

$\therefore$ Poles $z=-i, i$ and order $=1$

Hence the solution. $\longrightarrow$ 2 Marks

③ State the Euler's formula

Sol:- The Fourier series for the function f(x) in the interval $\alpha < x < \alpha + 2\pi$ is given by

$$f(x) = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos nx + \sum_{n=1}^{\infty} b_n \sin nx$$

where $a_0 = \frac{1}{\pi}\int_{\alpha}^{\alpha+2\pi} f(x)\,dx,$

$$\left.\begin{aligned} a_n &= \frac{1}{\pi}\int_{\alpha}^{\alpha+2\pi} f(x)\cos nx\,dx \\ b_n &= \frac{1}{\pi}\int_{\alpha}^{\alpha+2\pi} f(x)\sin nx\,dx \end{aligned}\right\} \rightarrow \text{2 Marks}$$

∴ These values of a_0, a_n, b_n are known as Euler's formula.

④ Find the Laplace transform of $1 - e^t$

Sol:- $f(t) = 1 - e^t$

$$L\{f(t)\} = L\{1\} - L\{e^t\}$$

$$= \frac{1}{s} - \frac{1}{s-a} \longrightarrow \text{2 Marks}$$

⑤ Define Inverse Fourier Transform

Sol:- The inverse Fourier transform of $F(p)$ is given by

$$f(x) = \frac{1}{2\pi}\int_{-\infty}^{\infty} F(p)\,e^{-ipx}\,dp$$

$\longrightarrow$ 2 Marks

Here $p = f(x)$

Hence the formula.

⑥ ⑥ If $f(z) = u+iv$ then $u-v = \dfrac{\cos x \overset{(3)}{} + \sin x - e^{-y}}{2(\cos x - \cosh y)}$

and $f\left(\frac{\pi}{2}\right) = 0$

Sol:- We know that

$$f(z) = u+iv \longrightarrow ①$$

$$i\,f(z) = iu - v \longrightarrow ②$$

$① + ② \Rightarrow$

$$(1+i)f(z) = (u-v)+i(u+v) \longrightarrow ③$$

Let $(1+i)\,f(z) = F(z)$, $u-v = U$

$u+v = V$ then ③ becomes as

$$F(z) = U + iV$$

It is given that

$$U = u-v = \frac{\cos x + \sin x - e^{-y}}{2\cos x - e^{y} - e^{-y}} \longrightarrow$$ 1 Mark

$$U = \frac{\cos x + \sin x - e^{-y}}{2(\cos x - \cosh y)}$$

$$\therefore \frac{\delta U}{\delta x} = \frac{(\cos x - \cosh y)(-\sin x + \cos x) - (\cos x + \sin x - e^{-y})(-\sin x)}{2(\cos x - \cosh y)^2}$$

and

$$\frac{\delta U}{\delta y} = \frac{(\cos x - \cosh y)e^{-y} - (\cos x + \sin x - e^{-y})(-\sinh y)}{2(\cos x - \cosh y)^2}$$

Now $F'(z) = \dfrac{\delta U}{\delta x} + i\dfrac{\delta V}{\delta x} = \dfrac{\delta U}{\delta x} - i\dfrac{\delta U}{\delta y}$

$$= \frac{\left[(\cos x - \cosh y)(-\sin x + \cos x) + \sin x(\cos x + \sin x - e^{-y}) - i\left[(\cos x - \cosh y)e^{-y} + (\cos x + \sin x - e^{-y})(\sinh y)\right]\right.}{2(\cos x - \cosh y)^2}$$ ④

By Milne – Thomson's method we express $F'(z)$ is in terms of z by putting $x = z$ and $y = 0$

$$\therefore F'(z) = \frac{(\cos z - 1)(-\sin z - \cos z) + \sin z(\cos z + \sin z - 1) - i(\cos z - 1)]}{2(\cos z - 1)^2}$$ ↓ 2 Mark

$$= \frac{\cos z(\cos z - 1) + \sin^2 z - i(\cos z - 1)}{2(\cos z - 1)^2}$$

$$= \frac{(1 - \cos z) - i(\cos z - 1)}{2(\cos z - 1)^2} = \frac{-1 - i}{2(\cos z - 1)}$$

Now $(1+i) f'(z) = \dfrac{-(1+i)}{2(\cos z - 1)}$

$$\therefore f'(z) = \frac{-1}{2(\cos z - 1)} = \frac{1}{2\left(1 - 2\sin^2 \frac{z}{2} - 1\right)}$$

$$= \frac{1}{4}\operatorname{cosec}^2\left(\frac{z}{2}\right)$$ ⟶ 2 Marks

$$\therefore f(z) = \frac{1}{4}\int \operatorname{cosec}^2\left(\frac{z}{2}\right)dz + c$$

$$= -\frac{1}{2}\cot\left(\frac{z}{2}\right) + c.$$ ⟶ 1 Mark

$\therefore f\left(\frac{\pi}{2}\right) = 0 \Rightarrow 0 = -\frac{1}{2}\cot\frac{\pi}{4} + c$ ⑤

$\therefore \quad c = \frac{1}{2}$

$\therefore f(z) = \frac{1}{2} - \frac{1}{2}\cot\left(\frac{z}{2}\right)$

$f(z) = \frac{1}{2}\left[1 - \cot\frac{z}{2}\right]$

Hence the solution Total: 6 Marks.

---

6 ⓑ If $f(z)$ is an analytic function with constant modulus then show that $f(z)$ is constant

Sol:- Suppose $f(z)$ is analytic in a domain D and $|f(z)| = k$ = constant in D.

Then we want to prove that

$f(z)$ = constant in D

$\therefore |f(z)| = k \Rightarrow u^2 + v^2 = k^2$

Diff w.r.t x and y we get

$$2u\frac{\partial u}{\partial x} + 2v\frac{\partial v}{\partial x} = 0 \text{ and}$$

$$2u\frac{\partial u}{\partial y} + 2v\frac{\partial v}{\partial y} = 0$$

$$\Rightarrow u\frac{\partial u}{\partial x} + v\frac{\partial v}{\partial x} = 0 \longrightarrow ①$$

$$u\frac{\partial u}{\partial y} + v\frac{\partial v}{\partial y} = 0 \longrightarrow ②$$

⟶ 2 Marks.

Now using by Cauchy-Riemann equation we get

$$\frac{\delta v}{\delta x} = -\frac{\delta u}{\delta y} \text{ and } \frac{\delta v}{\delta y} = \frac{\delta u}{\delta x}$$ ⑥

we get $u\frac{\delta u}{\delta x} - v\frac{\delta u}{\delta y} = 0$ and → ③

$$u\frac{\delta u}{\delta y} + v\frac{\delta u}{\delta x} = 0 \longrightarrow ④$$

Again eliminating $\frac{\delta u}{\delta y}$ from ③ & ④

we get $(u^2+v^2)\frac{\delta u}{\delta x} = 0$ and

again eliminating $\frac{\delta u}{\delta x}$, we get

$(u^2+v^2)\frac{\delta u}{\delta y} = 0$. → 2 Marks.

If $k^2 = u^2+v^2 = 0$ then $u = 0$ and $v = 0$

and hence $f(z) = 0$

If $k \neq 0$ then $u_x = 0$, $u_y = 0$.

From Cauchy-Riemann equations

$V_x = V_y = 0$.

All these yield u = constant

and V = constant so that

$f(z)$ is constant → 2 Marks

Hence the solution

Total : 6 Marks.

⑦ Find the regular function whose imaginary ⑦
ⓐ part is $v = \log(x^2+y^2) + x - 2y$.

Sol:- Let $v = \log(x^2+y^2) + x - 2y \rightarrow ①$

$$\therefore \frac{\delta v}{\delta x} = \frac{2x}{x^2+y^2} + 1 \; ; \; \frac{\delta v}{\delta y} = \frac{2y}{x^2+y^2} - 2$$

Let $f(z) = u + iv$. Then

$$f'(z) = \frac{\delta u}{\delta x} + i\frac{\delta v}{\delta x} = \frac{\delta v}{\delta y} + i\frac{\delta v}{\delta x}$$

$$= \left(\frac{2y}{x^2+y^2} - 2\right) + i\left(\frac{2x}{x^2+y^2} + 1\right) \rightarrow \text{2 Marks}$$

By Milne-Thomson's method $f'(z)$ is expressed in terms of $z = x$, $y = 0$.

$$f'(z) = -2 + i\left(\frac{2z}{z^2} + 1\right)$$

$$= -2 + i\left(\frac{2}{z} + 1\right) \rightarrow \text{2 Marks}$$

Integrating we get

$$f(z) = \int\left[-2 + i\left(\frac{2}{z} + 1\right)\right] dz + c$$

$$= -2z + i(2\log z + z) + c.$$

$$= 2i\log z - (2-i)z + c \rightarrow \text{2 Marks}$$

Hence the solution. Total : 6 Marks.

7(b) Use cauchy's integral formula to calculate $\oint_C \frac{\sin \pi z + \cos \pi z}{(z-1)(z-2)}\, dz$, where ⑧

c is $|z| = 4$

Sol:- Given that $\oint_C \frac{\sin \pi z + \cos \pi z}{(z-1)(z-2)}\, dz$

Here $f(z) = \sin \pi z + \cos \pi z$ is analytic with in the circle $|z| = 4$ and the two singular points $z = 1$ and $z = 2$ lie inside the circle

2 Marks

$$\oint_C \frac{f(z)}{(z-1)(z-2)}\, dz = \oint_C \frac{\sin \pi z + \cos \pi z}{(z-2)(z-1)}\, dz$$

$$= \oint_C (\sin \pi z + \cos \pi z)\left[\frac{1}{z-2} - \frac{1}{z-1}\right] dz$$

2 Marks

$$= \oint_C \frac{\sin \pi z + \cos \pi z}{z-2}\, dz - \int \frac{\sin \pi z + \cos \pi z}{z-1}\, dz$$

$$= 2\pi i\left[\sin \pi(2) + \cos \pi(2)\right] - 2\pi i\left[\sin \pi(1) + \cos \pi(1)\right]$$

$$= 2\pi i\,[0+1] - 2\pi\,[0-1]$$

$$= 4\pi i.$$ → 2 Marks.

Hence the solution

Total: 6 Marks.

8) a) Find the Laurents' series expansion of ⑨

$f(z) = \frac{7z-2}{(z+1)z(z-2)}$ in the region

$1 < z+1 < 3$..

Sol:- Given $f(z) = \frac{7z-2}{(z+1)(z)(z-2)}$

put $z+1 = \omega$. Then $z = \omega - 1$

$$\therefore f(z) = \frac{7(\omega-1)-2}{\omega(\omega-1)(\omega-1-2)}$$

$$= \frac{7\omega-9}{\omega(\omega-1)(\omega-3)} \text{ by P.F}$$

$$\therefore f(z) = \frac{-3}{\omega} + \frac{1}{\omega-1} + \frac{2}{\omega-3}. \rightarrow 2 \text{ Marks}$$

$$= -\frac{3}{\omega} + \frac{1}{\omega(1-\frac{1}{\omega})} + \frac{-2}{3(1-\frac{\omega}{3})}$$

$$= -\frac{3}{\omega} + \frac{1}{\omega}\left[1+\frac{1}{\omega}+\frac{1}{\omega^2}+\cdots\right] - \frac{2}{3}\left[1+\frac{\omega}{3}+\frac{\omega^2}{9}+\cdots\right] \rightarrow 2 \text{ Marks}$$

$$= \left(-\frac{2}{\omega}+\frac{1}{\omega^2}+\frac{1}{\omega^3}\right) - \frac{2}{3}\left(1+\frac{\omega}{3}+\frac{\omega^2}{9}+\cdots\right)$$

$$= -\frac{2}{z+1} + \frac{1}{(z+1)^2} + \frac{1}{(z+1)^3}. -\frac{2}{3}\left[1+\left(\frac{z+1}{3}\right) + \left(\frac{z+1}{3}\right)^2+\cdots\right]$$

The above series is valid in the region $\left|\frac{1}{\omega}\right| < 1$ and $\left|\frac{\omega}{3}\right| < 1$

$\Rightarrow 1 < |z+1| < 3. \rightarrow$ 2 Marks

Total: 6 Marks.

⑧ ⓑ Find the residue of the function
$$f(z) = \frac{ze^z}{(z-1)^2} \text{ at each pole.}$$

(10)

Sol:- Let $f(z) = \frac{ze^z}{(z-1)^2}$, poles of $f(z)$ are obtained by putting the denominator equal to zero

$\therefore$ $z = 1$ is a pole of $f(z)$ of order $= 2$

$$[Res\, f(z)]_{z=a}$$
$$= \frac{1}{(m-1)!} \underset{z\to a}{Lt} \frac{d^{m-1}}{dz^{m-1}} [(z-a)^m f(z)]$$

here $a = 1$ and $m = 3$. → 2 Marks

$$\therefore [Res\, f(z)]_{z=1} = \frac{1}{2!} \underset{z\to 1}{Lt} \frac{d^2}{dz^2}(ze^z) \rightarrow 2\text{ Marks}$$

$$= \frac{1}{2} \underset{z\to 1}{Lt} [ze^z + e^z + e^z]$$

$$= \frac{3e}{2} \longrightarrow 2 \text{ Marks}$$

Hence the solution. Total: 6 Marks

---

⑨ ⓐ Evaluate $\oint_C \frac{z^2+4}{z^3+2z^2+2z} dz$, where

C, is the circle given by (i), $|z| = 1$

(ii), $|z+1-i| = 1$, (iii), $|z+i+1| = 1$

(iv) $|z-1| = 5$.

For each 1 Mark (1×4 = 4)

Problem: 2 Marks

Total: 6 Marks

Sol:-

Q) (b) Find the Laurent's series of $f(z) = \frac{e^z}{z(1-z)}$ ⑫ about $z=1$. Find the region of convergence.

Sol:- Let $f(z) = \frac{e^z}{z(-z+1)} = \frac{e^{z-1}e}{z(1-z)}$

$$= -\frac{e}{z-1} \cdot \frac{e^{z-1}}{(z-1+1)}$$

$$= -\frac{e}{z-1} e^{z-1}\left[1+(z-1)\right]^{-1} \rightarrow \text{2 Marks.}$$

$$= -\frac{e}{z-1}\left[\sum_{n=0}^{\infty}\frac{(z-1)^n}{n!}\right]\left[\sum_{n=0}^{\infty}(-1)(z-1)^n\right]$$

if $|z-1|<1$

$$= \frac{-e}{z-1}\left[1+(z-1)+\frac{(z-1)^2}{2!}+\frac{(z-1)^3}{3!}+\cdots\right]$$
$$+\left[1-(z-1)+(z-1)^2-(z-1)^3+\cdots\right]$$

$$= -\frac{e}{z-1}\left[1-(z-1)+(z-1)^2+(z-1)-(z-1)^2 + \frac{(z-1)^2}{2} - \cdots\right]$$

↳ 2 Marks.

$$= -\frac{e}{z-1}\left[1+\frac{(z-1)^2}{2}-\frac{1}{3}(z-1)^3-\frac{2}{3}(z-1)^4 \cdots\right]$$

$$= e\left[\frac{-1}{z-1}-\frac{1}{2}(z-1)+\frac{1}{3}(z-1)^2 - \frac{2}{3}(z-1)^3+\cdots\right]$$

→ 2 Marks.

Hence the solution. Total: 6 Marks.

(10) (a) Find a Fourier Series to represent $x - x^2$ from $x = -\pi$ to $x = \pi$. (13)

Sol:- $f(x) = \{ x - x^2 \text{ for } (-\pi, \pi) \}$.

The Fourier series $f(x) = x - x^2$ in $(-\pi, \pi)$ is given $f(x) = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos nx$

$+ \sum_{n=1}^{\infty} b_n \sin nx$ → ① Mark

Let $a_0 = \frac{1}{\pi} \int_{-\pi}^{\pi} f(x)\, dx$

$$= \frac{1}{\pi} \int_{-\pi}^{\pi} (x - x^2)\, dx$$

$$= \frac{1}{\pi} \left[\frac{x^2}{2} - \frac{x^3}{3} \right]_{-\pi}^{\pi}$$

$$= \frac{1}{\pi} \left[\left(\frac{\pi^2}{2} - \frac{\pi^3}{3} \right) - \left(\frac{\pi^2}{2} + \frac{\pi^3}{3} \right) \right]$$

$$= \frac{1}{\pi} \left(\cancel{\frac{\pi^2}{2}} - \frac{\pi^3}{3} - \cancel{\frac{\pi^2}{2}} - \frac{\pi^3}{3} \right]$$

$$= \frac{1}{\pi} \left(-\frac{2\pi^3}{3} \right) = -\frac{2\pi^2}{3} \longrightarrow \text{(I)}$$

↳ 2 Marks

Let $a_n = \frac{1}{\pi} \int_{-\pi}^{\pi} f(x) \cos nx\, dx$

$$= \frac{1}{\pi} \int_{-\pi}^{\pi} (x - x^2) \cos nx\, dx$$

$$= \frac{1}{\pi} \left[(x - x^2) \int \cos nx\, dx - \int \left[\frac{d}{dx} (x - x^2) \int \cos nx\, dx \right] dx \right.$$

$$= \frac{1}{\pi} \left[(x - x^2) \left(\frac{\sin nx}{n} \right) - \int \left[(1 - 2x) \left(\frac{\sin nx}{n} \right) dx \right]_{-\pi}^{\pi} \right.$$

$\therefore a_n = \frac{4}{n^2}(-1)^n$

Now $b_n = \frac{1}{\pi}\int_{-\pi}^{\pi} f(x)\sin nx\,dx$

$= \frac{1}{\pi}\left[\int_{-\pi}^{\pi}(x-x^2)\sin nx\,dx\right]$

$= \frac{1}{\pi}\left[\int_{-\pi}^{\pi} x\sin nx\,dx - \int_{-\pi}^{\pi} x^2\sin nx\,dx\right]$

$= +\frac{2}{\pi}\int_{0}^{\pi} x\sin nx\,dx$

$\Rightarrow \frac{2}{n}(-1)^n = b_n$ → 2 Marks

$$\therefore f(x) = -\frac{2\pi^2}{3} + 4\sum_{n=1}^{\infty}\frac{(-1)^n}{n^2}\cos nx + 2\sum_{n=1}^{\infty}\frac{(-1)^n}{n}\sin nx$$ → 1 Mark

Total: 6 Marks.

Hence the solution.

---

10 (b) Expand $f(x) = \sqrt{1-\cos x}$ when $0 < x < 2\pi$ in a Fourier series. Hence evaluate

$\frac{1}{1\cdot 3} + \frac{1}{3\cdot 5} + \frac{1}{5\cdot 7} + \cdots$

Sol:- Given that $f(x) = \sqrt{1-\cos x} = \sqrt{2}\sin x/2$

$$f(x) = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n\cos nx + \sum_{n=1}^{\infty} b_n\sin nx$$ → 1 Mark

$$a_0 = \frac{1}{\pi}\int_0^{2\pi}\sqrt{2}\sin x/2\,dx = \frac{4\sqrt{2}}{\pi}$$

→ 2 Marks

$$a_n = \frac{1}{\pi}\int_0^{2\pi} \sin x/_2 \cos nx\, dx$$

$$= \frac{\sqrt{2}}{2\pi}\int_0^{2\pi} 2\cos nx\, dx \sin x/_2\, dx$$

$$= \frac{1}{\sqrt{2\pi}}\int_0^{2\pi} \sqrt{2}\sin\frac{x}{2}\cos nx\, dx$$

$$= \frac{\sqrt{2}}{\pi}\left[\frac{2}{2n+1} - \frac{2}{2n-1}\right] = \frac{4\sqrt{2}}{\pi(n^2-1)}$$

→ 2 Marks

$$b_n = \frac{1}{\pi}\int_0^{2\pi} \sqrt{2}\sin x/_2 \sin nx\, dx$$

$$= \frac{\sqrt{2}}{2\pi}\int_0^{2\pi} 2\sin nx \sin x/_2\, dx$$

$$= \frac{\sqrt{2}}{\pi}\left[\frac{1}{(2n-1)}\{\sin(2n-1)\pi - 0\} - \frac{2}{2n+1}\sin\left(\frac{2n+1}{2}\right)x\right]_0^{2\pi}$$

put a_0, a_n, b_n values the formula

$$\therefore \sqrt{(1-\cos x)} = \frac{2\sqrt{2}}{\pi} - \sum_{n=1}^{\infty}\frac{4\sqrt{2}}{(4n^2-1)\pi}\cos nx$$

when $x=0$, we have

$$0 = \frac{2\sqrt{2}}{\pi} - \frac{4\sqrt{2}}{\pi}\sum_{n=1}^{\infty}\frac{1}{4n^2-1}$$

→ 1 Mark

$$\therefore \frac{1}{1\cdot 3} + \frac{1}{3\cdot 5} + \frac{1}{5\cdot 7} + \cdots\infty = \frac{1}{2}$$

Total 6 Marks

Hence the solution.

11 ⓐ Find the Fourier series expansion of ⑮
$f(x) = 2x - x^2$ in $(0,3)$ and hence deduce that $\frac{1}{1^2} + \frac{1}{2^2} + \cdots + = \frac{\pi}{2}$

Sol:- Here $2l = 3 \therefore l = \frac{3}{2}$

∴ Formula for this problem is

$$f(x) = 2x - x^2 = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos\left(\frac{n\pi x}{l}\right) + \sum_{n=1}^{\infty} b_n \sin\left(\frac{n\pi x}{l}\right) \rightarrow 1 \text{ Mark}$$

$$= \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos\left(\frac{2n\pi x}{3}\right) + \sum_{n=1}^{\infty} b_n \sin\left(\frac{2n\pi x}{3}\right)$$

$$\therefore a_0 = \frac{1}{l}\int_0^{2l} f(x)\,dx = \frac{2}{3}\int_0^{3}(2x - x^2)\,dx$$

$$= 0 \longrightarrow 1 \text{ Mark}$$

Then to find a_n

$$\therefore a_n = \frac{1}{l}\int_0^{2l} f(x)\cos\left(\frac{n\pi x}{l}\right)dx$$

$$= \frac{2}{3}\int_0^{3}(2x - x^2)\cos\left(\frac{2n\pi x}{3}\right)dx$$

after simplification we get

$$= \frac{-9}{n^2\pi^2} \longrightarrow 1 \text{ Mark}$$

Now to find b_n

$$\therefore b_n = \frac{1}{l}\int_0^{2l} f(x)\sin\left(\frac{n\pi x}{l}\right)dx$$

$$b_n = \frac{1}{l}\int_0^{2l} f(x)\sin\left(\frac{n\pi x}{l}\right)dx$$

$$= \frac{2}{3}\int_0^{3}(2x-x^2)\sin\left(\frac{2n\pi x}{3}\right)dx$$

$$= \frac{2}{3}\left[-\frac{3}{2n\pi}(-3)\right] = \frac{3}{n\pi} \rightarrow \text{1 Mark}$$

put a_0, a_n, b_n values in formula

$$2x - x^2 = \frac{-9}{\pi^2}\sum_{n=1}^{\infty}\left(\cos\frac{2n\pi x}{3}\right) + \frac{3}{\pi}\sum_{n=1}^{\infty}\frac{1}{n}\sin\left(\frac{2n\pi}{3}\right. \rightarrow \text{1 Mark}$$

After simplification we get the result.

and we get → 1 Mark Total 6 Marks.

$$= \frac{1}{1^2} + \frac{1}{2^2} + \cdots = \frac{\pi}{2}.$$

---

11 (b) Obtain a half-range cosine series for consider $f(x) = \pi - x$ in $(0, \pi)$

Hence show that $\sum_{n=0}^{\infty}\left(\frac{1}{(2n+1)^2}\right) = \frac{\pi^2}{8}$

$(x = \pi)$.

Sol:- Given that $f(x) = \pi - x$ in $(0, \pi)$

$$f(x) = \pi - x = \frac{a_0}{2} + \sum_{n=1}^{\infty} a_n \cos nx \rightarrow \text{1 Mark}$$

$$\therefore a_0 = \frac{2}{\pi}\int_0^{\pi} f(x)\,dx$$

$$= \frac{2}{\pi}\int_0^{\pi}(\pi - x)\,dx = \pi \longrightarrow \text{1 Mark}$$

Now to find a_n

$$\therefore a_0 = \frac{2}{\pi}\int_0^{\pi} f(x)\,dx = \frac{2}{\pi}\int_0^{\pi} (\pi - x)\,dx = \pi$$

$$a_m = \frac{2}{\pi}\int_0^{\pi} f(x)\cos mx\,dx$$

$$= \frac{2}{\pi}\int_0^{\pi} (\pi - x)\cos mx\,dx$$

$$= \frac{2}{\pi}\left[\frac{1-\cos m\pi}{m^2}\right] = \frac{2}{\pi m^2}\left[1-(-1)^m\right]$$

$$\therefore a_m = \begin{cases} 0 & \text{for } m \text{ is even} \\ \dfrac{4}{\pi m^2} & \text{for } m \text{ is odd} \end{cases} \rightarrow \text{1 Mark}$$

put a_0, a_m values in formula we get

$$f(x) = \frac{a_0}{2} + \sum_{m=1}^{\infty} a_m \cos mx \rightarrow \text{1 Mark}$$

$$\pi - x = \frac{\pi}{2} + \frac{4}{\pi}\sum_{m=1,3,5}^{\infty} \frac{1}{m^2}\cos mx$$

$$= \frac{\pi}{2} + \frac{4}{\pi}\left[\frac{1}{1^2}\cos x + \frac{1}{3^2}\cos 3x + \cdots\right]$$

after simplification we get

$$\sum_{l=0}^{\infty} \frac{1}{(2l+1)^2} = \frac{\pi^2}{8} \rightarrow \text{2 Marks}$$

Total: 6 Marks

Hence the solution

(12) (a) Find the Laplace transform of ⑲

$$f(t) = \frac{\cos at - \sin at}{t} + t \sin at$$

Sol:- Given $f(t) = \frac{\cos at - \sin at}{t} + t \sin at$

$$f(t) = \frac{\cos at}{t} - \frac{\sin at}{t} + t \sin t$$

$$= L\{\cos at\} = \frac{s}{s^2+a^2} ; L\{\sin at\} = \frac{a}{s^2+a^2}$$ ↳ 2 Marks

$$L\left\{\frac{1}{t}\cos at\right\} = a\int_s^\infty \frac{s}{s^2+a^2}\,ds$$

and $L\left\{\frac{1}{t}\sin at\right\} = a\int_s^\infty \frac{1}{s^2+a^2}\,ds$ → 2 Marks

2 Marks

$$= \tan^{-1}\left(\frac{s}{a}\right) - \cot^{-1}\left(\frac{s}{a}\right) + \frac{s^2-a^2}{(s^2+a^2)^2}$$

Hence the solution. Total: 6 Marks.

(12) (b) Apply convolution theorem to solve

$$\bar{f}(s) = \frac{s^2}{(s^2+a^2)(s^2+b^2)}$$ → 1 Mark

Ans:- Given that $\frac{s^2}{(s^2+a^2)(s^2+b^2)}$ 2 Marks

$$\therefore L^{-1}\left\{\frac{s}{s^2+a^2} \cdot \frac{s}{s^2+b^2}\right\} = \int_0^t \frac{\cos au \sin a(t-u)}{a}\,du$$

$$= \frac{1}{2}\left[\int_0^t [\cos((a-b)u + bt] + \cos[(a+b)u - bt]\right]du$$

$$= \frac{a \sin at - b \sin bt}{a^2-b^2}$$ → 3 Marks

Hence the solution. 6 Marks

⑬ (a): Find the inverse Laplace Transform of $\frac{s}{s^4+4a^4}$ (H)

Sol:- Given that

$$\bar{f}(s) = \frac{s}{s^4+4a^4} \longrightarrow ①$$

Let us take s^4+4a^4 can be written as

2 Marks

$$s^4+4a^4 = \frac{As+B}{s^2+2as+2a^2} + \frac{Cs+D}{s^2-2as+2a^2}$$

$$\left[\because s^4+4a^4 = (s^2+a^2)^2-(2as)^2\right] \longrightarrow ②$$

$$= (s^2+a^2+2as)(s^2+a^2-2as)$$

$$= (s^2+2as+a^2)(s^2-2as+a^2)$$

This step using in ① we get

$$\frac{s}{s^4+4a^4} = \frac{As+B}{s^2+2as+2a^2} + \frac{Cs+D}{s^2-2as+2a^2}$$ 2.0

After solving we get the values of A, B, C, D.

$\therefore A = C = 0$ and

$B = \frac{-1}{4a}$, $D = \frac{1}{4a}$ → 2 Marks

$$L^{-1}\left\{\frac{s}{s^2+4a^4}\right\} = \frac{-1}{4a} L^{-1}\left\{\frac{1}{s^2+2as+2a^2}\right\} + \frac{1}{4a} L^{-1}\left\{\frac{1}{s^2-2as+2a^2}\right\}$$

$$= \frac{-1}{4a} L^{-1}\left\{\frac{1}{(s+a)^2+a^2}\right\} + \frac{1}{4a} L^{-1}\left\{\frac{1}{(s-a)^2+a^2}\right\}$$

$$= \frac{1}{2a^2} \sin at \left(\frac{e^{at}-e^{-at}}{2}\right)$$ → 2 Marks.

$$= \frac{1}{2a^2} \sin at \sinh at$$

Hence the solution.

Total: 6 Marks

13 (b) Using the Laplace transform method to solve $(D^2-3D+2)y = 4e^{2t}$; $y(0)=3$, $y'(0)=5$

Sol:- Given $(D^2-3D+2)y = 4e^{2t}$ (2L)

$$\frac{d^2y}{dt^2} - 3\frac{dy}{dt} + 2y = 4e^{2t}$$

Taking L.T on both sides we get

$$[\bar{y}s^2 - y(0)s - y'(0)] - 3[\bar{y}(s) - y(0)] + 2\bar{y} = \frac{4}{s-2}$$ → 2 Marks

$$[\bar{y}(s^2) - 3s - 5] - 3[\bar{y}(s) - 3] + 2\bar{y} = \frac{4}{s-2}$$

$$\bar{y}[s^2 - 3s + 2] = \frac{4}{s-2} + 3s + 5 + 9$$

$$\therefore \bar{y} = \frac{4}{(s-2)(s^2-3s+2)} + \frac{3s}{s^2-3s+2} + \frac{14}{s^2-3s+2}$$

$$= \frac{4}{(s-2)(s-1)(s-2)} + \frac{3s}{(s-2)(s-1)} + \frac{14}{(s-1)(s-2)}$$

$$= \frac{4}{(s-1)(s-2)^2} + \frac{3s}{(s-1)(s-2)} + \frac{14}{(s-1)(s-2)}$$ → 2 Marks

$$= \frac{4}{(s-1)(s-2)^2} + \frac{3s+14}{(s-1)(s-2)}$$

= After solving we get the solution. → 2 Marks

Total: 6 Marks.

14 ⓐ Find the Fourier cosine transform of x

㉓

Sol:- Let $f(x) = x$

Formula for this problem is

$f(x) = x$ and

$$F_c\{f(x)\} = \int_0^\infty f(x) \cos px \, dx$$

$$= \int_0^\infty x \cos px \, dx \longrightarrow \text{2 Marks}$$

$$= x\int \cos px \, dx - \int \left[\frac{d}{dx}(x) \int \cos px \, dx\right] dx$$

$$= x \frac{\sin px}{p} - \int \frac{\sin px}{p} \, dx$$

$$= \left[x \frac{\sin px}{p} + \frac{\cos px}{p^2}\right]_0^\infty \longrightarrow \text{2 Marks}$$

$$= \left[\infty \frac{\sin \infty}{p} + \frac{\cos p\infty}{p^2}\right] - \left[0 + \frac{1}{p^2}\right]$$

$$= (\infty + \infty) - \frac{1}{p^2} = \text{Ans} \longrightarrow \text{2 Marks}$$

Hence the solution. Total: 6 Marks.

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14 ⓑ Find the Fourier Transform of

$$f(x) = \begin{cases} a^2 - x^2 ; & |x| \le a \\ 0 : & |x| > a \end{cases}.$$ Hence

deduce that $\int_0^\infty \frac{\sin t - t \cos t}{t^3} \, dt = \frac{\pi}{4}$

14 (b) Ans : Fourier transform of a function f(x) is given by $F\{f(x)\} = \int_{-\infty}^{\infty} f(x) e^{ipx} dx$ ④

$$= \int_{-\infty}^{-a} f(x) e^{ipx} dx + \int_{-a}^{a} f(x) e^{ipx} dx + \int_{a}^{\infty} f(x) e^{ipx} dx.$$

$$= 0 + \int_{-a}^{a} f(x) e^{ipx} dx + 0$$

$$= \int_{-a}^{a} (a^2 - x^2) e^{ipx} dx \longrightarrow \text{2 Marks}$$

After solving we get

$$= F\{f(x)\} = \frac{4}{p^3}(\sin ap - ap \cos ap) = F(p)$$

Deduction :-

$$\therefore f(x) = \frac{4}{\pi} \int_0^{\infty} \frac{\sin ap - ap \cos ap}{p^3} \cos px \, dp$$

↳ 2 Marks

put a = 1 and x = 0 we get

$$1 = \frac{4}{\pi} \int_0^{\infty} \frac{\sin p - p \cos p}{p^3} dp$$

$$\Rightarrow \int_0^{\infty} \frac{\sin x - x \cos x}{x^3} dx = \frac{\pi}{4}$$

put $\boxed{p = x}$ → 2 Marks

Hence the solution. Total: 6 Marks.

(15) (a) Find the cosine Fourier transform of $e^{-\alpha x}$, (2S) and hence evaluate $\int_0^\infty \frac{\cos \alpha x}{x^2+\alpha^2}\,dx$

Sol:- Given that $f(x) = e^{-\alpha x}$ then

$$F_c\{f(x)\} = \int_0^\infty f(x)\cos px\,dx \longrightarrow$$ 2 Marks

$$= \int_0^\infty e^{-\alpha x}\cos px\,dx$$

$$= \left[\frac{e^{-\alpha x}}{\alpha^2+p^2}(-\alpha\cos px + p\sin px)\right]_0^\infty$$

$$F_c\{e^{-\alpha x}\} = \frac{\alpha}{\alpha^2+p^2}$$ and. → 2 Marks

Now by the Inverse Fourier cosine transform, we get

$$f(x) = \frac{2}{\pi}\int_0^\infty F_c\{f(x)\}\cos px\,dp$$

$$= \frac{2}{\pi}\int_0^\infty \frac{\alpha}{\alpha^2+p^2}\cos px\,dp.$$

$$\therefore \int_0^\infty \frac{\cos \alpha x}{\alpha^2+p^2} = \frac{\pi}{2\alpha}e^{-\alpha x}$$

↳ 2 Marks

Hence the solution. Total: 6 Marks.

(5) (b) Show that the inverse finite Fourier sine transform of (26)

$F_m\{x\} = \frac{1}{\pi}\left[1+\cos m\pi - 2\cos\frac{m\pi}{2}\right]$ is

$$f(x) = \begin{cases} 1 & \text{for } 0 < x < \frac{\pi}{2} \\ -1 & \text{for } \frac{\pi}{2} < x < \pi \end{cases}$$

Sol:- From the Inverse Finite sine transform f(x) if we have

$$f(x) = \frac{2}{l}\sum_{m=1}^{\infty} F_s(m)\sin\left(\frac{m\pi x}{l}\right) \rightarrow 2 \text{ Marks}$$

$$= \frac{2}{\pi}\sum_{m=1}^{\infty}\left(\frac{1}{\pi}\left[1+\cos m\pi - 2\cos\frac{m\pi}{2}\right)\sin\left(\frac{m\pi x}{l}\right)$$

As per the $f(x) = \begin{cases} 1 & \text{for } 0 < x < \frac{\pi}{2} \\ -1 & \text{for } \frac{\pi}{2} < x < \pi \end{cases}$ → 2 Marks

After solving we get the solution. → 2 Marks

Total : 6 Marks.

Hence the solution.

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Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

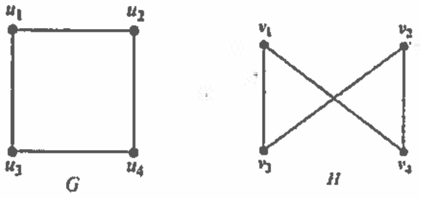
Degree	B. Tech.	Program	CSE, CSE(AI & ML), CSE(DS)	Academic Year	2022 - 2023
Course Code	20BSX16	Test Duration	3 Hrs. Max. Marks 70	Semester	III
Course	Mathematical Foundations of Computer Science				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Conjunction. Explain with truth table and suitable example	20BSX16.1	L1
2	Show that the relation divides (/) is a partial ordering on the set of integers	20BSX16.2	L2
3	What are the quotient and remainder when −22 is divided by 3?	20BSX16.3	L1
4	Find the first three terms in the sequence defined by the recurrence relation $a_n = 3\ a_{n-1}$ with initial condition $a_0 = 2$	20BSX16.4	L2
5	How many edges are there in a graph with 10 vertices each of degree six?	20BSX16.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Define Tautology, Contradiction, contingency with examples	6M	20BSX16.1	L2
6 (b)	Construct the truth table of the compound proposition $(p \vee \sim q) \rightarrow (p \wedge q)$	6M	20BSX16.1	L3
	OR			
7 (a)	(i) Obtain the Disjunctive Normal form of $\sim P \rightarrow (Q \wedge R)$ (ii) Obtain the Conjunctive Normal form of $P \wedge (P \rightarrow Q)$	6M	20BSX16.1	L2
7 (b)	If there was a ball game, then traveling was difficult. If they arrived on time, then traveling was not difficult. They arrived on time. Therefore, there was no ball game'. Show that these statements constitute a valid argument	6M	20BSX16.1	L3
8 (a)	Write the matrix representation and directed graph of the relation on the set A= {1,2,3,4} where R= {(1,1), (1,2), (2,1), (2,2), (2,4), (3,3), (3,1), (4,3), (4,1), (3,2)}	6M	20BSX16.2	L3
8 (b)	Construct the Hasse diagram for the partial ordering {(A, B) | A ⊆ B} on the power set P (S) where S = {a, b, c}	6M	20BSX16.2	L2
	OR			
9 (a)	Let S = {1, 2, 3, 4} and let f = $\begin{pmatrix} 1 & 2 & 3 & 4 \\ 2 & 1 & 4 & 3 \end{pmatrix}$ and g = $\begin{pmatrix} 1 & 2 & 3 & 4 \\ 4 & 1 & 2 & 3 \end{pmatrix}$. Find f °g, g ° f, f^{-1} and g^{-1} in the permutation form.	6M	20BSX16.2	L3
9 (b)	Define group and prove that Fourth root unity G = {1, -1, i, -i} is an abelian group.	6M	20BSX16.2	L2
10 (a)	State and prove Euler's theorem.	6M	20BSX16.3	L2
10 (b)	Find the prime factorization of 243, 125 and 289.	6M	20BSX16.3	L3
	OR			
11 (a)	State and prove Fermat's theorem.	6M	20BSX16.3	L2
11 (b)	Find the gcd of 1001 and 1331 , and find the integers x and y such that gcd (1001, 1331)= 1001x+1331y.	6M	20BSX16.3	L3

12 (a)	Find the solution of the recurrence relation $a_{n+2} - 4a_{n+1} + 4a_n = 2n$?	6M	20BSX16.4	L3
12 (b)	Find the solution to the recurrence relation $a_n = 6a_{n-1} - 8a_{n-2}$ for $n \geq 2$, $a_0 = 4$, $a_1 = 10$	6M	20BSX16.4	L3
	OR			
13 (a)	Solve the recurrence relation $a_n = 6a_{n-1} - 9a_{n-2}$, with $a_0 = 1$, $a_1 = 6$	6M	20BSX16.4	L2
13 (b)	Find the explicit formula for the Fibonacci numbers with $F_0 = 0$, $F_1 = 1$	6M	20BSX16.4	L1
14 (a)	Define Eulerian and Hamiltonian Graphs with suitable examples.	6M	20BSX16.5	L1
14 (b)	Show that the following two graphs G and H are isomorphic 	6M	20BSX16.5	L2
	OR			
15 (a)	Explain Prim's algorithm to find minimal spanning tree of the graph with suitable example	6M	20BSX16.5	L2
15 (b)	Explain Kruskal's algorithm to find minimal spanning tree of the graph with suitable example	6M	20BSX16.5	L2

① 14-11

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N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

III Sem

Mathematical foundations of Computer Science. 20 BSX 16.

1. Conjunction of two Statements is defined as P and Q it is denoted by $P \wedge Q$ — 1M

Truth table.

P	Q	$P \wedge Q$
T	T	T
T	F	F
F	T	F
F	F	F

Eg: P: Today is Friday

Q: It is rainy

Today is friday and it is rainy. — 1M

2M

2. Divides: $a|b \Rightarrow b = ka$, where $k \in Z^+$

$a|a \Rightarrow R$ is Reflexive — 1M

$a|b$ and $b|a \Rightarrow a = b$ ∴ R is Antisymmetric

$a|b$ and $b|c \Rightarrow a|c$ ∴ R is Transitive — 1M

∴ Divides is Reflexive, antisymmetric, Transitive ∴ '|' is poset

2M

3. $\frac{-22}{3}$ $3\,)\,-22\,(\,-7$; -21 ; -1

$-22 = -7 \times 3 + (-1)$

Quot -21 -1

Quotient: -7, remainder: -1 — 2M

4. $a_n = 3a_{n-1}$, $a_0 = 2$

$a_1 = 3a_0$ $a_1 = 3 \cdot 2 = 6$

$a_2 = 3a_1$ $a_2 = 3(6) = 18$

$a_3 = 3a_2$ $a_3 = 3(18) = 54$

2, 6, 18, 54. — 2M

5. number of vertices : 10

degree of each vertex : 6 — 1M

Total degree $10(6) = 60$.

By Hand Shaking thm Total degree $= 2|e|$

$60 = 2 \cdot e$ — 1M

$e = 30$

∴ The total number of edges: 30

2M

(4)

6(a) Tautology: A Compound proposition that is true for all possible truth values of the propositions that occur in it, in otherwords contain only 'T' in the last column of truth table is called Tautology. $p \vee \sim p$ is Tautology.

p	~p	p ∨ ~p
T	F	T
F	T	T

— 2M

Contradiction: A compound proposition that is always False for all possible truth values of the propositions that occur in it, in otherwords Contain only 'F' in last column of truth table is called a Contradiction.

$p \wedge \sim p$ is a Contradiction.

p	~p	p ∧ ~p
T	F	F
F	T	F

— 2M

Contingency: A Compound proposition is neither tautology nor Contradiction is called Contingency. $P \oplus Q$ is Contingency

P	Q	P ⊕ Q
T	T	F
T	F	T
F	T	F
F	F	F

— 2M

6M

6(b). Truth table of $(P \vee \sim Q) \wedge (P \wedge Q)$

P	Q	~Q	P ∨ ~Q	P ∧ Q	(P ∨ ~Q) → (P ∧ Q)
T	T	F	T	T	T
T	F	T	T	F	F
F	T	F	F	F	T
F	F	T	T	F	F

— 6M.

7(a) Obtain the DNF of $\sim P \rightarrow (Q \wedge R)$

(i) $\sim P \rightarrow (Q \wedge R) \cong \sim(\sim P) \vee (Q \wedge R)$

$\cong P \vee (Q \wedge R)$

Sum of elementary product — 3M

(ii) CNF $P \wedge (P \rightarrow Q) \cong$ ~~P ∨~~ $P \wedge (\sim P \vee Q)$

Product of elementary sum. — 3M

6M

7(b) · P: There was a ball game

Q: Travelling was difficult.

R: Arrived on time

The given statements are: $P \to Q$

$R \to \sim Q$

R

$\therefore \sim P$. ——— 2 M

Premise	step	statement formula	Rule	Justification.
[1]	(1)	$P \to Q$	P	Hyp.
[2]	(2)	$R \to \sim Q$	P	Hyp.
[2]	(3)	$\sim(\sim Q) \to \sim R$	Rule T	C.P. Step 2
[2]	(4)	$Q \to \sim R$	"	"
[1,2]	(5)	$P \to \sim R$	Rule T	H.S. (Transitive)
[6]	(6)	$\sim(\sim R) \to \sim P$ R	Rule P	step (1) & step (4) $P \to Q$ $Q \to \sim R$ $\therefore P \to \sim R$
[1,2,6]	(7)	$\therefore \sim P$.		

— 4 M

6 M.

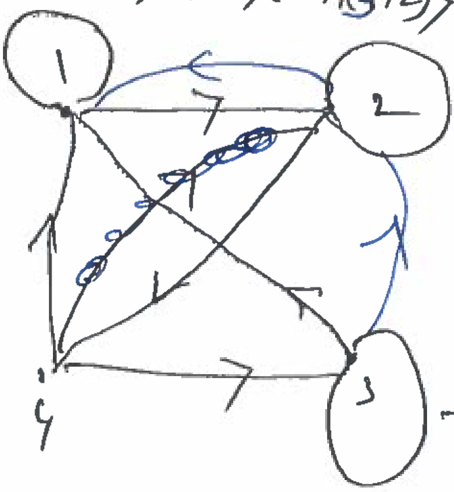
---

8)
(a) $A = \{1,2,3,4\}$

$R = \{(1,1), (1,2), (2,1), (2,2), (2,4, (3,3), (3,1), (4,3), (4,1), (3,2)\}$

$$M_{A_{ij}} = \begin{array}{c|cccc} & 1 & 2 & 3 & 4 \\ \hline 1 & 1 & 1 & 0 & 0 \\ 2 & 1 & 1 & 0 & 1 \\ 3 & 1 & 0 & 1 & 0 \\ 4 & 0 & 0 & 1 & 0 \end{array}$$

— 3 M

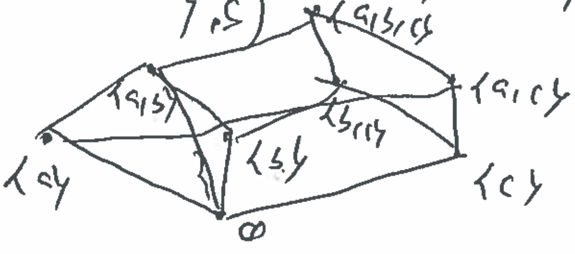


— 3 M
6 M

8) b) $S = \{a,b,c\}$ — 1 M

$P(S) = \{\phi, \{a\}, \{b\}, \{c\}, \{a,b\}, \{a,c\}, \{b,c\}, \{a,b,c\}\}$ — 1 M

$\{P(S), \subseteq\} = (\{$ $\}, \subseteq)$



— 4 M

④

9.(i) $S = \{1, 2, 3, 4\}$

$f = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 2 & 1 & 4 & 3 \end{pmatrix}$, $g = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 4 & 1 & 2 & 3 \end{pmatrix}$ — 2M

$f \circ g = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 3 & 2 & 1 & 4 \end{pmatrix}$, — 1M

$g \circ f = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 1 & 4 & 3 & 2 \end{pmatrix}$ — 1M

$f^{-1} = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 2 & 1 & 4 & 3 \end{pmatrix}$ — 1M

$g^{-1} = \begin{pmatrix} 1 & 2 & 3 & 4 \\ 2 & 3 & 4 & 1 \end{pmatrix}$ — 1M

6M

9(b) $G = \{1, -1, i, -i\}$

·	1	-1	i	-i
1	1	-1	i	-i
-1	-1	1	-i	i
i	i	-i	-1	1
-i	-i	i	1	-1

← 4M.

1st row = 1st column
2nd row = 2nd column
3rd " = 3rd "
4th " 4th "

∴ it is Closure
∴ it is Associat.
1 is identity

$(-1)^{-1} = -1$, $(1)^{-1} = 1$, $(i)^{-1} = -i$, $(-i)^{-1} = i$. ∴ it is Inverse

∴ it is Commutative. ← 6M

1(c) State and Prove Eulers theorem

Euler's Theorem: ~~Let~~ a and $m > 0$ are integers such that $\gcd(a, m) = 1$, then $a^{\varphi(m)} \equiv 1 \pmod{m}$ — 2M

Proof Let $\gcd(a, m) = 1$

Let $r_1, r_2, r_3 \ldots, r_{\varphi(m)}$ be Reduced residue (Complete) System modulo m.

Then $ar_1, ar_2, \ldots, ar_{\varphi(m)}$ are also a Complete Reduced residue system modulo m.

Hence corresponding each r_i there is ~~and~~ one and only one ar_j such that $r_j \equiv ar_j \pmod{m}$

i.e., the numbers $ar_1, ar_2, ar_3, \ldots, ar_{\varphi(m)}$ are just residue modulo m of $r_1, r_2, \ldots r_{\varphi(m)}$ not necessarily in same order

multiply the above $ar_1, ar_2, \ldots ar_{\varphi(m)} \equiv r_1 \cdot r_2 \ldots r_{\varphi(m)} \bmod m$ — 2M

$$\prod_{j=1}^{\varphi(m)} ar_j \equiv \prod_{j=1}^{\varphi(m)} r_j \pmod{m}$$

$$a^{\varphi(m)} \prod_{j=1}^{\varphi(m)} r_j \equiv \prod_{j=1}^{\varphi(m)} r_j \pmod{m}$$ ⇒ Since $(r_j, m) = 1$

⇒ $a^{\varphi(m)} \equiv 1 \pmod{m}$. — 2M

6M

10(b) Prime factorization

$3 \underline{|243}$, $3 \underline{|81}$, $3 \underline{|27}$, $3 \underline{|9}$, 3 — 3^5

$5 \underline{|125}$, $5 \underline{|25}$, 5 — 5^3

$17 \underline{|289}$, 17 — 17^2

— 6M

⑤

11(a) **Fermat's Theorem:** If p is a prime number and a is an integer such that p does not divides a (gcd)(a, p) = 1 then $a^{p-1} \equiv 1 \pmod{m}$ — 2M

$a^p \equiv a \pmod{m}$

Proof: Let us consider the first (p−1) positive multiples of a.

i.e., 1.a, 2.a, 3.a, .., (p−1)a

Since gcd (a, p) = 1, p does not divides a, thus remainders are 1, 2, 3, .. p−1 not ~~not~~ necessarily in the same order

let $a \equiv r_1 \pmod{p}$

$2a \equiv r_2 \pmod{p}$

$3a \equiv r_3 \pmod{p}$ — 2M

⋮

$(p-1)a \equiv r_{p-1} \pmod{p}$

But $r_1, r_2, \ldots r_{p-1}$ are the remainders when a, 2a, 3a, .., (p−1)a are divided by p.

Consider $a.2a.3a \ldots (p-1)a \equiv r_1.r_2.r_3 \ldots r_{p-1} \pmod{p}$

$a^{p-1}.1.2.3 \ldots (p-1) \equiv 1.2.3 \ldots (p-1) \bmod p)$

$a^{p-1}.(p-1)! \equiv (p-1)! \pmod{p}$ — 2M

$a^{p-1} \equiv 1 \pmod{p}$.

$a^p \equiv a \pmod{p}$

6M

11(b) gcd of (1001, 1331) = 11.

By Euclid algorithm

$11 = 1001 - 330(3)$

$= 1001 - (1331 - 1.1001)3$

$= -3.1331 + 4.1001$ — 2M

$11 = 1001.x + 1331y$

$1001.4 + 1331(-3)$

$x = 4, y = -3.$ — 2M

6M

$1001 \overline{)1331}(1$
1001
$330 \overline{)1001}(3$
990
$11 \overline{)330}(30$
330
0 — 2M

$1331 \times 3 = 3993$

$4004 - 3993 = 11$

12) a) Find the solution of $a_{n+2} - 4a_{n+1} + 4a_n = 2n$.

The given Recurrence relation is $a_{n+2} - 4a_{n+1} + 4a_n = 2n$

which is non homogeneous R.R of degree 2.

Characteristic equation is $r^2 - 4r + 4 = 0$

$(r-2)^2 = 0$

roots are $r_1 = r_2 = 2$ real and equal.

C.F $a_n^{(H)} = (\alpha_1 + n\alpha_2)(2)^n$. — 3M

Since $f(n) = 2n$, Assume $a_n^{(P)} = An + B$.

$a_{n+1}^{(P)} = A(n+1) + B$

$a_{n+2}^{(P)} = A(n+2) + B$

$An + 2A + B - 4An - 4A - 4B + 4An + 4B = 2n$

$An = 2n \Rightarrow A = 2$, $-2A + B = 0$

$-4 + B = 0$ $B = 4$ — 2M

$a_n^{(P)} = 2n + 4$

Solution is $a_n = a_n^{(H)} + a_n^{(P)} = (\alpha_1 + n\alpha_2)(2)^n + 2n + 4$ — 1M

6M

12 ⓑ Given R.R $a_n = 6a_{n-1} - 8a_{n-2}$

H.R.R of degree 2

$a_n - 6a_{n-1} + 8a_{n-2} = 0$

Che eqn is $r^2 - 6r + 8 = 0$

$(r-4)(r-2) = 0$

$r_1 = 2$, $r_2 = 4$

$a_n^{(H)} = \alpha_1 2^n + \alpha_2 4^n$

$a_n = \alpha_1 2^n + \alpha_2 4^n$ given that $a_0 = 4$, $a_1 = 10$

$a_0 = \alpha_1(2)^0 + \alpha_2(4)^0$

$\alpha_1 + \alpha_2 = 4$

$a_1 = \alpha_1(2)^1 + \alpha_2(4)^1$

$10 = 2\alpha_1 + 4\alpha_2$

$2\alpha_1 + 4\alpha_2 = 10$

$2\alpha_1 + 2\alpha_2 = 8$

$2\alpha_2 = 2$

13 (c). Solve $a_n = 6a_{n-1} - 9a_{n-2}$, $a_0 = 1$, $a_1 = 6$ ⑦

~~H.R.R of degree 2~~

$a_n - 6a_{n-1} + 9a_{n-2} = 0.$

Che. eqn $r^2 - 6r + 9 = 0$

$(r-3)^2 = 0$

$r_1 = r_2 = r = 3$, real equal. —— 3M

$a_n = (\alpha_1 + n\alpha_2)(3)^n$ ——

$a_0 = (\alpha_1 + 0\cdot\alpha_2)(3^0)$

$1 = \alpha_1$ — 1M

$a_1 = (\alpha_1 + 1\alpha_2)3^1$

$6 = (1 + \alpha_2)3$

$6 = 3 + 3\alpha_2$

$3\alpha_2 = 3$

$\alpha_2 = 1$ — 1M

Solution is $a_n = (1 + n)(3)^n$ — 1M

$\underline{6M}$

13 (b) fibonacci recurrence relation $f_n = f_{n-1} + f_{n-2}$

$f_n - f_{n-1} - f_{n-2} = 0$

H.R.R of degree 2

Che. $r^2 - r - 1 = 0$

$r = \frac{1 \pm \sqrt{1-4(1)(-1)}}{2} = \frac{1 \pm \sqrt{5}}{2}$ $r_1 = \frac{1+\sqrt{5}}{2}$, $r_2 = \frac{1-\sqrt{5}}{2}$ —— 2M

$f_n^{(H)} = \alpha_1\left(\frac{1+\sqrt{5}}{2}\right)^n + \alpha_2\left(\frac{1-\sqrt{5}}{2}\right)^n$ —— 1M

Since Homoge. $f_n = \alpha_1\left(\frac{1+\sqrt{5}}{2}\right)^n + \alpha_2\left(\frac{1-\sqrt{5}}{2}\right)^n$

$f_0 = \alpha_1\left(\frac{1+\sqrt{5}}{2}\right)^0 + \alpha_2\left(\frac{1-\sqrt{5}}{2}\right)^0$

$0 = \alpha_1 + \alpha_2 \Rightarrow \alpha_2 = -\alpha_1$

$f_1 = \alpha_1\left(\frac{1+\sqrt{5}}{2}\right)^1 + \alpha_2\left(\frac{1-\sqrt{5}}{2}\right)^1$

$1 = \alpha_1\left(\frac{1+\sqrt{5}}{2}\right) - \alpha_1\left(\frac{1-\sqrt{5}}{2}\right)$

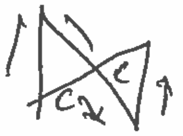
$1 = \alpha_1\left(\frac{2\sqrt{5}}{2}\right)$ $\alpha_1 = \frac{1}{\sqrt{5}}$, $\alpha_2 = \frac{-1}{\sqrt{5}}$ — 1M

Solution is $\frac{1}{\sqrt{5}}\left(\frac{1+\sqrt{5}}{2}\right)^n + -\frac{1}{\sqrt{5}}\left(\frac{1-\sqrt{5}}{2}\right)^n$ is the required explicit f.r. — 1M

14 (a) Eulerian graph.

An Eulerian circuit in a graph G is an Eulerian path in graph G whose end points are identical

A path in graph G that includes each edge of G exactly once and intersects each vertex of G at least once.



— 3M

Eulerian graph even vertices degree of.

Hamiltonian graph.

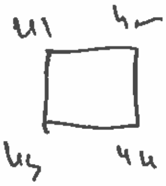
A path in a graph G is called Hamiltonian path if it contains every vertex of G.

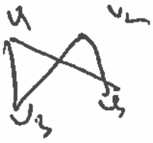
A cycles is said to be a Hamiltonian cycle, it contains every vertex

— 3M

6M

14 (b)





— 2M

|V1| = 4, |V2| = 4

|E| = 4, |E2| = 4

degree sequence 2,2,2,2 — 2M

Define f(u1) = v1

f(u2) = v4

f(u3) = v3

f(u4) = v2

(u1, u2) ∈ E(G1) ⇒ (f(u1), f(u2)) = (v1, v4) ∈ E(G2) — 2M

$A(G_1) = \begin{pmatrix} 0 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 1 & 0 & 0 & 1 \\ 0 & 1 & 1 & 0 \end{pmatrix}$ $A(G_2) = \begin{pmatrix} 0 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 1 & 0 & 0 & 1 \\ 0 & 1 & 1 & 1 \end{pmatrix}$

8M

15 a) Prims algorithm.

procedure — 2M

steps

Example — 4M

6M

b) Kruskals algorithm — Procedure steps — 2M

14 (a) Eulerean graph.

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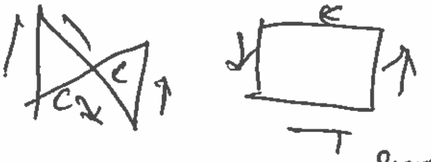


— 3 M

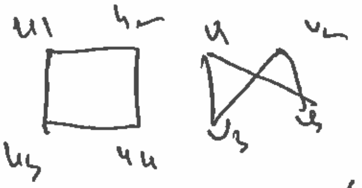
Eulerian graph even degree of vertices

Hamiltonian graph.

A path in a graph G is called Hamiltonian path if it contains every vertex of G. — 3 M

A cycle is said to be a Hamiltonian cycle, it contains every vertex — 6 M

14 (b)



— 2 M

$|V_1| = 4$, $|V_2| = 4$

$|E_1| = 4$, $|E_2| = 4$

degree sequence 2,2,2,2 — 2 M

Define $f(u_1) = v_1$

$f(u_2) = v_4$

$f(u_3) = v_3$

$f(u_4) = v_2$

$(u_1, u_2) \in E(G_1) \Rightarrow (f(u_1), f(u_2)) = (v_1, v_4) \in E(G_2)$ — 2 M

$A(G_1) = \begin{bmatrix} 0 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 1 & 0 & 0 & 1 \\ 0 & 1 & 1 & 0 \end{bmatrix}$ $A(G_2) = \begin{pmatrix} 0 & 1 & 1 & 0 \\ 1 & 0 & 0 & 1 \\ 1 & 0 & 0 & 1 \\ 0 & 1 & 1 & 1 \end{pmatrix}$

8 M

15 a) Prims algorithm. –

procedure – 2 M

steps

Example – 4 M

6 M

b) Kruskals algorithm – procedure steps – 2 M

Semester End Regular/Supplementary Examination, Dec. /Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	Program	Civil Engineering			Academic Year	2022 – 2023
Course Code	20CE302	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	Building Planning And Drawing						

Part A (Short Answer Questions 14 x 3 = 42 Marks)

No.	Questions (1 through 5)	Marks	Learning Outcome (s)	DoK
1 (a)	Explain the importance of dimensioning & conventional representations in the drawing and draw the conventional signs for the following :- i) Stone ii) Concrete iii) Plywood iii) Glass iv) Steel	7 M	20CE302.1	L2
1 (b)	Draw 1 ½ brick English bond with plan and elevation	7 M	20CE302.1	L2
2 (a)	Write in detail about for guidelines for planning the buildings to suit their functional requirements	7 M	20CE302.2	L2
2 (b)	List out the bye- laws and regulations which include all the features of various types of buildings	7 M	20CE302.2	L2
3 (a)	Write in detail about the orientation of buildings	7 M	20CE302.3	L2
3 (b)	In any residential building, what are the essentials which need to be considered while planning?	7 M	20CE302.3	L2
4 (a)	Name any four types of truss and sketch any one	7 M	20CE302.4	L2
4 (b)	Draw the elevation of a glazed window	7 M	20CE302.4	L3
5 (a)	Draw the layout for the residential building	8 M	20CE302.5	L3
5 (b)	Explain the principle of planning a hospital	6 M	20CE302.5	L2

Part B (Long Answer Questions 1 x 28 = 28 Marks)

No.	Questions (6 through 7)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw the layout of panelled door.	10 M	20CE302.4	L3
6 (b)	Draw the king post truss for a span of 10 m.	18 M	20CE302.4	L3
7	For the given line diagram from fig.1. Develop a plan and elevation of residential building. Take external wall thickness as 30 cm and internal wall 20 cm.	28 M	20CE302.5	L3

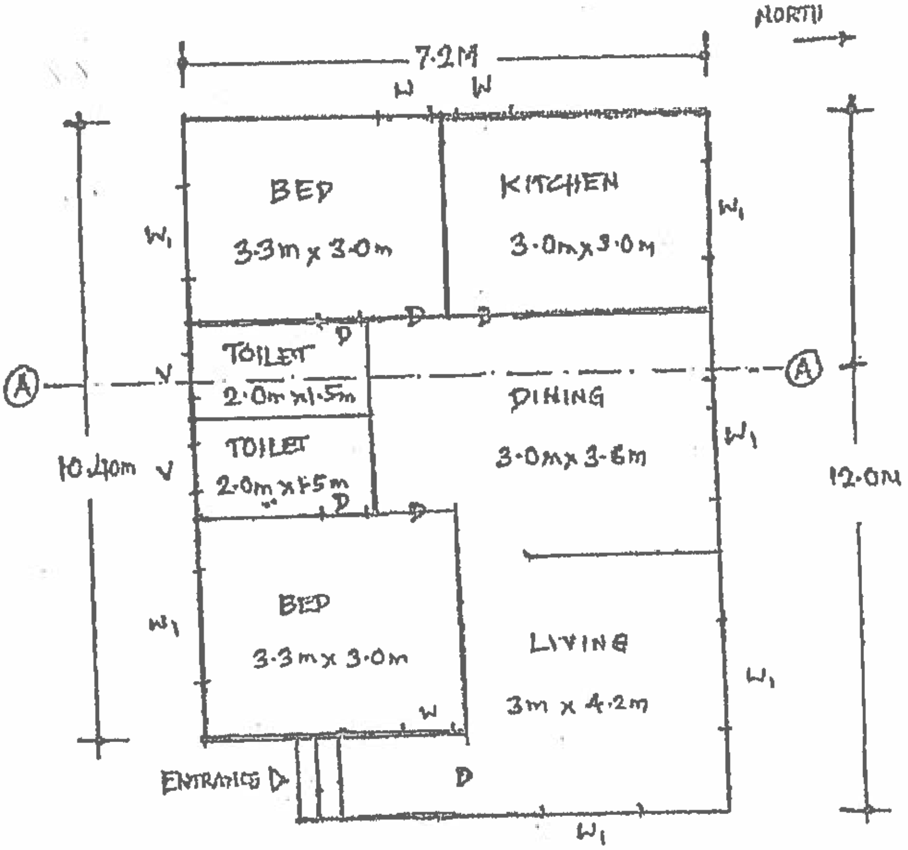

NORTH
7.2M
W
W
BED
3.3m x 3.0m
KITCHEN
3.0m x 3.0m
W1
W1
D
D
D
TOILET
2.0m x 1.5m
A
A
V
V
DINING
3.0m x 3.6m
W1
10.40m
12.0M
TOILET
2.0m x 1.5m
D
D
BED
3.3m x 3.0m
W1
LIVING
3m x 4.2m
W1
W
ENTRANCE D
D
W1

Fig.Q1

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ANSWER KEY AND SCHEME OF EVALUATION

28/11/22

Scheme of Evaluation

(2022-23 A.Y).

20CE302 – Building planning & Drawing, III sem

Part A: (14×3 = 42 M) (Answer any three).

1(a) Drawing of conventional signs for i) Stone ii) Concrete iii) Plywood iv) Glass v) Steel (each 1M)
Importance of Dimensioning in detail (2M). } 7M

1(b) Drawing of 1½ brick English bond } 7M
1) plan (4M)
2) Elevation (3M)

2(a) Guidelines for planning of buildings (4M) and suitability of functional requirements (3M) } 7M

2(b) Bye-Laws and regulations (5M)
features of various types of Buildings. (2M) } 7M.

3(a) Detailed explanation for orientation of Buildings 7M.

3(b) Essentials and considerations for residential buildings 7M.

4(a) List of some types of trusses (2M).
Drawing of any one type of truss (5M) } 7M

(P.T.O).

4(b) Drawing of elevation of glazed window with Dimensions 7M.

5(a) Drawing of layout of residential buildings. and mention scale & dimensions 8m

5(b) Explanation of the principles regarding the hospital building planning 6M

PART B. (1 X 28 = 28M). (Answer any one)

6(a) Drawing of layout for panelled door with Dimensions and scale 10 M

6(b). Drawing of king post truss for span of 10 M 18 m

} 28M

(or)

(7) For the given Dimensions & specifications Drawing of plan and elevation for residential building and assumptions for other dimensions 28m

70 marks

28/12/22

Faculty Signature

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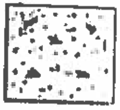
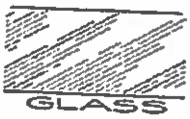
ANSWER KEY AND SCHEME OF EVALUATION

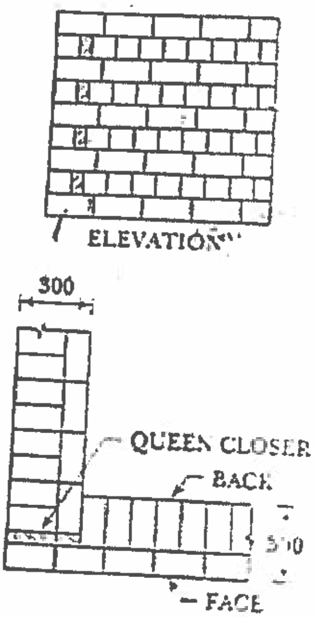
PART A

1A

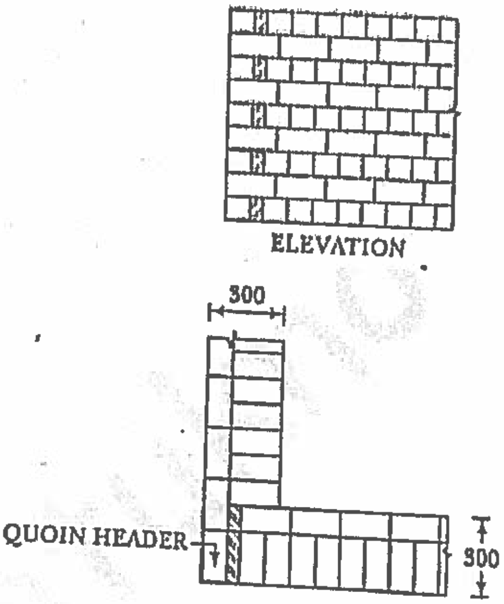


1B



1·20

ONE BRICK WALL



2a

Important fundamental requirements, a building should satisfy the following requirements in its design and construction:

1. Comfort and convenience
2. Durability
3. Heat or thermal insulation
4. Moisture or damp prevention
5. Security against burglary
6. Sound insulation
7. Strength and stability
8. Dimensional stability
9. Economy
10. Fire protection
11. Light and ventilation

Some other requirements are

- planning regulations.
- building line and control line.
- distances of building lines and control lines.
- built-up area.
- open space requirements.
- size of the rooms.
- the specifications for a maximum height of buildings

2B

Building bye laws are **a set of rules and regulations, that are followed to ensure a hassle-free construction process**. The development authority does not approve a building plan if it fails to adhere to the bye laws, hence it is important to adhere to the building bye laws. Building Bye-Laws are legal tools used to **regulate coverage, height, building bulk, and architectural design and construction aspects of buildings so as to achieve orderly development of an area. Building bye laws include norms related to the following:**

- Floor Area Ratio (FAR) and ground coverage.
- Density.
- Basement and parking spaces.
- Setbacks and projections.
- Area and its usage.
- Building height and other service spaces.
- Provision for lifts and basement area.

3a

Orientation is how a **building is positioned in relation to the sun's paths in different seasons, as well as to prevailing wind patterns**. In passive design, it is also about how living and sleeping areas are designed and positioned, either to take advantage of the sun and wind, or be protected from their effects. **Factors Affecting Building Orientation**

- Solar radiation and temperature. The intensity of solar radiation depends on the direction of sunrays
- Clouds and Rains. Clouds and rains have comparatively less importance in orientation of building.
- Humidity
- Humidity design consideration
- Prevailing winds.

While deciding the building orientation, one must also **take into consideration the location of landscape feature in a plot**, i.e. trees, planters, etc. which will affect the building depending on sun direction and sun path.

3B

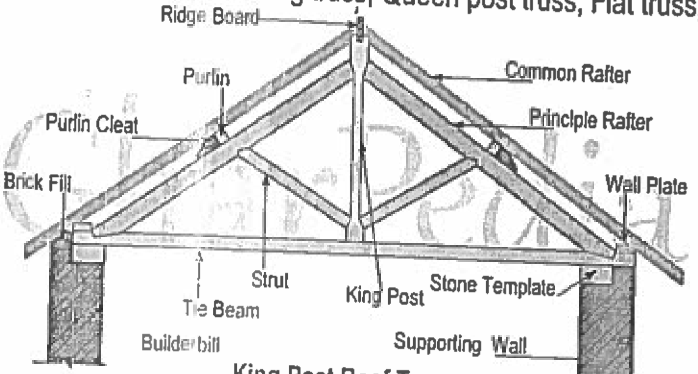
Main essentials are

- Aspects: The aspect of the house should be such that it enables the family members to live comfortably
- Prospect: Prospect is the view from outside of a house. ...
- Privacy: It is an important principle while planning a residential building. ...
- Grouping:
- Roominess
- Sanitation

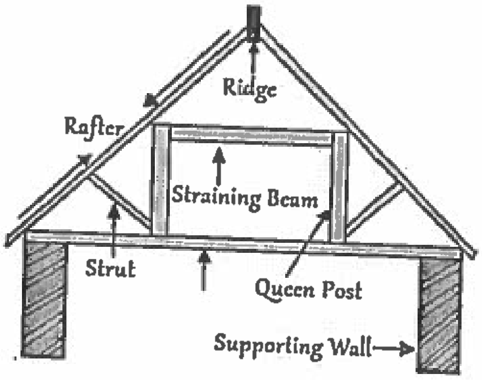
- Communication

4A

King post truss, Bowstring truss, Queen post truss, Flat truss, Lenticular truss are some other forms of trusses



King Post Roof Truss



Queen Post Truss

4b

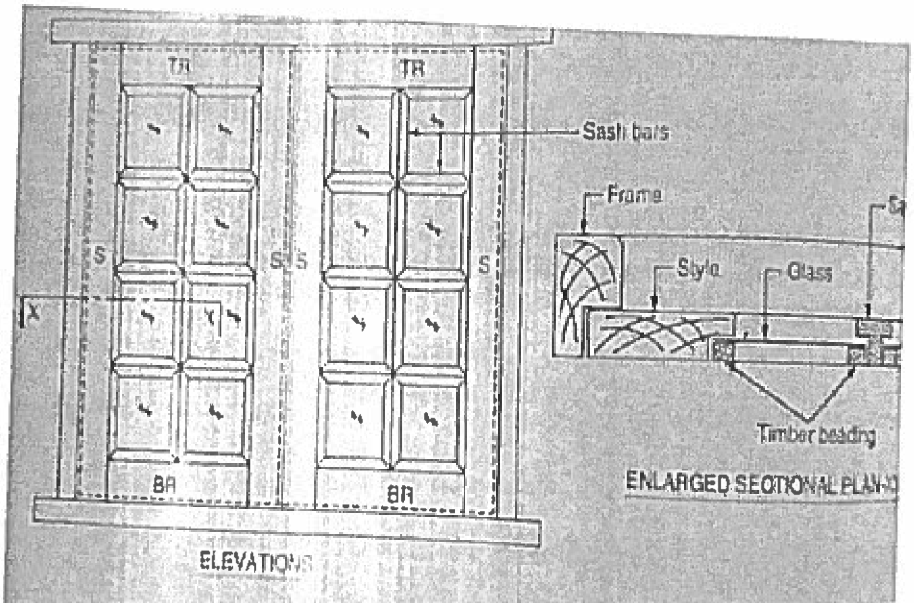
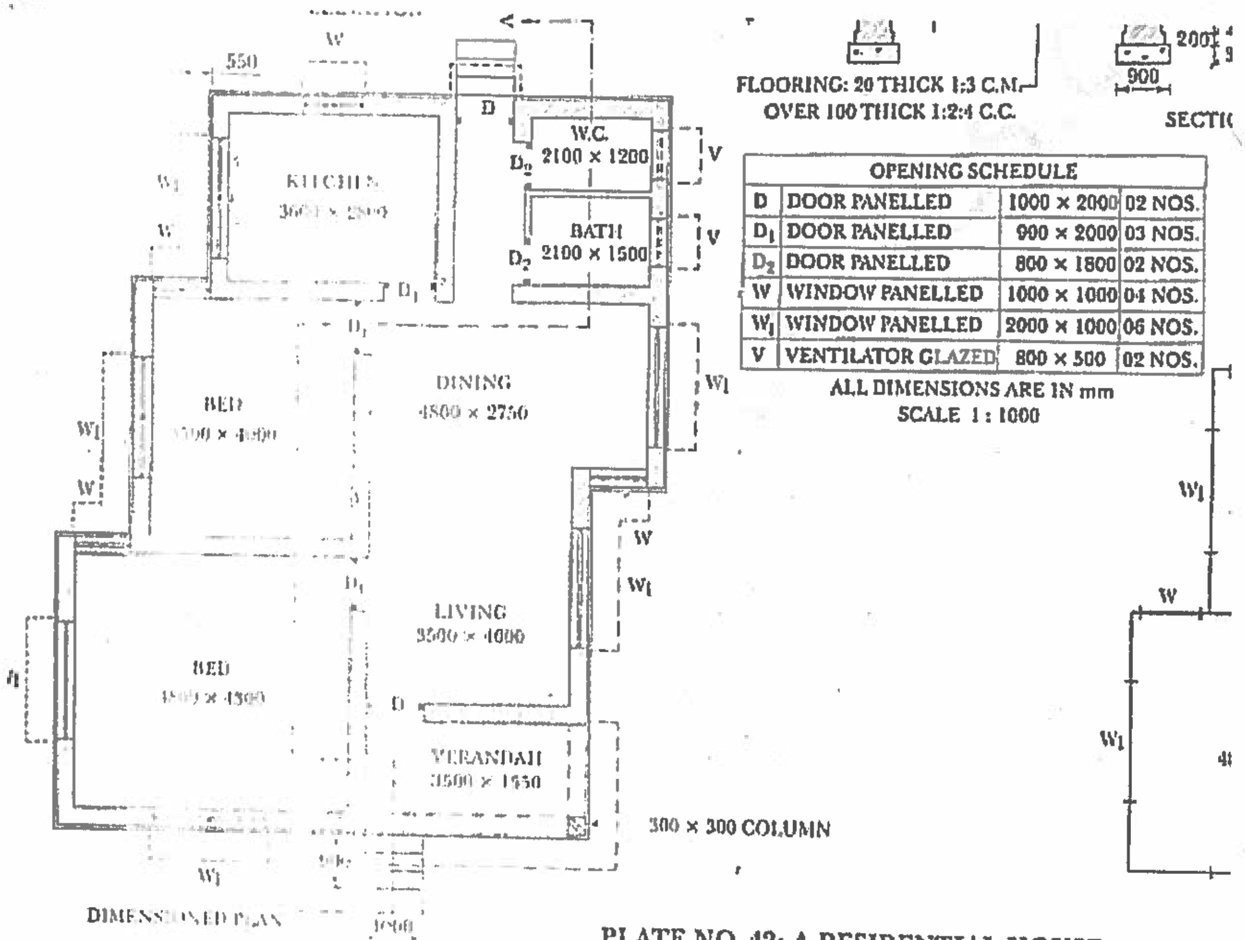


Fig. 6.18 Glazed or Sash Windows

5a



OPENING SCHEDULE

D	DOOR PANELLED	1000 × 2000	02 NOS.
D_1	DOOR PANELLED	900 × 2000	03 NOS.
D_2	DOOR PANELLED	800 × 1800	02 NOS.
W	WINDOW PANELLED	1000 × 1000	04 NOS.
W_1	WINDOW PANELLED	2000 × 1000	06 NOS.
V	VENTILATOR GLAZED	800 × 500	02 NOS.

ALL DIMENSIONS ARE IN mm
SCALE 1 : 1000

PLATE NO. 42: A RESIDENTIAL HOUSE

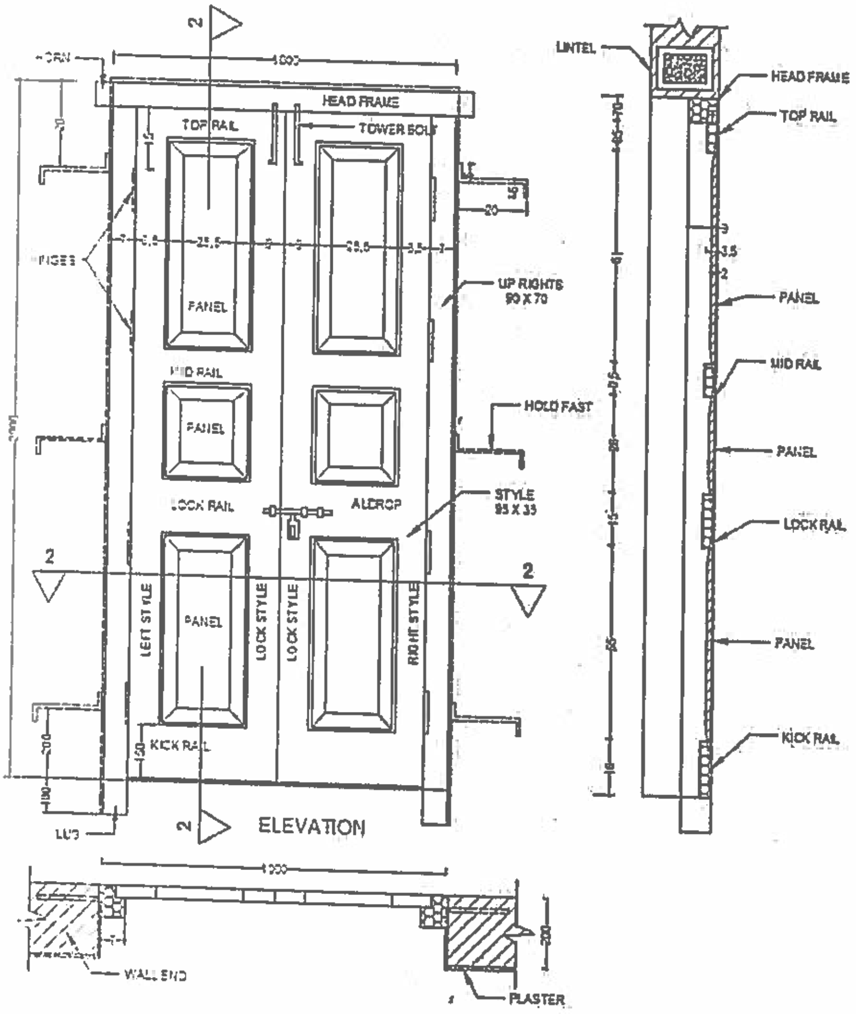
5b

Design an Efficient Space

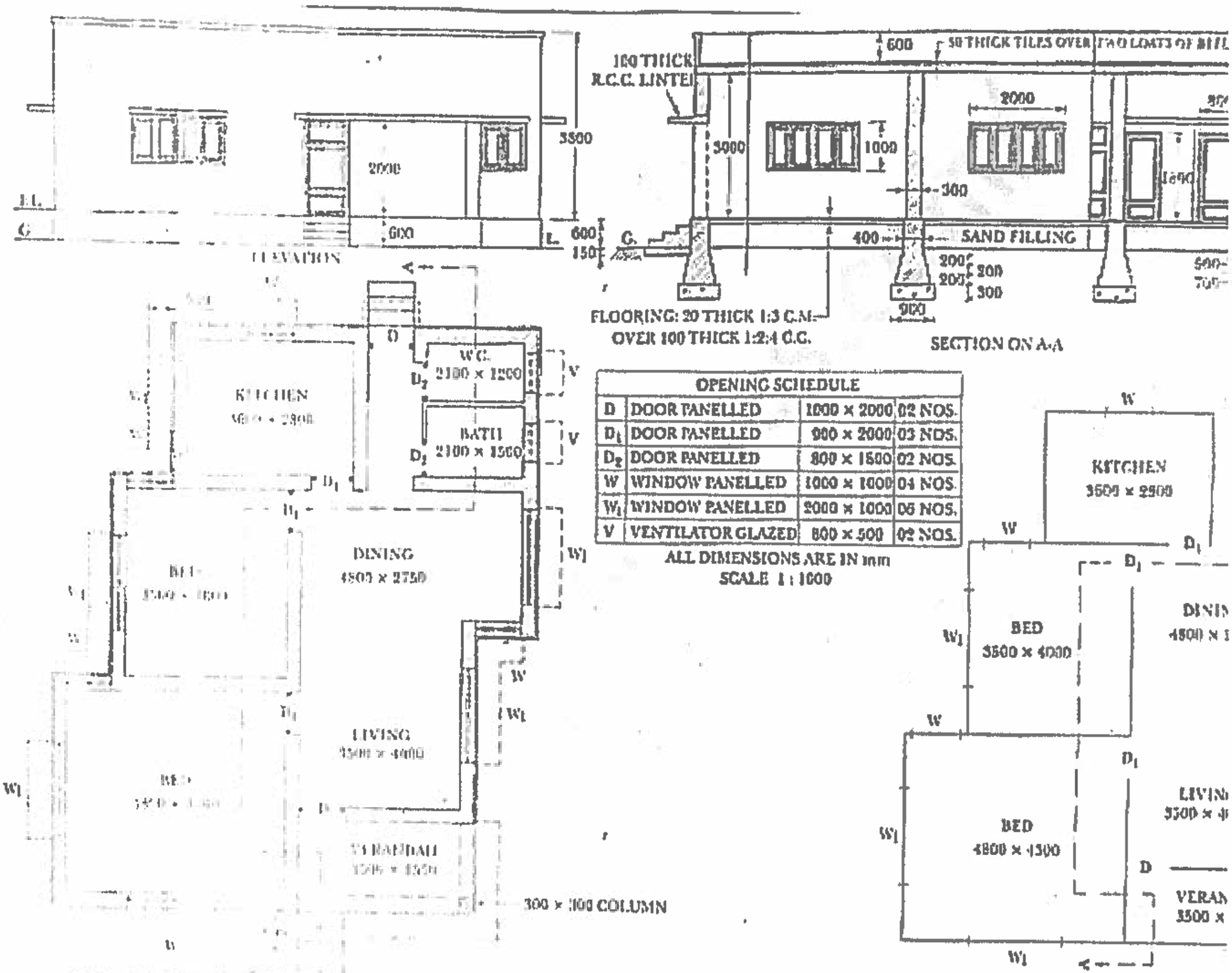
- Compliance and Expandability. The modes of treatments and medical needs change continuously in hospital therefore, a modular space planning and layout would be a great option. ...
- Choose The Right Materials. ...
- Security Concern. ...
- Waste Management System.
  - Hospital building differs from other building types in the complex functional relationship that exist between the various parts of the hospital.
  - Apart from providing right environment for patients and care providers, it should also be sensitive to the needs of visitors.
  - It is thus imperative to examine the emerging issues, analyze the challenges, appreciate the emerging trends and study the various strategic options available for planning, designing and construction of a hospital.

PART B

6.



PANELLED DOOR



OPENING SCHEDULE			
D	DOOR PANELLED	1000 × 2000	02 NOS.
D₁	DOOR PANELLED	900 × 2000	03 NOS.
D₂	DOOR PANELLED	800 × 1500	02 NOS.
W	WINDOW PANELLED	1000 × 1000	04 NOS.
W₁	WINDOW PANELLED	2000 × 1000	06 NOS.
V	VENTILATOR GLAZED	800 × 500	02 NOS.

ALL DIMENSIONS ARE IN mm
SCALE 1 : 1000

PLATE NO. 40: A RESIDENTIAL HOUSE

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	Program	Mechanical Engineering			Academic Year	2022 - 2023
Course Code	20ME301	Test Duration	3 Hrs.	Max. Marks	70	Semester	III
Course	Thermodynamics						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define an open and closed system	20ME301.1	L1
2	Define the term Internal energy and enthalpy	20ME301.2	L1
3	Define the term COP	20ME301.3	L1
4	Define Boiling point and Melting point	20ME301.4	L1
5	What is compressibility factor?	20ME301.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	With neat sketch, explain about constant volume gas thermometer	6M	20ME301.1	L2
6 (b)	Derive the pdV - work expression for isothermal ($pV = C$) and polytrophic ($pV^n = C$) quasi -static processes	6M	20ME301.1	L3
	OR			
7 (a)	A gas expands from an initial state with $p_1 = 340$ kPa and $V_1 = 0.0425$ m^3 to a final state where $p_2 = 136$ kPa. If the pressure-volume relationship during the process is PV^2 = Constant, Determine the work in kJ	7M	20ME301.1	L3
7 (b)	A platinum resistance thermometer has aresistance of 2.8 Ω at 0 °C and 3.8 Ω at 100 °C. Calculate the temperature when the resistance indicated is 5.8 Ω	5M	20ME301.1	L3
8 (a)	Air enters a compressor with a velocity of 60 m/s , pressure 100 kPa, temperature 40°C and leaves the compressor with a velocity of 90 m/s, 500 kPa and 120 °C. Consider the system is adiabatic. Find the power of motor for the mass flow rate of 40 kg/min. Write the assumption made.	10M	20ME301.2	L3
8 (b)	Write down the simplified steady flow energy equation for steam turbine	2M	20ME301.2	L1
	OR			
9 (a)	Using first law of thermodynamics, prove that the difference in specific heat capacities (C_p-C_v) = gas constant (R)	8M	20ME301.2	L3
9 (b)	What is PMM1? Why it is impossible?	4M	20ME301.2	L1
10 (a)	With neat sketch, state Kelvin-Planck and Clausius statement of Second law of thermodynamics	6M	20ME301.3	L2
10 (b)	A refrigerator removes heat from a refrigerated space at 2 °C at a rate of 300 kJ/min and rejects heat to kitchen air at 26 °C at a rate of 345 kJ/min. Verify whether it violates II law of thermodynamics by Clausius inequality and Carnot theorem	6M	20ME301.3	L3
	OR			
11 (a)	Two Carnot engines A and B are operated in series. The first one (A) receives heat at 870 K and rejects to a reservoir at temperature T. The second engine (B) receives the heat rejected by the first engine and in turn rejects to a heat reservoir at 300 K. Calculate the temperature T in °C for the following cases. (i) The work output of the two engines are equal (ii) The efficiencies of the two engines are equal	8M	20ME301.3	L3
11 (b)	A Carnot engine works between 300 °C and 30 °C . The heat supplied to the engine is 20 kJ.Determine (i) efficiency (ii) Work output and (iii) Heat rejection	4M	20ME301.3	L3

12 (a)	A boiler generates steam at 3 bar and 0.85 dry from water at 45 °C, 540 kJ/s heat is added during the evaporation. Calculate the amount of steam generated per hour.	4M	20ME301.4	L3
12 (b)	Draw the p-v diagram of pure substances and explain various regions of the diagram in detail	8M	20ME301.4	L2
	OR			
13 (a)	Ten kg of water at 45 °C is heated at a constant pressure of 10 bar until it becomes superheated vapour at 300 °C. Find the change in volume , enthalpy ,internal energy, and entropy.	4M	20ME301.4	L3
13 (b)	Discuss about h-s and T-s diagram for a pure substance	8M	20ME301.4	L2
14 (a)	Compute the specif volume of steam at 0.9 bar and 550 K using Vander Waal's equation. Take critical temperature of steam is 647.3 K and critical pressure is 220.9 bar, Molecular weight of steam is 18 g/mol	5M	20ME301.5	L3
14 (b)	A tank of 1 m^3 capacity originally contains O_2 at a pressure of 5 bar and 350 K. Nitrogen is introduced without change in temperature until the pressure in the tank becomes 12 bar. Determine the mass of each gas in the tank and partial volume of each gas.	7M	20ME301.5	L3
	OR			
15 (a)	Discuss about vander-Waal's equation of state and its limitations	6M	20ME301.5	L2
15 (b)	Define the following: (i) Specific Humidity (ii) Relative Humidity (iii) Absolute Humidity	6M	20ME301.5	L1

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY
(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Degree:- B.Tech. Branch:- Mech. Course Code:- 20ME301
3rd Semester. Subject :- Thermodynamics AY → 2022-23

PART – A

1) Open System :- In this system, both energy and mass transfer takes place.

closed system :- In this system, Energy transfer takes place, where the mass remains constant.

2) Internal energy(U):- Under Internal energy, the system is pocessed in closed system and the energy is created within it due to increase in pressure (or) temperature.

Enthalpy (H):- In this parameter, the energy is created by the addition of Internal energy (U) and pressure and volume.

$$\boxed{H = U + PV}$$

3) COP (Coefficient of Performance)

The coefficient of performance is factor applicable for the Heat pump and Refrigerator systems. Under this analysis the heat absorption and heat delivered is determined, which is greater than 1.

4) **Boiling Point** :- The Boiling point of a substance is defined as the increase in temperatiere the convers of liquid state to vapour state.

Eg:- water converting into vapour at 100°c

Melting point :- The conversion of solid state to liquid state is ~~de~~ simply defined as the melting point.

Eg:- Conversion of solid ice to water at 0°c.

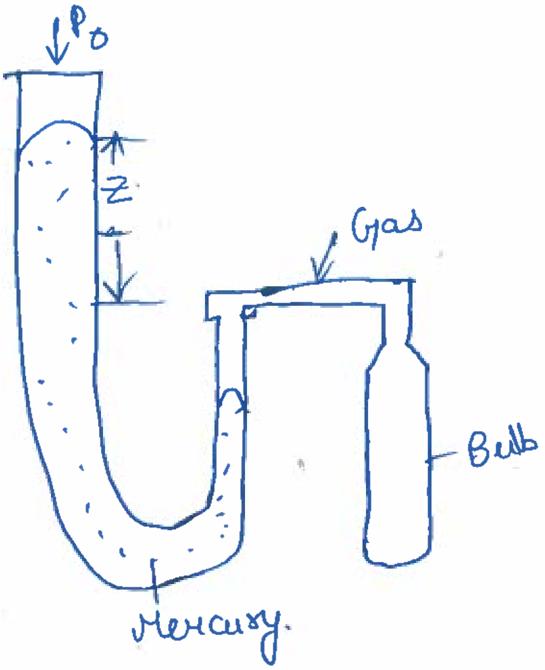
5). • The compressibility factor can be defined as the molar volume of gas to an ideal gas. For an ide ideal gas, the compressibility factor will always = 1

PART-B

6(a) **Constant Volume gas thermometer.** (6M)

In constant volume gas thermometry a small amount of gas is enclosed in bulb "B", which is connected to capillary tube with two limbs filled up with mercury, while the other limb of manometer is open to the atmosphere and can be moved vertically to adjust mercury level.

The pressure of the gas with the bulb is given by $\boxed{P_g = P_m + P_{atm}}$ where $P = \rho g h$.



According to the ideal gas equation we know that

$$\boxed{PV = mRT}$$

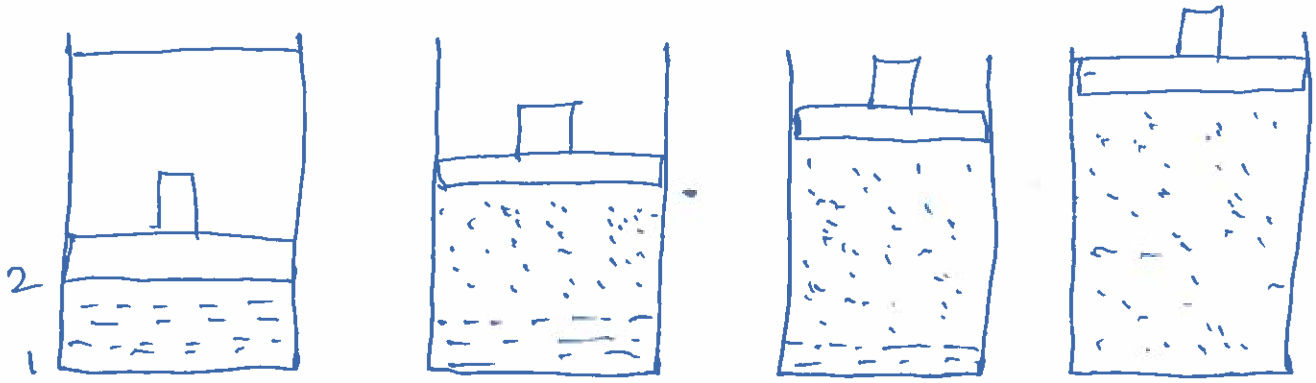
∵ constant pressure gas thermometer

$$V = \frac{mRT}{P}$$

$$\boxed{dV = \frac{mRdT}{P}}$$

6(b) ·PdV – work expansion. (6M)

Let the gas in the cylinder be a system having initially the pressure P_1 and volume V_1. The system is in thermodynamic equilibrium, the of which is described by the coordinates P_1, V_1. The piston is the only boundary which moves due to gas pressure.



Let the piston move to a new position 2, which is also a thermodynamic equilibrium state specified by pressure p_2 and volume V_2. When the piston moves an infinitesimal distance dl, and if 'a' be the area of the piston, the force F acting on the piston F=pa.

$$dW = F \cdot dl = pa\,dl = PdV$$

where $dV = a\,dl$ = infinitesimal displacement volume.

The piston moves out of position 1 to position 2 with the volume changing from V_1 to V_2.

$$W_{1-2} = \int_{V_1}^{V_2} P\,dV.$$

7(a). Given Data. $P_1 = 340 kPa$

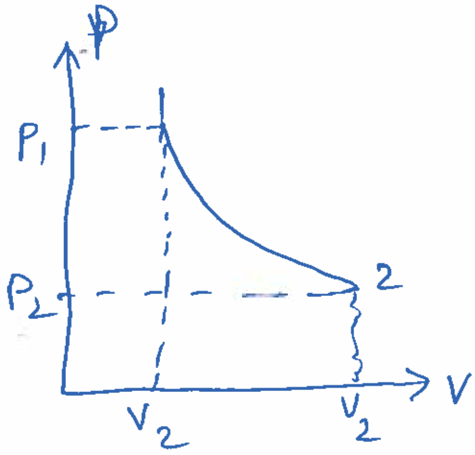
$P_2 = 136 kPa$ } 1M

$V_1 = 0.0425 m^3$

$\boxed{P_1 V_1 = P_2 V_2}$ — 1M

$\boxed{dQ = dU + dW}$ — 1M.

we need to find Out workdone



calculation :— 3M.

Solution result — 1M

from the relation $PV^2 = c$

$$P_1 V_1^2 = P_2 V_2^2$$

$$V_2 = \sqrt{\frac{P_1 V_1^2}{P_2}} = \sqrt{\frac{340 \times 0.0425^2}{136}}$$

$$= 0.0672 m^3$$

$$W_{1-2} = \int_1^2 P\,dV$$

$$W_{1-2} = \int_{V_1}^{V_2} \frac{c}{V_2^2}\,dv = c \int_{V_1}^{V_2} V^{-2}\,dV = -c\left[\frac{1}{V}\right]_{V_1}^{V_2}$$

$$= c\left[\frac{1}{V_1} - \frac{1}{V_2}\right] = P_1 V_1^2 \left[\frac{1}{V_1} - \frac{1}{V_2}\right]$$

$$= 340 \times 0.0425^2 \left[\frac{1}{0.0425} - \frac{1}{0.0672}\right]$$

$$\boxed{W_{1-1} = 5.31 KJ}$$

7(b). $R = R_0 (1+\alpha t)$

$t = 0°C$; $R = R_0 = 2.8$ ohm

$t = 100°C$, $R_{100} = 3.8$ ohm

$R_{100} = R_0 (1+\alpha \times 100)$

$3.8 = 2.8 (1+\alpha \times 100)$

$\alpha = 0.357 \times 10^{-2}$

$R = 5.8$ ohm

$5.8 = 2.8 (1+0.357 \times 10^{-2} \times t)$

$\boxed{t = 300°C}$

8(a) Given

The steam ~~expands~~ Compresses

the assumptions made

1. are with steady flow with no change with time

2. Potential energy negligible.

$E_{in} - E_{out} = \Delta E_{system}$

$W_{in} + mh_1 = Q_{out} + mh_2$.

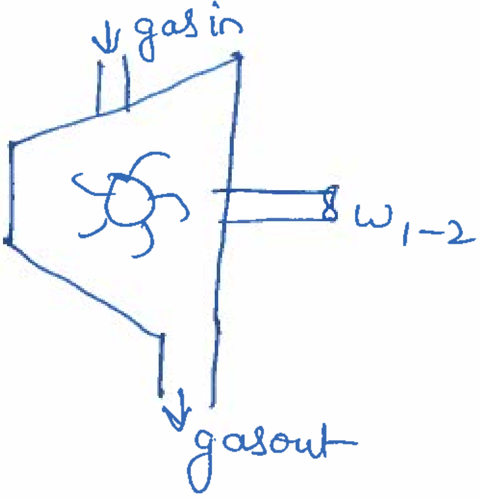
$W_{in} - Q_{out} = m (h_2 - h_1) = mc_p (T_2 - T_1)$

$W = Q_{out} + mc_p (T_2 - T_1)$

$\boxed{W = 965 \text{ kW}}$

8(b) Steady flow equation for turbine

The turbine is device converts the energy of the working substance into the workdone with in it



$$\cancel{q_{1-2}}^{\,0} - w_{1-2} = d\cancel{(KE)}^{\,0} + \cancel{d(PE)}^{\,0} + dh$$

$$-w_{1-2} = h_2 - h_1$$

$$\boxed{w_{1-2} = h_1 - h_2}$$

9(a) According to first Law of Thermodynamics

$$dQ = dU + dw$$

$$mc_p dT = mc_v dT + d(PV)$$

$$mc_p(\Delta T) = mc_v \Delta T + P_2V_2 - P_1V_1$$

$$mc_p(\Delta T) = mc_v \Delta T + mRT_2 - mRT_1$$

$$mc_p dT = mc_v dT + mR(dT)$$

$$\cancel{m}c_p \cancel{\Delta t} = \cancel{m}\cancel{dT}(C_v + R)$$

$$C_p = C_v + R$$

$$\boxed{C_p - C_v = R}$$

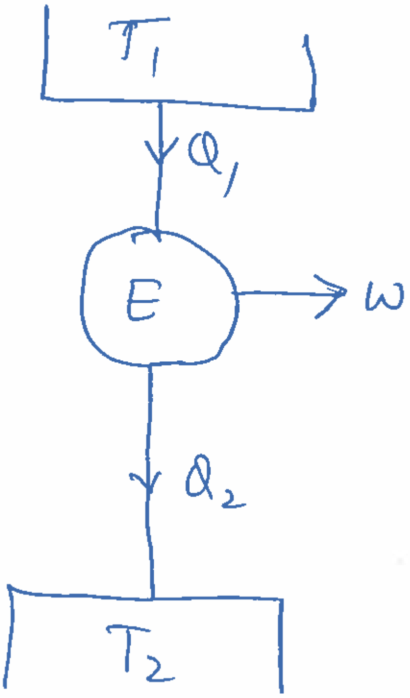
9(b) PMM-I

Perpetual Motion Machine of the first kind would be one to violate the 1st law of Thermodynamics to permanently produce useful energy without any energy source or to produce more energy than consumed

Thus generate energy from nowhere such a machine is impossible to construct as any machine cannot produce energy on their own without any supply of energy to it initially.

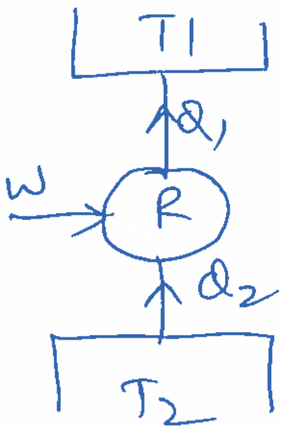
10(a) Kelvin Plank statement

It is impossible to construct a machine working in a cyclic process with a single reservoir..



Clausius Statement

The machine working in a cyclic process, which is impossible to flow from low temperature reservoir to higher temperature reservoir without external agency.



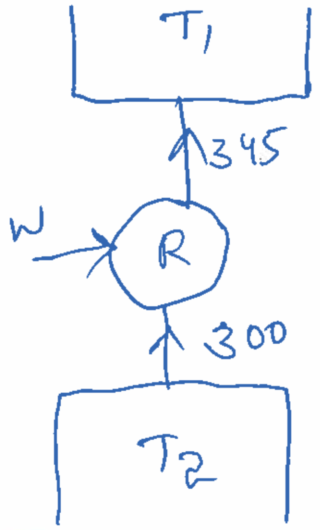
10(b) Given $T_2 = 2°C$; $Q_2 = 300$ kJ/min ; $Q_1 = 345$ kJ/min

$T_1 = 26°C$.

$$W = Q_1 - Q_2 = 45\,kW$$

$$COP = \frac{Q_2}{W}$$

$$COP = \frac{300}{45} = 6.66$$



It is to be noted that the above statement proves the clausius inequality & satisfies towards it

11(a) Given $T_1 = 800\,K$; $T_3 = 400\,K$..

Let the output of both engines be W.

Let the engine A take Q_1 heat as input at temperature T_1 and gives out heat Q_2 at temperature T the second engine B recieve Q_2 as input and give out Q_3 at temperature T_2 to the sink

workdone by Engine, A ; $W = Q_1 - Q_2$

" " " B ; $W = Q_2 - Q_3$

Thus

$$Q_1 - Q_2 = Q_2 - Q_3$$

Dividing both sides by Q_1

$$1 - \frac{Q_2}{Q_1} = \frac{Q_2}{Q_1} - \frac{Q_3}{Q_1}$$

$$1 - \frac{T}{T_1} = \frac{Q_2}{Q_1}\left(1 - \frac{Q_3}{Q_2}\right)$$

$$1 - \frac{T}{T_1} = \frac{Q_2}{Q_1}\left(1 - \frac{T_3}{T}\right)$$

$$1 - \frac{T}{T_1} = \frac{T}{T_1}\left(1 - \frac{T_3}{T}\right)$$

$$\frac{T_1}{T} - 1 = 1 - \frac{T_3}{T}$$

$$\frac{T_1}{T} + \frac{T_3}{T} = 2$$

$$\frac{1}{T}(T_1+T_3)=2$$

$$T=\frac{(T_1+T_3)}{2}=650\text{K}$$ is the temperature when output of the engines are equal

Let the efficiency of both engines be η, Now considering both engines efficiency are equal.

$$1-\frac{T}{T_1}=1-\frac{T_3}{T}$$

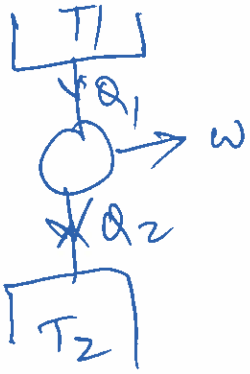
$$\frac{T}{T_1}=\frac{T_3}{T}$$

$$T^2=T_1T_3$$

$$T^2=800\times400=320000$$

$T=565{\cdot}68\text{ K}$ is the temperature when efficiencies of the both the engines are equal.

11(b) Given $T_1 = 300°C$; $T_2 = 30°C$; $Q_1 = 20\text{kJ}$



$$\eta=\frac{\cancel{T_1}}{T_1-T_2}=1-\frac{T_2}{T_1}=1-\frac{30}{300}=1-0{\cdot}03$$

$$= ~~30~~\ 96{\cdot}7\%$$

$$\eta=\frac{W}{Q_1}=96{\cdot}7\times20=W$$

$$W=1934\text{ W}$$

~~$W=(Q_2-Q_1$~~ $W=Q_1-Q_2$

12 (a) Steam generation at

$P_1 = 3$ bar and $x = 0.85$

$T = 45°C$; $\dot{Q} = 540$ kJ/s.

from steam tables take the values of

$P_1 = 3$ bar ; for V_f, V_g, S_f, S_g.

$$x = \frac{m_v}{m_f + m_v}$$

$$v = v_f + x v_g \; ; \; s = S_f + x \cdot S_{fg}$$

13 (a) From steam tables, corresponding to 45°C

$V_1 = V_{f_1} = 0.001010 \, m^3/kg$; $h_1 = h_{f_1} = 188.4$ kJ/kg

$S_1 = S_{f_1} = 0.638$ kJ/kgK

from steam tables, corresponding to 10 bar & 300°C

$h_2 = 3052.1$ kJ/kg ; $S_2 = 7.125$ kJ/kgK

$V_2 = 0.258 \, m^3/kg$

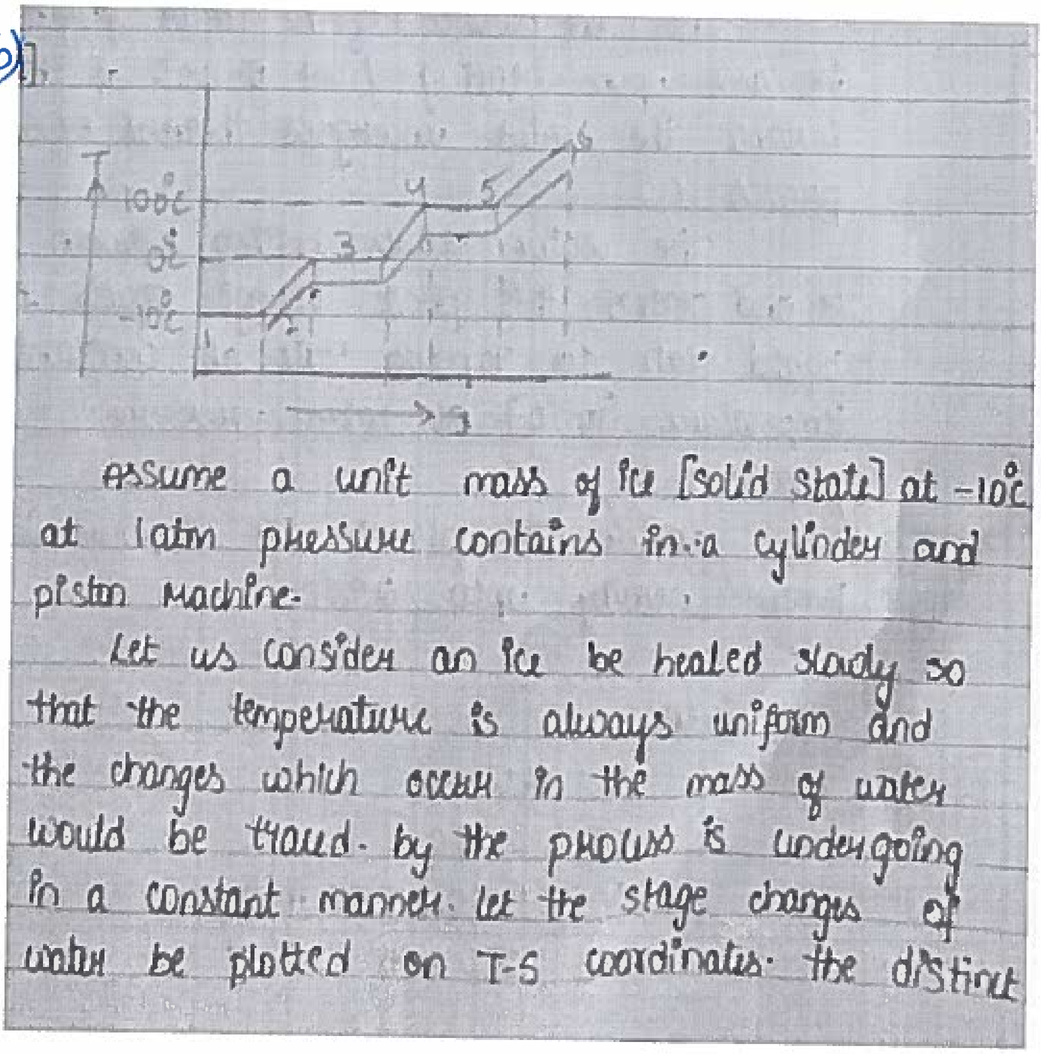
Change in volume ⇒ $v = v_f + x v_{fg}$

enthalpy ⇒ $h = h_f + x h_{fg}$

Internal energy ⇒ $u = u_f + x u_{fg}$

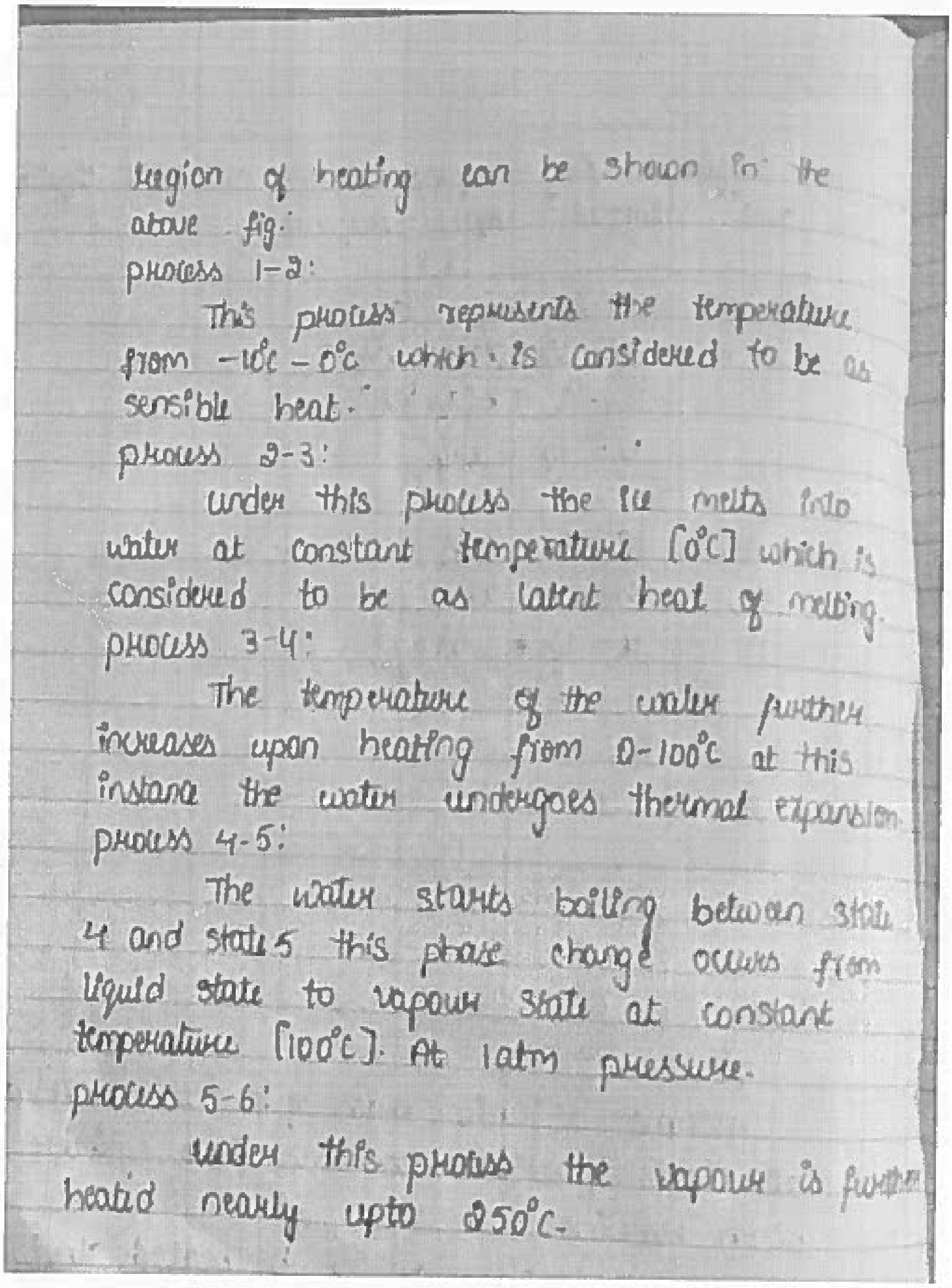
Entropy ⇒ $s = s_f + x s_{fg}$.

12(b)



Assume a unit mass of ice [solid state] at -10°c at 1atm pressure contains in a cylinder and piston machine.

Let us consider an ice be heated slowly so that the temperature is always uniform and the changes which occur in the mass of water would be traced. by the process is undergoing in a constant manner. Let the stage changes of water be plotted on T-S coordinates. The distinct



region of heating can be shown in the above fig.

process 1-2:

This process represents the temperature from -10°c - 0°c which is considered to be as sensible heat.

process 2-3:

under this process the ice melts into water at constant temperature [0°C] which is considered to be as latent heat of melting.

process 3-4:

The temperature of the water further increases upon heating from 0-100°c at this instance the water undergoes thermal expansion

process 4-5:

The water starts boiling between state 4 and state 5 this phase change occurs from liquid state to vapour state at constant temperature [100°C]. At 1atm pressure.

process 5-6:

under this process the vapour is further heated nearly upto 250°C.

B(b)

⑮

Pure Substance:

A pure substance is a compound of const chemical composition throughout its mass constitutin of single component system it may exist in one or more phases

Dryness fraction:

If 1kg of mass of liquid + vapour mixture x kg is the mass of the vapour and (1-x) kg is the mass of the liquid the "x" is represented as the dryness fraction (or) quality of substance

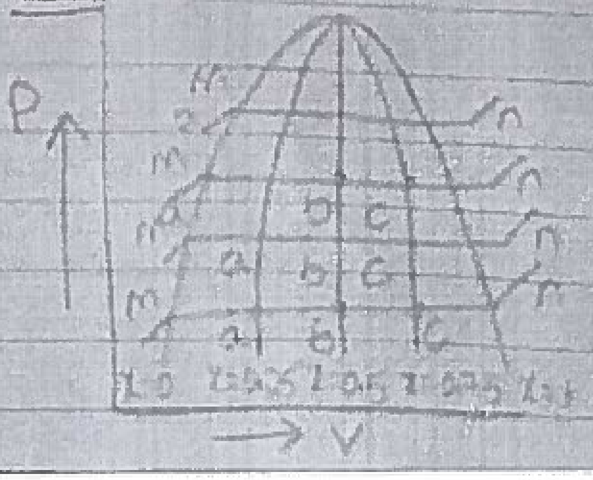
$$x = \frac{m_v}{m_v + m_f}$$

m_v & m_f represents the mass of the vapour and liquid respectively in the mixture. The value of the 'x' varies between 0 and 1 for saturated water the dryness fraction is '0'.

And for saturated vapour the dryness fractio x=1

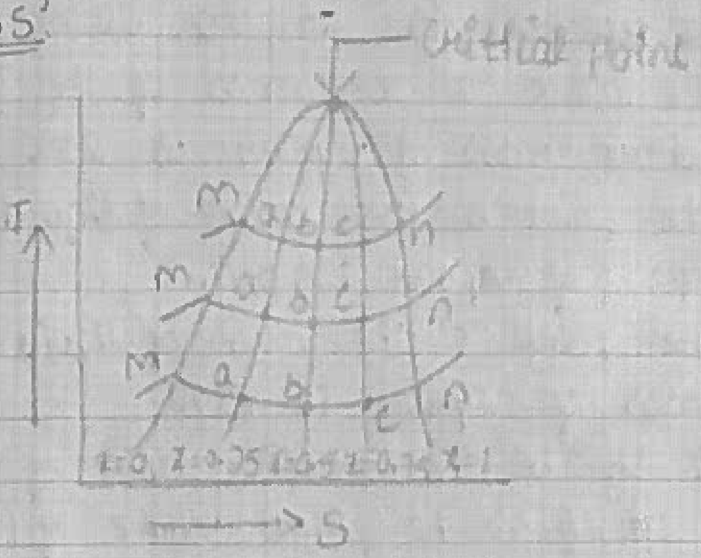
For which the vapour is set to dry saturated.

P.V:

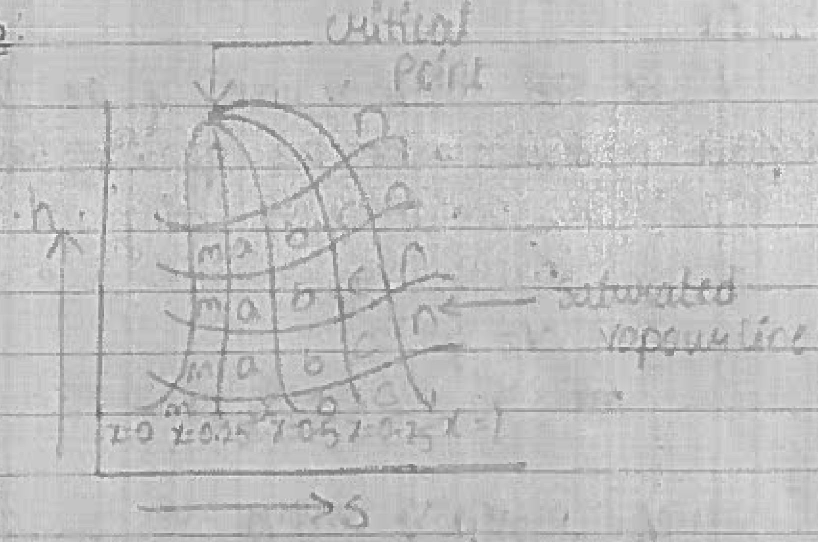


16

T.S:



h.S:



m_v and m_f are mass of the "vapour and fluid" respectively.

The point 'm' indicates the saturated liquid state with "x = 0".

The point 'n' indicates the saturated vapour state with

'x = 1'

The lines m-n transition of liquid to vapour points a, b, c represents the pressure and temp. variations with respective to the mass of the vapour reaches to 25, 50, 120%

→ At the point 'a' the mass of the liquid is 75% where mass of vapour is 25%

→ At the point 'b' the mass of liquid is 50% and mass of vapour is 50%.

→ At the point 'c' the mass of liquid is 25% and mass of vapour is 75%.

Let V_g be the specific volume of saturated vapour. Let V_f be the specific volume of saturated liquid.

"V" be the total volume of the liquid cum vapour mixture under the quality index 'x'.

The corresponding masses being m, m_f, m_g.

$$\boxed{m = m_f + m_g}$$

$$\boxed{V = V_f + V_g}$$

$$\boxed{mv = m_f V_f + m_g V_f}$$

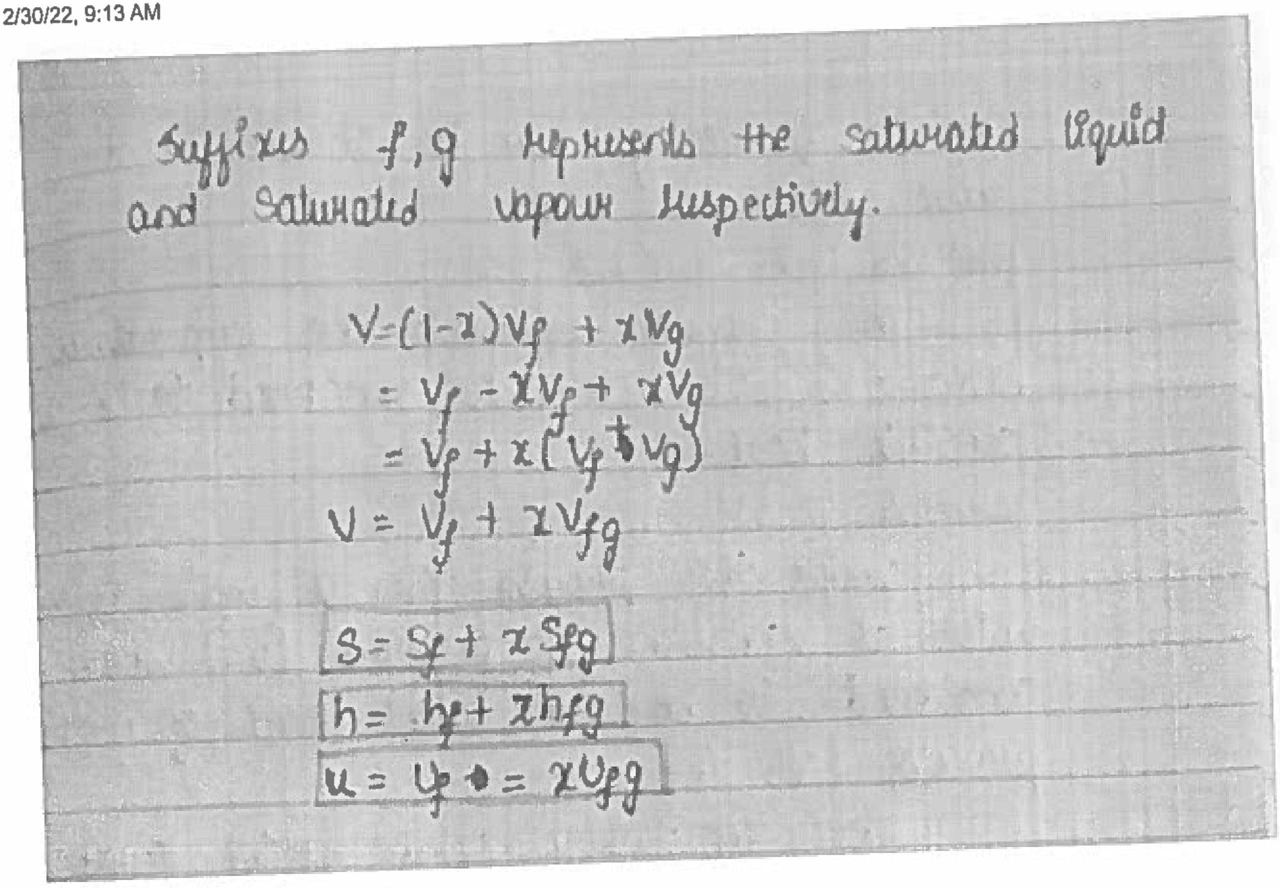
$$mV = (m - m_g)V_f + m_g V_g$$

$$\boxed{V = (1-x)V_f + x V_g} \rightarrow \text{specific volume}$$

$$\boxed{S = (1-x)S_f + xS_g} \rightarrow \text{entropy}$$

$$\boxed{h = (1-x)h_f + x h_g} \rightarrow \text{enthalpy}$$

The terms s, v, h, u are the entropy, specific volume, enthalpy and internal energy with respect to the mixture of quality ~~x~~ "x"



Suffixes f, g represents the saturated liquid and saturated vapour respectively.

$$v = (1-x)v_f + x v_g$$
$$= v_f - x v_f + x v_g$$
$$= v_f + x(v_f + v_g)$$
$$v = v_f + x v_{fg}$$

$$\boxed{s = s_f + x s_{fg}}$$
$$\boxed{h = h_f + x h_{fg}}$$
$$\boxed{u = u_f = x u_{fg}}$$

14(a) · Given data ; steam at 0·9 bar. (P_1)

$T_1 = 550k.$; $T_2 = 647.3k.$

$P_2 = 220.9$ bar.

$M = 18$ g/mol.

~~$V = 1m^3$~~

~~$P_3 = 5$ bar ; $T_3 = 350k.$~~

from steam table take the values retated ·

$$v = v_f + x\, v_{fg}.$$

14(b) $P_1 = 5$ bar ; $T_1 = 350$ k. ; $V = 1m^3$; $P_2 = 12$ bar.

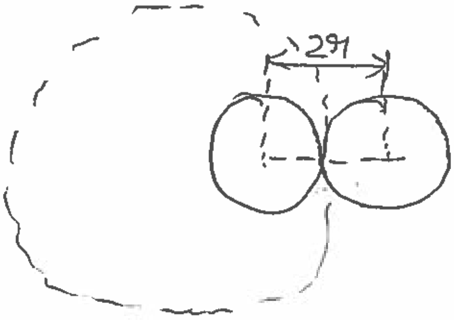
mass of the gas $\Rightarrow x = \frac{m_f}{m_f + m_g}.$

$$v = v_f + x \cdot v_{fg}.$$

15(a) <u>Vander waal's equation</u>

It states that the ideal gases doesnot exhibit the conditions of real gases

In the real gases the above 2 statements are not suitable and correction need to be made



Under the volume with respect to the real gases the occupation of single molecule is considered to be as

$$V = \frac{4}{3}\pi r^3$$

for two molecules

$$V = \frac{4}{3}\pi (2r)^3$$

$$= \frac{8\left(\frac{4}{3}\pi r^3\right)}{2}$$

~~=~~

$$= 4\left(\frac{4}{3}\pi r^3\right)$$

for n molecules

$$P \propto \frac{n^2}{V^2}$$

$$P = a\frac{n^2}{V^2} = P + a\frac{n^2}{V^2}$$

$$\because PV = nRT$$

$$\boxed{\left(P + a\frac{n^2}{V^2}\right)(V - nb) = nRT}$$

15 (b) (i) Specific Humidity

The Mass of water vapour present is 1kg of dry air which is donated in terms of gm/kg

(ii) Absolute Humidity

The mass of water vapour present is volume of dry air and is measured in the terms of gm/m³

(iii) Relative Humidity

The ratio of the mass of the water vapour is the moist air to the mass of the water vapour is the saturated air

K. Abdul

N

HoD ME
30/12/2022

Semester End Regular/Supplementary Examination, Dec. /Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	EEE & ECE			**Academic Year**	2022 - 2023
Course Code	20EC302	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Electronic Devices and Circuits**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Write any 2 applications of PN-junction diode.	20EC302.1	L1
2	Draw Zener Diode Characteristics.	20EC302.2	L1
3	Compare CE, CC and CB configurations.	20EC302.3	L1
4	What is thermal run away?	20EC302.4	L1
5	Mention small signal parameters of JFET.	20EC302.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive the current diode equation.	6M	20EC302.1	L2
6 (b)	What is the P-N junction? Discuss the behavior of a P-N junction under forward and reverse bias.	6M	20EC302.1	L2
	OR			
7 (a)	Describe the current components in P-N diode.	6M	20EC302.1	L2
7 (b)	What is the effect of temperature on P-N junction diode?	6M	20EC302.1	L1
8 (a)	Compare the characteristics of PN junction diode, Zener Diode and Tunnel diode.	6M	20EC302.2	L2
8 (b)	Draw the equivalent circuit of UJT and discuss its working from the circuit.	6M	20EC302.2	L2
	OR			
9 (a)	Explain the operation of Full Wave Rectifier with necessary graphs	6M	20EC302.2	L2
9 (b)	Explain the operation of (i) Inductor filter (ii) capacitor filter	6M	20EC302.2	L2
10 (a)	Explain the drain and transfer characteristics of a N-Channel JFET.	5M	20EC302.3	L2
10 (b)	Sketch the family of CE output characteristics for a transistor. Explain cutoff, active, saturation region.	7M	20EC302.3	L2
	OR			
11 (a)	Define α and β of a transistor and derive the relationship between them.	4M	20EC302.3	L2
11 (b)	Explain the operation of n-p-n BJT with CE input and output characteristics.	8M	20EC302.3	L2
12 (a)	Obtain an expression of stability factor for fixed bias.	5M	20EC302.4	L2
12 (b)	What is Biasing? Explain the need of it. List out different types of biasing methods.	7M	20EC302.4	L2
	OR			
13 (a)	In a Silicon transistor circuit with a fixed bias, VCC=9V, RC=3KΩ, RB=8KΩ, β=50, VBE=0.7V. Find the operating point and Stability factor.	6M	20EC302.4	L3
13 (b)	Explain about Thermistor and Sensistor bias compensation techniques.	6M	20EC302.4	L2
14 (a)	For the Common Source Amplifier, calculate the value of the voltage gain, given i) rd=100KΩ, RL=10KΩ, gm=300μ and RO=9.09KΩ. ii) If CDS=3pF, determine the output impedance at a signal frequency of 1 MHz.	7M	20EC302.5	L3

14 (b)	Discuss the analysis of small signal model of JFET.	5M	20EC302.5	L2
	OR			
15 (a)	Give the comparison of BJT, JFET and MOSFET.	4M	20EC302.5	L1
15 (b)	Obtain the expression for voltage gain and current gain of a small signal low frequency Common Emitter amplifier.	8M	20EC302.5	L2

NSRIT (AUTONOMOUS) ①

Subject name: Electronic Devices and Circuits
Code: (20EC302)

	Scheme of Valuation	Marks Division	Total Marks
PART-A	(5×2=10 Marks)		
Q.NO			
1	P-N Junction diode Any two Applications (valid points)	2 Marks.	
2	zener diode characteritics	2 marks.	
3	Defination CB, CE, CC Any two Diffence of CB, CE, CC	1 Mark 1 Mark	
4.	Thermal runeway defination	2 Marks	
5.	Any two Parameters of JFET	2 Marks.	
	PART-B		
6(a)	Diode current equation $P_n(x) = P_n - P_n(0)$ upto $P_n(x) = P_n + P_n(0)e^{-x/Lp}$	2M	6M
	Solvation upto $I_{pn}(x) + I_{np}(x)$ $I_{pn}(0) + I_{np}(0)$	2M	
	Final Expression $I = I_o(e^{V/nV_T} - 1)$ soluation.	2M	
6(b)	P-N Junction diode Defination	2M	6M
	Forward circuit diagram with graph	1M	
	operation & working	1M	
	Reverse circuit diagram with graph	1M	
	working	1M	

7(a)	Current component diagram list out the all the current components. I_{pn}, I_{np}, I_n, I_p $I_{pn}(x)$, $I_{pn}(0)$, $I_{np}(x)$, $I_{np}(0)$ Abrivations & working	2M 2M 2M	6M
7(b)	Effects of temperature of P-N Junction diode characteristis Explination Formations and Expressions	2M 2M 2M	6M
8(a)	Difference between PN, zener, Tunnel diodes Abrivation and full form Symbols diffence Other diffence	2M 1M 3M	6M
8(b)	UJT Defination UJT symbol UJT Equivalent circuit diagram working of UJT	1M 1M 2M 2M	6M
9(a)	Fullwave rectifier Defination circuit diagram working wave forms	1M 2M 2M 1M	6M
9(b)	inductor filter, capacitor filter Derivations circuit diagrams working & wave forms	2M 2M 2M	6M

Q.NO				
10(a)	N-channel Defination J-FET	— 1M	5M	
	N-channel circuit diagram of JFET	— 1M		
	Drain & Trans- Characteritics -for	— 2M		
	JFET Explination	— 1M		
10(b)	Common Emitter Defination	— 1M	7M	
	BJT CE Circuit diagram	— $1\frac{1}{2}$M		
	input output wave form of CE	— $1\frac{1}{2}$M		
	CE Explination	— 3M		
11(a)	Transistor General equations & - Derive & Derivation process	— 2M — [illegible]	4M	
	output and final Expression $\alpha = \frac{\beta}{1+\beta}$, $\beta = \frac{\alpha}{1-\alpha}$	— 2M		
11(b)	BJT CE Defination	— 2M	8M	
	BJT CE Circuit diagram	— 1M		
	Input & output wave forms of CE	— $1\frac{1}{2}$M		
	BJT CE circuit operation & BJT CE working	— $3\frac{1}{2}$M		
12(a)	Stability Factor Defination	— 1M	5M	
	Circuit diagram	— $1\frac{1}{2}$M		
	Expression and Derivations	— $2\frac{1}{2}$M		
	and working	— [illegible]		

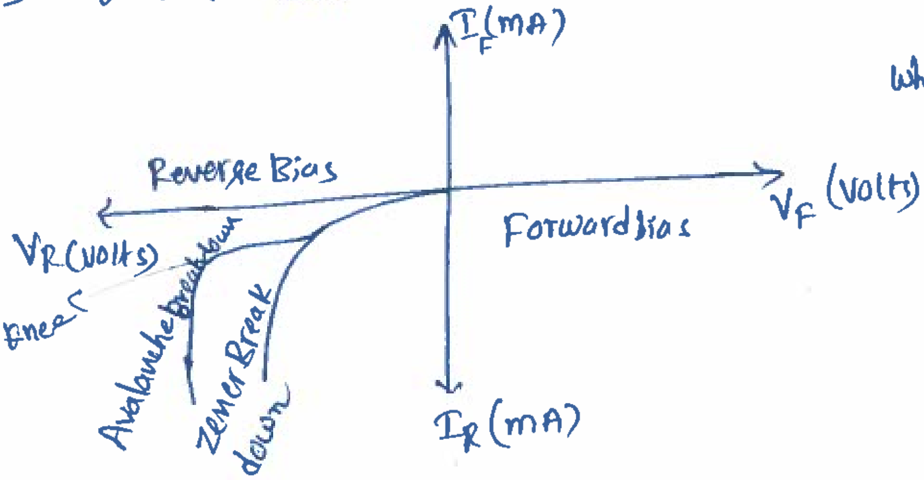
Q.NO			
12(b)	biasing Defination — 2M listout the types of biasing methods — 2M Circuit diagram of Needfor biasing — 1½ 2M Explanation of biasing methods — 1½ M	7M	
13(a)	Given data of fixed bias — 1 M formula's for operating points, Stability factor — 2 M Calculation — 2 M Final Answer — 1 M	6 M	
13(b)	Definations of thermistor, Sensistor — 2 M Circuit diagram of campersion — 2 M Explination of thermistor & Sensistor — 2 M	6 M	
14(a)	Given that CS Amplifier — 1M Formule's for Voltage gain — 2 M Calculation Current gain — 2 M final Answer — 2 M	7M	
14(b)	Defination Small signal model — 2M J-FET Circuit diagram — 1½ M J-FET working — 1½ M	5M	
15(a)	Avbrivations (BJT, FET, mosFET) — 1M symbols BJT, FET, mosFET 1M Any other diffence (atleast) '2' — 2M	4M	
15)(b)	Small signal circuitdiagram — 2M voltage gain — 2M current gain — 2M explination — 2 M	8 M.	

PART-A (Short Answer Questions 5×2 = 10 Marks).

① Any 2 Applications of P-N Junction Diode:-

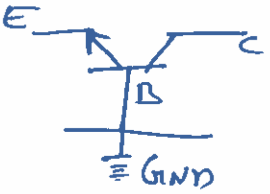
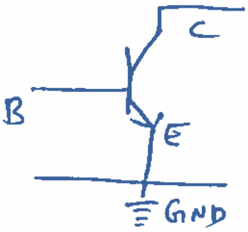
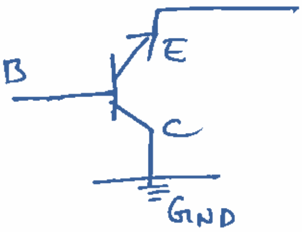
(i) It is used in DC power Supplies
(ii) It is used in rectifiers
(iii) It is used in Solar cell
(iv) It is used in computers, etc...
(v) It is used as a switch.

② Draw the zener diode characteristics:-



where:
I_F = Forward Current (mA)
V_F = Forward voltage in volts
I_R = Reverse Current (mA)
V_R = Reverse Voltage in Volts.

③ Compare CE, CC and CB configuration

S.NO	Parameter	CB	CE	CC
1	Full form	Common Base	Common Emitter	Common collector
2	Symbol's			
3	Current Amplification factor	α	β	γ
4	Current Gain	< unity	High	High
5.	Application	high frequency circuits	Audio frequency circuits.	Impedence matching circuits

4. What is thermal runaway ?

The self Distoration of unstablized transistor, the collector junction of a transistor withstand maximum temperature is called thermal runaway in this condition the transistor CE configuraton those parameters I_B, β, I_{co} when every 10°c the I_{co} (leakage current) value is doubled $\therefore I_c = \beta I_B + (1+\beta) I_{co}$, this process called "thermal runaway".

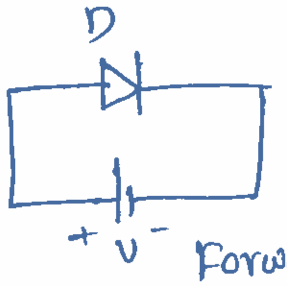
5. Mention Small signal Parameters of FET ?

i) Input impedence / Drain resistance (r_d)
ii) Current Gain / Amplification factor ($\mu = g_m \times r_d$)
iii) Trance Conductance (g_m)
iv) Voltage Gain (A_v).

PART-B (Long Answers & Questions 5×12=60 Marks):

6(a) Derive the current diode Equation :

let us assume the channel width (x) of the depletion region is 'zero'. the concentration of holes in N-type.



Forward biased P-n junction Diode

$$P_n(x) = P_n - P_{no} \longrightarrow ①$$

$$P_n(x) = P_{no} + P_n(0) e^{-x/Lp} \longrightarrow ②$$

$$P_n(x) = P_n(0) e^{-x/Lp}$$

At the condition where the distance $x = 0$ in eq ① & ②

$$P_n(x) = P_n$$

$$P_n(0) = P_n$$

$$P_n(0) = P_{no} + P_n(0) e^{-0/Lp} \longrightarrow ③$$

$$P_n(0) = P_n(0) - P_{no}$$

$$P_p = P_n\, e^{V_B/V_T} \longrightarrow ④$$

where V_B = Barrer voltage

V_T = thermal voltage

the total diode current and voltage of Barrier

$V_B = V_0 - V$ Applied in forward bias

$$P_p = P_{P_0}, \quad P_n = P_{no}$$

$$P_{P_0} = P_{no}\, e^{(V_0 - V)/V_T} \longrightarrow ⑤$$

Here the potential voltage 'V' is equal to zero

$$P_{P_0} = P_{no}\, e^{V_0/V_T} \longrightarrow ⑥$$

eq ⑤ = eq ⑥

$$P_{no}\left(e^{V_0 - V/V_T}\right) = P_{no}\, e^{V_0/V_T} \longrightarrow ⑦$$

$$\boxed{P_n = P_{no}\, e^{V/V_T}} \longrightarrow ⑧$$

Substituting the value of 'P_n' in equation ① from equation ①,

$$P_n(x) = P_n - P_{no}$$

$$P_n(x) = P_{no}\, e^{V/V_T} - P_{no}$$

$$P_n(0) = P_{no}\left(e^{V/V_T} - 1\right) \longrightarrow ⑨.$$

The total diode current is equal to sum of $I_{pn}(x)$ and $I_{np}(x)$

$$I = I_{pn}(x) + I_{np}(x)$$

$$I_{pn}(x) = -AeD_p \frac{d}{dx}(P_n(x))$$

$$I_{np}(x) = -AeD_p \frac{d}{dx}\left(P_{no} + P_n(0)\bar{e}^{x/L_p}\right)$$

$$I_{pn}(x) = Ae\frac{D_p}{L_p}\left(P_{no}\, e^{V/V_T} - 1\right) \longrightarrow ⑩$$

Similar

$$I_{np}(x) = \frac{AeD_n}{L_n} N_{po}\left(e^{V/V_T} - 1\right) \longrightarrow ⑪$$

Total current

$$I = I_{pn}(0) + I_{np}(0)$$

$$I = \frac{AeD_p}{L_p} P_{no}\left(e^{V/V_T} - 1\right) + \frac{AeD_n}{L_n} N_{po}\left(e^{V/V_T} - 1\right) \longrightarrow ⑫$$

$$\boxed{\therefore\ I = I_o\left(e^{V/V_T} - 1\right)}$$

where I = Diode current

I_o = Reverse Sturation current

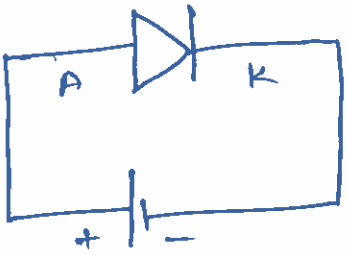
V = Applied voltage

η = constant 1 for Germanium
2 for silicon

V_T = thermal voltage $\left(V_T = \frac{kT}{q}\right)$

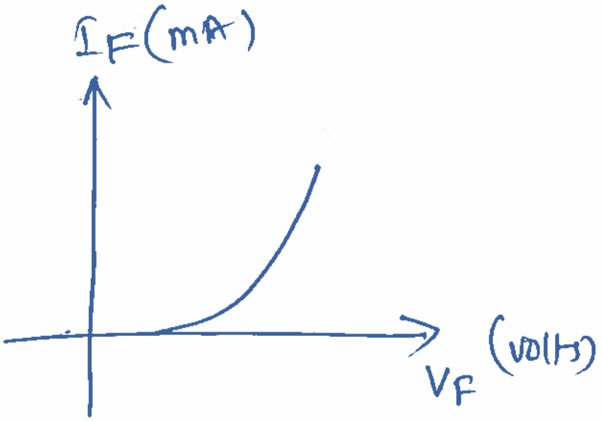
6(b) what is the P-N Junction ? the behaviour of a P-N Junction under forward and reverse bias ?

The P-n junction diode is two terminal Semiconductor device the combination of two Extrinsic material P-type and N-type material Physically Join together forming a new device is called "Pn junction Diode".



Forward Biasing:-

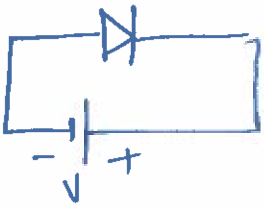
In this condition the battery of the positive terminal is connected to Anode and, battery -ve. terminal is connected to Cathode this process is called P-n junction diode forward biasing



Forward biasing characteristics:-

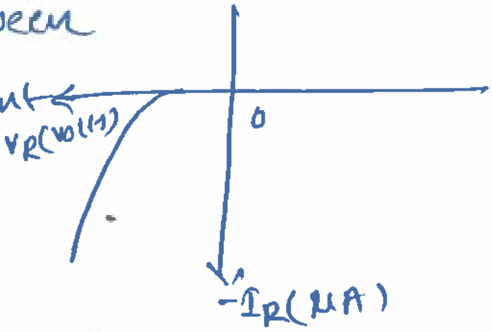
the graph drawn between forward voltage (V_F) and forward current (I_F) in mA.

Reverse biasing :-

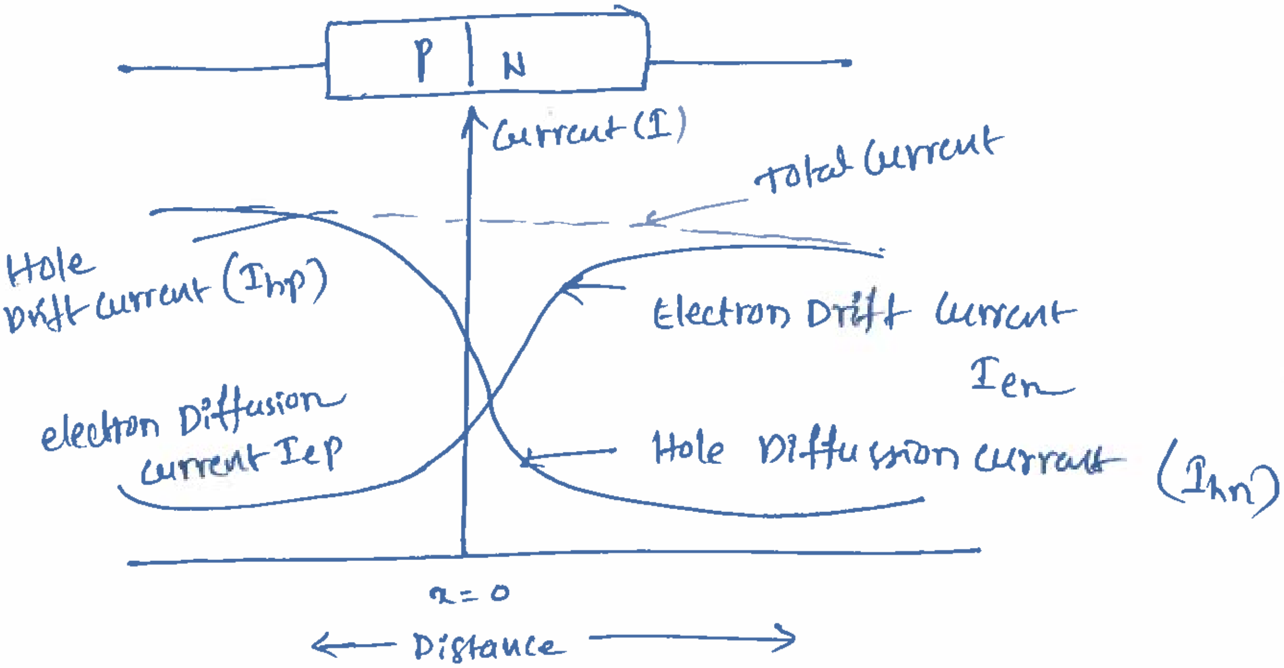


the battery of the -ve terminal is connected to Anode and battery of the +ve terminal is connected to Cathode this process is called "Reverse biasing in PN Diode".

Characteristics:- the graph drawn between reverse voltage (V_R) and reverse current (I_R) in µA



7(a) Describe the current components in P-N diode :-



the current componts in Pnjunction diode :

I_{np} = it is due to electrons in p region

I_{nn} = due to electrons in N-region

I_{pp} = due to holes in P-region

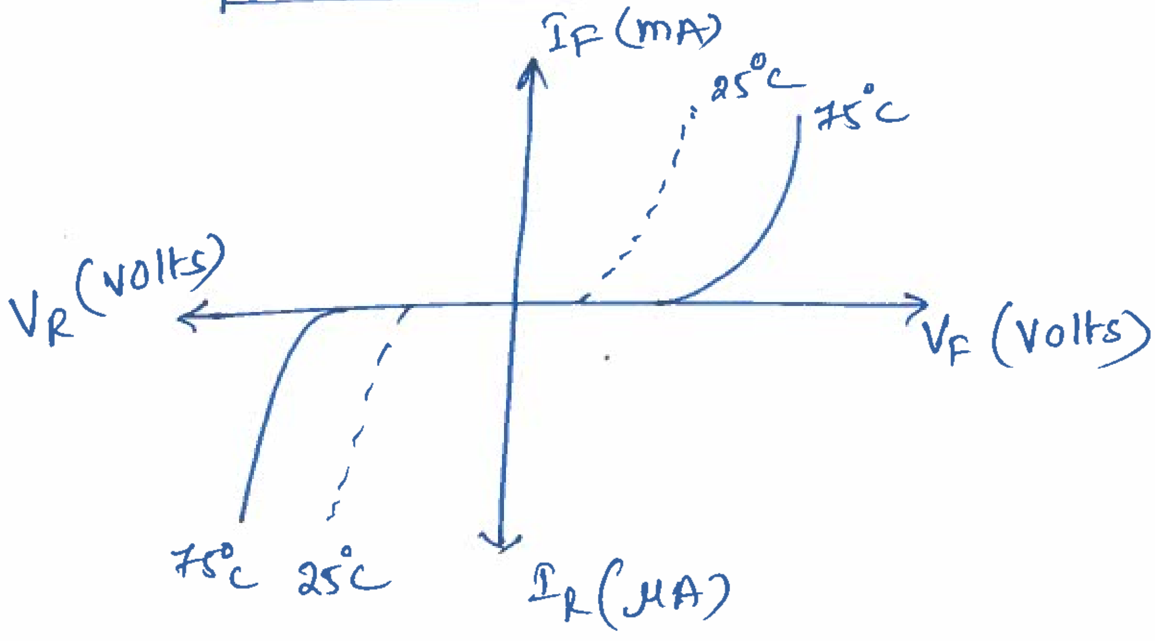
I_{pn} = due to holes in N-region.

7(b) what is the effect of temperature on P-N Junction diode·

→ In the P-n junction diode the rise in temperature -increases hole pairs in Semi conductors increases

→ the conductivity also increased as the result of current through

$$\boxed{I = I_0 \left(e^{V/\eta V_T} - 1\right)}$$

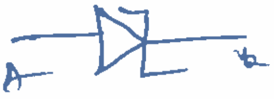
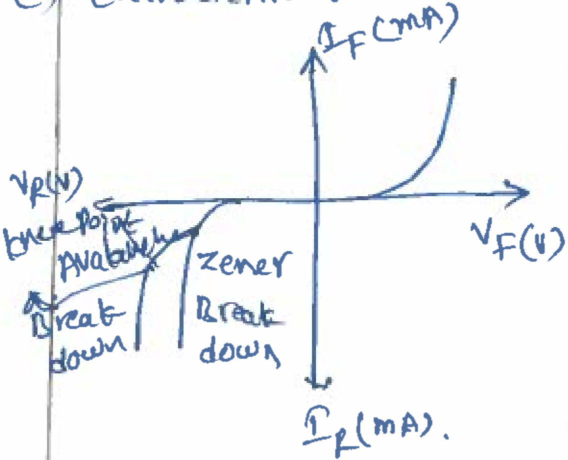
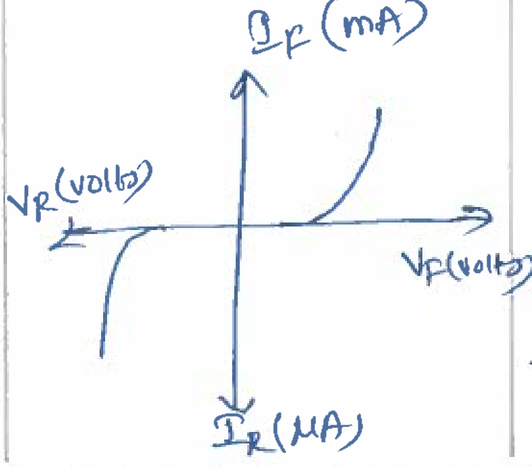
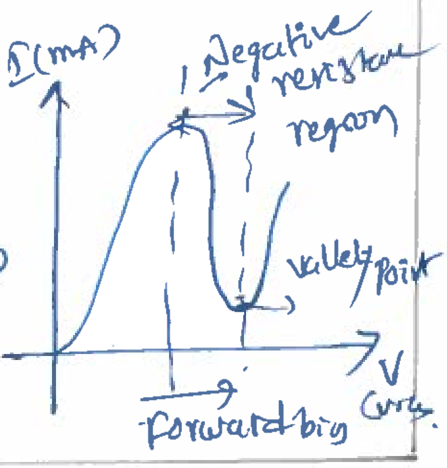


In this phenomena every 10°c of the leakage current value is doubled, either, Germenium and silicon $\frac{dv}{dT} = -2.5mV/°c$, in order to maintain the current (I) to a constant value $I_{o2} = I_{o1} \times 2^{(T_2-T_1)/10}$

I_{O_1} = Saturation Current of the diode at temperature (T_1)

I_{O_2} = Saturation current of the diode at temperature (T_2).

8(a) Campare the characteritics of PN Junction diode, zenerdiode, and Tunnel diode.

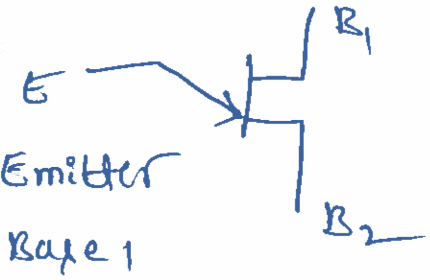
S.No	Zener diode	PN junction Diode	Tunnel diode
(i)	it is a two terminal semi conductor device	(i) it is a two terminal semi conductor device	(i) it is a two terminal Semi Conductor device.
(ii)	it is a heavily doped	(ii) it is lightly doped ordinary diode	(ii) Ekasi diode is a Tunneling Occours.
(iii)	it's operated in Reverse bias zener mechanisum based worked	(iii) it's conduct's both Forward and reverse biased.	(iii) it is working in Forward bias only in Negative resistence region only
(iv)	symbol 		
(v)	Characteritics:- 		

8(b) Draw the Equivalent Circuit of UJT and discuss its working from the Circuit?

Equivalent Circuit diagram of UJT :-

Unijunction Transistor (UJT) :

symbol



→ it is a three terminal semi conductor device

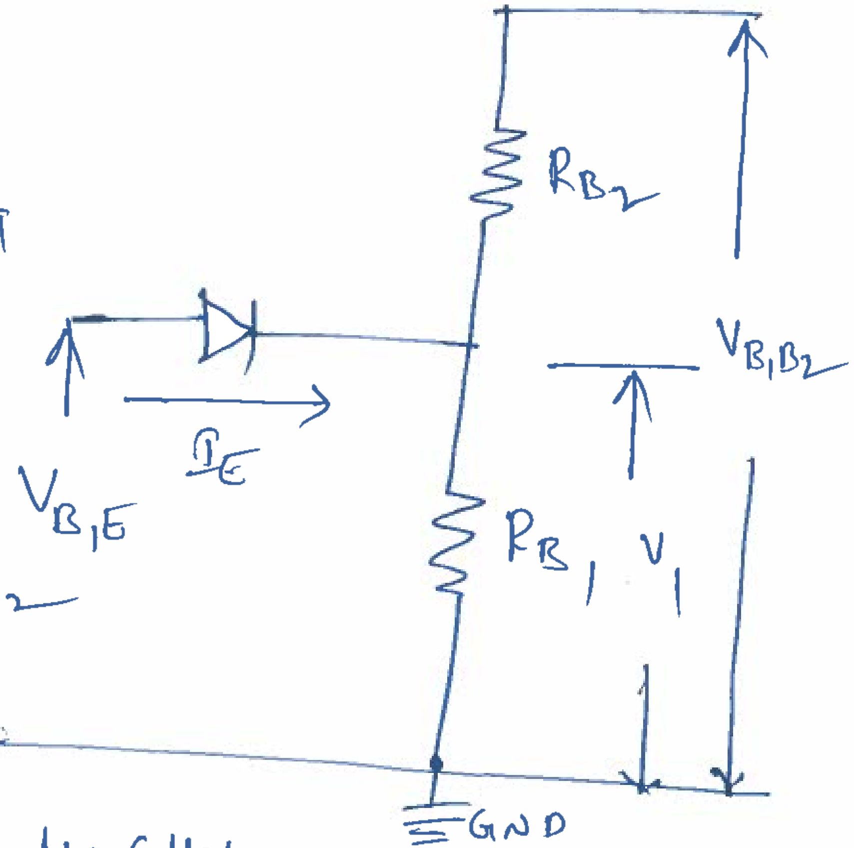
E - Emitter

B_1 - Base 1

B_2 - Base 2

→ The emitter reactance value is repiddly decreases

equivalent ckt diagram of UJT



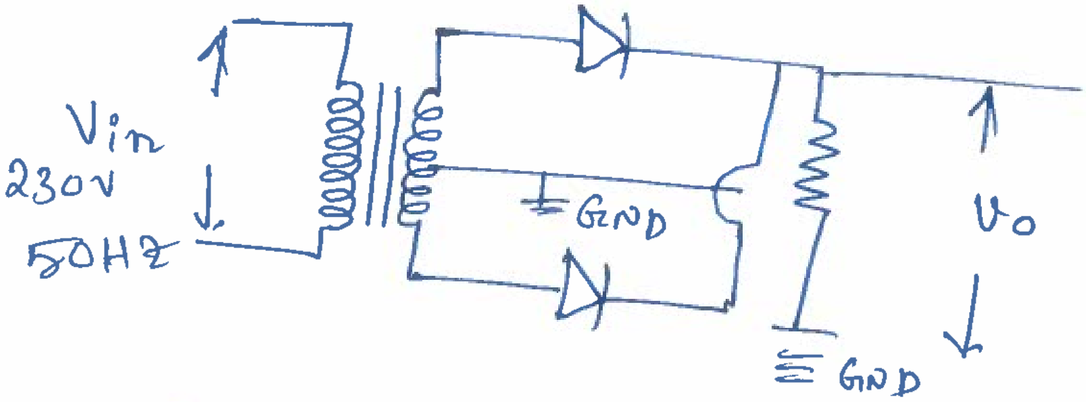
$$\text{Voltage } (V_1) = \frac{V_{B_1B_2} \times R_{B_1}}{R_{B_1}+R_{B_2}}$$

where $R_{BB} = R_{B_1}+R_{B_2}$

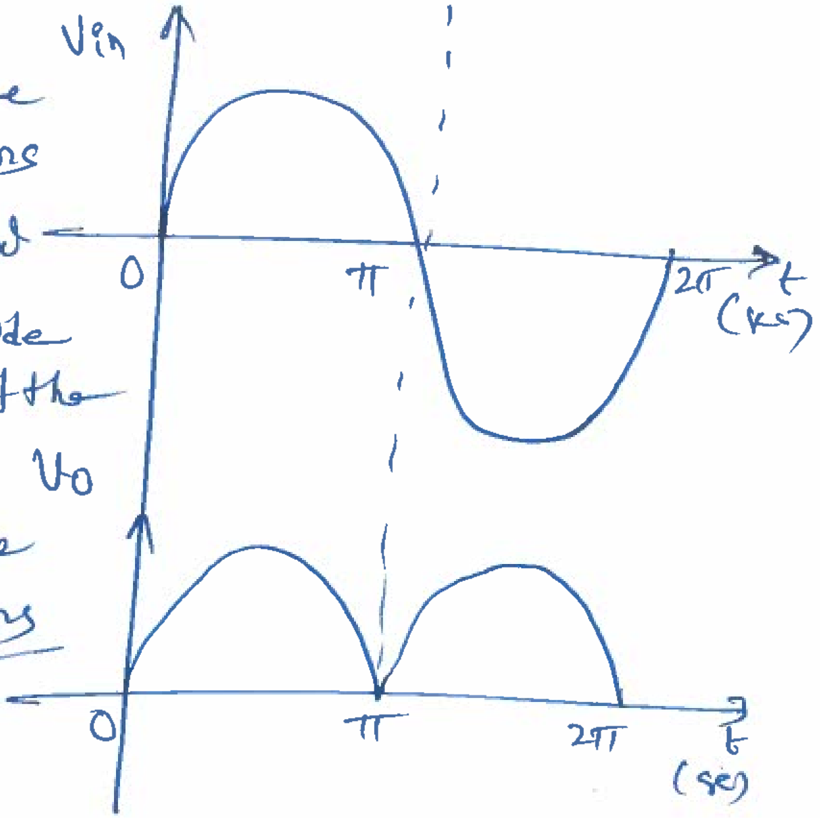
the unijunction transistor also Called as double base diode) Consistes of lightly doped n-type silicon with a small heavily doped P-type material joined to one side the end terminal of the bar are designated base (B_1) and Base (B_2).

9(a) Explain the operation of Fullwave Rectifier with necessary graphs.

which converts A·C voltag/current in to pulsating D·C voltage/current is called Fullwave Rectifier. In this rectifier, we can using two diodes, which is used to conducting the +ve & -ve cycles and also used there center taped Transformer stepdown transformer used to generate a input signals.



input/out put graph's & wave forms:



when D_1 conduct's the positive cycle of the input signal with respective cathode, and diode D_2 conducts the negative cycle of the input signal with respect Diode D_1 dosn't conduct.

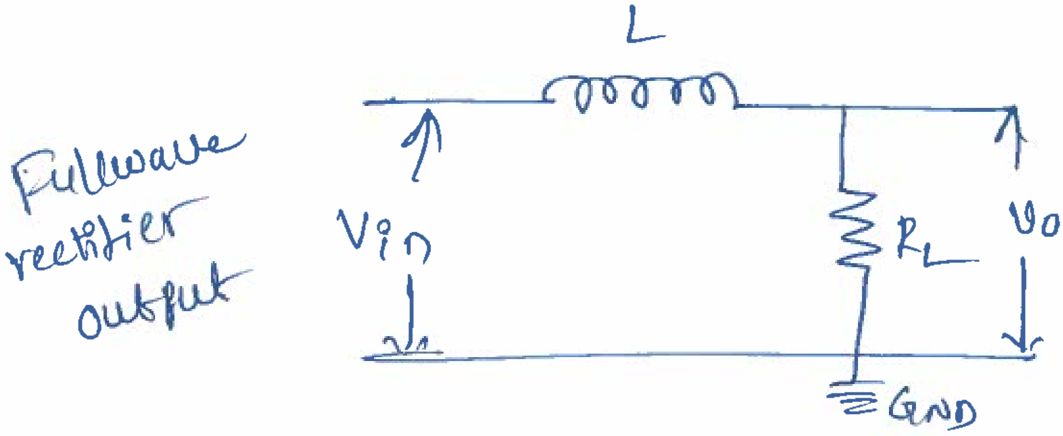
∴ Efficiency(η) = 81·2%.

ripple factor(γ) = 0·48.

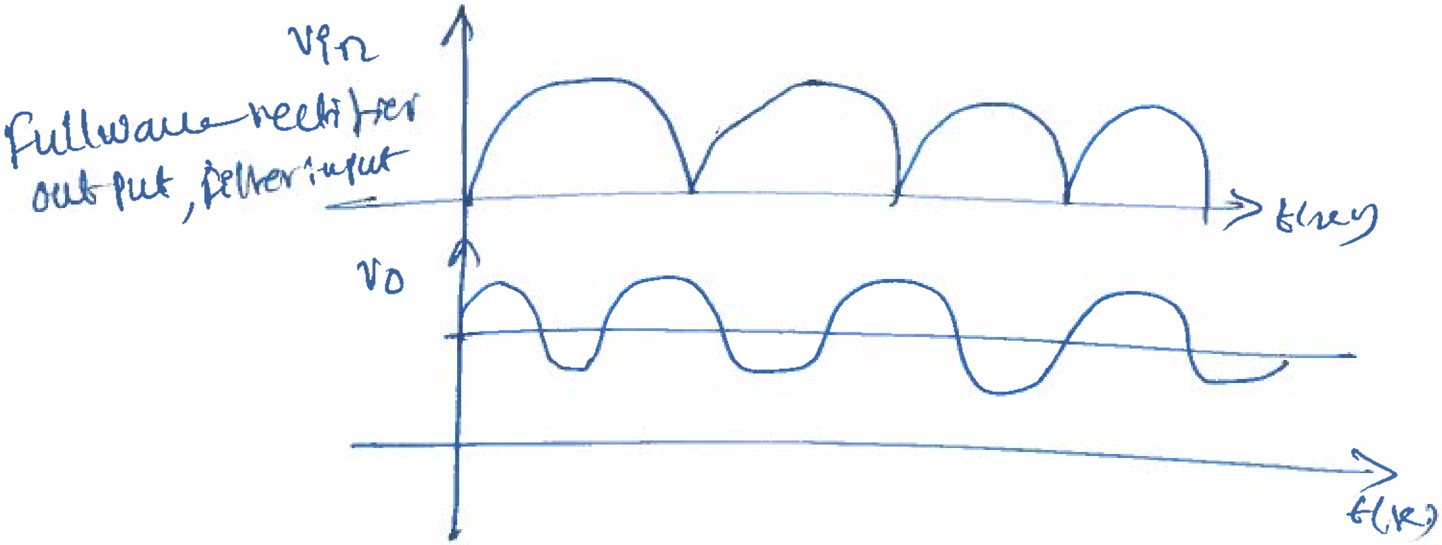
9(b) Explain the operation of (i) inductor Filter (ii) Capacitor Filter

when the output of the fullwave rectifier is input of the filter, which can removes the unwanted noise signal, this process is called filter.

(i) <u>inductor filter</u>: — In this filter inductor is connected with load resistance (R_L), the input of the inductor filter is fullwave rectifier output.

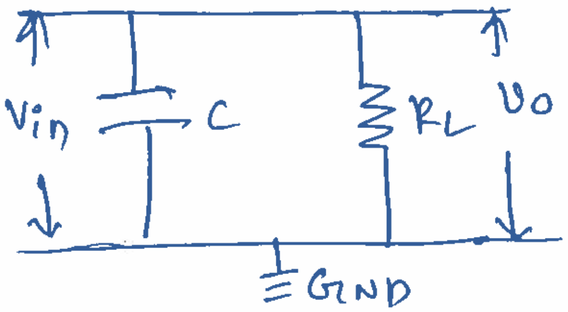


Input output wave forms :-



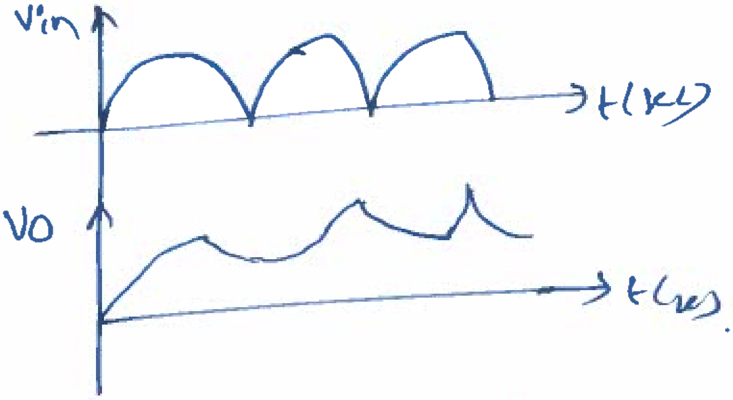
(iii) Capacitor Filter:-

the operation of Capacitor filter which blocks the DC voltage input of the Capacitor Filter



fullwave rectifier output is input of the capacitor filter.

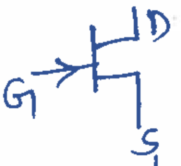
input output wave forms:-



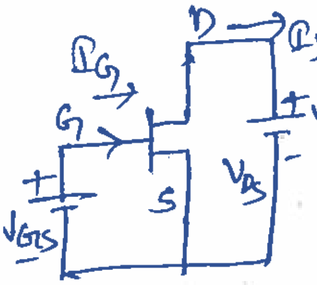
10(a) Explain the drain and transfer characteristics of a N-channel JFET:-

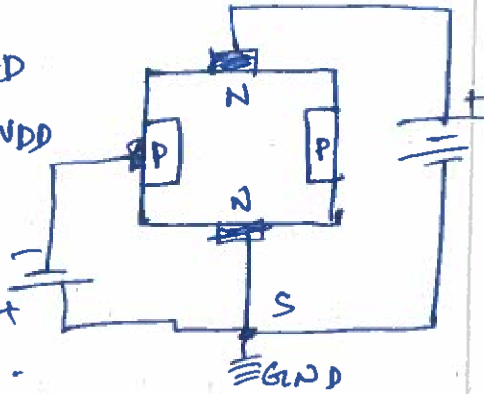
10(a) Drain and transfer characteristics of N-channel JFet:–

D-Drain, G-Gate,
S-Source

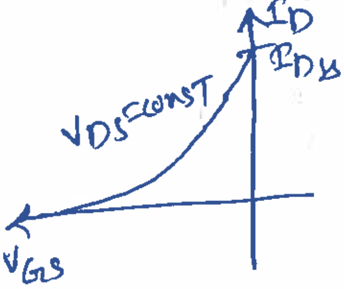


N-channel JFet

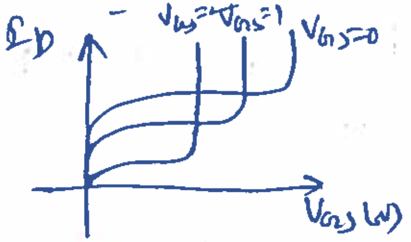




Transfer characteristics:–



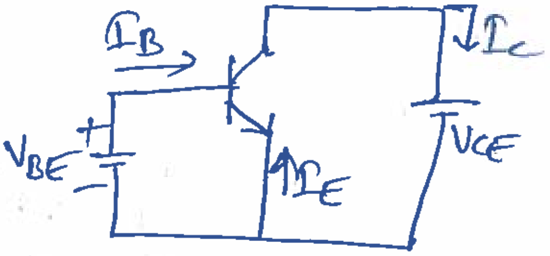
Drain characteristics.



→ which is used to control the voltage

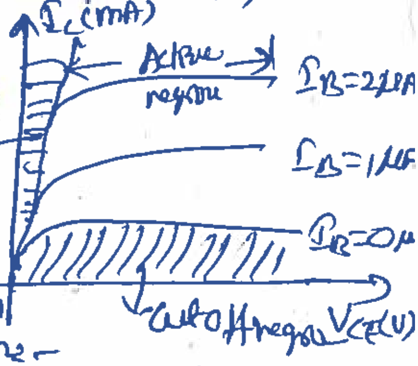
→ . this is voltage variable resistor.

10(b)



In this configuration Emitter is common to the both input and output ends.

Output characteristics:– the graph drawn between output current (I_C) and collector-Emitter voltage (V_{CE}), at constant input current (I_B)



(i) Active region: In this region transistor Act as a Amplifier

(ii) Saturation region:–
In this region transistor $\beta = \frac{I_C}{I_B}$.
acts as a On switch

(iii) Cut off region:– In this region transistor Acts as off switch.

11(a):– relation between α & β:–

$$\alpha = \frac{\Delta I_C}{\Delta I_E}, \quad \beta = \frac{\Delta I_C}{\Delta I_B}$$

$$\alpha = \frac{\Delta I_C}{\Delta I_B + \Delta I_C} \longrightarrow ① \qquad \because \Delta I_E = \Delta I_B + \Delta I_C$$

eq ① with $\frac{\Delta I_B}{\Delta I_B}$

$$\alpha = \frac{\Delta I_C/\Delta I_B}{\cancel{\Delta I_B}/\cancel{\Delta I_B} + \Delta I_C/\Delta I_B}$$

$$\alpha = \frac{\Delta I_C/\Delta I_B}{1 + \Delta I_C/\Delta I_B} \Rightarrow \boxed{\alpha = \frac{\beta}{1+\beta}}$$

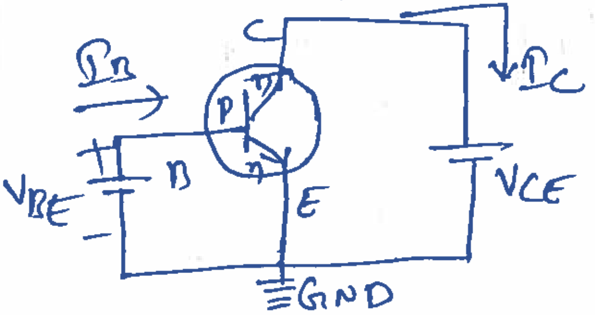
$$\beta = \frac{\Delta I_C}{\Delta I_B} = \frac{\Delta I_C}{\Delta I_E - \Delta I_C} \longrightarrow ②$$

eq ② with $\Delta I_C/\Delta I_C$

$$\beta = \frac{\Delta I_C/\Delta I_C}{\Delta I_E/\Delta I_C - 1} = \frac{1}{\frac{1}{\alpha} - 1}$$

(11)(b) n-p-n BJT CE configuration:–

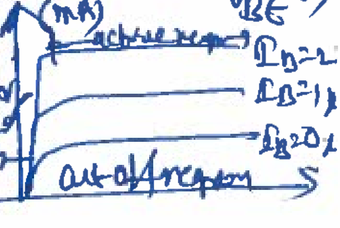
In this configuration emitter is common to the both the input and output



Input characteristics:– the graph drawn between input current (I_B) and input voltage (V_{BE}) at that time output voltage must be constant

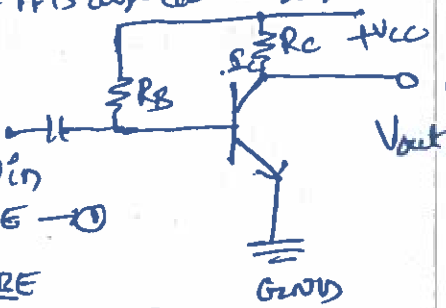
Output characteristics:–

The graph drawn between output current (I_C) and output voltage (V_{CE}) at



2(a) Expression of stability factor for

<u>fixed bias</u>:- it is also called base resistor method



$V_{CC} = I_B R_B + V_{BE}$ → ①

$I_B = \frac{V_{CC} - V_{BE}}{R_B}$ → ②

both sides differentiating with $\frac{d}{dI_C}$

$\frac{dI_B}{dI_C} = 0$

Stability factor $S = \frac{1+\beta}{1-\beta\left(\frac{dI_B}{dI_C}\right)}$

$S = \frac{1+\beta}{1-\beta(0)}$

$\boxed{\therefore S = 1+\beta}$

13(a) Given that

$V_{CC} = 9V$, $R_C = 3k\Omega$, $R_B = 8k\Omega$

$\beta = 50$, $V_{BE} = 0.7V$

$I_C = ?$, $V_{CE} = ?$, $S = ?$

operating points: $\boxed{I_C = V_{CC}/R_C}$

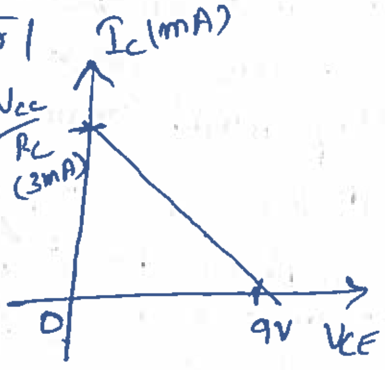
$I_C = \frac{9}{3 \times 10^3}$

$I_C = 3mA$

$V_{CE} = V_{CC} = 9V$

$S = 1+\beta = 1+50$

$S = 51$



Dc load line

operating points

(I_C & V_{CE})

= (3mA, 9V)

12(b) What is biasing:- the proper flow of zero signal collector current and the maintainance of proper collector current and Emitter voltage during the passing of signal is known as transistor biasing.

$I_C = \beta I_B + I_{CO}(1+\beta)$

there are 3 types :-

(i) Base resistor method

(ii) collector feed back method

(iii) voltage divider (or) self bias method.

13(b): there are two types of compensation techniques:.

14(a) Given that

Common source Amplifier

drain resistance $(r_d) = 100k\Omega$

load resistance $(R_L) = 10k\Omega$

mutual conductance $(g_m) = 300\mu$

Output Resistance $(R_o) = 9.09k\Omega$

Capacitance $(C_{ds}) = 3PF$

output impedence $(Z_o) = ?$

at freuey $f = 1MHz$

Voltage gain $(A_v) = ?$

useful formula's

$$A_v = \frac{V_o}{V_i} = \frac{V_o}{V_{gs}} = -\frac{\mu R_D}{R_D + r_d} = \frac{-g_m r_d R_D}{R_D + r_d}.$$

$$i_d = \left(\frac{r_d}{R_D + r_d}\right) \times g_m V_{gs}.$$

$$V_o = -i_d R_D = -\left(\frac{r_d}{R_D + r_d}\right) \times (g_m) \times R_D \times V_{gs}$$

$$r_L = \left(\frac{r_d R_D}{R_D + r_d}\right),$$

$$\boxed{A_V = -g_m r_L}$$

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech. (U. G.)	**Program**	CSE			**Academic Year**	2022 - 2023
Course Code	20CS302	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Design and Analysis of Algorithms**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define time complexity.	20CS302.1	L1
2	Write the basic principle of Divide and Conquer strategy	20CS302.2	L2
3	State the principle of optimality.	20CS302.3	L1
4	What is NP-Hard?	20CS302.4	L2
5	Define State space tree.	20CS302.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain Asymptotic Notations with examples and graphs.	12M	20CS302.1	L2
	OR			
7 (a)	Discuss the steps in mathematical analysis for recursive algorithm. Do the same for finding the factorial of a number	8M	20CS302.1	L2
7 (b)	What is an algorithm? List the 3 characteristics of algorithm.	4M	20CS302.1	L1
8 (a)	Sort the records with the following index values in the ascending order using quick sort algorithm. 2, 3, 8, 5, 4, 7, 6, 9, 1.	6M	20CS302.4	L3
8 (b)	Write Merge Sort algorithm.	6M	20CS302.5	L2
	OR			
9 (a)	Construct a maxHeap with 2, 3, 8, 5, 4, 7, 6, 9, 1.	8M	20CS302.4	L3
9 (b)	Write Strassen's Matrix Multiplication algorithm.	4M	20CS302.5	L2
10	State the Job – Sequencing with deadlines problem. Find an optimal sequence to the n = 5 Jobs where profits (P1, P2, P3, P4, P5) = (20, 15, 10, 5, 1) and deadlines (d1, d2, d3, d4, d5) =(2, 2, 1, 3, 3).	12M	20CS302.5	L3
	OR			
11 (a)	Differentiate between greedy method and dynamic programming	4M	20CS302.3	L2
11 (b)	Describe the 0/1 Knapsack Problem. Find an optimal solution using dynamic programming 0/1 knapsack instance for n=3, m=6, profits are (p1, p2,p3) = (1,2,5), weights are (w1,w2,w3)=(2,3,4).	8M	20CS302.4	L3
12 (a)	Explain class of P, NP, NP complete problems	6M	20CS302.5	L2
12 (b)	Explain Cook's theorem	6M	20CS302.5	L2
	OR			
13 (a)	Explain the non-deterministic sorting problem.	6M	20CS302.5	L2
13 (b)	Differentiate between NP-complete and NP-Hard.	6M	20CS302.5	L2
14	Construct complete state space tree for the subset sum problem for n = 5, d = 15, S = {1, 3, 5, 7, 11}	12M	20CS302.3	L3
	OR			
15 (a)	Explain 8-queen's problem and apply back tracking to solve this problem	8M	20CS302.4	L3
15 (b)	Write the Graph – coloring problem. And draw the state space tree for m= 3 colors n=4 vertices graph.	4M	20CS302.4	L3

1

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ANSWER KEY AND SCHEME OF EVALUATION

Course code : 20CS302   Acy : 2022-23.
Course Title : Design and Analysis of algorithms

Part A.

1. Time complexity is the total amount of CPU time taken by the algorithm to give output.

2. Divide and conquer strategy divides the problem into sub problems and solves the subproblems individually and integrates the solutions to get solution to given problem

3. Principle of optimality states that if there exists a sub optimal solution to subproblem, optimal solution is guranteed.

4. NP Hard is a class of problems for which the verification can be done in polynomial time with deterministic algorithm.

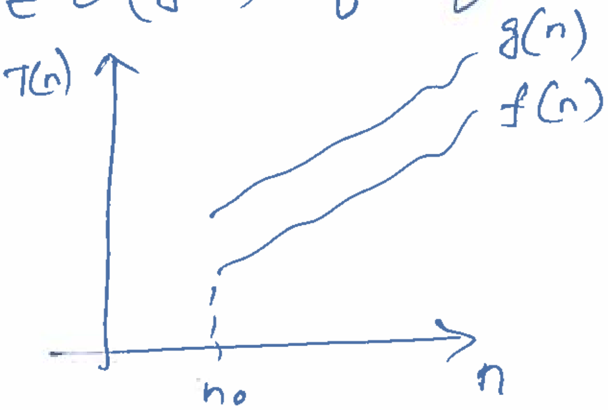
5. State space Tree is the tree constructed during construction of solutions by BT and BB algorithms. Each node indicates a decision and root represents empty solution. Solution is at least level

Part B.

6. Asymptotic notations

1. Big oh (4M)

$f(n) \in O(g(n))$ if $f(n) \le c . g(n)$

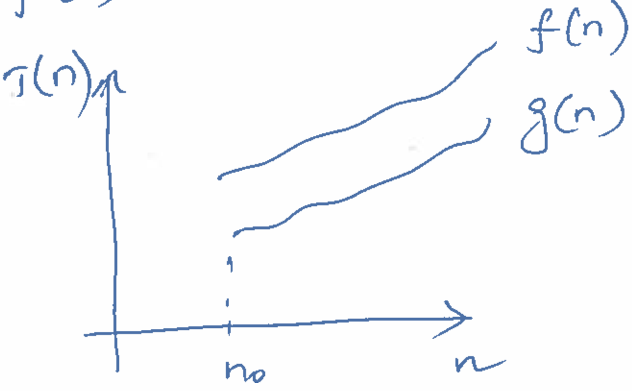


2. Big theta (4M)

$f(n) \in \Theta(g(n))$ if $C_1 . g(n) \le f(n) \le C_2 . g($

3. Big Omega (4M)

$f(n) \in \Omega(g(n))$ if $f(n) \ge c . g(n)$



7.a) Analysis of Recursive algorithm

1. Decide on input size n (4M)
2. Identify basic operation
3. Count on basic operation
4. Check if the count depends on input. If so find best case, worst case, average case analysis
5. Count of basic operation is a recurrence relation; Solve it to find time complexity

Algorithm : RFact (n) (4M)

```
begin
        if n==0 return 1
        else return (n * RFact(n-1))
end
```

Analysis

input size = n

basic operation = multiplication

$M(n) = M(n-1) + 1 \quad , \quad n > 0$

$M(0) = 0$

$M(n) = M(n-1) + 1$

$M(n-1) = M(n-2) + 1$

Using it back, $M(n) = M(n-2) + 2$

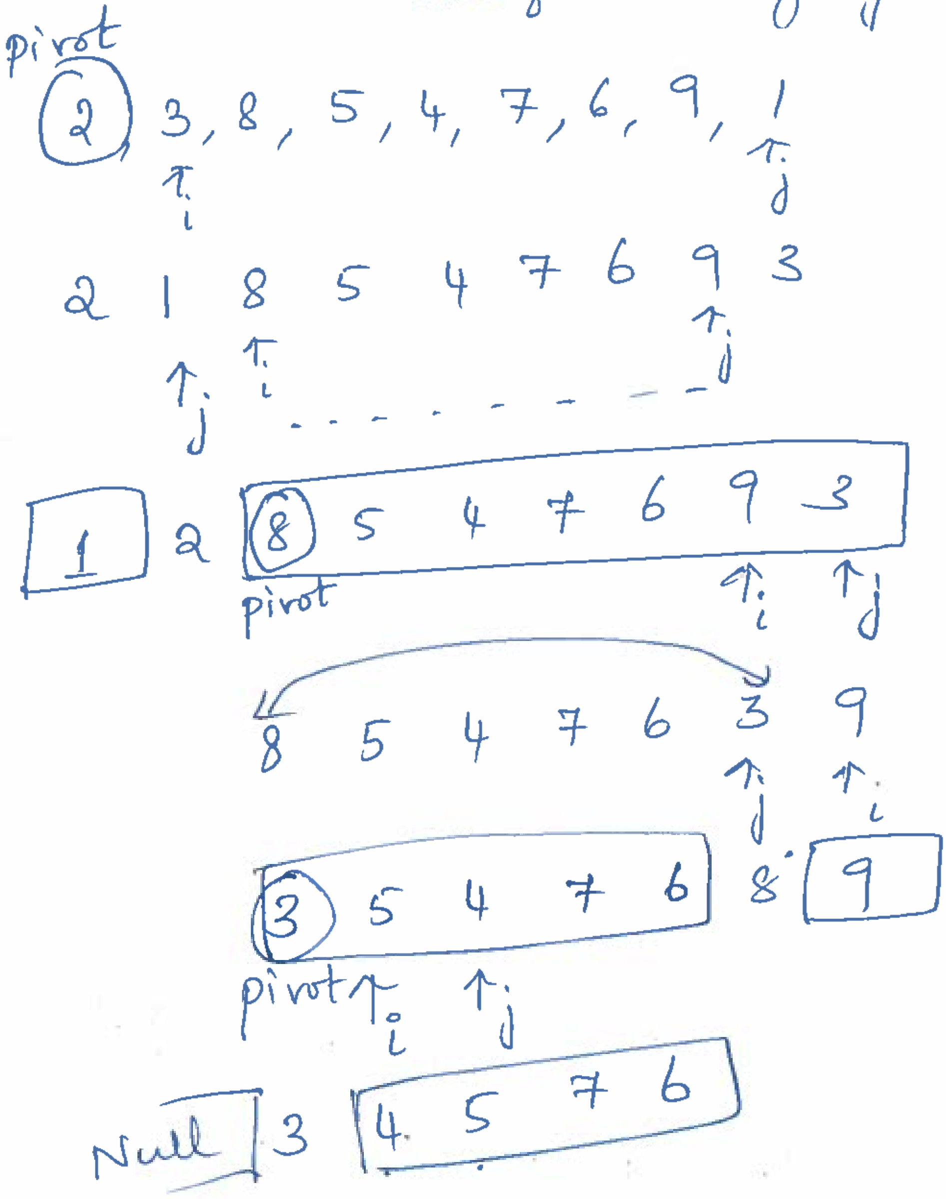
In general, $M(n) = M(n-i) + i \quad n > 0$

Put $i = n \Rightarrow M(n) = n \Rightarrow \boxed{T(n) = n}$

7. b. Algorithm: Step-by-step procedure designed to solve a problem (2M)

Characteristics Definiteness (2M)
Finiteness
Free from ambiguity

8. a)



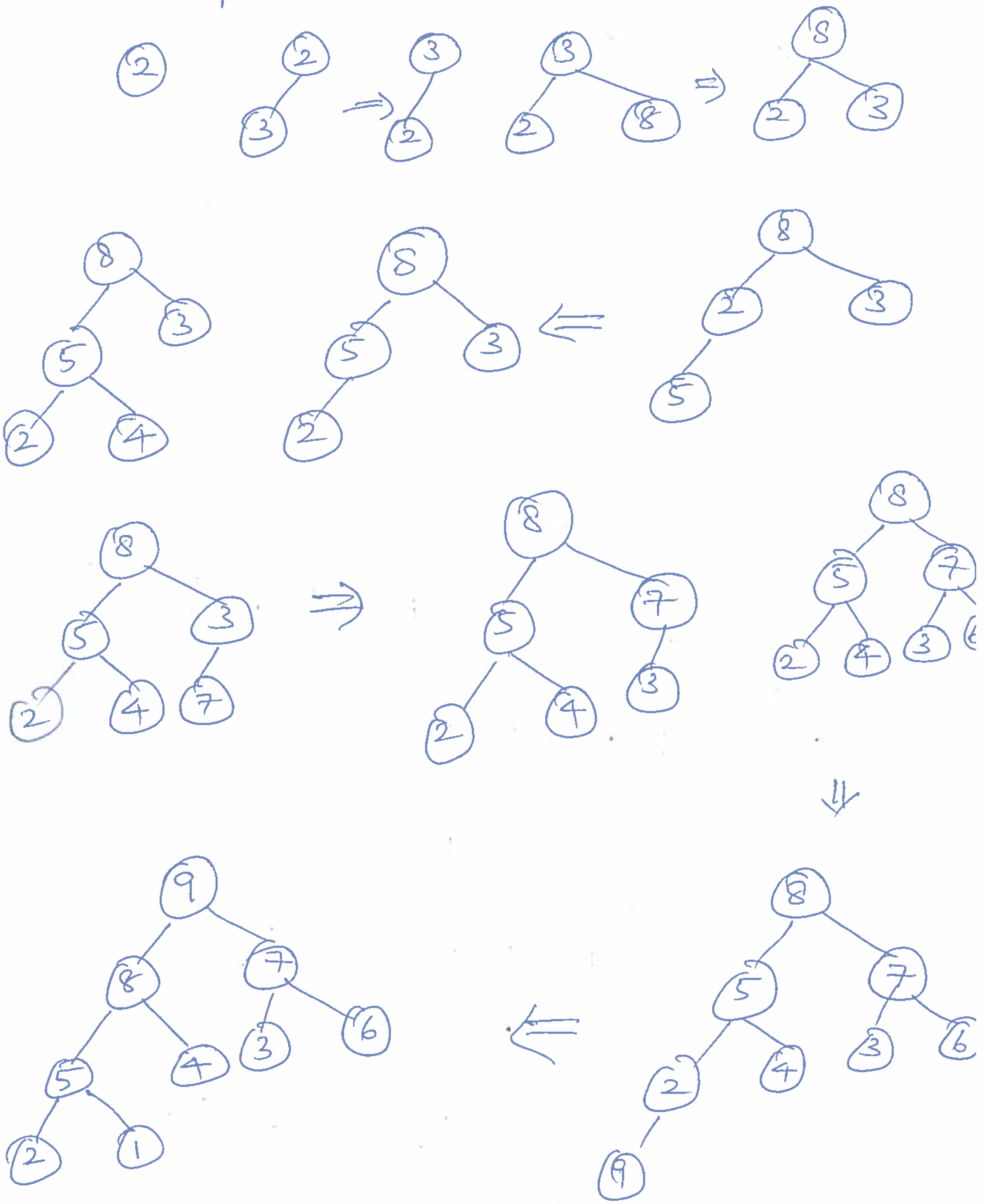
⋮

1 2 3 4 5 6 7 8 9.

8. b. Mergesort algorithm. ⑤

```
Algorithm: Mergesort (A, n, l, r)                    (3M)
begin
    if (l < r)
        copy A(0..n/2) to B
        copy B(n/2+1...n) to C
        Mergesort (B, n/2, 0, n/2 ~~-1~~)
        mergesort (C, n/2, n/2+1, n)
        merge (B, C)
end
Algorithm: Merge(B, C, p, q)                         (3M)
begin
    ~~i= pivot =~~
    i=0; j=0; k=0;
    ~~do while B[i] <~~
    do
    {
        if B[i] < C[j]
            D[k] = B[i]; i++
        else
            D[k] = C[j]; j++
          k=k+1
    while(i<p or j<q)
      if (i < p) copy remaining from B to D
      if (j<q) copy remaining from C to D
end.
```

9. a) Maxheap 2, 3, 8, 5, 4, 7, 6, 9, 1

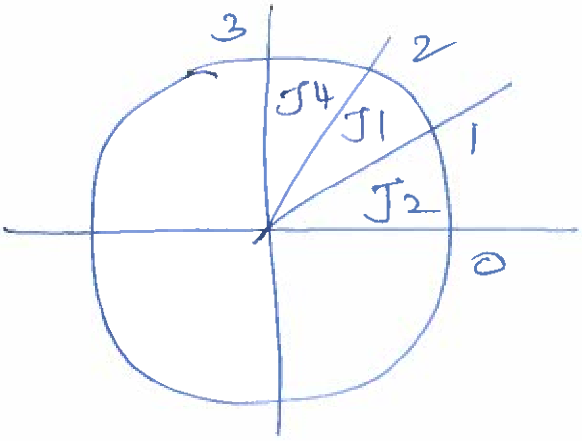


9. b) Strassen Matrix Multiplication

* divide matrix into 2 sections
* Repeatedly do this until a singleton matrix is formed
* Multiply smallest matrix and start to combine solution

10. $n = 5$

	J1	J2	J3	J4	J5
D	2	2	1	3	3
P	20	15	10	5	1



∴ < J2 J1 J4 > with profit $20+15+5 = 40$.

11 a) Greedy method	Dynamic Programming
→ Constructs solution in phases	→ repeated subproblems
→ takes one greedy decision at a time and finds optimal solution	→ solves smallest subproblem; records solution; uses it to find solution to given problem.

b) n=3 m=6 P: (1,2,5)
W: (2,3,4)

		0	1	2	3	4	5	6
	0	0	0	0	0	0	0	0
(1,2)	1	0	0	1	1	1	1	1
(2,3)	2	0	0	1	2	2	3	3
(5,4)	3	0	0	1	2	5	3	$\boxed{6}$

Subset : { i1, i3 }

profit : 6

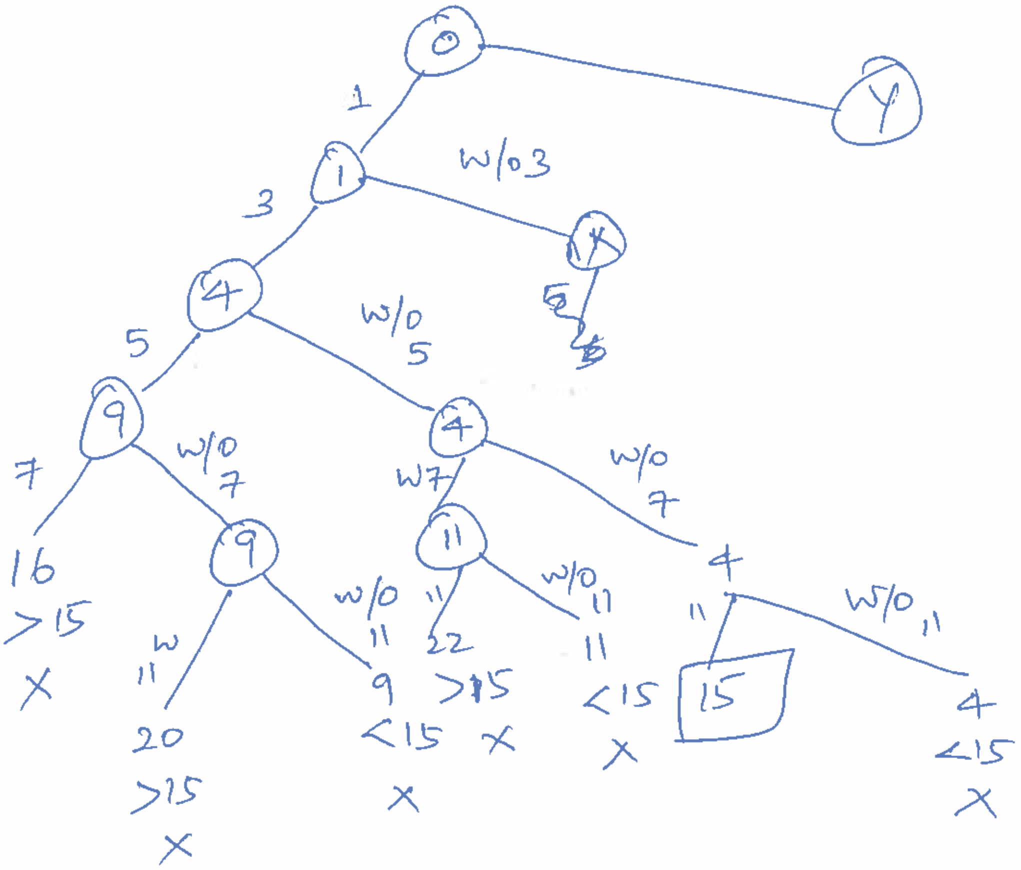
12 a) Definitions P (2)

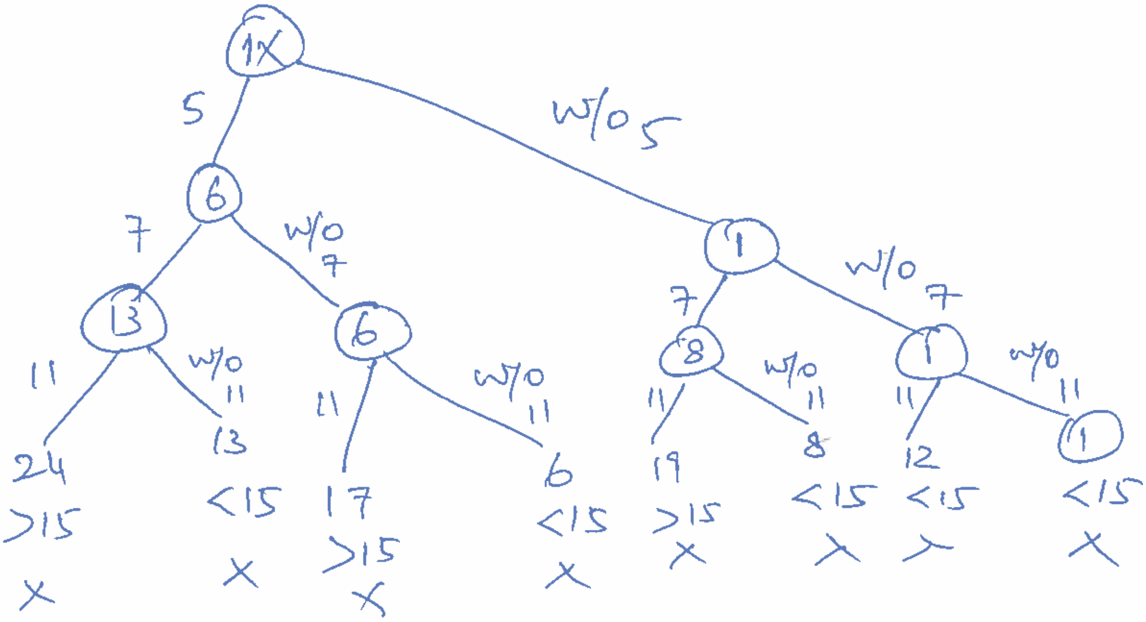
NP (2)

NP Complete (2)

b) Cook's theorem

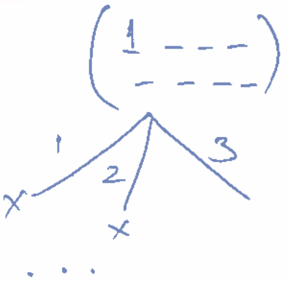
14. $n=5$ $d=15$ $S=\{1,3,5,7,11\}$



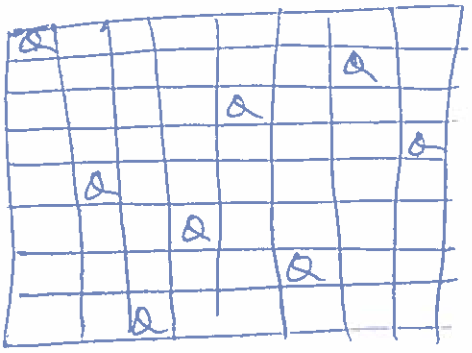


15/a 8 Queen's problem

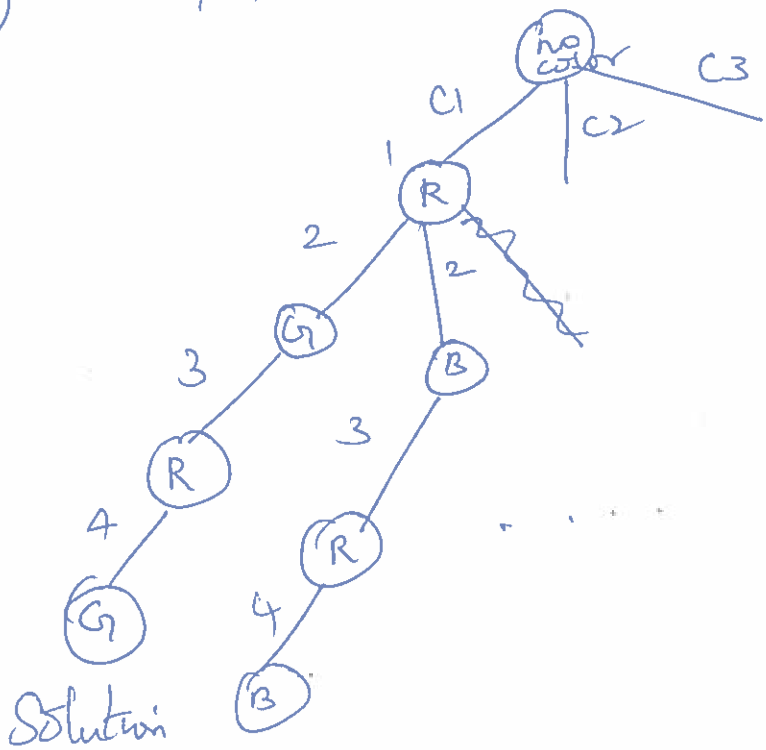
- to place 8 queens, one in each row on a 8×8 board s.t no 2 queens get attacking positions
- can be solved using backtracking

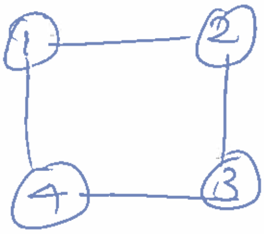


Solution 1



b) n=4, m=3 Graph coloring problem





Semester End Regular/Supplementary Examination, Dec./Jan., 2022/2023

Degree	B. Tech. (U. G.)	**Program**	CSE (AI & ML)			**Academic Year**	2022 - 2023
Course Code	20AI302	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Artificial Neural Networks**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What are the Learning Rules in Neural Network?	20AI302.1	L1
2	What is the difference between single layer and multilayer feedforward networks?	20AI302.2	L1
3	What is the function of Linear Least-Square Filters?	20AI302.3	L1
4	Where are the convolutional networks used?	20AI302.4	L1
5	How the Hetero Associative memory differ from Auto Associative memory?	20AI302.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Discuss about the building blocks of ANN.	12M	20AI302.1	L1
	OR			
7	Discuss about the basic learning laws.	12M	20AI302.1	L1
8	What is McCulloh Pitts neuron model? Draw the diagrammatic representation of the model. Generate the output of OR and NOR function using the model.	12M	20AI302.2	L2
	OR			
9	What is supervised and unsupervised learning? How the unsupervised learning works? List any 4 differences between supervised and unsupervised learning.	12M	20AI302.2	L2
10	Define learning rate. What is learning rate annealing? Discuss about the methods of learning rate annealing.	12M	20AI302.3	L2
	OR			
11	What is Gaussian Bayes classifier? Discuss about the relation between the Perceptron and Bayes Classifier for a Gaussian Environment.	12M	20AI302.3	L2
12	What do you mean by XOR problem? Discuss about the Heuristics for Making the Back-Propagation Algorithm Perform Better.	12M	20AI302.4	L2
	OR			
13 (a)	Discuss the back propagation and differentiation.	6M	20AI302.4	L2
13 (b)	Discuss the cross validation.	6M	20AI302.4	L2
14	What is content addressable memory? Draw the architecture of auto associative memory network. Write the steps of testing algorithm for auto associative memory.	12M	20AI302.5	L2
	OR			
15	What do you mean by pattern association? What is the purpose of bi-directional associative memory? Draw the architecture and discuss the algorithm of BAM.	12M	20AI302.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

COURSE: ARTIFICIAL NEURAL NETWORKS
CODE:20AI302
III-SEM B.TECH CSE(AI/ML)

	PART-A	
1	**What are the Learning Rules in Neural Network?** Learning rule or Learning process is a method or a mathematical logic. It improves the Artificial Neural Network's performance and applies this rule over the network. Thus, learning rules updates the weights and bias levels of a network when a network simulates in a specific data environment. Applying learning rule is an iterative process. It helps a neural network to learn from the existing conditions and improve its performance.	2M
2	**What is the difference between single layer and multilayer feedforward networks?** **Single Layer Feedforward Neural Network:** • It consists of only input and output layer • In this network signals travel in only one direction. **Multilayer Feedforward Neural Network:** • It consists of input, output layer and more than 2 hidden layers. • In this network also signals travel in only one direction.	2M
3	**What is the function of Linear Least-Square Filters?** $W(n+1) = X(n). d(n)$	2M
4	**Where are the convolutional networks used?** Convolutional neural networks are usually used for visual imagery, helping the computer identify and learn from images, image processing and for other autocorrelated data.	2M
5	**How the Hetero Associative memory differ from Auto Associative memory?** **Auto-associative Networks:** • Auto assosiative networks are special kind of networks used to simulate associative processes. • They are capable to retrieve piece of data with the partial information and also capable for remembering from small portion of data. • The input and output both will be same. **Hetero-associative Networks:** • Hetero associative networks stores input-output pattern pairs to recall stored output pattern by receiving noisy or incomplete version. • In each of the pairs, an input pattern should differ from an output pattern.	2M
	PART-B	
6	**Discuss about the building blocks of ANN.** Processing of Artificial neural network depends upon the given three building blocks: • Network Topology • Adjustments of weights or learning • Activation functions **1. Network Topology:** The topology of a neural network refers to the way how Neurons are associated, and it is a significant factor in network functioning and learning. It consists of 3 types:	12M

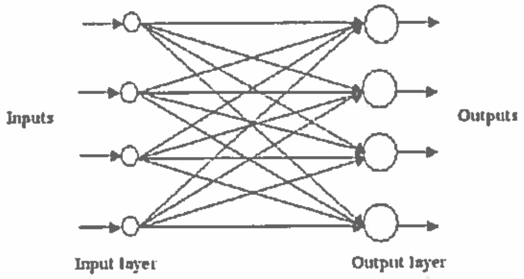
Feedforward Neural Network:

In this neural network signals travel in only one direction.

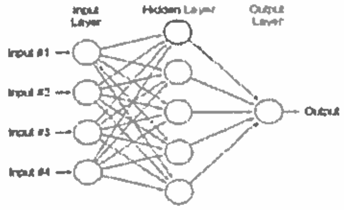
This neural network consists of two categories:

- Single layer neural network- It consists of only input layer and output layer.
- Multilayer neural network- It consists of input, output layers and hidden layers.

Singlelayer feed forward network



Multilayer feed forward network



Can be used to solve complicated problems

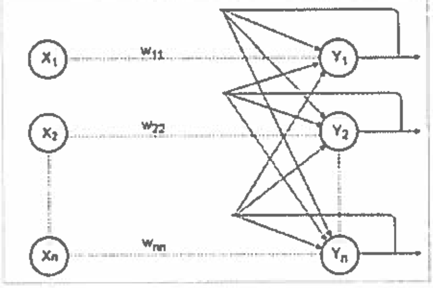
Feedback Neural Network:

In this neural network signals travel in both the directions.

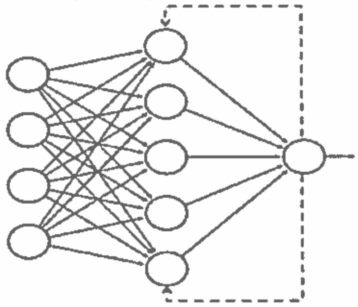
These are also known as recurrent neural network.

This neural network consists of two categories:

- Single layer neural network- It consists of only input layer and output layer.



- Multilayer neural network- It consists of input, output layers and hidden layers.



2. **Adjustments of Weights or Learning:**

	Learning in ANN is the technique for changing the weights of associations between the neurons of a specified network. Learning in artificial neural networks can be characterized into two different categories, namely supervised learning, unsupervised learning, and reinforcement learning. • Supervised learning- The model is learned using labelled data. • Unsupervised learning- The model is learned using unlabelled data. Weight Adjustment: If the neural network has any error, then the error will be rectified with the help of backpropagation algorithm where the paramenters like weight and bias will be adjusted and then the model has to be retrained. **3. Activation Functions:** The activation functions are used to get the final output of the neural network. These are of two types, linear and non-linear activation function. **Linear Activation Function:** The equation of the linear activation function is the same as the equation of a straight line i.e. Y= MX+ C **Non-linear activation functions:** These are of the following: • Binary activation function • Bipolar activation function • Sigmoid activation function • ReLU activation function	
7	**Discuss about the basic learning laws.** • **Hebbian learning rule** – It identifies, how to modify the weights of nodes of a network. The Hebbian rule was the first learning rule. In 1949 *Donald Hebb* developed it as learning algorithm of the unsupervised neural network. We can use it to identify how to improve the weights of nodes of a network. The Hebb learning rule assumes that – If two neighbor neurons activated and deactivated at the same time. Then the weight connecting these neurons should increase. For neurons operating in the opposite phase, the weight between them should decrease. If there is no signal correlation, the weight should not change. • **Perceptron learning rule** – Network starts its learning by assigning a random value to each weight. As you know, each connection in a neural network has an associated weight, which changes in the course of learning. According to it, an example of supervised learning, the network starts its learning by assigning a random value to each weight.Calculate the output value on the basis of a set of records for which we can know the expected output value. This is the learning sample that indicates the entire definition. As a result, it is called a learning sample.The network then compares the calculated output value with the expected value. Next calculates an error function ∈, which can be the sum of squares of the errors occurring for each individual in the learning sample. • **Delta learning rule** – Modification in sympatric weight of a node is equal to the multiplication of error and the input. Developed by *Widrow* and *Hoff*, the delta rule, is one of the most common learning rules. It depends on supervised learning.This rule states that the modification in sympatric weight of a node is equal to the multiplication of error and the input. • **Correlation learning rule** – The correlation rule is the supervised learning. The correlation learning rule based on a similar principle as the Hebbian learning rule. It assumes that weights between responding neurons should be more positive, and weights between neurons with opposite reaction should be more negative.Contrary to the Hebbian rule, the correlation rule is the supervised learning. Instead of an actual response, oj, the desired response, dj, uses for the weight-change calculation. • **Outstar learning rule** – We can use it when it assumes that nodes or neurons in a network arranged in a layer. We use the Out Star Learning Rule when we assume that nodes or neurons in a network arranged in a layer. Here the weights connected to a certain node should be equal to the desired outputs for the neurons connected through those weights. The out-star rule produces the desired response t for the layer of n nodes.	12M
9	**What is supervised and unsupervised learning? How the unsupervised learning works? List any 4 differences between supervised and unsupervised learning.** • **Supervised learning-** The model is learned using labelled data. • **Unsupervised learning-** The model is learned using unlabelled data.	12M

Unsupervised learning working:

Unsupervised learning, also known as unsupervised machine learning, uses machine learning algorithms to analyze and cluster unlabeled datasets. These algorithms discover hidden patterns or data groupings without the need for human intervention. Its ability to discover similarities and differences in information make it the ideal solution for exploratory data analysis, cross-selling strategies, customer segmentation, and image recognition. Unsupervised learning models are utilized for three main tasks—clustering, association, and dimensionality reduction

Supervised Learning	Unsupervised Learning
Supervised learning algorithms are trained using labeled data.	Unsupervised learning algorithms are trained using unlabeled data.
Supervised learning model takes direct feedback to check if it is predicting correct output or not.	Unsupervised learning model does not take any feedback.
Supervised learning model predicts the output.	Unsupervised learning model finds the hidden patterns in data.
In supervised learning, input data is provided to the model along with the output.	In unsupervised learning, only input data is provided to the model.
The goal of supervised learning is to train the model so that it can predict the output when it is given new data.	The goal of unsupervised learning is to find the hidden patterns and useful insights from the unknown dataset.
Supervised learning needs supervision to train the model.	Unsupervised learning does not need any supervision to train the model.
Supervised learning can be categorized in Classification and Regression problems.	Unsupervised Learning can be classified in Clustering and Associations problems.

10 **Define learning rate. What is learning rate annealing? Discuss about the methods of learning rate annealing.** 12M

Learning Rate- It determines the size of the jumps your model makes and, as a result, how rapidly it learns.

Learning Rate Annealing- Changing the learning rate for your stochastic gradient descent optimization technique can improve performance while also cutting down on training time. This is also known as adaptable learning rates or learning rate annealing. This method is referred to as a learning rate schedule since the default schedule updates network weights at a constant rate for each training period.

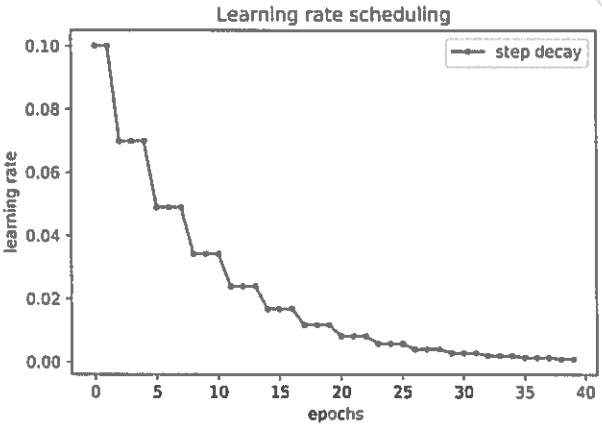
Techniques that reduce the learning rate over time are the simplest and arguably most commonly used modification of the learning rate during training. These have the advantage of making big modifications at the start of the training procedure when larger learning rate values are employed and decreasing the learning rate later in the training procedure when a smaller rate and hence smaller training updates are made to weights.

Methods of Learning Rate Annealing

1. **A Learning Rate Decay**

The way things will evolve throughout time is defined by a schedule. The learning rate schedule, in general, defines a learning rate for each epoch and batch. For scheduling global learning rates, there are two sorts of methods: decay and cyclical. The learning rate annealing approach, which is scheduled to progressively decay the learning rate during the training process, is the most popular method.

In order to get a stronger generalization effect, a somewhat big step size is preferred in the early stages of training. The stochastic noise is reduced when the learning rate decreases. This helps the algorithm converge by avoiding oscillation at the ideal spot.



2. Adaptive Learning Rate

It is desirable to automatically calculate the step size in gradient-based optimization based on the loss gradient that indicates the convergence of each of the unknown parameters. To this end, parameter-wise adaptive learning rate scheduling algorithms have been developed, such as AdaGrad, AdaDelta, RMSprop, and Adam, which allow the process to quickly converge in practice. The adaptive technique has recently been used to combine Adam with SGD, to automatically select learning rate methods, and to develop an efficient loss-based method.

In supervised learning, such as image classification using standard shallow models, the adaptive technique is frequently inferior to SGD in the accuracy for unknown data. Due to the benefits of generalization and training advantage, SGD with planned annealing outperforms adaptive approaches in practice. As a result, the hand-crafted schedule remains an important method for solving optimization challenges.

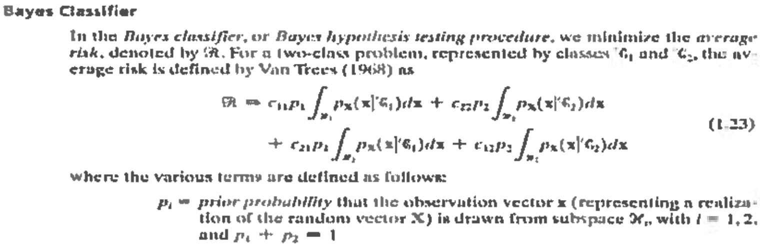
3. Learning-Rate Warmup

For example, the learning rate warmup is a relatively new method that uses a short step size at the start of the training. In the first few epochs, the learning rate is increased linearly or nonlinearly to a given value, and then it decreases to zero.

11 **What is Gaussian Bayes classifier? Discuss about the relation between the Perceptron and Bayes Classifier for a Gaussian Environment.** 12M

RELATION BETWEEN THE PERCEPTRON AND BAYES CLASSIFIER FOR A GAUSSIAN ENVIRONMENT :

The perceptron bears a certain relationship to a classical pattern classifier known as the Bayes classifier. When the environment is Gaussian, the Bayes classifier reduces to a linear classifier.This is the same form taken by the perceptron. However, the linear nature of the perceptron is not contingent on the assumption of Gaussianity. In this section, we study this relationship and thereby develop further insight into the operation of the perceptron. We begin the discussion with a brief review of the Bayes classifier.



Bayes Classifier

In the *Bayes classifier*, or *Bayes hypothesis testing procedure*, we minimize the *average risk*, denoted by $\mathcal{R}$. For a two-class problem, represented by classes $\mathcal{C}_1$ and $\mathcal{C}_2$, the average risk is defined by Van Trees (1968) as

$$\mathcal{R} = c_{11}p_1 \int_{\mathcal{H}_1} p_{\mathbf{X}}(\mathbf{x}|\mathcal{C}_1)d\mathbf{x} + c_{22}p_2 \int_{\mathcal{H}_2} p_{\mathbf{X}}(\mathbf{x}|\mathcal{C}_2)d\mathbf{x} + c_{21}p_1 \int_{\mathcal{H}_2} p_{\mathbf{X}}(\mathbf{x}|\mathcal{C}_1)d\mathbf{x} + c_{12}p_2 \int_{\mathcal{H}_1} p_{\mathbf{X}}(\mathbf{x}|\mathcal{C}_2)d\mathbf{x} \tag{1.23}$$

where the various terms are defined as follows:

p_i = *prior probability* that the observation vector $\mathbf{x}$ (representing a realization of the random vector $\mathbf{X}$) is drawn from subspace $\mathcal{H}_i$, with $i = 1, 2$, and $p_1 + p_2 = 1$

c_{ij} = cost of deciding in favor of class $\mathscr{C}_i$ represented by subspace $\mathscr{X}_i$ when class $\mathscr{C}_j$ is true (i.e., observation vector **x** is drawn from subspace $\mathscr{X}_j$), with $i, j = 1, 2$

$p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_i)$ = conditional probability density function of the random vector **X**, given that the observation vector **x** is drawn from subspace $\mathscr{X}_i$, with $i = 1, 2$.

The first two terms on the right-hand side of Eq. (1.23) represent *correct* decisions (i.e., correct classifications), whereas the last two terms represent *incorrect* decisions (i.e., misclassifications). Each decision is weighted by the product of two factors: the cost involved in making the decision and the relative frequency (i.e., *prior* probability) with which it occurs.

The intention is to determine a strategy for the *minimum average risk*. Because we require that a decision be made, each observation vector **x** must be assigned in the overall observation space $\mathscr{X}$ to either $\mathscr{X}_1$ or $\mathscr{X}_2$. Thus,

$$\mathscr{X} = \mathscr{X}_1 + \mathscr{X}_2 \tag{1.24}$$

Accordingly, we may rewrite Eq. (1.23) in the equivalent form

$$\begin{aligned}\mathfrak{R} = {} & c_{11}p_1 \int_{\mathscr{X}_1} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1)d\mathbf{x} + c_{22}p_2 \int_{\mathscr{X}-\mathscr{X}_1} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2)d\mathbf{x} \\ & + c_{21}p_1 \int_{\mathscr{X}-\mathscr{X}_1} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1)\mathbf{x} + c_{12}p_2 \int_{\mathscr{X}_1} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2)d\mathbf{x}\end{aligned} \tag{1.25}$$

where $c_{11} < c_{21}$ and $c_{22} < c_{12}$. We now observe the fact that

$$\int_{\mathscr{X}} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1)d\mathbf{x} = \int_{\mathscr{X}} p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2)d\mathbf{x} = 1 \tag{1.26}$$

Hence, Eq. (1.25) reduces to

$$\begin{aligned}\mathfrak{R} = {} & c_{21}p_1 + c_{22}p_2 \\ & + \int_{\mathscr{X}_1} [p_2(c_{12} - c_{22})\, p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2) - p_1(c_{21} - c_{11})\, p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1)]d\mathbf{x}\end{aligned} \tag{1.27}$$

The first two terms on the right-hand side of Eq. (1.27) represent a fixed cost. Since the requirement is to minimize the average risk $\mathfrak{R}$, we may therefore deduce the following strategy from Eq. (1.27) for optimum classification:

1. All values of the observation vector **x** for which the integrand (i.e., the expression inside the square brackets) is negative should be assigned to subspace $\mathscr{X}_1$ (i.e., class $\mathscr{C}_1$), for the integral would then make a negative contribution to the risk $\mathfrak{R}$.
2. All values of the observation vector **x** for which the integrand is positive should be excluded from subspace $\mathscr{X}_1$ (i.e., assigned to class $\mathscr{C}_2$), for the integral would then make a positive contribution to the risk $\mathfrak{R}$.
3. Values of **x** for which the integrand is zero have no effect on the average risk $\mathfrak{R}$ and may be assingned arbitrarily. We shall assume that these points are assigned to subspace $\mathscr{X}_2$ (i.e., class $\mathscr{C}_2$).

On this basis, we may now formulate the Bayes classifier as follows:

If the condition

$$p_1(c_{21} - c_{11})\, p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1) > p_2(c_{12} - c_{22})\, p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2)$$

holds, assign the observation vector **x** *to subspace* $\mathscr{X}_1$ *(i.e., class* $\mathscr{C}_1$*). Otherwise assign* **x** *to* $\mathscr{X}_2$ *(i.e., class* $\mathscr{C}_2$*).*

To simplify matters, define

$$\Lambda(\mathbf{x}) = \frac{p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_1)}{p_{\mathbf{X}}(\mathbf{x}|\mathscr{C}_2)} \tag{1.28}$$

and

$$\xi = \frac{p_2(c_{12} - c_{22})}{p_1(c_{21} - c_{11})} \tag{1.29}$$

12 **What do you mean by XOR problem? Discuss about the Heuristics for Making the Back-Propagation Algorithm Perform Better.** 12M

The XOR Problem

The XOR, or "exclusive or", problem is a classic problem in ANN research. It is the problem of using a neural network to predict the outputs of XOR logic gates given two binary inputs. An XOR function should return a true value if the two inputs are not equal and a false value if they are equal. All possible inputs and predicted outputs are shown below.

Input 1	Input 2	Output
0	0	0
0	1	1
1	1	0
1	0	1

XOR is a classification problem and one for which the expected outputs are known in advance. It is therefore appropriate to use a supervised learning approach.

Heuristics for Making the Back-Propagation Algorithm Perform Better.

- Stochastic Versus Batch Learning
- Shuffling the Examples
- Normalizing the Inputs
- The Sigmoid
- Choosing Target Values
- Initializing the Weights
- Choosing Learning Rates
- Radial Basis Function vs Sigmoid

1. **Stochastic Versus Batch Learning**

Stochastic learning is generally the preferred method for basic backpropagation for the following three reasons:

- Stochastic learning is usually much faster than batch learning.
- Stochastic learning also often results in better solutions.
- Stochastic learning can be used for tracking changes.

2. **Shuffling the Examples:**

Networks learn the fastest from the most unexpected sample. Therefore, it is advisable to choose a sample at each iteration that is the most unfamiliar to the system.

3. **Normalizing the Inputs:**

This tip highlights the importance of data preparation prior to training a neural network model.
The authors point out that neural networks often learn faster when the examples in the training dataset sum to zero. This can be achieved by subtracting the mean value from each input variable, called centering.
Convergence is usually faster if the average of each input variable over the training set is close to zero.

4. **Sigmoid Activation function:**

onlinear activation functions are what give neural networks their nonlinear capabilities. One of the most common forms of activation function is the sigmoid

5. **Choosing target values:**

In the case of binary classification problems, target variables may be in the set {0, 1} for the limits of the logistic activation function or in the set {-1, 1} for the hyperbolic tangent function when using the cross-entropy or hinge loss functions respectively, even in modern neural networks.

6. **Initializing the weights :**

The starting values of the weights can have a significant effect on the training process. Weights should be chosen randomly but in such a way that the sigmoid is primarily activated in its linear region.

7. **Choosing learning rates:**

Most of those schemes decrease the learning rate when the weight vector "oscillates", and increase it when the weight vector follows a relatively steady direction.

13(a) **Discuss the back propagation and differentiation.** 6M

Back-propagation is an automatic differentiation algorithm that can be used to calculate the gradients for the parameters in neural networks. Together, the back-propagation algorithm and Stochastic Gradient Descent algorithm can be used to train a neural network.

(b) **Discuss the cross validation.** 6M

Cross-validation is a machine learning technique where the training data is split into two parts: A training set and a test set. The training set is used to build the model, and the test set is used to evaluate how well the model performs when in production.

Types of Cross-Validation techniques:

1. Holdout Method
2. K-Fold Cross-Validation
3. Stratified K-Fold Cross-Validation
4. Leave-P-Out Cross-Validation

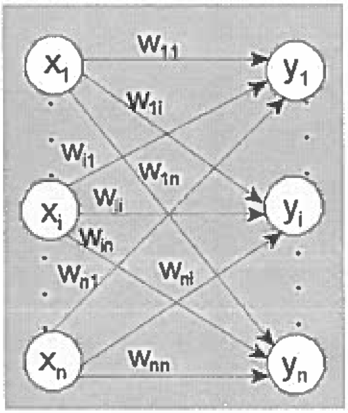
14 **What is content addressable memory? Draw the architecture of auto associative memory network. Write the steps of testing algorithm for auto associative memory.** 12M

Content Addressable Memory:

An associate memory network refers to a content addressable memory structure that associates a relationship between the set of input patterns and output patterns. A content addressable memory structure is a kind of memory structure that enables the recollection of data based on the intensity of similarity between the input pattern and the patterns stored in the memory.

Auto associative memory network:

An auto-associative memory recovers a previously stored pattern that most closely relates to the current pattern. It is also known as an auto-associative correlator.



Testing Algorithm for auto-associative memory network:

Step 1 – Set the weights obtained during training for Hebb's rule.
Step 2 – Perform steps 3-5 for each input vector.
Step 3 – Set the activation of the input units equal to that of the input vector.
Step 4 – Calculate the net input to each output unit **j = 1 to n**

$$y_{inj} = \sum_{i=1}^{n} x_i w_{ij}$$

Step 5 – Apply the following activation function to calculate the output

$$y_j = f(y_{inj}) = \begin{cases} +1 & if\ y_{inj} > 0 \\ -1 & if\ y_{inj} \leqslant 0 \end{cases}$$

15 **What do you mean by pattern association? What is the purpose of bi-directional associative memory? Draw the architecture and discuss the algorithm of BAM.** 12M

Pattern Association:

Pattern association is the process of memorizing input-output patterns or recalling of the patterns *in a neural network.*

Bidirectional Associative Memory:

Bidirectional Associative Memory (BAM) is a supervised learning model in Artificial Neural Network. This is hetero-associative memory, for an input pattern, it returns another pattern which is potentially of a different size. This phenomenon is very similar to the human brain. Human memory is necessarily associative. The main objective to introduce such a network model is to store hetero-associative pattern pairs. This is used to retrieve a pattern given a noisy or incomplete pattern.

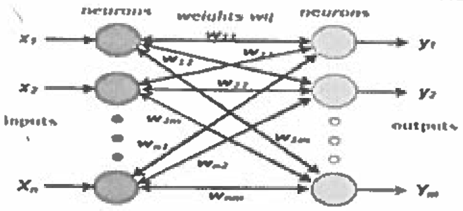


Fig. Bidirectional Associative Memory model

Algorithm:

1. **Storage (Learning):** In this learning step of BAM, weight matrix is calculated between M pairs of patterns (fundamental memories) are stored in the synaptic weights of the network following the equation

$$W = \sum_{m=1}^{M} X_m Y_m^T$$

2. **Testing:** We have to check that the BAM recalls perfectly for corresponding and recalls for corresponding. Using,

$$Y_m = \text{sign}\left(W^T X_m\right), \quad m = 1.2, \ldots, M$$

$$X_m = \text{sign}\left(W Y_m\right), \quad m = 1.2, \ldots, M$$

All pairs should be recalled accordingly.

3. **Retrieval:** For an unknown vector ***X*** (a corrupted or incomplete version of a pattern from set ***A*** or ***B***) to the BAM and retrieve a previously stored association:

- ***Initialize the BAM:*** $X(0) = X, \quad p = 0$
- ***Calculate the BAM output at iteration*** :

$$Y(p) = \text{sign}\left[W^T X(p)\right]$$

- ***Update the input vector*** : $X(p+1) = \text{sign}[W Y(p)]$
- ***Repeat*** the iteration until convergence, when input and output remain unchanged.

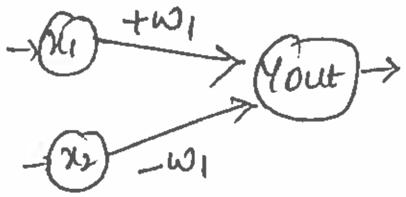
⑧ McCulloch Pitts Neuron Model :- [12M]

* It is also known as MP neuron model. and it is the earliest neural network model.

* In this neural network model there are two types of weighted inputs - Excitatory ⇒ weights of positive magnitude
- Inhibitory ⇒ weights of negative magnitude.

* These are generally used to implement logic gates.

* Here threshold acts like an activation function to get the final output.



* Conditions :- Let threshold = θ

$$Y_{out} = \begin{cases} 1, & \text{if } y_{sum} \geq \theta \\ 0, & \text{if } y_{sum} < \theta \end{cases}$$

OR Gate Implementation:

STEP 1: Truth table of OR gate with inputs x_1, x_2 & output Y.

x_1	x_2	Y
0	0	0
0	1	1
1	0	1
1	1	1

⇒ The output will be high if atleast one of the input is high.

STEP 2: Assume weights and calculate net input signal.

$$\Rightarrow \boxed{y_{sum} = \sum x_i w_i}$$

Case 1 : $w_1 = w_2 = 1$	Case 2 :-	Case 3 :	Case 4
	$w_1 = -1, w_2 = 1$	$w_1 = 1, w_2 = -1$	$w_1 = w_2 = -1$
$(0,0) = 0$	$(0,0) = 0$	$(0,0) = 0$	$(0,0) = 0$
$(0,1) = 1$	$(0,1) = 1$	$(0,1) = -1$	$(0,1) = -1$
$(1,0) = 1$	$(1,0) = -1$	$(1,0) = 1$	$(1,0) = -1$
$(1,1) = 2$	$(1,1) = 0$	$(1,1) = 0$	$(1,1) = -1$

STEP 3 : Upon comparing these cases with Y values, case 1 matches with Y values as $(0,1), (1,0), (1,1)$ cases are high.

So consider threshold value as '1' and '2'. & weights are $w_1 = w_2 = 1$

STEP 4 : Assume $\theta = +1$

$w_1 = w_2 = 1$

$y_{sum} \geq 1$

$0 \geq 1 \ (F) \Rightarrow 0$

$1 \geq 1 \ (T) \Rightarrow 1$

$1 \geq 1 \ (T) \Rightarrow 1$

$2 \geq 1 \ (T) \Rightarrow 1$

Assume $\theta = 2$

$w_1 = w_2 = 1$

$y_{sum} \geq 2$

$0 \geq 2 \ (F) \Rightarrow 0$

$1 \geq 2 \ (F) \Rightarrow 0$

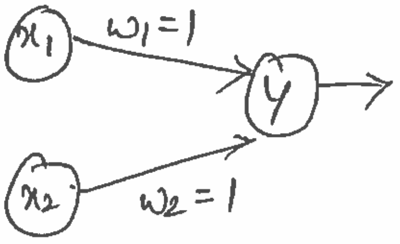
$1 \geq 2 \ (F) \Rightarrow 0$

$2 \geq 2 \ (T) \Rightarrow 1$

From these two cases, case 1 satisfies/matches with Y values,

$\therefore \ \boxed{\theta = +1}$ and $\boxed{w_1 = w_2 = 1}$

$$Y = f(y_{sum}) = \begin{cases} 1, & y_{sum} \geq 1 \\ 0, & y_{sum} < 1 \end{cases}$$



NOR Gate Implementation

STEP 1 :- truth table

x_1	x_2	y
0	0	1
0	1	0
1	0	0
1	1	0

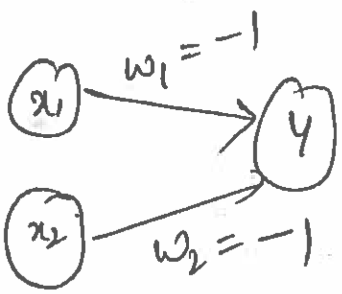
STEP 2 :- Assume weights & calculate net input signal.

$$\Rightarrow \boxed{y_{sum} = \Sigma x_i w_i}$$

	Case 1 :- $w_1 = w_2 = 1$	Case 2 :- $w_1 = -1, w_2 = 1$	Case 3 :- $w_1 = 1, w_2 = -1$	Case 4 : $w_1 = -1, w_2 = -1$
$(0,0) =$	$\boxed{0}$	$\boxed{0}$	$\boxed{0}$	$\boxed{0}$
$(0,1) =$	1	1	-1	-1
$(1,0) =$	1	-1	1	-1
$(1,1) =$	2	0	0	2

STEP 3 :-

$$y = f(y_{in}) = \begin{cases} 0 , & y_{in} < 0 \\ 1 , & y_{in} \geq 0 . \end{cases}$$



$\therefore \theta = 0.$

$w_1 = -1, \; w_2 = -1$

Nadimpalli Satyanarayana Raju Institute of Technology (Autonomous): IQAC: Quality Management System (QMS)

NSRIT

Semester End Regular/Supplementary Examination, Dec. /Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	CSE (DS)			**Academic Year**	2022 - 2023
Course Code	20DS302	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Foundations of Data Science**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Differentiate volume and variety	20DS302.1	L2
2	Define list in python	20DS302.2	L1
3	Distinguish supervised and unsupervised learning	20DS302.3	L2
4	What is data visualization dash board?	20DS302.4	L1
5	Where we use data wrangling	20DS302.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Compare data scientists and data engineers	6M	20DS302.1	L2
6 (b)	Explain Digging into MapReduce	6M	20DS302.1	L2
	OR			
7 (a)	Compare the Types of data analytics	6M	20DS302.1	L2
7 (b)	Explain Taking Action on Business Insights	6M	20DS302.1	L2
8 (a)	Explain various types of loops and functions in python	6M	20DS302.2	L2
8 (b)	Explain R programming for data science	6M	20DS302.2	L2
	OR			
9 (a)	Discuss about SQL in data science	6M	20DS302.2	L2
9 (b)	Explain using KNIME for advanced analytics	6M	20DS302.2	L2
10 (a)	Discuss about clustering with k-means algorithm	6M	20DS302.3	L2
10 (b)	Explain any two techniques for detecting outliers	6M	20DS302.3	L2
	OR			
11 (a)	Explain lingo and spark streaming for IoT	6M	20DS302.3	L2
11 (b)	Differentiate Random forest and Decision tree methods	6M	20DS302.3	L2
12 (a)	Explain the designing meet the needs of your target audience	6M	20DS302.4	L2
12 (b)	Write about D3.js web applications	6M	20DS302.4	L2
	OR			
13 (a)	Explain designing data visualizations for collaboration	6M	20DS302.4	L2
13 (b)	Analyze map projections and co ordinate systems	6M	20DS302.4	L2
14 (a)	How to Finding and telling your data story in journalism	6M	20DS302.5	L2
14 (b)	Explain modeling natural resources in Raw	6M	20DS302.5	L2
	OR			
15 (a)	Explain Angling in on analytics in E-commerce	6M	20DS302.5	L2
15 (b)	Explain spatial crime prediction and monitoring	6M	20DS302.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

FOUNDATIONS OF DATA SCIENCE

1. Differentiate volume and variety

The lower limit of big data volume starts as low as 1 terabyte, and it has no upper limit. If your organization owns at least 1 terabyte of data, it's probably a good candidate for a big data deployment.

This high-variety data comes from a multi-tude of sources. The most salient point about it is that it's composed of a combination of datasets with differing underlying structures

2. Define list in python

A list is a sequence of numbers and/or strings. To create a list, you simply enclose the elements of the list (separated by commas) within square brackets. Here's an example of a basic list:

```
>>> variable2=["ID","Name","Depth","Latitude","Longitude"]
>>> depth=[0,120,140,0,150,80,0,10]
>>> variable2[3]
'Latitude'
```

3. Distinguish supervised and unsupervised learning

Learning with supervised algorithms

Supervised learning algorithms require that input data has labeled features. These algorithms learn from known features of that data to produce an output model that successfully predicts labels for new incoming, unlabeled data points. You use supervised learning when you have a labeled dataset composed of historical values that are good predictors of future events. Use cases include survival analysis and fraud detection, among others. Logistic regression is a type of supervised learning algorithm.

Learning with unsupervised algorithms

Unsupervised learning algorithms accept unlabeled data and attempt to group observations into categories based on underlying similarities in input features, as shown in Figure 4-1. Principal component analysis, k-means clustering, and sin-gular value decomposition are all examples of unsupervised machine learning algorithms. Popular use cases include recommendation engines, facial recognition systems, and customer segmentation.

4. **What is data visualization dash board?**
   data analytics dashboards are one of the more popular methods for delivering such information. Acting as a (hopefully) user-friendly soft-ware interface, such dashboards can provide a single-page, easy-to- understand summary of information that's vital to organizational and managerial decision making

5. **Where we use data wrangling**

» Data extraction

» Data preparation

» Data governance

» Data architecture

6.a. Compare data scientists and data engineers

Parameters	Data Scientist	Data Engineer
Goal	Answer questions, Reduce costs,Create Effeciences and data increase revenue.	Develop more robust and efficient data systems and warehouse solutions
Nature	Analytical	Operational
Output	Data product	Data flow,storage and retrieval
Also know as	Statistician, Data Manager	Database administrator,Data Architect
Function	End to end project management predictive modelling solutions, storytelling and visualization	Large dataset structures predictive modelling, Implementation, Automation.

6. b. Explain Digging into MapReduce

MapReduce is a parallel distributed processing framework that can be used to pro-cess tremendous volumes of data in-batch — where data is collected and thenprocessed as one unit with processing completion times on the order of hours or days.

MapReduce works by converting raw data down to sets of tuples and then combining and reducing those tuples into smaller sets of tuples (with respect to MapReduce, tuples refer to key-value pairs by which data is grouped, sorted, and processed). In layman's terms, MapReduce uses parallel distributed computing to transform big data into manageable-size data. Parallel distributed processing refers to a powerful framework where data is processed very quickly via the distribution and parallel processing of tasks across clusters of commodity servers.

MapReduce jobs implement a sequence of map- and reduce-tasks across a dis-tributed set of servers. In the map task, you delegate data to key-value pairs, transform it, and filter it. Then you assign the data to nodes for processing. In the reduce task, you aggregate that data down to smaller-size datasets. Data from the reduce step is transformed into a standard key-value format — where the key acts as the record identifier and the value is the value being identified by the key. The clusters' computing nodes process the map tasks and reduce tasks that are defined NM NMby the user.

MapReduce as a batch-processing tool, to boil down and begin to make sense of a huge volume, velocity, and variety of data by using map and reduce tasks to tag the data by (key, value) pairs, and then reduce those pairs into smaller sets of data through aggregation operation.

This work is done in two steps:

1. Map the data.

The incoming data must first be delegated into key-value pairs and divided into fragments, which are then assigned to map tasks. Each computing cluster (a group of nodes that are connected to each other and perform a shared computing task) is assigned a number of map tasks, which are subsequently distributed among its nodes. Upon processing of the key-value pairs, interme-

diate key-value pairs are generated. The intermediate key-value pairs are sorted by their key values, and this list is divided into a new set of fragments. Whatever count you have for these new fragments, it will be the same as the count of the reduce tasks.

2. Reduce the data.

Every reduce task has a fragment assigned to it. The reduce task simply processes the fragment and produces an output, which is also a key-value pair. Reduce tasks are also distributed among the different nodes of the cluster. After the task is completed, the final output is written onto a file system.

7.a. Compare the Types of data analytics

Types of analytics:-

» Descriptive analytics: This type of analytics answers the question, "What happened?" Descriptive analytics are based on historical and current data. A business analyst or a business-centric data scientist bases modern-day business intelligence on descriptive analytics.

» Diagnostic analytics: You use this type of analytics to find answers to the question, "Why did this particular something happen?" or "What went wrong?" Diagnostic analytics are useful for deducing and inferring the success or failure of subcomponents of any data-driven initiative.

» Predictive analytics: Although this type of analytics is based on historical and current data, predictive analytics go one step further than descriptive analytics. Predictive analytics involve complex model-building and analysis in order to predict a future event or trend. In a business context, these analyses would be performed by the business-centric data scientist.

» Prescriptive analytics: This type of analytics aims to optimize processes, structures, and systems through informed action that's based on predictive analytics — essentially telling you what you should do based on an informed estimation of what will happen. Both business analysts and business-centric data scientists can generate prescriptive analytics, but their methods and data sources differ.

7.b. Explain Taking Action on Business Insights

Business-centric data science is multidisciplinary and incorporates the following elements:

» Quantitative analysis: Can be in the form of mathematical modeling, multivariate statistical analysis, forecasting, and/or simulations.The term multivariate refers to more than one variable. A multivariate statistical analysis is a simultaneous statistical analysis of more than one

variable at a time.

» Programming skills: You need the necessary programming skills to analyze raw data and to make this data accessible to business users.

» Business knowledge: You need knowledge of the business and its environ-ment so that you can better understand the relevancy of your findings.Data science is a pioneering discipline. Data scientists often employ the scientific method for data exploration, hypotheses formation, and hypothesis testing (through simulation and statistical modeling). Business-centric data scientists

generate valuable data insights, often by exploring patterns and anomalies in business data.

» Internal and external datasets: Data science is flexible. You can create business data mash-ups from internal and external sources of structured and unstructured data fairly easily. (A data mash-up is combination of two or more data sources that are then analyzed together in order to provide users with a more complete view of the situation at hand.)

» Tools, technologies, and skillsets: Examples here could involve using cloud-based platforms, statistical and mathematical programming, machine learning, data analysis using Python and R, and advanced data visualization.

8. a. Explain various types of loops and functions in python

- **Loops in python**

You can use looping to execute the same block of code multiple times for a sequence of items. Consequently, rather than manually access all elements one by one, you simply create a loop to automatically iterate (orpass through in successive cycles) each element of the list.
Two types of loops in Python: the for loop and the while loop. The most often used looping technique is the for loop — designed especially to iterate through sequences, strings, tuples, sets, and dictionaries. The other available looping technique in Python is the while loop. Use a while loop to perform actions while a given condition is true.

- **Functions in Python**

Functionsare the crucial building blocks of almost every programming language. They provide a way to build organized, reusable code. Functions are blocks of code that take an input, process it, and return an output. Function inputs can be numbers, strings, lists, objects, or functions. Python has two types of functions: built-in and custom. Built-in functions are predefined inside Python.

```
>>> def average(any_list):return(sum(any_list)/len(any_list))
```

This code snippet defines a function named average, which takes any list as input and calculates the average of its elements. The function is not executed yet, but the code defines what the function does when it later receives some input values.Executing a function is straightforward. You can use functions to do the same thing repeatedly, as many times as you need, for different input values. The beauty here is that, once the functions are constructed, you can reuse them without having to rewrite the calculating algorithm.

8. b. Explain R programming for data science

- **R's Basic Vocabulary**

» Non-interactive: You run your R code by executing it as a .r file (the .r file extension is the one that's assigned to script files created for execution by the R program) directly from the command line.

» Interactive: You generally work in a software application that interacts with you by prompting you to enter your data and R code. In an R session within interactive mode, you can import datasets or enter the raw data directly; assign names to variables and data objects; and use functions, operators, and built-in iterators to help you gain some insight into your source data.

- **R works with the following main object types:**

» **Vector**: A vector is an ordered list of the same mode — character (alphanu-meric), numeric, or Boolean. Vectors can have any number of dimensions. For instance, the vector A = ["a", "cat", "def"] is a 3-dimensional vector of mode character. B = [2, 3.1, -5, 33] is a 4-dimensional vector of mode numerical. To identify specific elements of these vectors, you could enter the following codes at the prompt in interactive mode to get R to generate the following returns: A[[1]] = "a" or A[[2]] = "cat" or A[[3]] = "def" or B[[1]] = 2 or B[[2]] = 3.1 or B[[3]] = -5 or B[[4]] = 33. R views a single number as a vector of dimension one. Because they can't be broken down further in R, vectors are also known as atomic vectorsR's treatment of atomic vectors gives the language tremendous advantages
with respect to speed and efficiency.
» **Matrix:** Think of a matrix as a collection of vectors. A matrix can be of any mode (numerical, character, or Boolean), but all elements in the matrix must be of the same mode. A matrix is also characterized by its number of dimensions. Unlike a vector, a matrix has only two dimensions: number of rows and number of columns.
» **List:** A list is a list of items of arbitrary modes, including other lists or vectors.Lists are sometimes also called generic vectors because some of the same operations performed on vectors can beperformed on lists as well.
» **Data frame:** A data frame is a type of list that's analogous to a table in a database. Technically speaking, a data frame is a list of vectors, each of which is the same length. A row in a table contains the information for an individual record, but elements in the row most likely will not be of the same mode. All elements in a specific column, however, are all of the same mode. Data frames are structured in this same way — each vector in a data frame corresponds to a column in a data table, and each possible index for these vectors is a row.

There are two ways to access members of vectors, matrices, and lists in R:
» Single brackets [ ] give a vector, matrix, or list (respectively) of the element(s) that are indexed.
» Double brackets [[ ]] give a single element.

- **Assigning an Object and Concatenating in R**

```
> EmployeeRoll <- data.frame(list(EmployeeName=c("Smith,
John","O'Bannon, Tom","Simmons, Sarah"),Grade=c(10,8,12),
Salary=c(100000,75000,125000), Union=c(TRUE, FALSE,
TRUE)))
> EmployeeRoll
 EmployeeName Grade Salary Union
1 Smith,John 10 100000 TRUE
2 O'Bannon, Tom 8 75000 FALSE
3 Simmons, Sarah 12 125000 TRUE
```

One other object within R is vitally important: the function. Functions use atomic vectors, matrices, lists, and data frames to accomplish whatever analysis or computation you want done.

- **Delving into Functions and Operators**

There are 2 methods during writing functions To illustrate the difference nobetween these two methods, consider again the EmployeeRoll dataset.

```
> #Method 1 of Calculating the Mean Salary
> MeanSalary1 <- mean(EmployeeRoll$Salary)
> MeanSalary1
[1] 1e+05
```

In this method, the mean( ) function calculates and saves the average salary, 100,000 (or 1e+05, in scientific notation) as an object (a vector, of course!) named MeanSalary1.
Method 2 illustrates a more complicated but possibly more useful approach. Rather than define only a single operation, as in Method 1, Method 2's function can define a series of separate operations if they're needed; therefore, the method can oftentimes get quite complex.

```
> #Method 2 of Calculating the Mean Salary
> #This method allows the user to create a custom set of
instructions for R that can be used again and again.
> MeanSalary2 <- function(x) {return(mean(x))}
>
> MeanSalary2(EmployeeRoll$Salary)
[1] 1e+05
```

- **Popular Operators**

This code snippet shows several examples of where operators are used as functions:

```
> "<"(2,3)
[1] TRUE
> "<"(100,10)
[1] FALSE
> "+"(100,1)
[1] 101
> "/"(4,2)
[1] 2
> "+"(2,5,6,3,10)
Error in `+`(2, 5, 6, 3, 10) : operator needs one or two
Arguments
```

In the preceding code, the Boolean operators less than (<) and greater than (>) return a value of either TRUE or FALSE. Also, do you see the error message that's generated by the last line of code? That error happened because the operator + can take only one or two arguments, and in that example, I provided three arguments more than it could handle.

- **Iterating in R**

Programming in R offers you an efficient way to handle loops and iterations.Essentially, R has built-in iterators that automatically loop over elements To better conceptualize this process, called vectorization, imagine that you want to add a constant c = 3 to a series of three numbers that you've stored as a vector, m = [10, 6, 9].

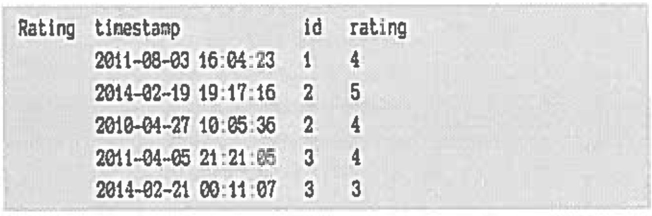
```
> c <- 3
> m <- c(10, 6, 9)
> m <- m + c
> m
[1] 13 9 12
```

9. a. Discuss about SQL in data science

When working with SQL commands, you use functions to perform tasks, and arguments to more narrowly specify those tasks. To query a particular set from within your data tables, for example, use the SELECT function. To combine separate tables into one, use the JOIN function. To place limits on the data that your query returns, use a WHERE argument. As I say in the preceding section, fewer than 20 commands are commonly used in SQL. This section introduces SELECT, FROM, JOIN, WHERE, GROUP, MAX(), MIN(), COUNT(), AVG(), and HAVING.

The most common SQL command is SELECT. You can use this function to generate a list of search results based on designated criteria. To generate a printout of all data FROM the Rating table, use the SELECT function.

Any function with SELECT is called a query, and SELECT functions accept different arguments to narrow down or expand the data that is returned. Since an asterisk (*) represents a wildcard, the asterisk in SELECT * tells the interpreter — the SQL component that carries out all SQL statements — to show every column in the table. You can then use the WHERE argument to limit the output to only certain values. For example, here is the complete Rating table:



```
Rating  timestamp            id  rating
        2011-08-03 16:04:23  1   4
        2014-02-19 19:17:16  2   5
        2010-04-27 10:05:36  2   4
        2011-04-05 21:21:05  3   4
        2014-02-21 00:11:07  3   3
```

Using SELECT, WHERE, and DATE() to Query Data

```
SELECT * FROM Rating
WHERE Rating.timestamp >= date('2014-01-01')
timestamp id rating
2014-02-19 19:17:16 2 5
2014-02-21 00:11:07 3 3
```

You can also use SQL to join columns into a new data table. Joins are made on the basis of shared (or compared) data in a particular column (or columns). There are several ways you can execute a join in SQL, but the ones listed here are probably the most popular:

» **Inner join:** The default JOIN type; returns all records that lie in the intersecting regions between the tables being queried

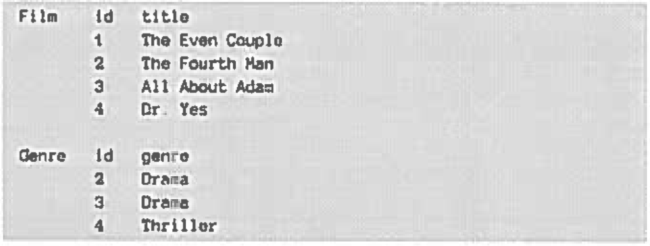
» **Outer join:** Returns all records that lie outside the overlapping regions between queried data tables

» **Full outer join:** Returns all records that lie both inside and outside the overlapping regions between queried data tables — in other words, returns all records for both tables

» **Left join:** Returns all records that reside in the leftmost table

» **Right join**: Returns all records that reside in the rightmost table

To differentiate between an inner join and an outer join because these functions handle missing data in different ways. As an example of a join in SQL, if you want a list of films that includes genres, you use an inner join between the Film and Genre tables to return only the results that intersect (overlap) between the two tables.



```
Film    id   title
        1    The Even Couple
        2    The Fourth Man
        3    All About Adam
        4    Dr. Yes

Genre   id   genre
        2    Drama
        3    Drama
        4    Thriller
```

An Inner JOIN Function

```
SELECT Film.id, Film.title, Genre.genre
FROM Film
JOIN Genre On Genre.id=Film.id
id          title               genre
2      The Fourth Man       Drama
3       All About Adam      Drama
4       Dr. Yes              Thriller
```

To avoid creating a duplicate id column in the table that results from the JOIN — one id from the Film table and one id from the Genre table. Since the default for JOIN is inner, and inner joins return only records that are overlapping or shared between tables, Film 1 is omitted from the results (due to its missing genre value).

A Full Outer JOIN

```
SELECT Film.id, Film.title, Genre.genre
FROM Film
FULL JOIN Genre On Genre.id=Film.id
id                 title                   genre
1      The Even Couple           NULL
2      The Fourth Man            Drama
3       All About Adam           Drama
4       Dr. Yes                    Thriller
```

If you want to return all rows, even ones with NULL values, simply do a full outer join.

Using a GROUP Statement to Aggregate Data

```
SELECT Film.title, AVG(rating) AS avg_rating
FROM Film
JOIN Rating On Film.id=Rating.id
GROUP BY Film.title
        title                  avg_rating
All About Adam                    3.5
The Even Couple                   4.0
The Fourth Man                    4.5
```

To aggregate values so that you can figure out the average rating for a film, use the GROUP statement. (GROUP statement commands include MAX(), MIN(), COUNT(), or AVG().)the average rating of films was returned; the AS statement was used in SELECT to rename the column, to make sure it was properly labeled. The Film and Ratings tables had to be joined, and because Dr. Yes had no ratings and an inner join was used, that film was left out.

A HAVING Clause to Narrow Results

```
SELECT Film.title, AVG(rating) AS avg_rating
FROM Film
JOIN Rating On Film.id=Rating.id
GROUP BY Film.title
HAVING avg_rating >= 4
        title                 avg_rating
The Even Couple                  4.0
The Fourth Man                   4.5
```

To narrow the results even further, add a HAVING clause at the end.The codelimits the data your query returns so that you get only records of titles that have an average rating greater than or equal to 4.

9. b. Explain using KNIME for advanced analytics

If you don't know how to code but still want the benefits that custom predictive analytics has to offer, you can download and install KNIME and use its visual environment to access these features. KNIME offers services, solutions, and open-source software to fulfill the advanced analytics requirements of today's data driven business enterprise. The company's purpose is to provide an open platform that meets the data-mining and analytics needs of the masses.

If you want data-mining software that you can install on your PC and use for predictive analytics, look no further than KNIME Analytics Platform. KNIME is easy to use, so even beginners who don't know how to code can use the program.

For more advanced users, however, KNIME offers plug-ins that can be used to integrate Weka's preconstructed analysis modules or to run R and Python scripts from within the application. Beginners and advanced users alike can use KNIME predictive analytics
» Upsell and cross-sell: Build cross-selling and upselling models that enable you to increase sales by making optimal recommendations of other products that customers are likely to be interested in purchasing as well.
» Churn reduction: Mine customer data and identify which customers you're most likely to lose and why.
» Sentiment and network analysis: Analyze the sentiment of people and organizations in your social networks, to help identify which areas of your business are performing well and which ones may need some work.
» Energy usage prediction and auditing: Perform time series analyses and build regression models from energy usage data.

10.a. Discuss about clustering with k-means algorithm

- Clustering with the k-means algorithm

The k-means clustering algorithm is a simple, fast, unsupervised learning algorithm that you can use to predict groupings within a dataset. The model makes its prediction based on the number of centroids present — represented by k, a model parameter that you must define — and the nearest mean values, measured by the Euclidean distance between observations. Because the features of a dataset are usually on different scales, the difference of scales can distort the results of this distance calculation. To avoid this problem, scale your variables before using k-means to predict data groupings.The quality of the clusters is heavily dependent on the correctness of the k value you specify. If your data is 2- or 3-dimensional, a plausible range of k values may be visually determinable. In the eyeballed approximation of clustering from the World Bank Income and Education data scatter plot (refer to Figure 6-2), a visual estimation of the k value would equate to three clusters, or k = 3.

If your dataset has more than three dimensions, however, you can use computational methods to generate a good value for k. One such method is the silhouette coefficient — a method that calculates the average distance of each point from all other points in a cluster and then compares that value with the average distance to every point in every other cluster. Luckily, because the k-means algorithm is efficient, it does not require much computer processing power, and you can easily calculate this coefficient for a wide range of k values.The k-means algorithm works by placing sample cluster centers on an n-dimensional plot and then evaluating whether moving them in any single direction would result in a new center with higher density — with more observations closer to it, in other words. The centers are moved from regions of lower density to regions of higher density until all centers are within a region of local maximum density — a true center of the cluster, where each cluster gets a maximum number of points closest to its cluster center. Whenever possible, you should try to place the centers yourself, manually. If that's not possible, simply place the centers randomly and run the algorithm several times to see how often you end up with the same clusters.

One weakness of the k-means algorithm is that it may produce incorrect results by placing cluster centers in areas of local minimum density. This happens when centers get lost in low-density regions (in other words, regions of the plot that have relatively few points plotted in them) and the algorithm-driven directional movement (the movement that's meant to increase point density) starts to bounce and oscillate between faraway clusters. In these cases, the center gets caught in a low-density space that's located between two high-point density zones. This results in erroneous clusters based around centers that converge in areas of low, local minimum density. Ironically, this happens most often when the underlying data is very well-clustered, with tight, dense regions that are separated by wide, sparse areas.

10.b. Explain any two techniques for detecting outliers

Detecting Outliers

Outliers are data points with values that are significantly different than the majority of data points comprising a variable. It is important to find and remove outliers, because, left untreated, they skew variable distribution, make variance appear falsely high, and cause a misrepresentation of intervariable ccorrelations.Most machine learning and statistical models assume that your data is free of outliers, so spotting and removing them is a critical part of preparing your data for analysis. Not only that, you can use outlier detection to spot anomalies that represent fraud, equipment failure, or cybersecurity attacks. In other words, outlier detection is a data preparation method and an analytical method in its own right.

Outliers fall into the following three categories:

» Point: Point outliers are data points with anomalous values compared to the normal range of values in a feature.

» Contextual: Contextual outliers are data points that are anomalous only within a specific context. To illustrate, if you are inspecting weather station data from January in Orlando, Florida, and you see a temperature reading of 23 degrees F, this would be quite anomalous because the average temperature there is 70 degrees F in January. But consider if you were looking at data from January at a weather station in Anchorage, Alaska — a temperature reading of 23 degrees F in this context is not anomalous at all.

» Collective: These outliers appear nearby to one another, all having similar values that are anomalous to the majority of values in the feature.

- Detecting outliers with univariate analysis

Univariate outlier detection is where you look at features in your dataset, and inspect them individually for anomalous values. There are two simple methods for doing this:

» Tukey outlier labeling

» Tukey boxplot

In comparison, a Tukey boxplot is a pretty easy way to spot outliers. Each boxplot has whiskers that are set at 1.5*IQR. Any values that lie beyond these whiskers are outliers. Figure 5-2 shows outliers as they appear within a Tukey boxplot.

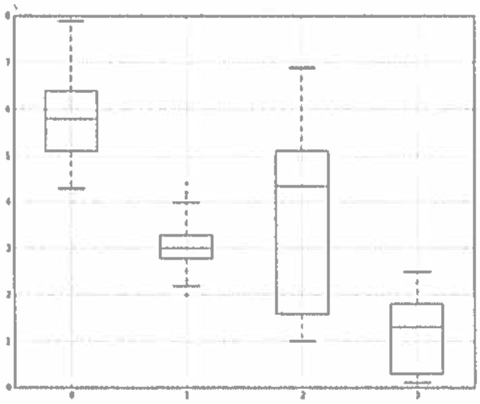


FIGURE 5-2: Spotting outliers with a Tukey boxplot.

Credit: Python for DS, Lynda.com

- Detecting outliers with multivariate analysis

Sometimes outliers only show up within combinations of data points from dispa-rate variables. These outliers really wreak havoc on machine learning algorithms, so it's important to detect and remove them. You can use multivariate analysis of outliers to do this. A multivariate approach to outlier detection involves considering two or more variables at a time and inspecting them together for outliers. There are several methods you can use, including

» Scatter-plot matrix
» Boxplot
» Density-based spatial clustering of applications with noise (DBScan) — as discussed in Chapter 6
» Principal component analysis.

11. a. Explain lingo and spark streaming for IoT

- Learning the lingo

Before delving into the data science and innovation that's related to IoT, you need a grasp of the fundamental vocabulary. The fog — or IoT cloud — is a network of cloud services that connect to IoT-enabled devices. Cloud-based big data processing and analytics requirements are supported by these IoT cloud services. They use cloud-based data processing and analytics to support the IoT by facilitating intelligent, adaptive, and autonomous device operations.

Edge devices are the IoT-enabled devices that are connected to the IoT cloud. Besides being connected to the fog, these devices all share one thing in common: They generate data through any number of appliances, including sensors, odometers, cameras, contact sensors, pressure sensors, laser scanners, thermometers, smoke detectors, microphones, electric meters, gas meters, water flow meters, and much more.

In fact, most edge devices come equipped with device-embedded applications that are capable of processing and deriving insights locally, using the data that's created by device appliances in real-time. Local data processing and analytic deployment is called edge processing, and it helps save resources by

» Detecting data that is useful to the analytics operations running on the device and discarding the rest: This lowers the data transfer and storage overhead.

» Handling analytic deployments locally, doing away with the need to transfer data to and from the cloud: A side benefit of these device-embedded analytics applications is that they return results faster than if the data is processed in the cloud.
An overview of popular machine learning methods for data science in IoT. Figure 8-1 illustrates some of these components to help pull them together into a conceptual schematic.

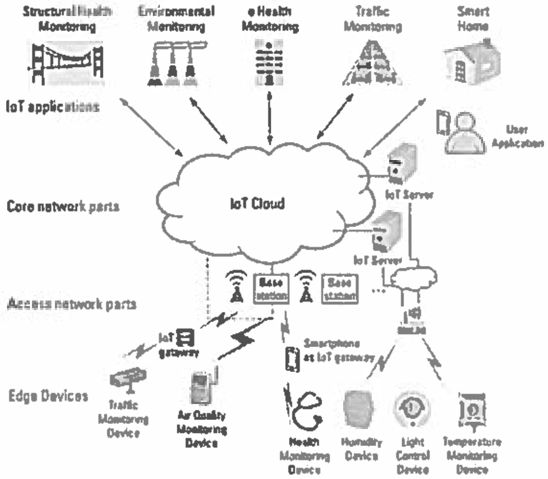


FIGURE 8-1: Conceptual schematic of the IoT network.

- Spark streaming for the IoT

Spark is an ideal framework for integrated real-time big data processing and analysis. With respect to the IoT, each IoT sensor stream can be transformed into Spark DStreams — discreet data streams that are the fundamental data abstraction in the Spark Streaming module (the module where data processing is carried out).You can use the Spark Streaming window operations on DStream sources to quickly and easily aggregate processing and alerting to any regular time intervals of your choosing. Lastly, for comparative analytics, you can use Spark's Resilient Distributed Datasets (RDD) — an immutable collection of objects, and a fundamental Spark data structure — to store any relevant historical datasets in-memory.

11. b. Differentiate Random forest and Decision tree methods

Categorizing Data with Decision Tree and Random Forest Algorithms

A decision tree algorithm works by developing a set of yes-or-no rules that you can follow for new data to see exactly how it will be characterized by the model. But you must be careful when using decision tree models, because they run the high risk of error propagation, which occurs whenever one of the model rules is incorrect. Errors are generated in the results of decisions that are made based on that incorrect rule, and then propagated through every other subsequent decision
made along that branch of the tree.To illustrate this type of algorithm, consider a dataset that's often used in machine learning demonstrations — the list of passenger names from the Titanic. Using a
simple decision tree model, you can predict that if a passenger was female or was a male child with a large family, he or she probably survived the catastrophe. Figure 6-6 illustrates this determination.

FIGURE 6-6:
A decision tree
model predicts
survival rates
from the Titanic
catastrophe

Lastly, random forest algorithms are a slower but more powerful alternative. Instead of building a tree from the data, the algorithm creates random trees and then determines which one best classifies the testing data. This method eliminates the risk of error propagation that is inherent in decision tree models.

12. a. Explain the designing meet the needs of your target audience
Focusing on Your Audience

To make a functional data visualization, you must get to know your target audience and then design precisely for their needs. But to make every design decision with your target audience in mind, you need to take a few steps to make sure you really understand your data visualization's target consumers.To gain the insights you need about your audience and purpose, follow this process:

✓ Brainstorm. Think about a specific member of your visualization's audience and make as many educated guesses as you can about that person's motivations.

✓ Define the purpose of your visualization. Narrow the purpose of your visualization by deciding exactly what action or outcome you want your audience members to make as a result of your visualization.

✓ Choose a functional design. Review the three main data visualization types (which I discuss in the preceding sections) and decide which type can best help you achieve your desired outcome.

The following sections delve into these processes in more detail.

Step one: Brainstorming about Brenda

To do a proper brainstorming, start by getting out a sheet of paper and really thinking about your imaginary audience member, Brenda. Answer the follow-ing questions to help you better understand her, and thus better understand and design for your target audience.Form a picture of what Brenda's average day looks like — what she does when she gets out of bed in the morning, what she does over her lunch hour, and what her workplace is like. Also consider how Brenda will use your visualization.

To form a comprehensive view of who Brenda is and how you can best meet her needs, consider the following questions:

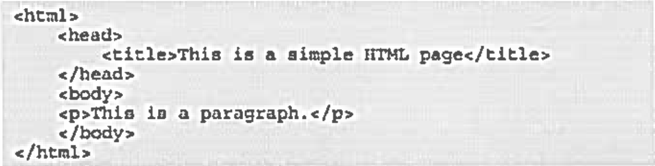
✓ Where does Brenda work? What does Brenda do for a living?

✓ What kind of technical education or experience, if any, does she have?

✓ How old is Brenda? Is she married? Does she have children? What does Brenda look like? Where does she live?

- Bringing in the HTML and DOM

HyperText Markup Language (HTML) is the backbone of a web page. It delivers the static content you see on many websites, especially the older ones. HTML is recognizable by its plain text and limited interactivity. The only interactive features you get with plain HTML websites are perhaps some hyperlinks that lead you to other boring static pages throughout the site.You can use HTML to display plain text with a series of tags that give instruc-tions to the client's browser. The following HTML code is pretty basic, but at least it gives you an idea of what's involved:



```
<html>
    <head>
        <title>This is a simple HTML page</title>
    </head>
    <body>
    <p>This is a paragraph.</p>
    </body>
</html>
```

In D3.js, the purpose of HTML is to provide a bare scaffold of static tags and web page content that can be interacted with via JavaScript's DOM to produce dynamic, changeable HTML pages. D3.js is built on top of a bare backbone of HTML. Although HTML is static, it becomes dynamic in D3.js if a programmer or user interaction causes predetermined scripts to make on-the-fly changes to the underlying HTML code. This means that the HTML that is displayed is often dynamic and different from that which was originally sent to the browser.

- Bringing in the JavaScript and SVG

Using the JavaScript language gives you a simple way to get work done client-side (on the user's machine). The slowest part of any interaction between a user and a website is in sending data from the server to the client's computer over the Internet. That interaction can be vastly accelerated if, instead of sending all of the information needed for a browser display, you send a much shorter, simpler chain of instructions that the client's web browser can use to re-create that information and then create the web page using the client computer's own processing speed. This is how client-side work is carried out.

In JavaScript, graphics rendering is based on Scalable Vector Graphics (SVG) — a vector image format that delivers images to interactive visualizations and web-based animations. In D3.js, SVG functions as a file format that stores vector graphics for use in two-dimensional, interactive web-based data visualizations. Vector graphics require a lot less bandwidth than images because vector graphics contain only instructions for how to draw them, as opposed to the final pixel-by-pixel raster renderings of images. If your goal is to rapidly deploy web-based graphics that also provide you lossless scaling capabilities, then SVG is a perfect solution. SVG is optimal for use in creating graphical elements such as bars, lines, and markers.

- Bringing in the Cascading Style Sheets (CSS)

The purpose of using Cascading Style Sheets (CSS) is to define the look of repeated visual elements, such as fonts, line widths, and colors. You can use CSS to specify the visual characteristics of your page elements all at one time, and then efficiently apply these characteristics to an entire HTML document (or to only the parts of the document defined in the DOM, if you wish). If you want to make sweeping, all-at-once changes to the look and feel of your web page elements, then use CSS.

As an example, the basic CSS for a simple web page might include the following:

✓ What social, political, caused-based, or professional issues are important to Brenda? What does Brenda think of herself?

✓ What problems and issues does Brenda have to deal with on a daily basis?

✓ How does your data visualization help solve Brenda's work problems or her family problems, or improve her self-esteem?

✓ Through what avenue will you present the visualization to Brenda — for example, through the Internet or in a staff meeting?

✓ What does Brenda need to be able to do with your data visualization? Spend some time thinking about Brenda (your target audience) and answering these questions. These answers can help you create a more functional and effective data visualization.

Step two: Defining your purpose

After you brainstorm about your typical audience member (see the preceding section), you can much more easily pinpoint exactly what you're trying to achieve with this data visualization. Are you attempting to get your consumer to feel a certain way about themselves or the world around them? Are you trying to make a statement? Are you seeking to influence organizational decision makers to make good business decisions? Or do you simply want to lay all the data out there, for all viewers to make sense of and deduce from what they will?

Returning now to our hypothetical Brenda . . . What decisions or processes are you trying to help her achieve? Well, you need to make sense of her data, and then you need to present it to her in a way that she can clearly understand. What's happening within the inner mechanics of her department?

Through your visualization, you seek to guide Brenda into making the most prudent and effective management choices.

Step three: Choosing the most functional visualization type for your purpose

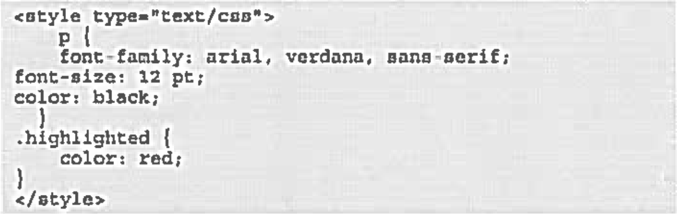
Keep in in mind that you have three main types of visualization from which to choose: data storytelling, data art, and data showcasing. If you're designing for organizational decision makers, then you're most likely going to want to use data storytelling to directly tell your audience what their data means with respect to their line of business. If you're designing for a social justice organization or a political campaign, then data art can best make a dramatic and impactful statement with your data. Lastly, if you're designing for engineers, scientists, or statisticians, then stick with data showcasing so that these analytical types have plenty of room to figure things out on their own.

Referring back to Brenda, because she's not extraordinarily analytical and because she's depending on you to help her make excellent data-drive decisions, you need to employ data storytelling techniques.

Create either a static or interactive data visualization with some, but not too much, context.

The visual elements of the design should tell a clear story so that Brenda doesn't have to work through tons of complexities to get the point of what you're trying to tell her about her data and her line of business.

12. b. Write about D3.js web applications



```
<style type="text/css">
    p {
    font-family: arial, verdana, sans-serif;
font-size: 12 pt;
color: black;
   }
.highlighted {
    color: red;
}
</style>
```

D3.js leverages CSS for drawing and styling text elements and drawn elements so that you can define and change the overall look of your visualization in one compact, easy-to-read block of code.

- Bringing in the web servers and PHP

While one of the main purposes behind using JavaScript and D3.js is to maximize the portion of work that's carried out on the client's machine, some work is just better carried out on the web server. (In case you aren't familiar with the term, a web server can be thought of as a server-based computer that sends information over the Internet to users when they visit a website.)In web programming, you commonly have a SQL database set up as a main information source and also a PHP program that defines the HTML code to be sent to the client's computer. You use PHP programs to query the SQL database and determine what information to send over to the client. PHP is a scripting language that's run on a server and produces on-the-fly HTML code in response to user interactions.

In pre-D3.js days, you'd have had to use a lot more time and bandwidth constructing web-based, interactive data visualizations. Due to the effectiveness of the PHP/D3.js combination, things are simpler now. In response to a request from the user, PHP selects information from the server and sends it to the client's computer in the form of HTML with embedded CSS, JavaScript, and D3.js code. At this point, the D3.js can take over and expand the HTML. If necessary, D3.js can even make on-the-fly HTML expansions in response to additional user interactions. This process uses only a fraction of the bandwidth and time that would have been required in a PHP-only or JavaScript-only setup.

Understanding More Advanced Concepts and Practices in D3□js

- Getting to know chain syntax

you can use D3.js to take a bare scaffold of HTML and turn it into a complex visualization by modifying page elements. The D3.js library uses a very efficient operator syntax called chain syntax. The purpose of chain syntax is to chain together several methods, thereby allowing you to perform multiple actions using only a single line of code. Instead of name-value pair syntax (like what you would see in CSS), D3.js chains multiple expressions together, each one creating a new object and selection for the next.

The fundamental concept behind D3.js is the Append, Enter, and Exit selec-tions. These methods select, add, or remove HTML tags that are bound, or assigned, to your data. The Append selection refers to existing data elements that are paired with existing HTML tags. When there are more data elements than tags, the Enter selection adds tags paired with the surplus data elements. When there are more tags than data elements, you can use the Exit selection to remove those tags.Taking another look at a section of Listing 10-1, notice the code block that draws the bars of the graph:

```
svg_container.selectAll("rect")
    .data(column_data, position)
    .enter()
    .append("rect")
    .attr("x", function (d, i) {
    return scale_x(i);
```

- Getting to know scales

In D3.js, the scale function plots input domains to output ranges so that the output data visualization graphics are drawn at appropriate, to-scale proportions. Looking at the following section of Listing 10-1, notice how scale vari-ables are defined using D3.js:

```
var scale_y = d3.scale.linear()
    .domain([0, d3.max(column_data, function (d) {
    return d.quantity;
})])
    .range([0, total_height]);

var scale_x = d3.scale.ordinal()
    .domain(d3.range(column_data.length))
    .rangeRoundBands([0, total_width], 0.05);
```

One of the main features of D3.js is its ability to do difficult and tedious calculations under the hood. A key part of this work is done in scaling plots. If you want to automatically map the range of your data to actual pixel dimensions in your graph, then you can use the scale function to change either or both parameters without having to do any manual recalculations.

- Getting to know transitions and interactions

The true beauty of D3.js is in how you can use it to easily incorporate dynamic elements into your web-based data visualization. If you want to encourage your users to really explore and analyze the data in your dynamic visualization, then create features that offer a lot of user interactivity. Also, incorporating transitions into your dynamic visualization can help you capture the interest of your audience. Transitions in D3.js build the aesthetic appeal of a data visualization by incorporating elements of motion into the design.As a prime example of D3.js interactivity, take a look at the following code from Listing 10-1.

```
<style type='text/css'>
    rect:hover {
    fill: brown;
}
  </style>
```

Here, a single piece of CSS code changes the color of the bars when the user hovers the cursor over them.And, looking at another snippet taken from Listing 10-1 (shown below), you can see code that defines a sort function and then creates buttons to transition the bars between sorted and unsorted states. If you tried for the same effect using vanilla JavaScript, it would be more tedious and time-consuming.

```
var sort = function () {
    bars_to_sort = function (a, b) {
        return b.quantity - a.quantity;
    };
```

13. a. Explain designing data visualizations for collaboration

Using Collaborative Data Visualization Platforms

Collaborative data visualization platforms are web-based platforms through which you can design data visualizations and then share those visualizations with other platform users to get their feedback on design or on the data insights conveyed.

Collaborative data visualization platforms have been described as the YouTube of data visualization, but actually, these platforms are far more interactive than YouTube. Collaborative data visualization platforms are like a version of YouTube that lets you instantly copy and edit every video using your own software tools, and then republish the video through your own social channels.

Collaborative platforms are very efficient and effective for working in teams. Instead of having to e-mail versions back and forth, or (heaven forbid) learn a dedicated version-control system like GitHub, you and your teammates can use the platform's sharing features to work on visualizations as a team.

- Working with IBM's Watson Analytics

The brand name "IBM" is synonymous with cutting-edge ingenuity, quality, and stability. IBM Watson Analytics is no exception to this rule. Watson Analytics is the first full-scale data science and analytics solution that's been made available as a 100% cloud-based offering. The application is available in a freemium version, or on a paid-subscription basis, to match all levels of needs and requirements. Although Watson Analytics is a full-scale data science accelerator, it also offers robust collaborative functionality so that users can share, interact, engage, and provide feedback on data insights that are generated.

IBM's purpose for building Watson Analytics was to democratize the power of data science by offering a platform across which users of all skill levels and backgrounds can access, refine, discover, visualize, report, and collaborate upon data-driven insights. No matter whether you're an advanced data scientist that's looking for automation functionality to make your daily work easier, or if you're an entry-level data analyst that's looking for a way to generate deep data insights without learning to code things up yourself, Watson Analytics has something for everyone. The pre-requisite skill level is simply that users know how to use a basic application like Microsoft Excel to do advanced data analysis — such as that which is done through pivot tables and custom VBA scripts.

- Visualizing and collaborating with Plotly

The Plotly collaborative platform aims to accommodate the data collaboration needs of professionals and non-professionals alike. This powerful tool doesn't stop at data visualization; it goes one step further by providing you the tools you need to make sense of your data through advanced statistical analysis. Plotly even offers seamless integration with dedicated programming environments like Python, MATLAB, and R.

If you want a quick and easy way to create interesting and attractive data visualizations, Plotly offers a great solution. Although Plotly focuses on traditional data chart types, you can much more easily portray variables by size or color in Plotly than in most other web applications. If you work in one of the STEM fields, Plotly may be particularly well-suited for your needs. In addition to standard bubble plots, line charts, bar charts, and area graphs, Plotly also offers histograms, two-dimensional histograms, heat maps, scatter charts, box plots, three-dimensional scatter charts, three-dimensional surface charts, and polar plots.Figures 11-3 and 11-4 show two visualizations created with Plotly — a simple bar chart of the in-county moving data and a scatter chart of that data com-pared to the area in square miles of each state.

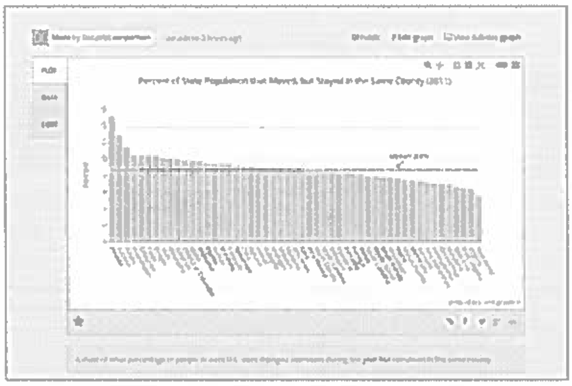


Figure 11-3: A bar chart on Plotly.

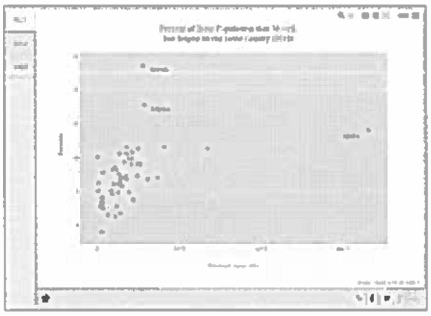


Figure 11-4: A scatter chart on Plotly.

- Visualizing Spatial Data with Online Geographic Tools

With the advent of online Geographic Information Systems (GIS, for short) like Google Maps, Open Street Map, and Bing Maps, geographic data visualization is no longer solely reserved for cartographers and GIS gurus. Web-based mapping applications have now made it possible for data enthusiasts from a wide range of backgrounds to quickly and easily analyze and map spatial data.

The purpose behind all web-based geographic applications is to visually present geographic data — quantitative and qualitative data that's associated with particular locations.

In web-based geographic data visualization, you're likely to represent areas using either a categorical fill or a chloropleth map. A categorical fill is a way of visually representing qualitative attributes of your spatial dataset. So, for example, when you are looking at a map that shows an election outcome,
states with a majority of Democrat votes are colored blue, and states with a majority of Republican votes are colored red. The categorical attribute is "Political Party", and the fill color is determined by whether the value of that categorical attribute is "Republican" or "Democrat". On the other hand,
choropleths are map representations where spatial areas are filled with a par-ticular hue or intensity of color to represent the comparative distribution of your data quantities across space.

Web-based geographic data visualizations depend heavily on geocoding — the automatic association of data points with geographic points, based on the location information that you provide. If you have a column of state names, or even street addresses, web applications generally can auto-map that data for you.

- Making pretty maps with OpenHeatMap

OpenHeatMap is a user-friendly service that allows you to upload and geocode spatial data. OpenHeatMap can automatically geocode spatial identifiers, requiring only minimal user oversight. It's not as versatile as Google Fusion Tables or CartoDB, but it's so easy to use that many people consider it their favorite web-based mapping application. A unique feature of OpenHeatMap is that it doesn't offer user accounts. Anyone and everyone can upload data and use the service anonymously.

- Map-making and spatial data analytics with CartoDB

If you're not a professional programmer or cartographer, CartoDB is about the most powerful online mapping solution that's available. People in information services, software engineering, media and entertainment, and urban development industries often use CartoDB for digital visual communications.
By using CartoDB, you can create a heat map simply by uploading or linking to a list of spatial coordinates. Likewise, if you want to create a choropleth map to show values for quantitative attributes, then simply upload or link to a set of spatial coordinates that includes attribute data.CartoDB allows you to overlay markers and shapes on all sorts of interesting base maps. You can use it to make anything from simple outline maps of geo-graphic regions to stylish, antiqued, magazine-style maps. You can even use it to generate street maps from satellite imagery. CartoDB's geocoding functionality is so well-implemented that you can drill down to a location using individual addresses, postal codes, or even IP addresses.

13. b. Analyze map projections and co ordinate systems

- Understanding map projections and coordinate systems

Map projections and coordinate systems give GIS a way to accurately represent a round Earth on a flat surface, translating the Earth's arced three-dimensional geometry into flat two-dimensional geometry. Projections and coordinate systems project spatial data. That is to say, they forecast and predict accurate spatial positions and geographic scale, depending on where those features are located on Earth. Although projection and coordinate systems are able to project most features rather accurately,they don't offer a one-size-fits-all solution. If features in one region are projected perfectly at scale, then features in another region are inevitably projected with at least a slight amount of distortion. This distortion is sort of like looking at things through a magnifying glass — you can see the object
in the center of the lens very accurately and clearly, but the objects on the outer edge of the lens always appear distorted. No matter where you move the magnifying glass, this fact remains unchanged. Similarly, you can't repre-sent all features of a rounded world accurately and to-scale on a flat map.Coordinate systems are referencing systems that are used to define a fea-ture's location on Earth. There are two types of coordinate systems:

✓ Projected Coordinate Systems: Also called map projections, projected coordinate systems are mathematical algorithms that you can use to transform the location of features on a round Earth to equivalent positions represented on a flat surface instead. The three common projec-tion types are cylindrical, conical, and planar.

✓ Geographic Coordinate Systems: A coordinate system that uses sets of numbers and/or letters to define every location on Earth. In geographic coordinate systems, location is often represented by latitude/longitude, decimal degrees, or degrees-minutes-seconds (if you're familiar with old fashioned surveying nomenclature).Figure 13-7 shows these three types in all their glory.

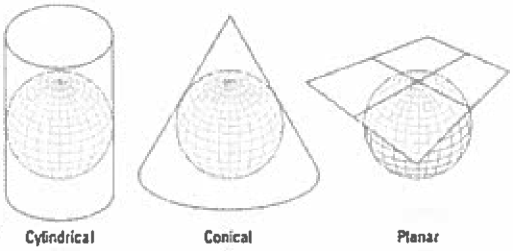


Figure 13-7: Three common projection types (left to right): cylindrical, conical, and planar.

In almost all cases, when you import a spatial dataset into GIS, it comes in with its own pre-defined coordinate system. The GIS software then adopts that coordinate system and assigns it to the entire project. When you add additional datasets to that project in the future, they may be using that same coordinate system or an alternative one. In cases where the new data is coming in with a coordinate system that's different from that of the project, the GIS software will transform the incoming data so that it is represented correctly on the map.

14. a. How to Finding and telling your data story in journalism

Finding and Telling Your Data's Story

- Spotting strange trends and outliers

One quick way to identify interesting stories in a dataset is to do a quick spotcheck for unusual trends or extreme outliers. These anomalies usually indicate that an external force is affecting a change that you see reflected in the data.If you want to do a quick-and-dirty spot-check for easy-to-identify stories, you can simply throw your data into an x-y scatter plot and visually inspect the result for obvious trends and outliers. After you spot these anomalies, look into reasons behind why the data behaves oddly. In doing so, you can usually uncover some juicy stories.

Illustrating this fact, consider the World Bank Global Development Indicator (GDI) open dataset. Looking at this data, you can easily see a clear correlation between a country's gross domestic product and the life expectancy of its citizens. The reason for this correlation is obvious: More affluent people can afford better healthcare.But say you're searching through the hundreds of GDI indicators for the year 2013 and you come across something less obvious — the survival rate of newborns is reasonably well-correlated with the percentage of employed females who receive wages or salaries instead of only performance-based remuneration, as illustrated in Figure 18-1.

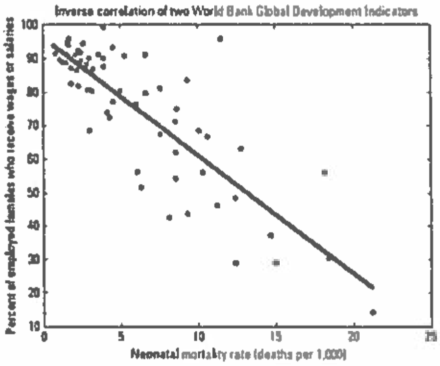


FIGURE 18-1: A scatter plot of the inverse correlation between two GDI indicators.

The relationship in this data is a little murky. Although you naturally expect two metrics based on health and economic well-being to be related, after analyzing your data a bit, you find a Pearson correlation coefficient of 0.86. That's quite high. Is there a story here? Does this qualify as a newsworthy trend? An effective and time-efficient way to explore answers to this question is to try to find the exception that proves the rule. In Figure 18-2, the simple least-squares best-fit line is in black, and the two data points that most differ (horizontally and verti-cally) from this line are indicated with light gray lines.

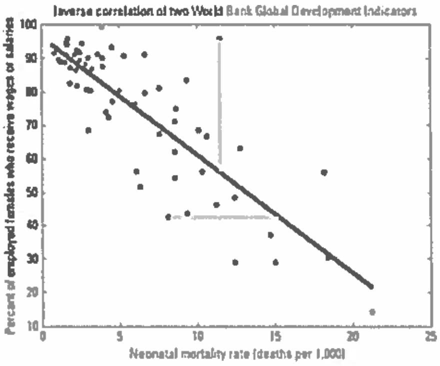


FIGURE 18-2: A least-squares line of the inverse correlation between two GDI indicators.

- Examining context to understand the significance of data

By pinpointing strange trends or outliers in your dataset, you can subsequently focus in on those patterns and look for the interesting stories about the external factors that cause them. If you want to cultivate the most thought-provoking story about what's happening in your source dataset, you need to further investigate, compare, and contrast these factors you've identified. By examining the context in which competing causative factors are creating extreme trends and outliers, you can then begin to get a solid understanding of your data's significance.

For example, think about the World Bank Global Development Indicator (GDI) in the preceding section. The topmost point in Figure 18-2 represents the country of Jordan. For a given level of child mortality, Jordan has an unusual number of women with predictable income. If you dig a little deeper into factors that might account for this outlier, you see that the overall employment rate for women in Jordan is among the lowest in the world. In a country where few women work, the women who do work in Jordan are earning relatively stable wages or salaries. This might indicate that the precariously paid work is being given mostly to men.

Maybe the underlying story here is about gender roles in Jordan? If so, what conclusions could you draw by looking at another outlier country in the dataset —
Peru, for example?Well, because only 41 percent of Peruvian women are employed with a stable income, Peru is near the bottom ranks in terms of employed women with stable incomes. Perhaps this is to be expected in a country with so much agriculture by hand. And, in all honesty, Peruvian men aren't much better off either, indicated by the fact that only 51 percent of them are reporting stable employment. But the Peruvian neonatal mortality rate is unusually low. Does Peru have an exceptionally well-funded healthcare system? Not really — it spends less, per capita, on healthcare than most of its neighbors.

- Emphasizing the story through visualization

When it comes to data-visualization design, you always want to pick the colors, styles, and data graphic types that most dramatically convey the visual story you're trying to tell. You want to make your visual stories as clear, concise, and easy to understand as possible. In making selections among visual elements for your data visualization, find the best way to visually convey your story without requiring your audience to have to strain and work to understand it.

Looking back to the World Bank Global Development Indicator (GDI) example from the preceding sections, imagine that you decide to go with the Peru story. Because the SIGI dataset is so relevant to the Peru story, you need to make a thorough study of that dataset, as well as any other datasets you've identified to be relevant. Time-series data on different statistics is likely to be quite informative because it should indicate several relevant metrics — the proportion of total income earned by women, survival rates of pregnant mothers, legal-based gender equality metrics, and so on.
After you gather and appraise the most relevant metrics that are available to you, pick and choose the metrics whose datasets show the most extreme trends. Subtle changes in data don't lend themselves to dramatic and easy-to-understand visual stories. After you select the most dramatic metrics to use in telling your story, it's time to decide the best way to represent that story visually.

- Creating compelling and highly focused narratives

As you well know, no one wants to wade through a bunch of needless, complicated words to try to figure out what your story says. It's frustrating, and it simply takes too much work. Presuming that your purpose for creating a data-driven story is to publish something that has impact and value in the lives of your readers, you must work hard to whittle your narrative down to its simplest, most-focused form.
Failure to do so decreases the impact and performance of your data-driven story.

Narrow each of your stories down to its hook and lede before going any further into the process of writing a full narrative. In journalism, a hook is a dramatic angle that cultivates interest in potential readers and draws them into your piece. A lede is the first sentence of your story — it introduces the story and shows readers why the story is newsworthy. After you go through your story and flesh out a hook, a lede, and a full narrative, you always need to go back through the piece once or twice and cut unnecessary words or restructure sentences so that they most directly express the ideas you seek to convey.

14. b. Explain modeling natural resources in Raw

Modeling Natural Resources in the Raw

You can use data science to model natural resources in their raw form. This type of environmental data science generally involves some advanced statistical modeling to better understand natural resources. You model the resources in the raw — water, air, and land conditions as they occur in nature — to better under-stand the natural environment's organic effects on human life.

- Exploring natural resource modeling

Environmental data science can model natural resources in the raw so that you can better understand environmental processes in order to comprehend how those processes affect life on Earth. After environmental processes are clearly understood, then and only then can environmental engineers step in to design systems to solve problems that these natural processes may be creating. The following list describes the types of natural-resource issues that environmental data science can model and predict:

» Water issues: Rainfall rates, geohydrologic patterns, groundwater flows, and groundwater toxin concentrations

» Air issues: The concentration and dispersion of particulate-matter levels and greenhouse gas concentrations

» Land issues: Soil contaminant migration and geomorphology as well as geophysics, mineral exploration, and oil and gas exploration.

- Dabbling in data science

Environmental processes and systems involve many different interdependent variables, most natural-resource modeling requires the use of incredibly complex statistical algorithms. The following list shows a few elements of data science that are commonly deployed in natural-resource modeling:

» Statistics, math, and machine learning: Bayesian inference, multilevel hierarchical Bayesian inference, multitaper spectral analysis, copulas, Wavelet Autoregressive Method (WARM), Autoregressive Moving Averages (ARMAs), Monte Carlo simulations, structured additive regression (STAR) models, regression on order statistics (ROS), maximum likelihood estimations (MLEs),
expectation-maximization (EM), linear and nonlinear dimension reduction, wavelets analysis, frequency domain methods, Markov chains, k-nearest neighbor (kNN), kernel density, and logspline density estimation, among other methods

» Spatial statistics: Generally, something like probabilistic mapping

» Data visualization: As in other data science areas, needed for exploratory analysis and for communicating findings with others

» Web-scraping: Many times, required when gathering data for environmental models

» GIS technology: Spatial analysis and mapmaking

» Coding requirements: Using Python, R, SPSS, SAS, MATLAB, Fortran, and SQL, among other programming languages.

- Modeling natural resources to solve environmental problems

The work of Columbia Water Center's director, Dr. Upmanu Lall, provides a world-class example of using environmental data science to solve incredibly complex water resource problems. Dr. Lall uses advanced statistics, math, coding, and a staggering subject-matter expertise in environmental engineering to uncover complex, interdependent relationships between global water-resource characteristics, national gross domestic products (GDPs), poverty, and national energy consumption rates.With respect to data science technologies and methodologies, Dr. Lall implements these tools:

» Statistical programming: Dr. Lall's arsenal includes multilevel hierarchical Bayesian models, multitaper spectral analysis, copulas, Wavelet Autoregressive Moving Averages (WARMs), Autoregressive Moving Averages (ARMAs), and Monte Carlo simulations.

» Mathematical programming: Tools here include linear and nonlinear dimension reduction, wavelets analysis, frequency domain methods, and nonhomogeneous hidden Markov models.

» Clustering analysis: In this case, Dr. Lall relies on the tried-and-true methods, including k-nearest neighbor, kernel density, and logspline density estimation.

» Machine learning: Here, Dr. Lall focuses on minimum variance embedding.

15. a. Explain Angling in on analytics in E-commerce

- Angling in on analytics

Web analytics can be described as the practice of generating, collecting, and making sense of Internet data in order to optimize web design and strategy. Configure web analytics applications to monitor and track absolutely all your growth tactics and strategies, because without this information, you're operating in the dark — and nothing grows in the dark.

Appraising popular web analytics applications

Data scientists working in growth hacking should be familiar with (and know how to derive insights from) the following web analytics applications:

» Google Analytics : A free, easy-to-use, powerful web analytics tool, Google Analytics is great for monitoring not only the volumes of traffic that come to your website over time but also the demographics and summary statistics on your visitors, your website referral sources, your visitor flow patterns, real-time visitor behavior analytics, and much more. Google Analytics can show you benchmarking analytics that provide insights about how your website's performance compares to the performance of other websites in your industry.

» Adobe Analytics : You can use Adobe Analytics for marketing attribution, mobile app performance, social media marketing performance, return-on-investment (ROI) investigation, and real-time visitor monitoring.

» IBM Digital Analytics : The perfect platform for integrating performance data from all your business's web channels — from data generated by website guests visiting using personal computers to mobile visitor statistics, and even social media channel performance — IBM Digital Analytics offers powerful analytics capabilities to keep you informed of real-time and historical visitor behaviors, as well as relevant cross-channel interactions. The platform also offers marketing attribution and tag management capabilities.

» Webtrends : Offering advanced multichannel analytics, real-time visitor behavior monitoring, and the technology you need to reclaim lost sales from shopping cart abandonment via email remarketing tactics, Webtrends is a powerhouse web analytics application. It even goes the extra mile by offering a campaign optimization feature that you can use to track, monitor, and optimize your search engine marketing efforts, as well as your search and social advertisement campaigns.

» Google Tag Manager : Website tags — code snippets that collect data for use in your third-party analytics applications — can help you measure and manage the effectiveness of your Internet marketing campaigns, but the process of deploying tags is error-prone and requires coding. Google Tag Manager is a free tag-management tool that offers a code-free interface and a rules-based system that allows you to easily manage and deploy your website marketing and tracking tags.

» Assorted social analytics tools: In addition to the more heavyweight offerings described in this list, you can find many free, easy-to-use social analytics applications to monitor and measure the effectiveness of your social media growth initiatives. These include Sendible , which has ample options for tracking statistics from your Twitter, Facebook Page, Instagram, and Google Analytics metrics on one custom dashboard; Facebook Page Insights ; Pinterest Analytics; Iconosquare Statistics for Instagram; and Google URL Shortener for link tracking .

15. b. Explain spatial crime prediction and monitoring

Spatial Crime Prediction and Monitoring

Spatial data is tabular data that's earmarked with spatial coordinate information for each record in the dataset. Many times, spatial datasets also have a field that indicates a date/time attribute for each of the records in the set — making it spatio-temporal data. If you want to create crime maps or uncover location-based trends in crime data, use spatial data analysis. You can also use spatial analysis methods to make location-based inferences that help you monitor and predict what crimes will occur where, when, and why. In the following sections, I show how you can use GIS technologies, data modeling, and advanced spatial statistics to build information products for the prediction and monitoring of criminal activity.

- Crime mapping with GIS technology

One of the most common forms of data insight that's used in law enforcement is the crime map. A crime map is a spatial map that visualizes where crimes have been committed during any given time interval. In olden days, you might have drawn this type of map out with pencil and paper, but nowadays you do the job using a GIS software, such as ArcGIS Desktop or QGIS.

- Going one step further with location-allocation analysis

Location allocation is a form of predictive spatial analytics that you can use for location optimization from complex spatial data models. For example, in law enforcement, location optimization can predict optimal locations for police stations so that dispatched officers can travel to an emergency in any part of the city within a 5-minute response-time window. To help your agency predict the best locations to position officers so that they can arrive immediately at any emergency in any part of town, use location-allocation analysis.

- Analyzing complex spatial statistics to better understand crime

29/12/2022

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

<table>
<tr><td>Degree</td><td>B. Tech. (U. G.)</td><td>Program</td><td colspan="3">Civil Engineering</td><td>Academic Year</td><td>2022- 2023</td></tr>
<tr><td>Course Code</td><td>20CE303</td><td>Test Duration</td><td>3 Hrs.</td><td>Max. Marks</td><td>70</td><td>Semester</td><td>III</td></tr>
<tr><td>Course</td><td colspan="7">SURVEYING</td></tr>
</table>

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Mention some of the points which chainmen should know, concerning the errors in chaining.	20CE303.1	L1
2	What principles are used in surveying?	20CE303.2	L1
3	List the types of curves.	20CE303.3	L1
4	Define vector.	20CE303.4	L1
5	What are the distinguishing features of objects?	20CE303.5	L1

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

<table>
<tr><th>No.</th><th>Questions (6 through 15)</th><th>Marks</th><th>Learning Outcome (s)</th><th>DoK</th></tr>
<tr><td>6</td><td>The fore and back bearings of a closed traverse conducted at a place are given below. Indicate which stations are affected by local attraction. Also find out the correct bearings.
<table>
<tr><th>Line</th><th>Fore Bearing</th><th>Back Bearing</th></tr>
<tr><td>AB</td><td>S54°20'E</td><td>N54°20'W</td></tr>
<tr><td>BC</td><td>N67°30'E</td><td>S66°20'W</td></tr>
<tr><td>CD</td><td>N48°30'W</td><td>S44°20'E</td></tr>
<tr><td>DA</td><td>S21°25'W</td><td>N19°40'E</td></tr>
</table></td><td>12 M</td><td>20CE303.1</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>7 (a)</td><td>Explain how details can be surveyed by offsets from survey lines. Discuss the relative merits of different types of offsets</td><td>9M</td><td>20CE303.1</td><td>L2</td></tr>
<tr><td>7 (b)</td><td>Distinguish between resection and intersection methods as applied to plane table surveying.</td><td>3M</td><td>20CE303.1</td><td>L1</td></tr>
<tr><td>8</td><td>The following series of readings of back sights and fore sights was taken in a fly leveling. The first reading was taken on a point of RL 100.000 m. Draw a page of leveling field book and enter readings on it. Find the reduced levels of all points using any method. Apply check.
2.228, 1.606, 0.988, 2.090, 2.864, 1.262, 0.602, 1.982, 1.044, 2.684.
The instrument having moved after 3, 6 and 8th readings</td><td>12M</td><td>20CE303.2</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>9 (a)</td><td>What is meant by the least count of a vernier? Draw a vernier scale to read 78°40'20". The main scale is graduated in 20' parts and vernier scale in 20".</td><td>6M</td><td>20CE303.2</td><td>L3</td></tr>
<tr><td>9 (b)</td><td>Explain how a subtense bar is used to determine horizontal angle.</td><td>6M</td><td>20CE303.2</td><td>L2</td></tr>
<tr><td>10</td><td>Explain the components of simple circular curve with a neat sketch.</td><td>12M</td><td>20CE303.3</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>11</td><td>It was required to obtain the elevation of the top of a tower located on the roof of a building. Since the direct measurement was not possible, the following data was obtained. A line AB, 120.0 m long was staked out and the horizontal angles to the</td><td>12M</td><td>20CE303.3</td><td>L3</td></tr>
</table>

	tower were observed at A as 56° and at B as 32°. At point B a back sight of 2.000 m was taken on a BM of elevation 100.000mand the vertical angle to the top of tower was found to be 58°. Calculate the elevation of the top of tower.			
12	What are the three segments of GPS? Describe them.	12M	20CE303.4	L2
	OR			
13 (a)	Determine the number of photographs required to cover 300 km^2 if the scale of photograph is 1 in 12000 and photograph format is 250 mm × 250 mm. The end overlap and side overlap are 60% and 30%.	6M	20CE303.4	L3
13 (b)	Explain the steps involved in creation of map projection.	6M	20CE303.4	L2
14 (a)	What is raster overlay operation? Explain.	6M	20CE303.5	L1
14(b)	What is supervised classification? What are the basic steps and stages involved in a typical supervised classification?	6M	20CE303.5	L1
	OR			
15 (a)	What do you understand by spatial data and how are they integrated to make a GIS?	6M	20CE303.5	L2
15 (b)	Discuss overlay using a decision table.	6M	20CE303.5	L1

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION ①

Couse Code: 20CE303; Course: surveying key

Part-A

1) Mention some of the Points which Chain men should know, concerning the errors in chaining.

Ans: → Errors due to inefficient Ranging
→ Errors due to inefficient straightening
→ Errors due to Careless holding and Markings
→ Errors due to Variation in Pull

} Any two
1×2 = 2marks

2Ans Priniciples of surveying

→ To work from Whole to Part
→ To locate a new station by at least two measurement (linear or angular) from fixed reference Point.

3Ans Types of curves:

1) Horizontal curves: 1) simple curve
2) compound curve
3) Reverse curve
4) Transition curve
5) Combined curve

2) Vertical curve: 1) summit curve
2) valley curve

4 Ans Vector - A Coordinate - based data model that represents geographic ②
features such as Points , lines and Polygons.

5 Ans The main features of objects

→ Points
→ nodes
→ arcs and.
→ Polygons
organized in to coverages.

Part-B convert WCB

6 Ans

Line	FB	BB
AB	$125^0 40'$ ✓	$305^0 40'$ ✓
BC	$67^0 30'$ ✓	$246^0 20'$
CD	$311^0 30'$	$135^0 40'$
DA	$201^0 25'$	$19^0 40'$

FB of BC = $67^0 30'$

BB of BC = $67^0 30' + 180^0 = 247^0 30'$

Correction at C = $247^0 30' - 246^0 20'$

= $1^0 10'$

FB of CD = $311^0 30' + 1^0 10' = 312^0 40'$

BB of CD = $312^0 40' - 180^0 = 132^0 40'$

Correction at D = $132^0 40' - 135^0 40' = -3^0 0'$

FB of DA = $201^0 25' - 3^0 0' = 198^0 25'$

BB of DA = $198^0 25' - 180^0 0' = 18^0 25'$

ANSWER KEY AND SCHEME OF EVALUATION

Corrected bearings table.

Line	observed bearing	Correction	Corrected bearing	Remarks
AB	125° 40'	0° Correction at A	125° 40'	
BA	305° 40'	0° correction at B	305° 40'	
BC	67° 30'	0° Correction at B	67° 30'	C and D
CB	246° 20'	1° 10' Correction at C	247° 30'	are effected
CD	311° 30'	"	312° 40'	by the local
DC	135° 40'	-3° Correction at D	132° 40'	Attraction.
DA	201° 25'	"	198° 25'	
AD	19° 40'	0° Correction at A	18° 25'	

Note:

7(a) The offsets are classified according to the direction and length. ④

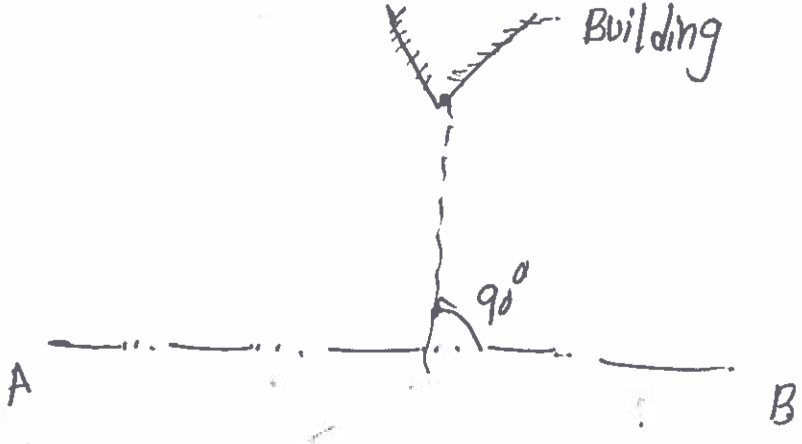
According to direction:-

→ Perpendicular offsets
→ Oblique offsets.

Perpendicular offsets:-

→ The distances measured at right angles to the chain line from the objects are known as Perpendicular offsights as shown in fig

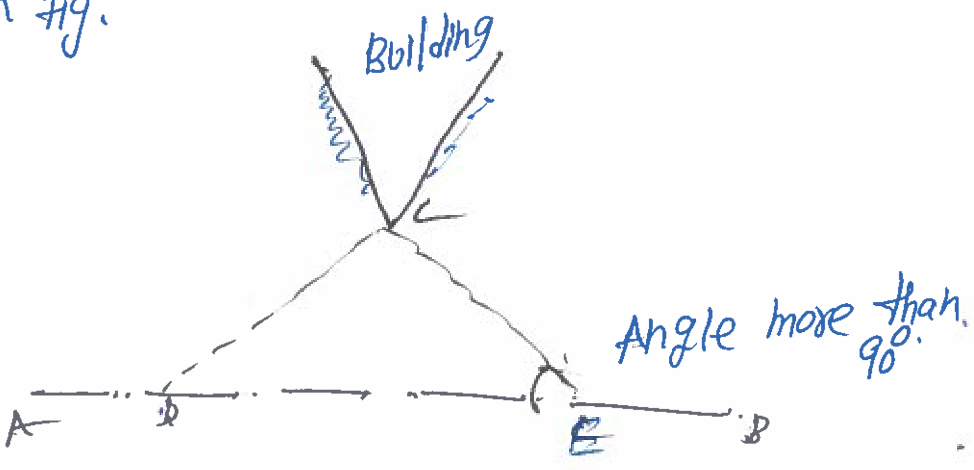
→ Usually the offsets are Perpendicular offsets. In the strict sense, an offset means a Perpendicular offset.



Oblique offsets:-

→ All offsets which are not at a right angles to main survey line are known as oblique ~~survey~~ offsets or tie line offsets as shown in fig.

→



According to the length

→ short offsets
→ long offsets

Short offsets: Generaly the offsets are called short when they are less than 15m in length.

long offsets: Generally the offsets are called ~~the~~ long, when they are more than 15m lengths.

Merits different types of offsets:

→ By using the chain, and and Ranging rod simple to use.
→ NO need of skilled labour.
→ NO need of levelling instruments.

7.b) Ans Resection:→ Resection is a method of Plane table surveying in which the location of the Plane table is unknown and it is determined by sighting to it known points or Plotted points.

→

→ It is also called the method of orientation and it can be conducted by two field conditions as follows

→ Three Point Problem
→ Two Point Problem

Intersection: In this method we can locate the point by plotting two rays from two known stations.

Sums

Height of instrument method:-

Back sight	IS	FS	HI	RL	Remarks
2.228			102.228	100.00	RL at BM.
	1.606			100.622	
2.090		0.988	103.330	101.240	CP or T.P
	2.864			100.466	
0.602		1.262	102.670	102.068	C.P (or) T.P
1.044		1.982	101.734	100.690	CP (or) T.P
		2.684		99.05	

or solve
Rise & fall metho
To solve any met
same RL's come.

(or)

$\Sigma BS = 5.964$

$\Sigma FS = 6.916$

Check: $\Sigma BS - \Sigma FS =$ Last RL - First RL

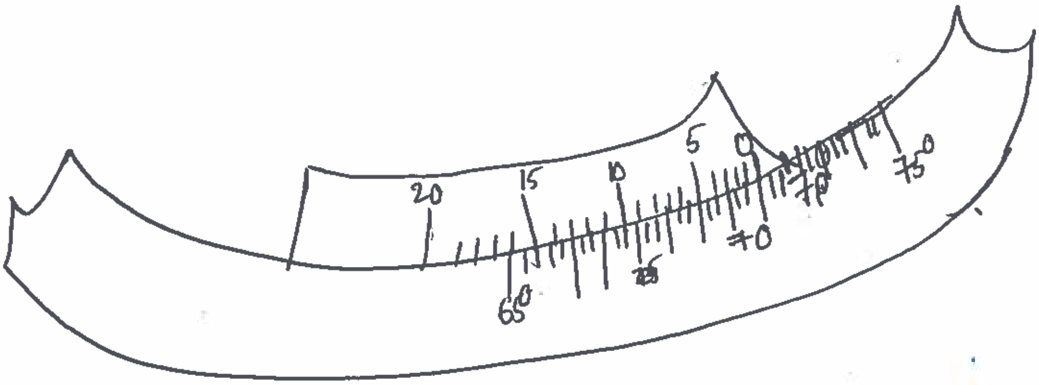
$5.964 - 6.916 = 99.05 - 100.00$

$= -0.952 = -0.95$ // Hence ok

⑦

9.(a) least count of a vernier: → The transit theodolite as it's telescope is transited i,e rotated through a whole revoluation regarding to its horizontal axis with in the vertical Plane.

→ It's least count varies from 10" to 20" depending upon the type of vernier used in it. However, most generally it is 20"



9.b Ans:. To find the horizontal angle. two methods are availble.

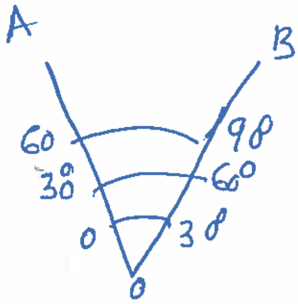
→ Repetation method

→ Reitration method.

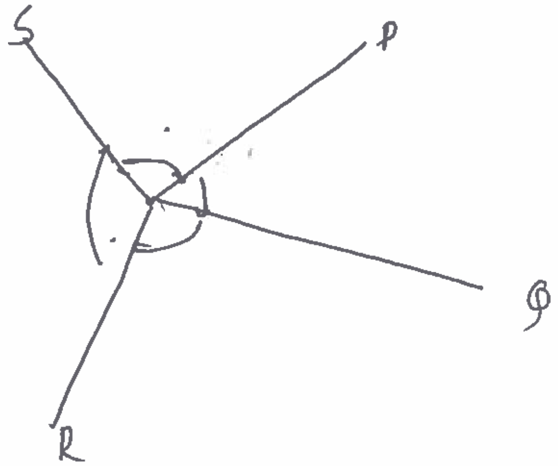
① Repetation method:. The Repetition method is used to improve Precision and accuracy of measurements of horizontal angles.

→ The same angle is measured multiple times, with the survey instrument rotated so the systematic errors tend to cancel.

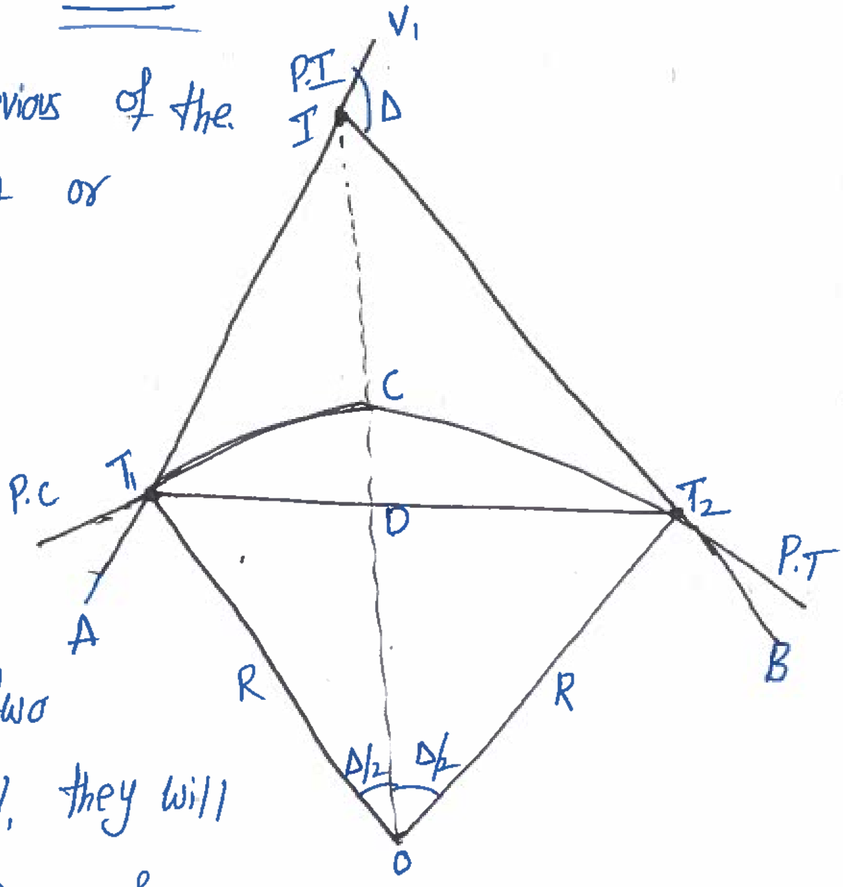
→ The arithmetic mean of these observations given true value of ∠ an angle.



Reiteration method: In this method, angles are measured successively starting from a point termed as initiation station. The angle between the terminating station and the initial station and the last observation during a set of measurement of horizontal angle. by method of reiteration



10Ans **Components of simple circular curve..**



Back tangent:- The tangent (AT_1) Previous of the curve is called the back tangent or first tangent.

Forward tangent:- The tangent (BT_2) following the curve is called the forward tangent.

Point of intersection:- If the two tangents AT_1 and BT_2 are Produced, they will meet in a Point called the Point of Intersection (PI).

Point of curve:- It is the begining of the curve where the alighnment changes from a tangent to a curve.

Point of tangency:-(PT) It is the end of the curve where the alighment changes from a curve to tangent.

Intersection angle:- The angle V'VB between the tangent AV Produced and VB is called the intersection angle(Δ) (or) the external deflection angle b/w the two tangents.

Tangent distance:- It is the distance b/w P.C to P.I or P.I to P.T

External distance (E) It is distance from the mid Point of the curve to P.I

Long chord:- It is chord Joining P.C to P.T

Mid ordinate:- It is the ordinate from the mid Point of the long chord to the mid Point of the curve.

Normal Chord:(C): A chord b/w two sucessive regular stations on a curve.

Sub chord (c):- sub chord is any chord shorter than the normal chord.

Right hand curve:- If the curve deflects to the right of the direction of the Progress of survey. It is called the right hand curve

Left hand curve: If the curve deflects to the left of the direction of the Progress of survey. It is called the left hand curve.

ANSWER KEY AND SCHEME OF EVALUATION

11

11. Ans

Given data

$b = 120m.$

$\alpha_1 = 56^0$

$\alpha_2 = 32^0$

$S = 2M$

$BM = 100m.$

Distance $(D) = \dfrac{b\tan\alpha_2}{\tan\alpha_1 - \tan\alpha_2}$

$= \dfrac{120 \tan 32^0}{\tan 56^0 - \tan 32^0}$

$= 87.42m$

$h = D\tan\alpha_1$

$= 87.42 \tan 56^0$

$= 129.16m$

Elevation of the object = RL at BM + Bs + h

$= 100 + 2 + 129.16$

$= 231.16m.$

ANSWER KEY AND SCHEME OF EVALUATION

(12)

12Ans **segments of GPS**

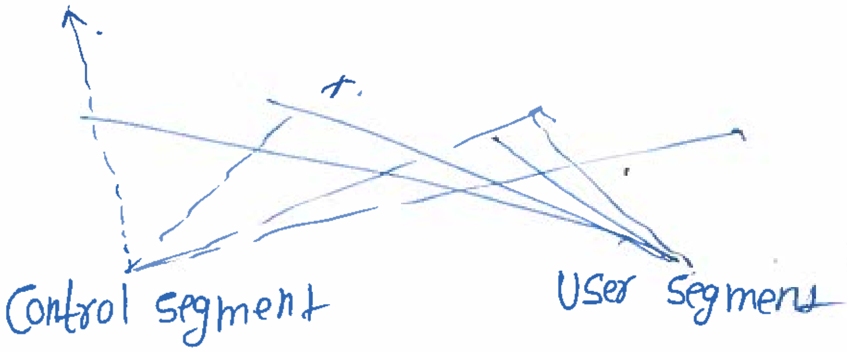
→ space segments

→ control segments &

→ user segments

space segments: The space segment contains 24 satellites, in 12hour near circular orbits at altitude of about 20,000km with inclination of orbit 55°.

→ The consellation ensure at least 4-satelites in view from any Point on the earth at any time for 3-D Positioning and navigation on world-wide basis



Control segment:

→ This has a master Control station (MCS), and an up a load station (ULS). The MSS are transportable shelters with receivers and computer.

→ All located in U.S.A. which passively track satellites accumlating ragging data from navigation signals.

→ Future navigation message are generated from this and loaded into satellite memory once a day via ULS which has a parabolic antenna, a transmitter and a computer.

→ Thus role of control segment is

1) To estimate satellite ephemerides and atomic block behavio
2) To predict SV positions and clock drifts
3) To upload this data to SV's

User segment:

→ The user equipment consists of an antenna, a receiver, a data processor with softward and a control / display unit.

→ The GPS receiver measures the pseudo range, phase and other data using navigation signals from minimum 4 + satellite and 3D position velocity and system time.

ANSWER KEY AND SCHEME OF EVALUATION

13.b) Given data ⑭

Area = $300 km^2$

Scale - 1 in 12000

Size = 250 mm × 250 mm

End over lap = 60%

Side over lap = 30%

No. of Photographs = $\boxed{N = \frac{A}{\left(\frac{1-Lo}{s}\right) l \times \left(\frac{1-So}{s}\right) b}}$

$A = 300 km^2 = 300 \times 10^6 m^2$

$Lo = 60\% = 0.6$

$So = 30\% = 0.3$

$S = 1/12000$

$l = 0.25 m$

$b = 0.25$

$$N = \frac{300 \times 10^6}{\left(\frac{1-0.6}{1/12000}\right) 0.25 \times \left(\frac{1-0.3}{1/12000}\right) 0.25}$$

$\boxed{N = 119 \text{ Photographs}}$

13. b) steps involved in creation of map projection

→ selection of a model for the shape of the earth or planetary body (usually choosing between a sphere or ellipsoid)

→ Transformation of geographical co-ordinates to plane co-ordinates

→ Reduction of the scale (it does not matter in what order the second and third steps are performed)

14. a) What is raster overlay operation? Explain.

Ans Raster overlay superimposes at least two input raster layers to produce an output layer

→ Each cell in the output layer is calculated from the corresponding pixels in the input layer

For example

Layer ①

3	2	0
1	2	3
7	1	0

Add

5	5	5
3	2	3
8	2	0

Layer ②

2	3	5
2	0	0
1	1	0

Subtrc

1	-1	-5
-1	2	3
6	0	0

→ Raster over lay frequently called map algebra is based on calculation which include arithmetic expression and set Boolen algebric operations.

Maximum

3	3	5
3	2	3
7	1	0

Layer 1>2
And
Layer 2>1

1	0	0
0	0	0
0	0	0

Minimum

2	2	0
1	0	0
1	1	0

Layer 1>2
(or)
Layer 2>1

1	1	1
1	0	1
1	0	0

Avg

2.5	2.5	2.5
1.5	1.0	15
4.0	10	00

14.b Supervised Classification

→ In a supervised classification, the signature file was created from known defined classes (for example, land-use-type) identified by pixels enclosed in polygons.

Basic steps

→ select training areas
→ Generate signature file.
→ classify

select training areas

→ In this step you find training samples for each land cover class you want to create.

Generate signature file

At this Point you should have training samples for each class The signature file is what holds all the training sample data. that you've collected up to this Point.

Classify : The most common supervised classification methods include → Maximum likelihood
→ Iso cluster
→ Class Probability
→ Principal Components
→ support Vector machine

In this step, the input is your signature file which has the training sample. If you run it and don't like the result then you may have to verify your training samples.

ANSWER KEY AND SCHEME OF EVALUATION

15.a) Spatial data :- Spatial data that which has physical dimensions and geographic location on the surface of the earth.

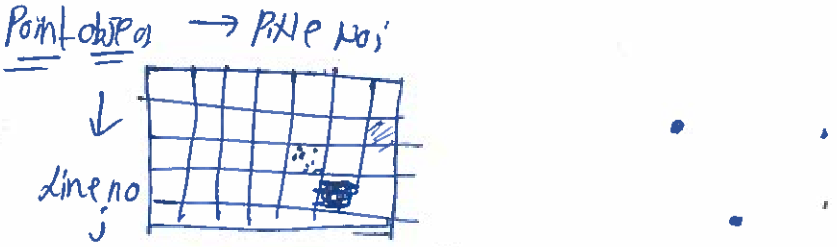
→ Some examples are a river, a state boundary, a lake, a state capital etc..

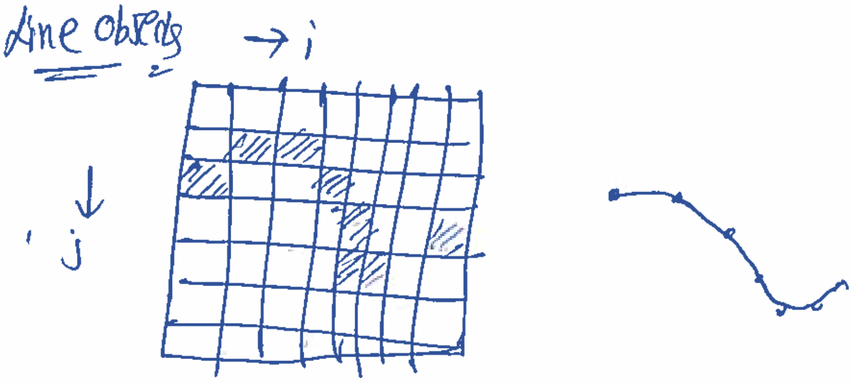
Representation of spatial information:

Geographical features are depicted on map by point, line & polygon

Point - A discrete location depicted by a special symbol or label. A single x,y ~~Co~~ co-ordinate.

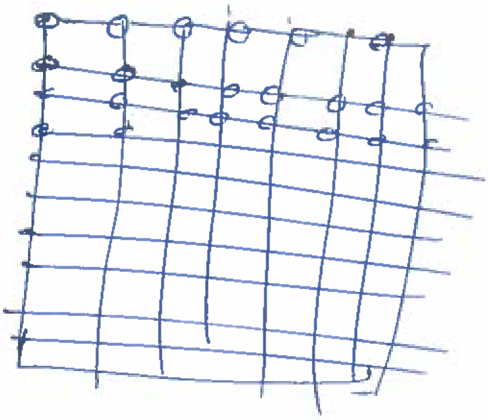
Line: Representation of a linear feature. A set ordered x,y coordinates





A	A	A	A	B	B	B	B
A	A	A	A	B	B	B	B
A	A	A	C	C	B	B	B
A	A	A	C	C	C	B	B
A	A	C	C	C	C	C	C
A	A	C	C	C	C	C	C

Geometry



Topology

Polygen feature: An area featre where boundary encloses a homogeneus Area

15.b) overlay using decision table.

→ overlay is an important technique for integrating data derived from Various sources and perhaps is the basic key function. In GIS data analysis and modelling source

→ Map overlay is a process by which it is possible to take two or more different thematic map layers of the same area.

→ overlayer them on top of all the other to form a composite new layer

→ There are some fundamentay difference in operations and analysis in the way map overlays are performed b/w the raster and vector world.

(D.V. Shan[illegible])

(T Prasanth)

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ANSWER KEY AND SCHEME OF EVALUATION

Scheme of valuation

Part- A

→① Errors in chaining: any two - 1×2=2

② Principles of surveying " " = 1×2=2

③ Types of curves – H.C, V.C } 2m.

④ Def vector - 2marks

⑤ features - any 2 } 1×2=2m.

Part- B

6Ans

Convert W.C.B – 2m.
Check difference – 2m.
Find the correction – 4m
Give the corrected bearing table – 4m.
} 12 marks

Note: BB of DA = N 19°40' E is given but this is wrong as per solution we got $\boxed{N 18°25' E}$

7.a) perpendicular offsets → 3m Explan } 9marks

2) oblike offsets → 3m.

Merits } — 3m.

7.b) Resection. def — 1.5marks } 3marks

Intersection def → 1.5marks

8 Ans ~~Ans~~ To solve RLs any one method

- HI (or) Rise & fall method.

Table – 10marks } 12marks.

Checks – 2marks

9.a) Def of least count — 3m. } 6marks

Drawing Verniee scale - 3m

b) horizontal Angle → Repetation → 3 } 6marks

Reiteration → 3

10 Ans Curve – Dia — 2m. } 12marks

Components at least 10nos 10m

11 Ans Given data - 4m } 12marks

formula - 4m

Calculation - 4m

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ANSWER KEY AND SCHEME OF EVALUATION

12Ans Segments of GPS →

Space segments — 4m
Control segments — 4m
User segments — 4m
} 12marks

13.a) Given data -2m.
Formula - 2m.
Calculaton -2m
} 6markss.

13.b) Steps involved in map creation } → 6marks

14.a) Def. Raster overlay Projection } 3marks
Diagram with layer } 3m
} 6marks

b) Def. Supervised Classification — 3marks
Basic step - 3 — 3marks
} 6marks

15.a) Spatial data GIS → 3 marks
Integrated with GIS → 3 marks
} 6 marks.

b) over lay using a decision table → 6 marks.

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Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engineering			**Academic Year**	2022 - 2023
Course Code	20ME303	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Material Science & Metallurgy**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Write the important characteristics of metals and alloys?	20ME302.1	L1
2	List any two applications of steel.	20ME302.2	L1
3	What is meant by normalizing?	20ME302.3	L1
4	What are the functions of compaction of metal powders?	20ME302.4	L1
5	List any two examples of cermets	20ME302.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Classify in detail the different types of crystal imperfections. Explain the edge dislocation with a neat sketch.	6M	20ME302.1	L2
6 (b)	Discuss selection criteria for materials used in engineering applications.	6M	20ME302.1	L2
	OR			
7	Explain Fe_3 - Fe_3C phase diagram with various reactions in it.	12M	20ME302.1	L2
8 (a)	Differentiate between grey cast iron and white cast iron with respect to properties and applications.	6M	20ME302.2	L2
8 (b)	Explain the structure and properties of plain carbon steels.	6M	20ME302.2	L2
	OR			
9 (a)	Write in detail about the properties and applications of copper and its alloys?	6M	20ME302.2	L2
9 (b)	Explain the properties and applications of phosphor bronze and aluminium bronze.	6M	20ME302.2	L2
10 (a)	Write full name of TTT diagram and explain how it is constructed.	6M	20ME302.3	L2
10 (b)	Distinguish between Normalizing and Annealing.	6M	20ME302.3	L2
	OR			
11 (a)	Explain the processes of Nitrading. When do you use it?	6M	20ME302.3	L2
11 (b)	Explain in detail about different types of carburizing methods?	6M	20ME302.3	L2
12 (a)	Explain the following processes (i) Infiltration (ii) Impregnation	6M	20ME302.4	L2
12 (b)	What is sintering in powder metallurgy? Explain.	6M	20ME302.4	L2
	OR			
13	Explain the important steps involved in the production of components by powder metallurgy technique.	12M	20ME302.4	L2
14 (a)	Sketch and explain different methods of processing ceramics.	6M	20ME302.5	L2
14 (b)	List the various types of glasses, enumerate its properties and applications.	6M	20ME302.5	L2
	OR			
15 (a)	Explain the term composite material with examples. State their advantages and limitations of composites in practice.	6M	20ME302.5	L2
15 (b)	Explain the typical material properties of nanomaterials.	6M	20ME302.5	L2

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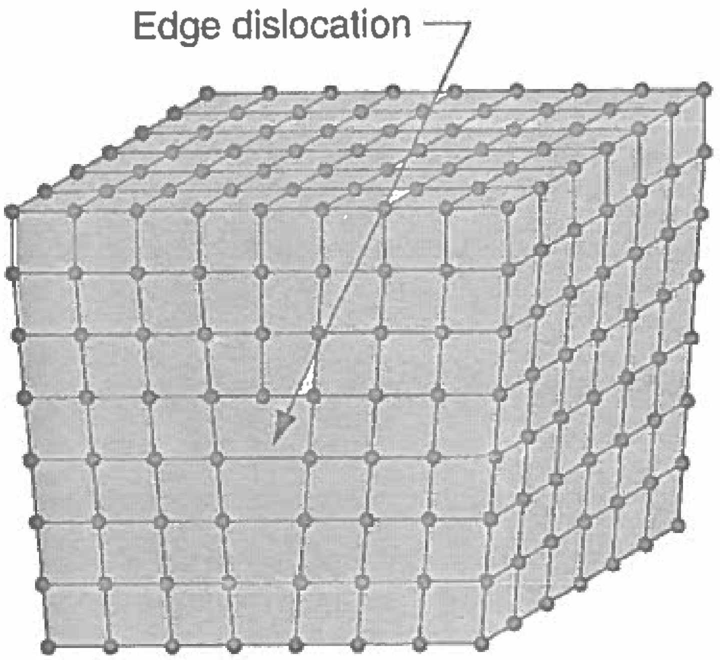
ANSWER KEY AND SCHEME OF EVALUATION

Degree	B.Tech (U.G.)			Year	II	**Academic Year**	2022 - 2023
Course Code	20ME303	**Test Duration**	3 Hrs	Max. Marks	70	**Semester**	III
Course	**Material Science & Metallurgy**						

Part A

No.	Answers	Marks
1.	High melting point. Good conductors of heat and electricity. Malleable (can be bent/shaped easily) Ductile (can be stretched easily without breakage) High density.	Any 2 properties of each - 2M
2.	Steel is used in **buildings, infrastructure, tools, ships, trains, cars, machines, electrical appliances, weapons, and rockets.**	Any two applications - 2M
3.	Normalizing involves **heating the steel to an elevated temperature, followed by slow cooling to room temperature.** The heating and slow cooling changes the microstructure of the steel. This reduces the hardness of the steel and will increases its ductility.	Definition - 2M
4.	Powder compaction is **the process of compacting metal powder in a die through the application of high pressures.** Typically the tools are held in the vertical orientation with the punch tool forming the bottom of the cavity. The powder is then compacted into a shape and then ejected from the die cavity.	Any 2 functions = 2M

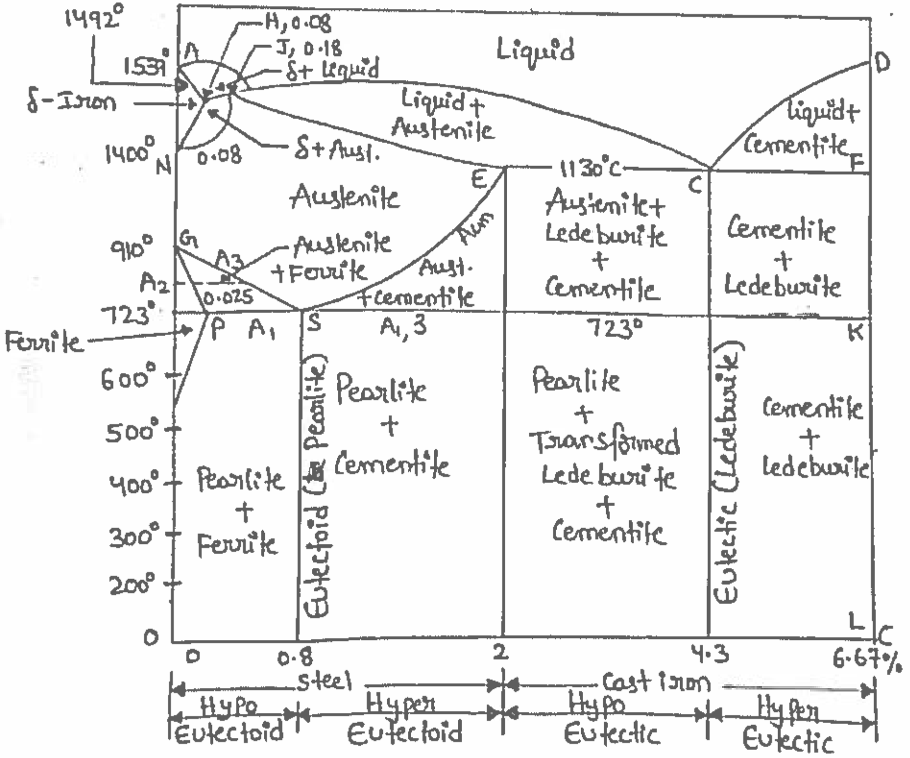
5.	Most commonly used cermets are: (1) **carbides based on TiC, such as TiC + Mo or Ni for better wear resistance and TaC + TiC + Co + WC for high fracture toughness,** (2) TiN-based carbides for higher toughness, and (3) TiCN-based carbides for better heat and wear resistance.	Any two examples = 2M
6.	**PART – B** **a)** **Types of crystal imperfections:** Point defects. Line defects. Planar defects. Edge dislocation - An edge dislocation is a **defect where an extra half-plane of atoms is introduced midway through the crystal, distorting nearby planes of atoms.**  **b)** Important characteristics of materials are : **strength, durability, flexibility, weight, resistance to heat and corrosion, ability to cast, welded or hardened, machinability, electrical conductivity, etc.** **OR**	Types with one type explanation = 6M Any 6 selection parameters = 6M

7

→ IRON – CARBON EQUILIBRIUM DIAGRAM

The Iron-Carbon equilibrium diagram is the basis of Steel and cast iron. It is constructed by plotting temperature along y-axis and percentage composition of the alloy along the X-axis. This diagram shows ranges of temperatures and compositions with in which the various phase changes are stable and also the boundaries at which the phase changes occur. It concerns transformations that occur in alloys having compositions from pure iron to cementite (6·67%) carbon.

It establishes a corelation between the microstructure and properties of steel and cast irons and provides a basis for the understanding of the principles of heat-treatment.



FE-FE3C DIAGRAM = 4M

Explanation = 8M

1. Pure iron melts at A i.e, at 1539°c and melting point of cementite is approximately 1560°c.
2. The curve ABCD is the liquidus line and AHJECF is the solidus line of the system.
3. Liquidus line define the temperature at which solidification starts, and solidus line indicates the temperature at which solidification just completed.
4. The region between these two lines represent liquid and solid phases.
5. The temperatures at which the transformation in the solid state occurs are called critical points or critical temperatures.
6. In hypo eutectoid steel,
   - line GS (A3) represents upper critical point
   - line PS (A1) represents lower critical point
7. In hyper eutectoid steel,
   - line SE(Acm) represents upper critical point
   - line SK (A1,3) represents lower critical point
8. The line HJB is temperature of transformation of δ-iron to γ-iron.
9. In the alloys with less than 2% C, the structure upon primary crystallisation is autenite (below NJE)
10. Alloys with more than 2% C, the structure is ledeburite plus excess austenite or cementite.
11. The primary austenite and the austenite in the eutectic mixture (ledeburite) contains maximum amount of carbon (2%) in the solution at the end of crystallisation.
12. From 1130°c carbon precipitates from austenite.
13. At 723°c the austenite and the austenite in ledeburite contain 0.8% C, and pearlite transformation takes place at that temperature.

---

14. Ledeburite below 723°c is a mixture of cementite and pearlite
15. The three horizontal lines HJB, ECF and PSK in the diagram indicates the three iso-thermal reactions at the fixed composition and temperature.

Peritectic Reaction at Point J :- The formation of the austenite (0.18% C) at conto constant temperature (1492°c) from the liquid and solid-phase is called peritectic reaction.

$$\text{Delta iron } (\delta) + \text{liquid} \underset{\text{Heating}}{\overset{\text{Cooling}}{\rightleftharpoons}} \text{Austenite.}$$

Eutectic Reaction at point C :- The solidification of liquid (4.3% c) at constant temperature (1130°c) into two phase mixture is called eutectic reaction.

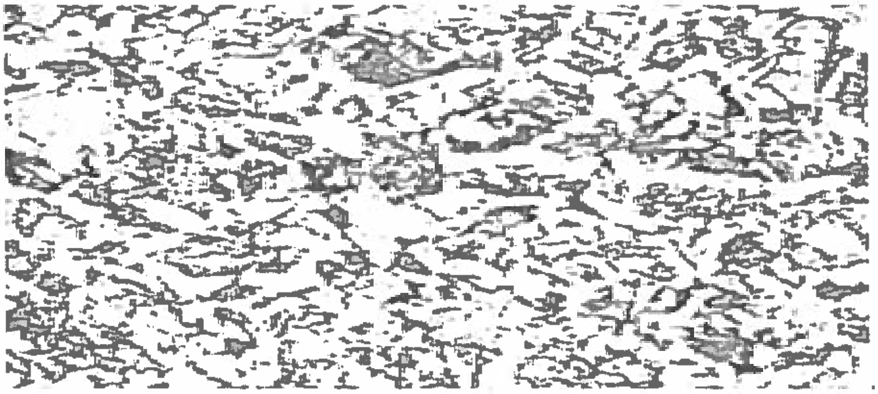
$$\text{Liquid} \underset{\text{Heating}}{\overset{\text{Cooling}}{\rightleftharpoons}} \underbrace{\text{Austenite + Cementite}}_{\text{Eutectic mixture (Ledeburite)}}$$

Eutectoid Reaction at point S:-

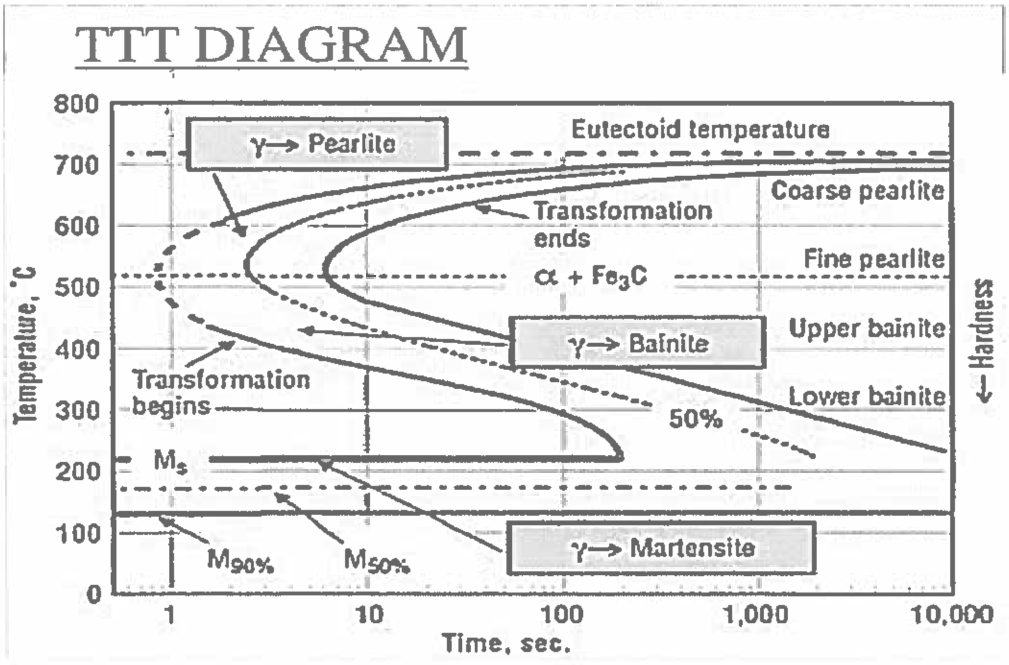
The decomposition of austenite (0.8% C) at constant temperature (723°c) into ferrite and cementite is known as eutectoid reaction.

$$\text{Solid (Austenite)} \underset{\text{Heating}}{\overset{\text{Cooling}}{\rightleftharpoons}} \underbrace{\text{Ferrite + Cementite}}_{\text{Eutectoid mixture (pearlite)}}$$

8	a) **Properties:** **Grey Cast Iron:** Grey cast iron has a higher compressive strength and high resistance to deformation. Its melting point is relatively low, 1140 °C to 1200 °C. It also has a greater resistance to oxidation; therefore, it rusts very slowly and this gives a permanent solution to the problem of corrosion. **White Cast Iron:** In white cast iron carbon is present in the form of carbide of iron. It is hard and brittle, has a greater tensile strength and extremely malleable (ability to hammer or press permanently out of shape without breaking or cracking). It also has high compressive strength and excellent wear resistance. It can maintain its hardness for limited periods, even up to a red heat. It cannot be easily cast as other irons because it has a relatively high solidification temperature. **Uses:** **Grey Cast Iron:** The most commonly used areas of grey cast iron are: in internal combustion engine cylinders, pump housings, electrical boxes, valve bodies and decorative castings. It is also used in cooking equipment and brake rotors. **White Cast Iron:** White cast iron is most extensively used in crushing, grinding, milling and handling of abrasive materials.	Any 6 differences = 6M

	b) **Properties and Uses:** (i) Low Carbon Steel: It possesses good formability and weldeability but lacks in hardness. It is used in making nuts, bolts, sheets, tubes and machine components not requiring much high strength. It is also used in making beams and channels. (ii) Medium Carbon steel: It has higher strength than low carbon steel and is harder due to increased content of carbon. Its properties can be improved by heat treatment processes and hence is very popular. It is used for making machine parts such as gears, axles, crank-shafts and parts for metal working mahinery. (iii) High Carbon Steel: It has low toughness and formalibility but hardness and wear resistance are high. It is used generally for making parts such as cutting tools, cables, springs,etc 	Structure = 2M and Any 4 properties = 4M

9	a) **Key Properties of Copper Alloys** Excellent heat conductivity. Excellent electrical conductivity. Good corrosion resistance. Good biofouling resistance. Good machinability. Retention of mechanical and electrical properties at cryogenic temperatures. The main applications of copper are in electrical wiring, roofing, plumbing, and industrial machinery. For many of these applications, copper is used in its pure form. However, it can be alloyed with other metals when increased levels of hardness are required.	**Any 6 Properties = 6M**
	b) **COMMON PHOSPHOR BRONZE PROPERTIES** Excellent resistance to corrosion and fatigue. Good electrical conductivity. Excellent product for strength performance. Low coefficient of friction. Fine grain. Excellent elasticity. The family of aluminum bronze alloys offers **high strength and hardness, excellent corrosion resistance, good wearing qualities and good fatigue resistance.** The alloys are well suited for service at elevated temperatures. Aluminum bronzes are used in **marine hardware, shafts and pump and valve components for handling seawater, sour mine waters, nonoxidizing acids, and industrial process fluids.** They are also used in applications such as heavy duty sleeve bearings, and machine tool ways. Most commonly, Phos Bronze is used in the **manufacturing of springs, fasteners, and bolts.** These parts must resist fatigue and wear while exhibiting high elasticity. Digital electronics, automatic controllers, and automobiles all contain parts made with Phosphor Bronze.	**Any 3 Properties = 3M** **Any 3 applications = 3M**
10	**a)** TTT diagram stands for "**time-temperature-transformation**" diagram. It is. also called isothermal transformation diagram. Definition: TTT diagrams give the kinetics of isothermal transformations. TTT DIAGRAM T (Time) T(Temperature) T(Transformation) diagram is a plot of temperature versus the logarithm of time for a steel alloy of definite composition. It is used to determine when transformations begin and end for an isothermal (constant temperature) heat treatment of a previously austenitized alloy. When austenite is cooled slowly to a temperature below LCT (Lower Critical Temperature), the structure that is formed is Pearlite. As the cooling rate increases, the pearlite transformation temperature gets lower. The microstructure of the material is significantly altered as the cooling rate increases. By heating and cooling a series of samples, the history of the austenite transformation may be recorded. T TT diagram indicates when a specific transformation starts and ends and it also shows what percentage of transformation of austenite at a particular temperature is achieved.	TTT DIAGRAM = 2M Explanation = 4M



b)

Basic Terms	Annealing	Normalizing
Meaning	It is a heat treatment method used to make metal ductile and less hard	It is a heat treatment technique only applicable to alloys of iron
Cooling Process	Either in air or quenching in water	Only in air
Grain Size	Not easy to attain uniform grain size	Quite easy to attain uniform grain size
The hardness of the final product	Ductile and less hard	Tend to remain harder
Cost	Too expensive	Comparatively affordable
Purpose	To refine the crystalline structure and remove residual stresses	To get a refined grain

Any 6 differences = 6M

OR

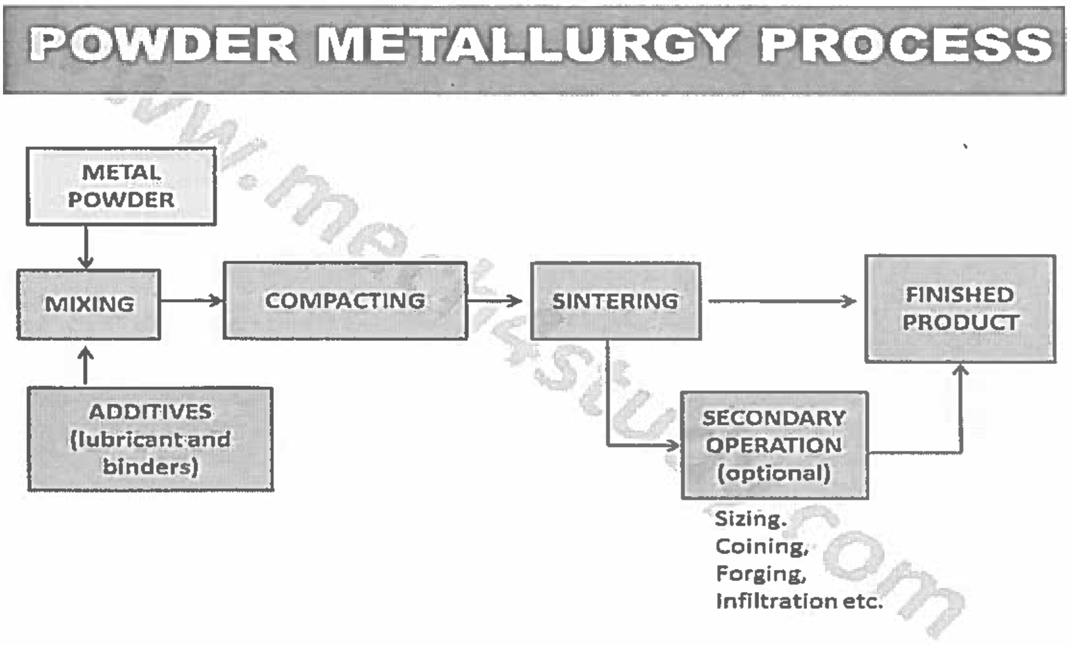
11 a)

Nitriding is a heat treating process that diffuses nitrogen into the surface of a metal to create a case-hardened surface. These processes are most commonly used on low-alloy steels. They are also used on titanium, aluminium and molybdenum.

Process = 6M

Nitriding can increase abrasion/wear resistance and improve bending and/or contact-fatigue

	properties. For example, nitriding increases the bending-fatigue strength of a 3% Cr-Mo steel from 480 to 840 MPa – a 75% improvement.	
	b) Carburising is a thermochemical process in which carbon is diffused into the surface of low carbon steels to increase the carbon content to sufficient levels so that the surface will respond to heat treatment and produce a hard, wear-resistant layer. There are three types of carburising commonly used: ***Gas carburising*** In gas carburising, the component is held in a furnace containing an atmosphere of methane or propane with a neutral carrier gas, usually a mixture of N_2, CO, CO_2, H_2 and CH_4. At the carburising temperature, methane (or propane) decomposes at the component surface to atomic carbon and hydrogen, with the carbon diffusing into the surface. ***Liquid carburising (or cyaniding)*** Liquid or cyanide carburising is carried out by placing the component in a salt bath at a temperature of 845 to 955°C. The salt is usually a cyanide-chloride-carbonate mixture and is highly toxic. The cyanide salts introduce a small amount of nitrogen into the surface which further improves its hardness. Although the fastest carburising process, it is suitable only for small batch sizes. ***Solid (pack) carburising*** In solid or pack carburising, the components are surrounded by a carburising medium and placed in a sealed box. The medium is usually coke or charcoal mixed with barium carbonate. The process is really one of gas carburisation since the CO produced dissociates into CO_2 and carbon which diffuses into the components' surface.	Any 3 methods with explanation = 6M
12	a) Infiltration is defined as the **unwanted outside air that migrates or leaks into the building through cracks and openings** in the building. Impregnation is the simplest way to prepare supported catalyst. In the first step of this process, **air is evacuated from the pores by using a deep vacuum.**	2 process = 6M
	b) The sintering process in powder metallurgy is **a form of heat treatment**. A conventional sintering process heats up the material to just below its melting point. A precise sintering temperature allows the metals to keep their beneficial properties while fusing them tightly together.	process = 6M
	OR	

13		Production Process = 12M
14	a) Some of the most common forming methods for ceramics include **extrusion, slip casting, pressing, tape casting and injection molding**. After the particles are formed, these "green" ceramics undergo a heat-treatment (called firing or sintering) to produce a rigid, finished product.	Any 3 methods with explanation = 6M
	b) Float Glass. Float glass manufactured from sodium silicate and calcium silicate so, it is also called as soda-lime glass. ... Shatterproof Glass. Shatterproof glass is used for windows, skylights, floors, etc. ... laminated Glass. ... Extra Clean Glass. ... Chromatic Glass. ... Tinted Glass. ... Toughened Glass. ... Glass Blocks. The main characteristics of glass are **transparency, heat resistance, pressure and breakage resistance and chemical resistance**. Glass has several strong points concerning optical properties: It can be produced in large and homogeneous panes. Its optical properties are not affected by ageing. Packaging (jars for food, bottles for drinks, flacon for cosmetics and pharmaceuticals) Tableware (drinking glasses, plate, cups, bowls) Housing and buildings (windows, facades, conservatory, insulation, reinforcement structures)	Types with explanation of properties and applications= 6M
	OR	
15	a) A composite material is **a combination of two materials with different physical and chemical properties**. When they are combined they create a material which is specialised to do a certain job, for instance to become stronger, lighter or resistant to electricity. Composite – Composite building material examples include **concrete, reinforced plastics, cement, steel–reinforced concrete, and composite wooden beams.** **Composites are more brittle than wrought metals and thus are more easily damaged.** Cast metals also tend to be brittle. 2. Repair introduces new problems, for the following reasons: Materials require	Definition and Types with advantages and limitations= 6M

<table>
<tr><td></td><td>refrigerated transport and storage and have limited shelf lives.</td><td></td></tr>
<tr><td></td><td>b)
<table>
<tr><td>Mechanical</td><td>Stiffness, strength, hardness, elongation at failure etc.</td></tr>
<tr><td>Electrical</td><td>Electrical conductivity, resistivity, etc.</td></tr>
<tr><td>Magnetic</td><td>Permittivity, permeability, etc.</td></tr>
<tr><td>Optical</td><td>Reflectivity, transmissivity, absorptivity, emissivity, etc.</td></tr>
<tr><td>Thermal</td><td>Heat capacity, thermal expansion, thermal conductivity, etc.</td></tr>
<tr><td>Chemical</td><td>Toxicity, reactivity, flammability, oxidation state, etc.</td></tr>
</table></td><td>Any 6 Properties = 6M</td></tr>
</table>

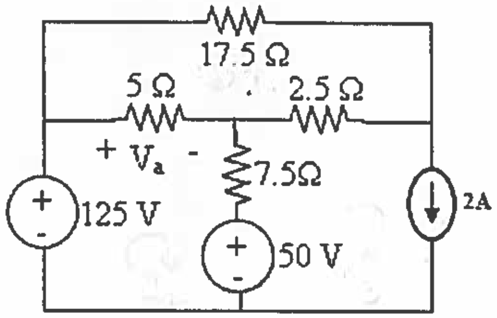
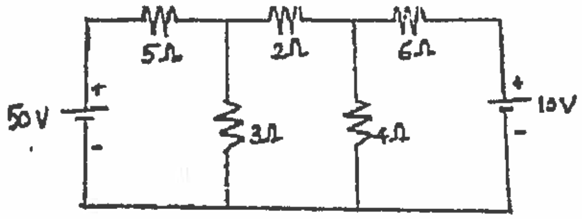
Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

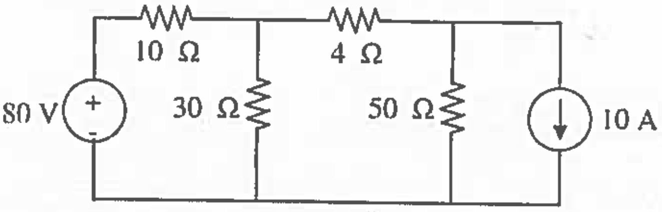
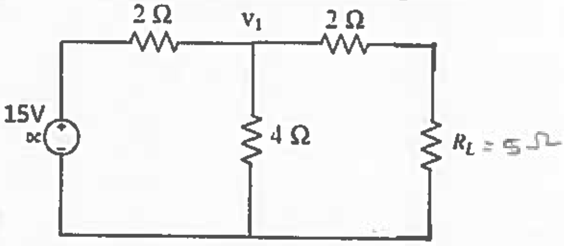
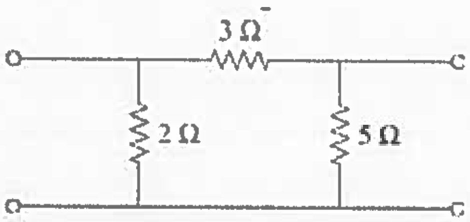
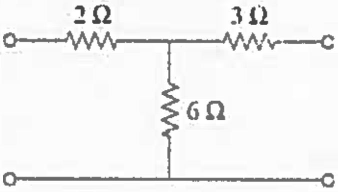
Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2022 - 2023
Course Code	20EE303	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**ELECTRICAL CIRCUIT ANALYSIS**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Compare any four features of AC and DC circuit.	20EE303.1	L2
2	State the relation between voltage and current in delta connected system	20EE303.2	L1
3	List any two advantages of Thevenins theorem.	20EE303.3	L1
4	Identify the symmetry and reciprocity of h parameters in two port networks.	20EE303.4	L1
5	Express the time constant for series RL and RC circuits.	20EE303.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Find the value of Va for the following circuit using KVL in figure .1  Figure.1	12M	20EE303.1	L3
	OR			
7 (a)	Distinguish between series and parallel circuit.	4M	20EE303.1	L2
7 (b)	Determine the power dissipation in the 4Ω resistor of the given circuit shown in figure.2 by using nodal analysis.  Figure.2	8M	20EE303.1	L3
8 (a)	The impedances of parallel circuit are Z1= (6+j8) ohms and Z2 = (8-j6) ohms. If the applied voltage is 120V, find (i) current and power factor of each branch (ii) overall current (iii) power consumed by each impedance. Draw the phasor diagram.	8M	20EE303.2	L3
8 (b)	Define power factor, apparent power, active power and reactive power.	4M	20EE303.2	L2
	OR			
9 (a)	A balanced star connected load of (4+j3) Ω per phase is connected to a balanced 3¢ 400V supply. Find a) active power b) reactive power c) Apparent power.	5M	20EE303.2	L3
9 (b)	Explain two wattmeter method with a neat sketch.	7M	20EE303.2	L2

10 (a)	Find the voltage across 4Ω resistance using superposition theorem  Figure.3	8M	20EE303.3	L3
10 (b)	State and explain Norton's theorem.	4M	20EE303.3	L2
	OR			
11 (a)	Find Thevenin's equivalent for the circuit shown in figure 4.  Figure.4	8M	20EE303.3	L3
11 (b)	Define maximum power transfer theorem and explain its importance.	4M	20EE303.3	L2
12 (a)	Find the Z- parameters for the following circuit.  	6M	20EE303.4	L3
12 (b)	Express h parameters in terms of ABCD parameters.	6M	20EE303.4	L2
	OR			
13 (a)	Find the Y- parameters for the following circuit.  	8M	20EE303.4	L3
13 (b)	Derive the relation between Z and Y parameters.	4M	20EE303.4	L2
14	Derive the expression of voltage across R and L for RL series circuit.	12M	20EE303.5	L3
	OR			
15	In Series RL Circuit with R=100 Ohms and L=20 Henry has a DC Voltage of 200 Volts applied through a switch at t=0. Find (i) Current and Voltage across each element (ii) Current at time t=0.5 Seconds (iii) Current at time t=1 Second (iv) Time at which $e_R=e_L$.	12M	20EE303.5	L3

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ANSWER KEY AND SCHEME OF EVALUATION

Electrical Circuit Analysis

1.A) (2M)

AC Circuit	DC Circuit.
1) In AC the electrons move forward and backward.	1) In DC the electrons flow steadily in a single direction.
2.) The voltage periodically changes from positive to negative.	2.) The voltage is constant.
3.) The direction of current changes periodically.	3.) The electricity flows in certain director.
4.) AC circuit is less expensive and easy to generate.	4.) DC circuit is expensive as well as hard to generate.

2.A) Delta Connected System :- (2M)

Line voltage $V_L = V_{ph}$

Line Current => $I_L = \sqrt{3}\ I_{ph}$.

Total active power, $P = \sqrt{3}\ V_L I_L \cos\phi$

3) Two advantages of thevenin theorem (2)

a) It reduces a complex circuit to simple circuit.

b) It is used especialy to determine the curret in a particular branch.

4) Symmetry and Reciprocity of h parameters :- (2)

Condition of symmetry : $h_{11}h_{22} - h_{12}h_{21} = \underline{1}$

Condition of Reciprocity :- $h_{12} = -h_{21}$

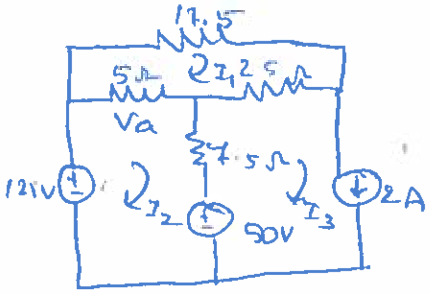
5) The time constants of series RC ~~hu~~and RL :- (2)

Time constant of RL circuit – $\tau = L/R$

Time constant of RC circuit – $\tau = RC$.

---

6) find Va by using KVL :- (12)



apply KVL in each loop.

Loop :1

$\Rightarrow 0 = 17.5 I_1 + 5(I_1 - I_2) + 2.5(I_1 - I_3)$

$\Rightarrow 0 = 17.5 I_1 + 5 I_1 - 5 I_2 + 2.5 I_1 - 2.5 I_3$

$\Rightarrow 0 = 25 I_1 - 5 I_2 - 2.5 I_3$ — ①

Loop 2 :

$125 = Va + \cancel{7.5}(50/\cancel{7.5})$

$125 = Va(50)$

$$\boxed{Va = \frac{125}{50} = 2.5V}$$

Loop 3 :

$50 = 2.5(I_3 - I_1)$

~~$50 = 2.5 I_5$~~

$20 = I_3 - I_1$ — ②

$\boxed{I_3 = 20 + I_1}$ substitute in ①

$25 I_1 - 5 I_2 - 2.5(20 + I_1) = 0$

$25 I_1 - 5 I_2 - 50 - 2.5 I_1 = 0$

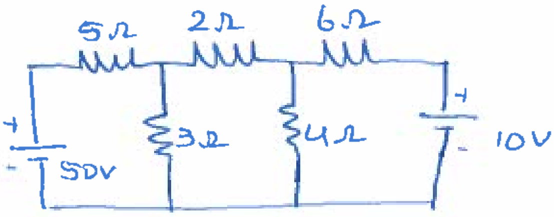
$27.5 I_1 - 5 I_2 = 50$

By solving Loop ② we got $\underline{\underline{Va = 2.5V}}$

7a) Difference between series and parallel circuit (4)

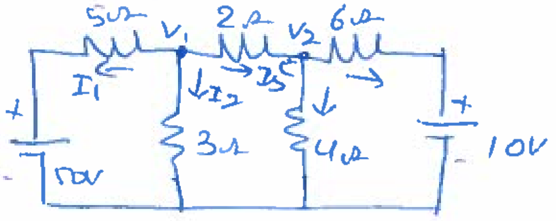
Series circuit	Parallel circuit.
1.) Second terminal of 1st resistor connected to 1st terminal of 2nd resistor.	1.) All 1st terminals connected to a single common point and all 2nd terminals connected to another common point.
2.) The total resistance is $R_{eq} = R_1 + R_2 + R_3$	2.) The total resistance is $R_{eq} : \frac{R_1 R_2 R_3}{R_1 R_2 + R_2 R_3 + R_3 R_1}$
3) The current direction is same as it is constant.	3.) The voltage remains constan
4). The voltage is $V_T = V_1 + V_2 + V_3$	4) The current will be $I_T = I_1 + I_2 + I_3$

7b)



(8)

applying Nodal Analysis



at node ①

$I_1 + I_2 + I_3 = 0$

$\Rightarrow \frac{V_1 - 50}{5} + \frac{V_1}{3} + \frac{V_1 - V_2}{2} = 0$

$\Rightarrow \frac{6V_1 - 300 + 10V_1 + 15V_1 - 15V_2}{30} = 0$

$\Rightarrow 31V_1 - 15V_2 = 300$ —①

at node ②

$I_4 + I_5 + I_6 = 0$

$\frac{V_2 - V_1}{2} + \frac{V_2}{4} + \frac{V_2 - 10}{6} = 0$

$\frac{6V_2 - 6V_1 + 3V_2 + 2V_2 - 20}{12} = 0$

$-6V_1 + 11V_2 = 20$ —②

solve ① & ②

$31V_1 - 15V_2 = 300 \times 6$

$-6V_1 + 11V_2 = 20 \times 31$

$\cancel{186V_1} - 90V_2 = 1800$

$-\cancel{186V_1} + 341V_2 = 620$

$251V_2 = 2420$

$\boxed{V_2 = 9.6V}$

Current through 4Ω is $I_{4\Omega} = \frac{V_2}{4} = \frac{9.6}{4} = 2.4A$

$\boxed{\text{Power at } 4\Omega = I^2R = (2.4)^2 \times 4 = 23.04W}$ (3)

8a) $Z_1 = (6+j8)\Omega$ ~~&~~ $Z_2 = (8-j6)\Omega$

$Z_1 = 6+j8 = 10\angle 53.1$

$|Z_1| = \sqrt{6^2+8^2} = \sqrt{36+64}$

$= \sqrt{100} = 10$

angle $\theta = \tan^{-1}\left(\frac{8}{6}\right)$

$\theta = 53.1^0$

$Z_2 = 8-j6\ \Omega = 10\angle 36.8$

$|Z_2| = \sqrt{8^2+6^2} = \sqrt{100} = 10$

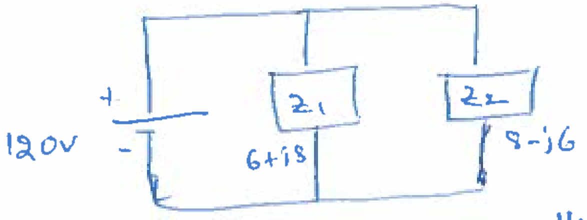
$\theta = \tan^{-1}\left(6/8\right)$

$\theta = 36.8^0$

apply Palallel Combination.

$\frac{R_1R_2}{R_1+R_2} = 100\angle 53.1\angle 36.9$

$\frac{Z_1Z_2}{Z_1+Z_2}$ ~~Z_2~~ $= 100\angle 89.9$



Sum = $14+j2$

$\sqrt{14^2+2^2} = 14.14\angle 8.13$

Total $I = V/Z = \frac{120}{100\angle 89.9} = 1.2\angle -89.9$ A

$\theta = -89.9$

$\boxed{\cos\theta = 1.7\times10^{-3}}$

Power $p = VI$

I through $Z_1 = 1.2\angle -89.9 \times \frac{10\angle 36.8}{10\angle 53.1 + 10\angle 36.8}$

I through $Z_1 = 1.2\angle -89.9 \times \frac{10\angle 36.8}{14.14\angle 8.13}$

$Z_1 = 1.2\angle -89.9 \times 0.707\angle 28.67$

$I_1 = 0.8484\angle -61.23$ A

Power $= I^2R = (0.8484\angle -61.23)^2 \times 10\angle 53.1$

$\boxed{\text{Power } P = 7.19\angle -69.36\, \omega}$

I through Z_2

Total $I = I_1 + I_2$

$I_2 = I - I_1$

$= 1.2\angle -89.9 - 0.8484\angle -61.2$

$I_2 = 0.35\angle -151.13$ A

$I^2R = (0.35\angle -151.13)^2\, 10\angle$

$(0.35\angle -151.3)$

$1.2\angle -26$

(4m)

8b) Power factor:- (1) It is defined as the ratio of average power to the apparant power.

$$\text{power factor} = \cos\phi = \frac{P_{avg}}{V_{eff}\, I_{eff}}$$

Apparent power:- (1) The total power flowing is known as the apparent power and is measured as the product of Voltage and current.

Active power:- (1) The active power is the amount of total electric power in an Ac electric circuit which actually consumed or utilized. measured in watts (w).

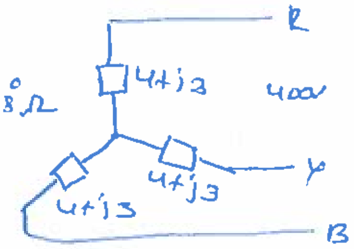
Reactive power:- (1) Reactive power is the part of complex power that corresponds to storage and retrieval of energy rather than Consumption

4

9a) given load $= 4+j3\,\Omega = 5\,\angle 36.8698^\circ\,\Omega$

line Voltage $= 400V$ $\phi = 36.8698^\circ$

given it is a star Connection.



Phase Voltage $= \frac{V_L}{\sqrt{3}} = \frac{400}{\sqrt{3}} = 230{\cdot}94V$

Power factor $= \cos\phi = \frac{R}{Z} = \frac{4}{5} = 0{\cdot}8$

$I_R = \frac{V_L}{\sqrt{3}Z} \angle{-\phi} = \frac{239.6}{5} \angle{-36.87} = 47.92 \angle{-36.87^\circ}A$

In star Connection $I_L = I_{ph}$.

i) Active power $= \sqrt{3} V_L I_L \cos\phi$

$= \sqrt{3} \times 400 \times 47{\cdot}92 \times \cos 36{\cdot}87$

$= 26559W = 26{\cdot}55kW$

Reactive power $= \sqrt{3} V_L I_L \sin\theta$

$= \sqrt{3} \times 400 \times 47.92 \times \sqrt{1-\cos^2\phi}$

$= \sqrt{3} \times 400 \times 47.92 \times \sqrt{1-\frac{16}{25}}$

$= \sqrt{3} \times 400 \times 47.92 \times \frac{3}{5}$

$= \sqrt{3} \times 400 \times 47.92 \times 0{\cdot}6$

$= 19919VAR = 19.9kVAR$

Apparent power $= \sqrt{3} V_L I_L$

$= \sqrt{3} \times 400 \times 47{\cdot}92$

$= 33199VA = 33{\cdot}199kVA$

9b) Two wattmeter method:-

Current through $W_1 = I_R$

Potential difference $V_{RY} = V_R - V_Y$

we know Power $P = VI$

Power measured by wattmeter $W_1 = VI = (V_R - V_Y) I_R$

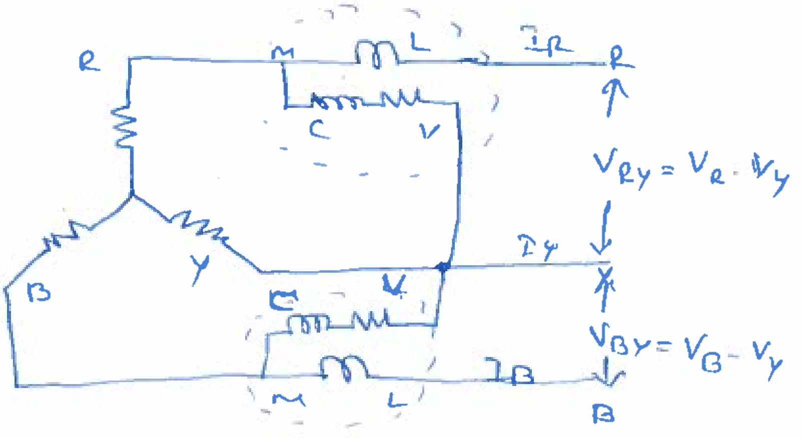
$W_1 = P_1 = V_R I_R - V_Y I_R$

Power measured by wattmeter $W_2 =$

$\Rightarrow P_2 = (V_B - V_Y) I_B$

$P_2 = V_B I_B - V_Y I_B$

$\boxed{\text{Total power} = W_1 + W_2}$



Total power measured by w_1 and w_2 is $P_1 + P_2 = V_R I_R - V_Y I_R + V_B I_B - V_Y I_B$

$= V_R I_R + V_B I_B - V_Y (I_R + I_B)$

$= I_R + I_Y + I_B = 0$

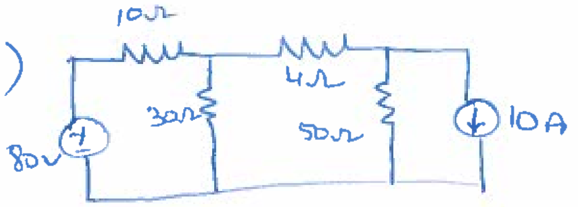
$I_R + I_B = -I_Y$

$= V_R I_R + V_B I_B - V_Y (-I_Y)$

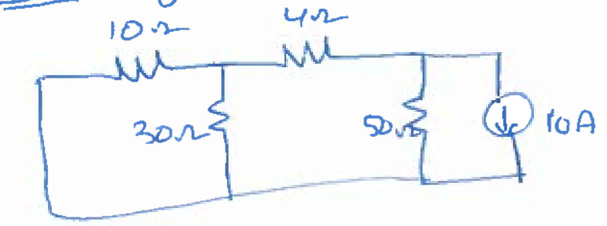
$W_1 + W_2 = V_R I_R + V_B I_B + V_Y I_Y$

⑤

10a)

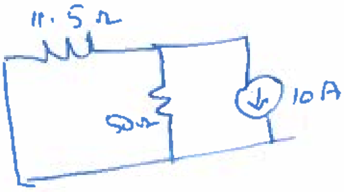


Step-1: Neglect(or) make 80V as zero.



$10\Omega // 30\Omega \Rightarrow \frac{10\times30}{10+30} = \frac{30\cancel{0}}{4\cancel{0}} = 7.5\Omega$

$7.5+4 = 11.5\Omega$

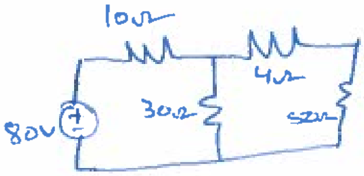


⇒ current division rule.

$\Rightarrow I_{11.5\Omega} = 10\times\frac{50}{61.5} = 8.13\text{ A}$

$\boxed{I_{4\Omega} = 8.13\text{ A}}$

Step 2: Neglect current source and replace with open circuit.



$4+50 = 54\Omega$

$54\Omega // 30\Omega$

$\Rightarrow \frac{54\times30}{54+30} = 19.2\Omega$

$19.2\Omega + 10\Omega = 29.2\Omega$

$I = V/R = \frac{80}{29.2} = 2.73\text{A}$

$I_{4\Omega} = 2.73 \times \frac{30}{84} = 0.97\text{A}$

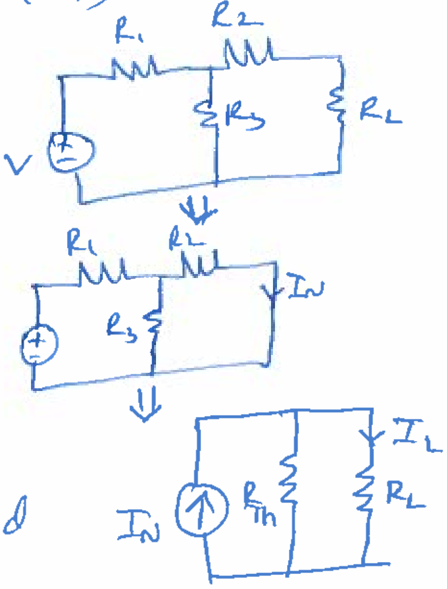
$I_{4\Omega} = 0.97\text{A}$

Total $I = 8.13+0.97$
$= 9.1\text{A}$

$\boxed{V_{4\Omega} = IR = 9.1\times4 = 36.4\text{ Volts}}$

(4m)

10b) Statement: In a linear bilateral network which contains of large number of voltage and current sources and lot number of resistance can be reduced to a single or simple network which contains of current source (I_N) in parallel to resistance (R_N).



Procedure:

Step-1. Remove R_L and make a short circuit.

step-2. find Norton's current I_N

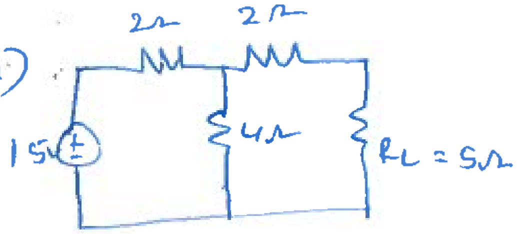
step-3 find Norton's resistance by replacing voltage source with short circuit and current source with open circuit.

step-4 Redraw the network which contains I_N and R_N

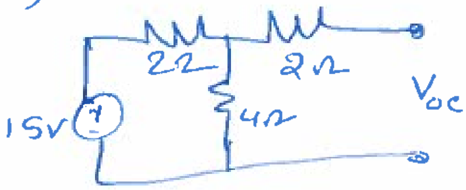
step-5 Attach the load resistance (R_L) and find current through it. (I_L)

⑥

11a)

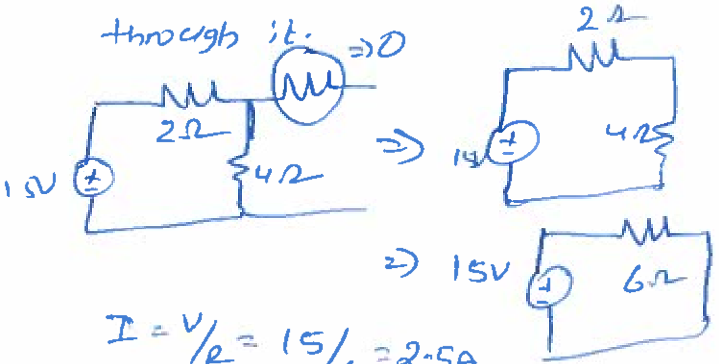


1) Remove RL



neglect 2Ω as current doesn't pass

through it.

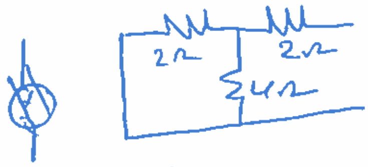


$I = V/R = 15/6 = 2.5A$

$V_{oc} = 2.5 \times 4$

$= 10V$

find R_{Th}, Replace voltage source with short circuit.

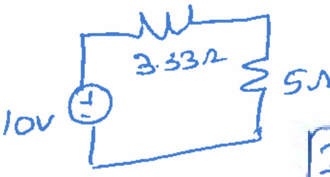


$2\Omega // 4\Omega + 2 = \frac{2\times4}{2+4} + 2$

$= \frac{8}{6} + 2 = 3.33\Omega$

$R_{Th} = 3.33\Omega$

$V_{oc} = 10V$



$\Rightarrow I_L = \frac{10}{3.33+5}$

$\boxed{I_L = \frac{10}{8.33} = 1.2A}$

(4m)

11b) Maximum power transfer theorem: (2m) It states that maximum power is delivered from a source to its load when load resistance equal to source resistance.

maximum power transfers when $\boxed{R_L = R_{source}}$

$P_{max} = \frac{V_{Th}^2}{4R_{Th}}$

Importance (2m)

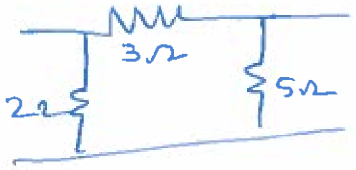
1) It can be implemented to linear N/w, along with R, L, C and resistance linear sources as elements.

2) It functions only when there is variable load.

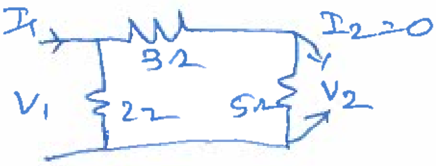
3) Large sound systems are built around this process.

4)

12a) z parameters



→ open circuit the o/p port.



$V_1 = z_{11}I_1 + z_{12}I_2 \Rightarrow V_1 = z_{11}I_1 + 0$

$V_2 = z_{21}I_1 + z_{22}I_2 \Rightarrow V_2 = z_{21}I_1 + 0$

$\boxed{\frac{V_1}{I_1} = z_{11}}$ $\boxed{\frac{V_2}{I_1} = z_{21}}$

$R_{eq} = 2 + \frac{3\times5}{3+5} = 2 + \frac{15}{8} = 3.875\Omega$

$V_1 = I_1\, 3.875$

$\boxed{\frac{V_1}{I_1} = 3.875\Omega = z_{11}}$

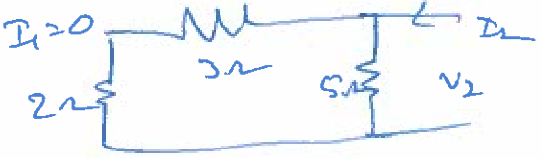
$V_2 = I R$

$V_2 = I_1 \times \frac{2}{\cancel{10}\,2} \times \cancel{8}$

$V_2 = I_1 \times 1$

$\boxed{\frac{V_2}{I_1} = 1\Omega = z_{21}}$

→ open circuit the input port



$V_1 = 0 + z_{12}I_2 \Rightarrow z_{12} = V_1/I_2$

$V_2 = 0 + z_{22}I_2 \Rightarrow z_{22} = V_2/I_2$

$R_{eq} = (2+3) // 5\Omega = \frac{5\times5}{5+5} = \frac{25}{10} = 2.5\Omega$

$V_2 = 2.5\, I_2$

$\boxed{\frac{V_2}{I_2} = 2.5\Omega = z_{22}}$

$V_1 = 2\times I \Rightarrow V_1 = \cancel{2}\times I_2 \times \frac{5}{\cancel{10}\,5}$

$V_1 = I_2 \times 1$

$\boxed{\frac{V_1}{I_2} = 1\Omega = z_{12}}$

12b) h parameters interms of ABCD:-

The defining equations of h parameters ⇒ $V_1 = h_{11}I_1 + h_{12}V_2$

$I_2 = h_{21}I_1 + h_{22}V_2$

ABCD parameters ⇒ $V_1 = AV_2 - BI_2$

$I_1 = CV_2 - DI_2$

Rewriting the second equ :- $I_2 = -\frac{1}{D}I_1 + \frac{C}{D}V_2$

Comparing with $I_2 = h_{21}I_1 + h_{22}I_2$

$\boxed{h_{21} = -\frac{1}{D}}$

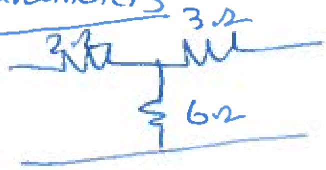
$\boxed{h_{22} = \frac{C}{D}}$

also $V_1 = AV_2 - B\left[-\frac{1}{D}I_1 + \frac{C}{D}V_2\right] = \frac{B}{D}I_1 + \left(\frac{AD-BC}{D}\right)V_2$

$V_1 = h_{11}I_1 + h_{12}V_2$

$\boxed{h_{11} = \frac{B}{D}}$ $\boxed{h_{12} = \frac{AD-BC}{D} = \frac{\Delta T}{D}}$ ⑧

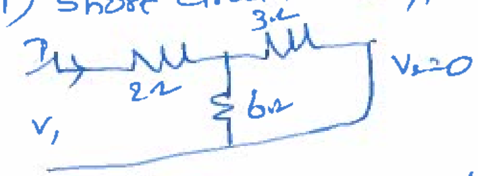
13a) Y-Parameters



$I_1 = Y_{11}V_1 + Y_{12}V_2$

$I_2 = Y_{21}V_1 + Y_{22}V_2$

1) Short Circuit the o/p Port



$I_1 = Y_{11}V_1 + 0 \Rightarrow Y_{11} = I_1/V_1$

$I_2 = Y_{21}V_1 + 0 \Rightarrow Y_{21} = I_2/V_1$

$R_{eq} = \frac{3\times6}{3+6} + 2 = \frac{18}{9} + 2$
$= 4\Omega$

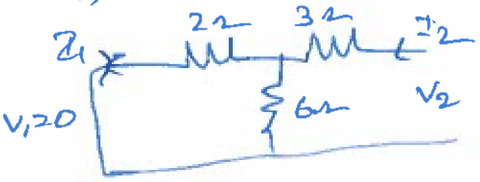
$V_1 = I_1 4 \Rightarrow \boxed{\frac{I_1}{V_1} = \frac{1}{4}\mho = Y_{11}}$

$\Rightarrow I_1 = V_1/4$

$I_2 = I_1 \times \frac{6}{6+3} \Rightarrow I_2 = \frac{V_1}{\cancel{4}_{2}} \times \frac{\cancel{6}^{3}}{9}$

$\boxed{\frac{I_2}{V_1} = \frac{3}{18}\mho = Y_{21}}$

2) Short Circuit the i/p Port.



$I_1 = 0 + Y_{12}V_2 \Rightarrow Y_{12} = I_1/V_2$

$I_2 = 0 + Y_{22}V_2 \Rightarrow Y_{22} = I_2/V_2$

$\frac{2\times6}{2+6} + 3 = \frac{12}{8} + 3 = 4.5\Omega$

$V_2 = I_2 4.5 \Rightarrow \boxed{I_2 = V_2/4.5}$

$\boxed{\frac{I_2}{V_2} = \frac{1}{4.5}\mho = Y_{22}}$

$I_1 = I_2 \times \frac{6}{6+2} = \frac{V_2}{4.5} \times \frac{6}{8}$

$\Rightarrow \frac{I_1}{V_2} = \frac{1}{4.5} \times \frac{6}{8} = 0.16$

$\Rightarrow \boxed{\frac{I_1}{V_2} = Y_{12} = 0.16\mho}$

13b) Relation between z and y parameters:

The defining equations

$I_1 = Y_{11}V_1 + Y_{12}V_2$

$I_2 = Y_{21}V_1 + Y_{22}V_2$

$V_1 = z_{11}I_1 + z_{12}I_2$

$V_2 = z_{21}I_1 + z_{22}I_2$

$$V_1 = \frac{\begin{vmatrix} I_1 & Y_{12} \\ I_2 & Y_{22} \end{vmatrix}}{\begin{vmatrix} Y_{11} & Y_{12} \\ Y_{21} & Y_{22} \end{vmatrix}} = \frac{Y_{22}I_1 - Y_{12}I_2}{Y_{11}Y_{22} - Y_{12}Y_{21}} = \frac{Y_{22}}{\Delta Y}I_1 - \frac{Y_{12}}{\Delta Y}I_2$$

$\Delta Y = Y_{11}Y_{22} - Y_{12}Y_{21}$

$V_1 = z_{11}I_1 + z_{12}I_2$

$\boxed{z_{11} = \frac{Y_{22}}{\Delta Y}}$

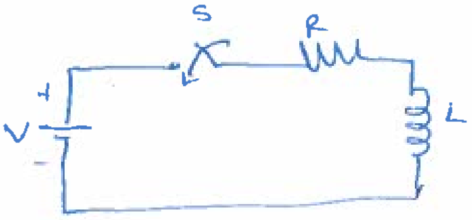
$\boxed{z_{12} = -\frac{Y_{12}}{\Delta Y}}$

then $V_2 = \frac{\begin{vmatrix} Y_{11} & I_1 \\ Y_{21} & I_2 \end{vmatrix}}{\Delta Y} = \frac{Y_{11}}{\Delta Y}I_2 - \frac{Y_{21}}{\Delta Y}I_1$

$V_2 = z_{21}I_1 + z_{22}I_2$

$\boxed{z_{22} = \frac{Y_{11}}{\Delta Y}}$ $\boxed{\Delta_{21} = -\frac{Y_{21}}{\Delta Y}}$

14) R-L Series Circuit:-



apply KVL to above circuit.

$$V = iR + L\frac{di}{dt}$$

$$\Rightarrow \frac{di}{dt} + \frac{R}{L}i = V/L$$

Compare with non-homogenous equ.

$$\frac{dx}{dt} + Px = k \Rightarrow x = e^{-Pt}\int ke^{Pt}dt + ce^{-Pt}$$

$$i = ce^{-(R/L)t} + e^{-(R/L)t}\int \frac{V}{L}e^{(R/L)t}dt$$

$$i = ce^{-(R/L)t} + V/R \quad -①$$

at $t=0$, $i=0$, substitute in equ①

$$0 = c + V/R$$

$$\boxed{C = -V/R}$$

Hence $i = (V/R - V/R\, e^{-(R/L)t})$

$$i = V/R(1-e^{-(R/L)t})$$

Voltage across Resistor:-

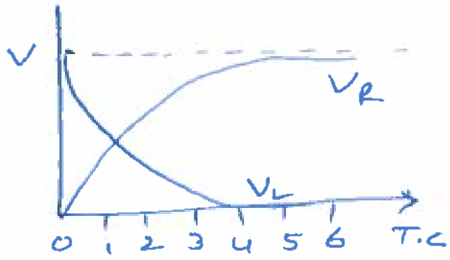
$$V_R = IR = \frac{V}{\cancel{R}}(1-e^{-(R/L)t})\cancel{R}$$

$$\boxed{V_R = V(1-e^{-(R/L)t})}$$

Voltage across inductor.

$$V_L = L\frac{di}{dt} = L\times\frac{V}{R}\times\frac{R}{L}e^{(-R/L)t}$$

$$\boxed{V_L = Ve^{-(R/L)t}}$$



Power, $P_R = V_R i = V(1-e^{-(R/L)t}) \times \frac{V}{R}(1-e^{-(R/L)t})$

$$P_R = \frac{V^2}{R}\left(1-2e^{-(R/L)t} + e^{-\left(\frac{2R}{L}\right)t}\right)$$

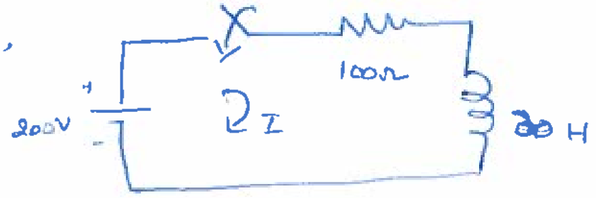
Power, $P_L = V_L i = Ve^{-(R/L)t} \times \frac{V}{R}(1-e^{-(R/L)t})$

$$P_L = \frac{V^2}{R}\left(e^{-(R/L)t} - e^{-\left(\frac{2R}{L}\right)t}\right)$$

15) given series RL Circuit,

$R = 100\Omega \quad L = 20H$

Voltage = 200V



apply KVL, we get

$$V = IR + L\frac{di}{dt} \Rightarrow 20\frac{di}{dt} + 100i = 120$$

$$\Rightarrow \frac{di}{dt} + 5i = 6 \qquad \left(\frac{d}{dt} + P\right)x = K$$

$$\Rightarrow \left(\frac{d}{dt} + 5\right)i = 6 \qquad P = 5 \; x = i, K = 6$$

$$i = ce^{-Pt} + e^{-Pt}\int K e^{Pt} dt \Rightarrow P = 5, K = 6$$

$$\Rightarrow i = ce^{-5t} + e^{-5t}\int 6e^{5t} dt$$

$$\Rightarrow i = ce^{-5t} + e^{-5t} \times \frac{e^{5t}}{5} \times 6$$

$$\Rightarrow i = ce^{-5t} + \frac{6}{5}$$

at $t = 0^+$, $i = 0$ substitute in above

$$\Rightarrow 0 = ce^{-5(0)} + \frac{6}{5}$$

$$\Rightarrow 0 = c + 6/5 \Rightarrow c = -6/5$$

$$\Rightarrow i = -\frac{6}{5}e^{-5t} + \frac{6}{5}$$

$$i = +\frac{6}{5}(-e^{-5t} + 1)$$

$$i = \frac{6}{5}(1 - e^{-5t})$$

Voltage across resistor $V_R = iR$

$$\Rightarrow V_R = \frac{6}{\cancel{5}}(1 - e^{-5t}) \times \overset{20}{\cancel{100}}$$

$$\Rightarrow \boxed{V_R = 120(1 - e^{-5t})}$$

voltage across inductor $V_L = L\frac{di}{dt}$

$$\Rightarrow V_L = 20\frac{d}{dt}\left(\frac{6}{5}(1 - e^{-5t})\right)$$

$$V_L = 20 \times \frac{6}{5}\frac{d}{dt}(1 - e^{-5t})$$

$$V_L = \overset{4}{\cancel{20}} \times \frac{6}{\cancel{5}}(-e^{-5t})$$

$$V_L = 24e^{-5t}(+5)$$

$$\boxed{V_L = +120e^{-5t}}$$

~~$V_L =$~~

ii) current at $t = 0.5$ sec

$$i = \frac{6}{5}\left(1 - e^{-5(0.5)}\right)$$

$$i = \frac{6}{5}(1 - e^{-2.5})$$

$$i = \frac{6}{5}(0.918) \Rightarrow i = 1.2(0.918)$$

$$\Rightarrow \boxed{i = 1.1016A}$$

iii) current at $t = 1$ sec

$$i = \frac{6}{5}(1 - e^{5(1)})$$

$$i = \frac{6}{5}(1 - e^{-5})$$

$$\boxed{i = 1.192\,A}$$

iv) when $V_R = V_L$

$$\Rightarrow 12\cancel{0}(1 - e^{-5t}) = 1\cancel{2}0e^{-5t}$$

$$\Rightarrow \quad 1 = e^{-5t} + e^{-5t}$$

$$\Rightarrow \quad 1 = 2e^{-5t}$$

$$\Rightarrow \quad \frac{1}{2} = e^{-5t}$$

apply log on bothsides.

$$\Rightarrow \log(\tfrac{1}{2}) = \cancel{\log}(\cancel{e})^{-5t}$$

$$\Rightarrow \quad -5t = \log(\tfrac{1}{2})$$

$$\Rightarrow \quad t = -\frac{1}{5}\log(\tfrac{1}{2})$$

$$t = -\frac{1}{5}(-0.30)$$

$$\boxed{t = 0.06 \text{ sec}}$$

Faculty
3/01/23

HoD - EEE
3/01/23.

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2022 - 2023
Course Code	20EC303	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	Signals and Systems						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define signal and mention its mathematical equation of Unit step signal	20EC303.1	L1
2	What is the Fourier transform of the impulse function δ(t)	20EC303.2	L1
3	Differentiate convolution and correlation	20EC303.3	L1
4	Draw the ideal filter characteristics of a Low pass filter	20EC303.4	L1
5	Find the Z-transform of the sequence $a^nU[n]$	20EC303.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Classify the systems with an example	6M	20EC303.1	L1
6 (b)	Give the mathematical equation and plot the waveform for the following functions a) δ(t) b) U(n) c) Analog Sinusoidal Signal	6M	20EC303.1	L2
	OR			
7	Check whether the following system is static or Dynamic, Linear or non Linear, Causal or Non causal, Time Variant or invariant i) y(t) = 10x(t) +5 ii) y(n) = x(n)+x(n-1) iii) y(t) = x(2t)	12M	20EC303.1	L2
8(a)	Develop the expression for mean square error using the expression of a function using orthogonal signal space	6M	20EC303.2	L3
8 (b)	Find the Fourier transform of a Full wave rectified output whose fundamental period is 2π	6M	20EC303.2	L1
	OR			
9	Prove the following properties of Fourier Transform 1) Linearity 3) Time Reversal 5) Frequency Shifting 2) Time shifting 4) Differentiation in Time Domain 6) Scaling	12M	20EC303.2	L2
10 (a)	Perform the convolution of the two sequences x[n]={-2,2,-2,2} and h[n]= {1,-1,1,-1}	6M	20EC303.3	L3
10 (b)	Explain cross correlation function, write any 4 properties and prove any two of them.	6M	20EC303.3	L4
	OR			
11 (a)	State and prove Parseval's theorem for energy / power signals	6M	20EC303.3	L2
11 (b)	Perform the convolution of $h(t)=e^{-2t}u(t)$ and $x(t)=e^{-3t}u(t)$	6M	20EC303.3	L2
12	Derive the Relationship between Bandwidth and Rise time	12M	20EC303.4	L2
	OR			
13	Enumerate the difference between Impulse sampling, Flat Top Sampling and Natural Sampling	12M	20EC303.4	L2
14	Obtain the Z-transform of $x(n) = a^nU(n) - b^nU(-n-1)$ indicate the ROC	12M	20EC303.5	L3
	OR			
15	Obtain the Laplace transform of the following signals i) Impulse function ii) unit step function iii) A $\sin w_0 t$ u(t)	12M	20EC303.5	L3

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Subject: Signals and Systems

Course code: 20EC303

PART A:
1 Definition of signal [1 Mark]
mathematical equation of unit step [1 mark]
2. Fourier transform of $\delta(t)$ formula & · [2, marks]
3. 2 Difference [2 marks]
4. Diagram and mathematical Expression [2 marks]
5. Z. transfer formula and derivation [2 marks].

PART B:
6)a) classification of system
write Any four with Example [4 × 1.5 = 6 marks]
6) b) Each Mathematical Equation and Diagram [3 × 2 = 6 marks]
7) For Each system check [1 mark]
3 systems given $3 \times 4 = 12$ marks
8) a) mean square Error Expression derivation using orthogonal signal space [5 + 1 = 6 marks]

8) b) Fourier Series Expression — 4 marks

Fourier transform — 2 marks

10) a) Analytical method — 6 marks

10) b) Cross Correlation definition — 1 marks
3 properties — 2 marks
proof of two properties — 3 marks
} 6 marks

11) a) statement and proof — 1 + 5 = 6 marks

b) F.T of Each signal + using Convolution property + partial fraction + IFT   1 + 1 + 1 + 3 = 6 marks

12) a) Lowpass filter Transfer function H(ω), + output Response Evaluation y(t) + Rise time Definition tr + Relation } 3+3+3+3 = 12 marks

13) Eight Differences   8 × 1.5 = 12 marks

Each Difference = 1.5 mark

14) $Z[a^n u[n]]$ and its ROC – 4 marks
$Z[-b^n u[-n-1]]$ and its ROC – 4 marks
Combination and analysis of ROC – 4 marks
} 12 marks

15) Laplace transform of Each signal carry 4 marks } 12 marks
3 signals = 3 × 4 marks

3/1/2023

HOD 3/1/23

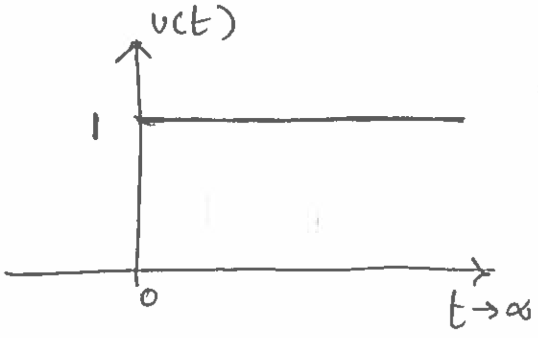
PART - A

1. Define Signal and mention its mathematical Equation of unit step signal. [2 Marks]

Sol **Signal:** A signal is a physical Quantity is a function of time (or) space which Conveys Some Information

Unit step function: $u(t) = \begin{cases} 1 & \text{for } t \geq 0 \\ 0 & \text{for } t < 0 \end{cases}$

The Amplitude of unit step function is equal to unity for t values greater than equal to 0, and Amplitude is Zero for t values less than zero.



2. What is the Fourier transform of the Impulse function δ(t)? [2 Marks]

Sol:

$$\mathcal{F}[\delta(t)] = \int_{-\infty}^{\infty} \delta(t)\, e^{-j\omega t}\, dt \qquad \delta(t) = \begin{cases} 1 & \text{for } t=0 \\ 0 & \text{Elsewhere} \end{cases}$$

$$= \int_{-\infty}^{\infty} \delta(t)\, dt \qquad \text{At } t=0$$

$$= e^{-j\omega t}\Big|_{t=0} = 1$$

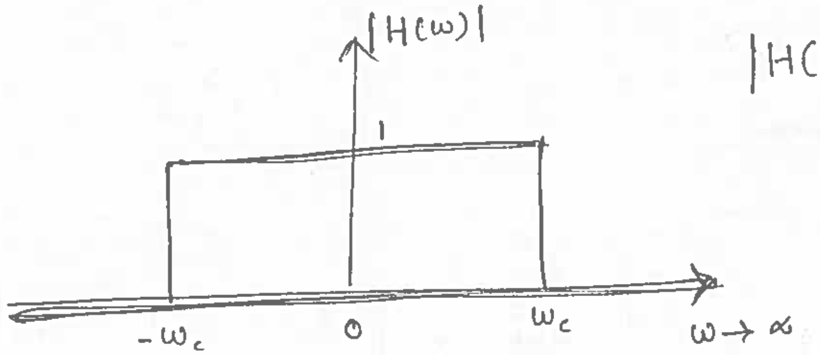
$$\mathcal{F}[\delta(t)] = 1 \quad \text{(or)} \quad \boxed{\delta(t) \longleftrightarrow 1}$$

③ Differentiate convolution and correlation? — $\boxed{\text{2 marks}}$

Convolution	Cross correlation
① $x_1(t) * x_2(t) = \int_{-\infty}^{\infty} x_1(\tau)x_2(t-\tau)d\tau$	① $R_{12}(\tau) = \int_{-\infty}^{\infty} x_1(t)x_2^*(t-\tau)dt$
② Convolution is a mathematical way of combining two signals to form a third signal.	② correlation also uses two signals to form a third signal.
③ Convolution is mathematical operation which is used to Express the Input-output relationship of an LTI system.	③ The signals may be compared on the Basis of similarity of waveforms

④ Draw the Ideal filter characteristics of a Low pass filter: $\boxed{\text{2-Marks}}$

Sol



$$|H(\omega)| = \begin{cases} 0, & |\omega| < \omega_c \\ 1 & |\omega| > \omega_c \end{cases}$$

Ideal LPF characteristics

⑤ Find the Z-transform of sequence $a^n u[n]$ $\boxed{\text{2-marks}}$

Sol

$$Z[a^n u[n]] = \sum_{n=-\infty}^{\infty} a^n u[n] z^{-n}$$

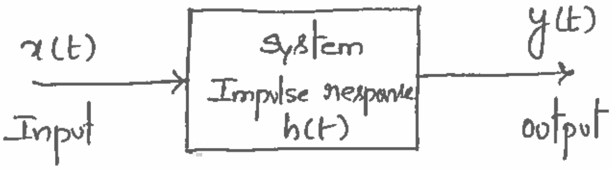
$$= \sum_{n=0}^{\infty} a^n z^{-n} = \sum_{n=0}^{\infty} (a'z^{-1})^n$$

$$= \frac{1}{1-az^{-1}} = \frac{z}{z-a}$$

6) a) classify the systems with an Example ?

Sol A system is defined as an Entity that acts on an Input signal and transform it into an output signal

⇒ A system is represented by a block diagram as shown in fig.



A system

⇒ An arrow entering the box is the input signal (also called Excitation, source or driving function and an arrow leaving the box is an output signal (also called response).

⇒ Generally, the input is denoted by x(t) and the output is denoted by y(t).

System may be classified as under

1. static (memoryless) and dynamic (memory) systems.
2. Causal and non-causal systems
3. Linear and non-Linear systems
4. Time-invariant and time varying systems.
5. stable and unstable systems.
6. Invertible and non-invertible system
7. FIR and IIR Systems.

Static and Dynamic systems

A System is said to be static (or) memoryless if the response is due to present input alone

Eg: $y(t) = x(t)$

A system is said to be Dynamic (or) memory system if the response depends upon past or future Inputs.

Eg: $y(t) = x(t-1)$

Causal and Non Causal systems

A system is said to be causal if the output of the system at any time t depends only on the present and past values of the Input but not on feutune Inputs.

Eg: $y(t) = x(t-2) + 2x(t)$

A system is said to be non-Causal if the output of the system at any time t depends on future Inputs. $y(t) = x(t+2) + x(t)$.

Linear and Non-Linear System

→ A system which obeys the principle of superposition and principle of homogeneity is called a linear system.

Eg: $\frac{d^2y(t)}{dt^2} + 2ty(t) = t^2x(t)$

→ A system which does not obey the principle of superposition and homogeneity is called a non-linear system

Eg: $2\frac{dy(t)}{dt} + 5y(t) = x^2(t)$

Time-invariant and Time-Varying System

Time-invariance is the property of a system which makes the behaviour of the system Independent of time

Eg: $y(t) = e^{2x(t)}$

⇒ A system is said to be time-invariant if its Input/output characteristics do not change with time,

⇒ A system not satisfying the above requirements is called a time-varying system.

Eg: $y(n) = x(n) + nx(n-2)$.

Stable and unstable Systems

A System is said to be bounded-input, bounded-output (BIBO) stable, if and only if Every bounded input produces a bounded output.

i.e if $y(t)$ is also bounded, then the System is BIBO stable. otherwise, the System is unstable.

Eg: $y(t) = e^{x(t)}$; $|x(t)| \leq 8$ ⟹ Stable Systems

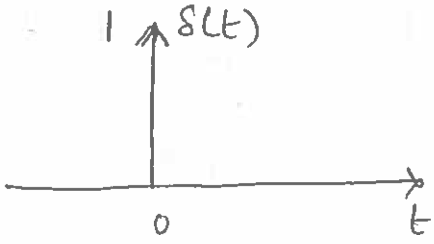
$y(t) = (t+5)\,u(t)$ ⟹ unstable System.

6(b) Give the mathematical equation and plot the waveform for the following function

Sol (i) $\delta(t)$ [unit Impulse function]

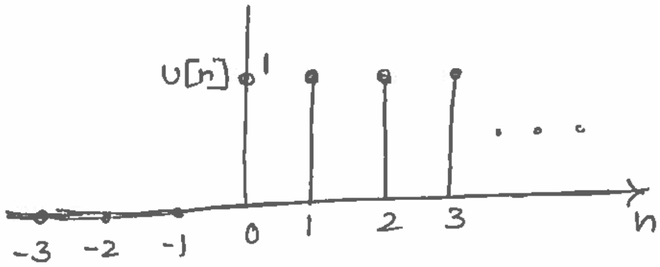
$$\delta(t) = 0 \quad \text{for } t \neq 0$$

i.e as $$\delta(t) = \begin{cases} 1 & \text{for } t = 0 \\ 0 & \text{for } t \neq 0 \end{cases}$$



unit Impulse

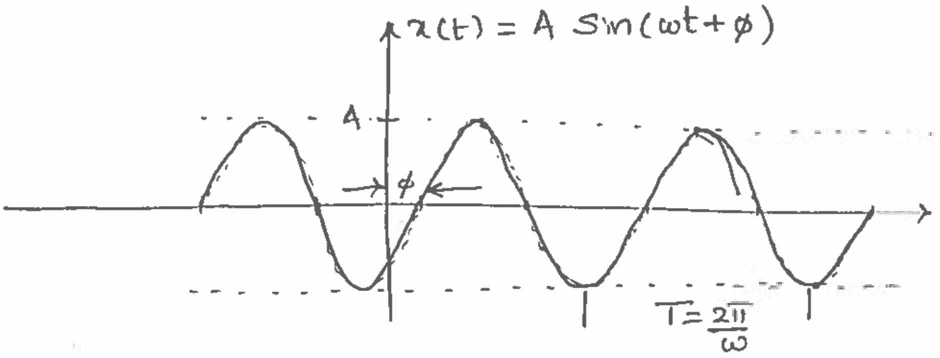
Unit step sequence: $u[n] = \begin{cases} 1 & \text{for } n \geq 0 \\ 0 & \text{for } n < 0 \end{cases}$



Sinusoidal Signal

A continuous-time sinusoidal signal in its most general form is given by $x(t) = A \sin(\omega t + \phi)$

where A = Amplitude
ω = Angular frequency in radians
ϕ = phase angle in radians.



Sinusoidal waveform

⑦ check whether the following system is static (or) Dynamic, linear or non-Linear, Causal or Non Causal, Time variant(or) Invariant

(i) $y(t) = 10\,x(t) + 5$

Sol

a) It is a static system

b) Given ~~x(t)~~ $y(t) = 10\,x(t) + 5$

for $x_1(t)$, $y_1(t) = 10x_1(t) + 5$)

for $x_2(t)$, $y_2(t) = 10x_2(t) + 5$.

⟹ weighted sum of output =

$$a y_1(t) + b y_2(t)$$

$$= a\left[10x_1(t) + 5\right] + b\left[10x_2(t) + 5\right]$$

$$= 10\left(\left[a x_1(t) + b x_2(t)\right] + 5\left[a+b\right]\right. \leftarrow ①$$

⟹ weighted sum of inputs

$$y_3[t] = T\left[a x_1(t) + b x_2(t)\right]$$

$$= 10\left[a x_1(t) + b x_2(t)\right] + 5 \leftarrow ②$$

Eq-① ≠ Eq-②

so The given system is not Linear system.

c) The given system is causal System.

d) The given System is Time Invariant.

Because $y(t-T) = y(t,T)$

i.e The output ~~delye~~ delayed by T sec is Equal to The output due to input delayed by T sec.

(ii) $y[n] = x[n] + x[n-1]$

a) It is ~~no~~ a dynamic system.

b) for $x_1[n]$, $y_1[n] = x_1[n] + x_1[n-1]$

for $x_2[n]$, $y_2[n] = x_2[n] + x_2[n-1]$

Weighted sum of output, i.e

$$a\,y_1(n) + b\,y_2[n] = a\,T[x_1[n]] + b\,T[x_2[n]]$$

$$= a\,x_1[n] + a\,x_1[n-1] + b\,x_2[n] + b\,x_2[n-1] \quad \text{—} ①$$

The output due to weighted sum of Input $y_3[n]$.

$$y_3[n] = T[a x_1[n] + b x_2[n]]$$

$$= a x_1[n] + b x_2[n] + a x_1[n-1] + b x_2[n-1]$$

The given System is Linear.

c) The given system is causal system.
Because the output depends on present and past values of time

d) The given system is Time Invariant.

(iii) $y(t) = x(2t)$

a) The given system is dynamic system.

b) The given system is Linear system, Because it abeys superposition theorem. $T[ax_1(t)+bx_2(t)] = aT[x_1(t)] + bT[x_2(t)]$

c) The given system is non causal system.
Because it the output depends on future values of time

d) The given system is Time Invariant.
It obeys $\boxed{y(t-T) = y(t,T)}$

8) a) Develop the Expression of mean square Error using the Expression of a function using orthogonal signal space.

Sol Lets find the value of ε when optimum values of coefficients $C_1, C_2, C_3 \ldots C_n$ are chosen according to the equation for $c_j = \dfrac{\int_{t_1}^{t_2} x(t) g_j(t)\,dt}{\int_{t_1}^{t_2} g_j^2(t)\,dt}$

By definition $$\varepsilon = \frac{1}{t_2 - t_1} \int_{t_1}^{t_2} \left[x(t) - \sum_{r=1}^{n} C_r g_r(t)\right]^2 dt$$

$$= \frac{1}{t_2-t_1}\left[\int_{t_1}^{t_2} x^2(t)\,dt + \sum_{r=1}^{n} C_r^2 \int_{t_1}^{t_2} g_r^2(t)\,dt - 2\sum_{r=1}^{n} C_r \int_{t_1}^{t_2} x(t) g_r(t)\,dt\right]$$

But from the equation for c_j, we have

$$\varepsilon = \frac{1}{t_2-t_1}\left[\int_{t_1}^{t_2} x^2(t)\,dt + \sum_{r=1}^{n} C_r^2 K_r - 2\sum_{r=1}^{n} C_r^2 K_r\right]$$

$$= \frac{1}{t_2-t_1}\left[\int_{t_1}^{t_2} x^2(t)\,dt - \sum_{r=1}^{n} C_r^2 K_r\right]$$

$$= \frac{1}{t_2-t_1}\left[\int_{t_1}^{t_2} x^2(t)\,dt - (C_1^2 K_1 + C_2^2 K_2 + \cdots + C_n^2 K_n)\right]$$

The mean square Error can therefore be Evaluated using this equation for ε.

8) b) Find the Fourier transform of a full wave rectified output whose fundamental period is 2π?

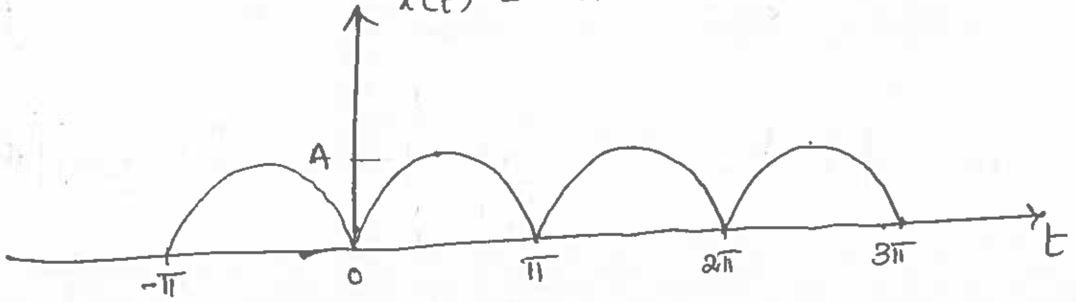
Sol

→ The fundamental period ~~$\omega_0 = \frac{2\pi}{2\pi}$~~ $= T = 2\pi$

fundamental frequency $\omega_0 = \frac{2\pi}{T} = \frac{2\pi}{2\pi} = 1.$

→ The fullwave rectified output

$x(t) = A \sin \omega_0 t = A \sin t \quad 0 \le t \le \pi$



$$\Longrightarrow \quad x(t) = \sum_{n=-\infty}^{\infty} C_n e^{jn\omega_0 t} \qquad \omega_0 = \frac{2\pi}{T}$$

$$= \cancel{\sum_{n=-\infty}^{\infty}} \qquad C_n = \frac{1}{2\pi}\int_0^{2\pi} A \sin t \, e^{-jn\omega_0 t}\, dt$$

$$= \frac{A}{2\pi} \; \frac{1}{jn^2\omega_0^2 + 1}$$

$$= \cancel{\frac{A}{2\pi} e^{-jn\omega_0 t} \cos t} \qquad C_n = \frac{1}{2\pi}\int_0^{2\pi} \left(\frac{e^{+jt} - e^{-jt}}{2j}\right)\left(e^{-jnt}\right) dt$$

$$C_n = \frac{1}{4\pi j}\int_0^{2\pi} \left[e^{j(1-n)t} - e^{-j(1+n)t}\right] dt$$

$$= \frac{1}{4\pi j}\left[\left[\frac{e^{j(1-n)t}}{j(1-n)}\right]_0^{2\pi} + \left[\frac{e^{j(1+n)t}}{j(1+n)}\right]_0^{2\pi} \right]$$

$$= \frac{1}{4\pi j}\left[\frac{e^{j(1-n)2\pi}-1}{j(1-n)} - \frac{e^{j(1+n)2\pi}-1}{j(1+n)}\right]$$

$$= \frac{1}{4\pi j}\left[\frac{-1}{j(1-n)} + \frac{1}{j(1+n)}\right]$$

$$= \frac{1}{4\pi j^2}\left[\frac{-1}{(1-n)} + \frac{1}{(1+n)}\right]$$

$$C_n = \left(\frac{-1}{4\pi}\right)\frac{-\cancel{1}-n+\cancel{1}-n}{(1-n^2)} = \frac{+2n}{4\pi(1-n^2)}$$

$$=$$

$$x(t) = \sum_{n=-\infty}^{\infty} \frac{2n}{4\pi(1-n^2)}\, e^{jnt}$$

$$\mathcal{F}[x(t)] = \sum_{n=-\infty}^{\infty} \frac{2n}{4\pi(1-n^2)} \int_{-\infty}^{\infty} e^{jnt}\, e^{-j\omega t}\, dt$$

$$f[x(t)] = \sum_{n=-\infty}^{\infty} \frac{2n}{4\pi(1-n^2)} \underline{\int_{-\infty}^{\infty} e^{j(n-\omega)t}\, dt}$$

$$\boxed{X(\omega) = \sum_{n=-\infty}^{\infty} \frac{2n}{4\pi(1-n^2)}\, X(n-\omega)}$$

9) prove the following properties of Fourier transform?

Sol

Linearity: It states that the fourier transform of a weighted sum of two signals is equal to the weighted sum of their individual fourier transform.

i.e if $x_1(t) \xleftrightarrow{FT} X_1(\omega)$ and $x_2(t) \xleftrightarrow{FT} X_2(\omega)$

Then $ax_1(t)+bx_2(t) \xleftrightarrow{FT} aX_1(\omega)+bX_2(\omega)$

where a and b are constants

proof: $$\mathcal{F}[ax_1(t)+bx_2(t)] = \int_{-\infty}^{\infty}[ax_1(t)+bx_2(t)]e^{-j\omega t}\,dt$$

$$= \int_{-\infty}^{\infty} ax_1(t)e^{-j\omega t}dt + \int_{-\infty}^{\infty} bx_2(t)e^{-j\omega t}dt$$

$$= a\int_{-\infty}^{\infty} x_1(t)\,e^{-j\omega t} + b\int_{-\infty}^{\infty} x_2(t)e^{-j\omega t}dt$$

$$= aX_1(\omega) + bX_2(\omega)$$

$$\therefore \boxed{ax_1(t)+bx_2(t) \xleftrightarrow{FT} aX_1(\omega)+bX_2(\omega)}$$

2) Time shifting: The Time shifting property states that if a signal $x(t)$ is shifted by t_0 sec, the spectrum is modified by a linear phase shift of slow $-\omega_0 t_0$, ie

If $x(t) \xleftrightarrow{FT} X(\omega)$

then $x(t-t_0) \xleftrightarrow{FT} e^{-j\omega t_0}X(\omega)$

proof: By definition $\mathcal{F}[x(t-t_0)] = \int_{-\infty}^{\infty} x(t-t_0)\, e^{-j\omega t}\, dt$

Let $t - t_0 = p$

$\therefore$ $t = p + t_0$ and $dt = dp$

$$\therefore \quad \mathcal{F}[x(t-t_0)] = \int_{-\infty}^{\infty} x(p)\, e^{-j\omega(p+t_0)}\, dp$$

$$= e^{-j\omega t_0} \int_{-\infty}^{\infty} x(p)\, e^{-j\omega p}\, dp$$

$$= e^{-j\omega t_0}\, X(\omega)$$

$$\therefore \quad \boxed{x(t-t_0) \xleftrightarrow{FT} e^{-j\omega t_0}\, X(\omega)}$$

3) Time Reversal property

The time reversal property states that

if $x(t) \xleftrightarrow{FT} X(\omega)$

Then $x(-t) \xleftrightarrow{FT} X(-\omega)$

proof: By definition $\mathcal{F}[x(-t)] = \int_{-\infty}^{\infty} x(-t)\, e^{-j\omega t}\, dt$

Replacing t by $-t$ in the RHS of the above Expression for $\mathcal{F}[x(-t)]$, we have

$$\mathcal{F}[x(-t)] = \int_{-\infty}^{\infty} x(t)\, e^{j\omega t}\, dt = \int_{-\infty}^{\infty} x(t)\, e^{-j(-\omega)t}\, dt = X(-\omega)$$

$$\boxed{x(-t) \xleftrightarrow{FT} X(-\omega)}$$

4) Differentiation in Time Domain property

The differentration in time domain property states that the differentration of a function in time domain is equivalent to the multiplication of its fourier transform by factor jω, i.e

If $x(t) \xleftrightarrow{FT} X(\omega)$

Then $\frac{d}{dt}x(t) \xleftrightarrow{FT} j\omega X(\omega)$

proof: By definition $x(t) = \frac{1}{2\pi}\int_{-\infty}^{\infty} X(\omega) e^{j\omega t} d\omega$

Differentiating both sides w.r.t t, we have

$$\frac{d}{dt}(x(t)) = \frac{1}{2\pi}\frac{d}{dt}\left[\int_{-\infty}^{\infty} X(\omega) e^{j\omega t} d\omega\right]$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} X(\omega)\frac{d}{dt}[e^{j\omega t}]d\omega = \frac{1}{2\pi}\int_{-\infty}^{\infty} X(\omega)\, j\omega e^{j\omega t} d\omega$$

$$= j\omega\left[\frac{1}{2\pi}\int_{-\infty}^{\infty} X(\omega) e^{j\omega t} d\omega\right]$$

$$= j\omega\, F^{-1}[X(\omega)]$$

$$\therefore F\left[\frac{d}{dt}x(t)\right] = j\omega X(\omega)$$

$$\therefore \boxed{\frac{d}{dt}x(t) \xleftrightarrow{FT} j\omega X(\omega)}$$

⑤ Frequency shifting property

Frequency shifting property states that the multiplication of a time domain signal $x(t)$ by $e^{j\omega t_0}$ results in the frequency spectrum shifted by ω_0, i.e

proof: If $x(t) \xleftrightarrow{FT} X(\omega)$

Then $e^{j\omega_0 t} x(t) \longleftrightarrow X(\omega - \omega_0)$

proof: By definition

$$\mathcal{F}\left[e^{j\omega_0 t} x(t)\right] = \int_{-\infty}^{\infty} e^{j\omega_0 t} x(t) e^{-j\omega t} dt$$

$$= \int_{-\infty}^{\infty} x(t) e^{-j(\omega - \omega_0)t} dt$$

$$= X(\omega - \omega_0)$$

$$\therefore \boxed{e^{j\omega_0 t} x(t) \xleftrightarrow{FT} X(\omega - \omega_0)}$$

⑥ Time Scaling property

Let $x(at)$ is a Compressed Version of $x(t)$ when $a > 1$ or Expanded Version of $x(t)$ when $a < 1$.

~~Sta~~: If $x(t) \xleftrightarrow{FT} X(\omega)$ Then $x(at) \xleftrightarrow{FT} \frac{1}{|a|} X\left(\frac{\omega}{a}\right)$

proof: By definition $\mathcal{F}\left[x(at)\right] = \int_{-\infty}^{\infty} x(at) e^{-j\omega t} dt$

Let $at = p$ $\therefore t = \frac{p}{a}$ and $dt = \frac{dp}{a}$

$$\mathcal{F}[x(at)] = \int_{-\infty}^{\infty} x(p)\, e^{-j\omega(P/a)} \frac{dp}{a}$$

$$= \frac{1}{a} \int_{-\infty}^{\infty} x(p)\, e^{-j(\omega/a)P} dp$$

10) a) perform the Convolution of the two sequences $x[n] = \{-2,2,-2,2\}$ and $h[n] = \{1,-1,1,-1\}$.

Sol The Convolution of two sequences represented by

$$y[n] = \sum_{K=-\infty}^{\infty} x[K]\, h[n-K]$$

⟹ The length of Two sequences is 4.

So the length of $y[n]$ sequence is $L = 4+4-1 = 7$.

$$y[0] = \sum_{K=-\infty}^{\infty} x[K]\, h[-K]$$

$$= \ldots\ldots + x[-1]\, x[1] + x[0]\, h[0] + x[1]\, h[-1]$$

$$= \qquad + -2\times 1 + \ldots = -2.$$

$$y[1] = \sum_{K=-\infty}^{\infty} x[K]\, h[1-K]$$

$$= \ldots\ldots x[-1]\, h[1+1] + x[0]\, h[1] + x[1]\, h[0] + \ldots\ldots$$

$$= \qquad + (-2)(-1) + (2)(1) + \ldots$$

$$= \ldots + 2 + 2 + \ldots = 4$$

$$y[2] = \sum_{K=-\infty}^{\infty} x[k]h[2-k]$$

$$= \ldots\ldots + x[-1]h[3] + x[0]h[2] + x[1]h[1] + x[2]h[0] + x[3]h[-1]$$

$$= \ldots + (-2)(1) + (2)(-1) + (-2)(1) + (2)(0)$$

$$= \cancel{=2+}\ -2-2-2 = -6$$

$$y[3] = \sum_{K=-\infty}^{\infty} x[k]h[3-k]$$

$$= \ldots\ldots + x[-1]h[4] + x[0]h[3] + x[1]h[2] + x[2]h[1] + x[3]h[0] + x[4]h[-1]$$

$$= + (-2)(-1) + (2)(1) + (-2)(-1) + (2)(1) + \ldots$$

$$= + (2) + (2) + (2) + (2) + \ldots$$

$$= 8$$

$$y[4] = \sum_{k=-\infty}^{\infty} x[k]h[4-k]$$

$$= \ldots + x[0]h[4] + x[1]h[3] + x[2]h[2] + x[3]h[1] + x[4]h[0] + \ldots$$

$$= + (-2)(0) + (2)(-1) + (-2)(1) + (2)(-1) + \ldots$$

$$= (-2) + (-2) + (-2) = -6$$

$$y[5] = \sum_{k=-\infty}^{\infty} x[k]h[5-k]$$

$$= x[0]h[5] + x[1]h[4] + x[2]h[3] + x[3]h[2] + x[4]h[1]$$

$$= (-2)(-1) + (2)(1) = \cancel{2x}\ 2+2=4.$$

$$y[6] = \sum_{k=-\infty}^{\infty} x[k]\,h[6-k]$$

$$= + x[0]h[6] + x[1]h[5] + x[2]h[4] + x[3]h[2]$$

$$+ x[4]h[2] + x[5]h[1] + x[6]h[0]$$

$$= 0 + 0 + 0 + (2)(1) + ---$$

$$= 2$$

$$y[n] = \{-2, 4, -6, 8, -6, 4, 2\}.$$

10) b) Explain Cross correlation function, write any 4 properties and prove any two of them.

Sol **Cross Correlation**

The Cross Correlation between two different waveform or signals is a measure of Similarity (or) match (or) relatedness (or) Coherence between one Signal and the time delayed version of another signal.

Cross Correlation of Energy Signals

Consider two general Complex signals $x_1(t)$ and $x_2(t)$ of finite Energy.

The Cross Correlation of these two Energy signals denoted by $R_{12}(\tau)$ is given by $R_{12}(\tau) = \int_{-\infty}^{\infty} x_1(t)\,x_2^*(t-\tau)\,dt$

properties of autocorrelation function of Energy signals

1. The ~~auto~~ cross correlation exhibit conjugate symmetry

$$R_{12}(\tau) = R_{21}^{*}(-\tau)$$

2. If $R_{12}(0) = 0$ i.e if $\int_{-\infty}^{\infty} x_1(t) x_2^{*}(t) dt = 0$

3. $R_{12}(\tau) \xleftrightarrow{FT} X_1(\omega) X_2^{*}(\omega)$.

→ proof of $R_{12}(\tau) = R_{21}^{*}(-\tau)$

proof The cross correlation of two signals $x_1(t)$ and $x_2(t)$ is given by $R_{12}(\tau) = \int_{-\infty}^{\infty} x_1(t) x_2^{*}(t-\tau) dt$

Let $t-\tau = n$ in the above equation for $R_{12}(\tau)$

$$\therefore \quad R_{12}(\tau) = \int_{-\infty}^{\infty} x_1(n+\tau) x_2^{*}(n) dn$$

Also we know that

$$R_{21}(\tau) = \int_{-\infty}^{\infty} x_2(t) x_1^{*}(t-\tau) dt$$

Let $t = n$ in the above equation $R_{21}(\tau)$

$$\therefore \quad R_{21}(\tau) = \int_{-\infty}^{\infty} x_2(n) x_1^{*}(n-\tau) dn$$

$$\therefore \quad R_{21}^{*}(\tau) = \int_{-\infty}^{\infty} x_2^{*}(n) x_1(n-\tau) dn$$

$$\therefore \quad R_{21}^{*}(-\tau) = \int_{-\infty}^{\infty} x_2^{*}(n) x_1(n+\tau) dn$$

Comparing the above two Equations for $R_{12}(\tau)$ and $R_{21}^{*}(-\tau)$, we can write $\boxed{R_{12}(\tau) = R_{21}^{*}(-\tau)}$

proof of $R_{12}(0) = 0$.

proof: we know $R_{12}(\tau) = \int_{-\infty}^{\infty} x_1(t)\, x_2^{*}(t-\tau)\, dt$ ← ①

Let $\tau = 0$

$$R_{12}(0) = \int_{-\infty}^{\infty} x_1(t)\, x_2^{*}(t-0)\, dt$$

$$= \int_{-\infty}^{\infty} x_1(t)\, x_2^{*}(t)\, dt \quad — ②$$

⟹ If $x_1(t)$ and $x_2^{*}(t)$ are orthogonal to each other in any period.

Then Eq-② becomes 0

so $\boxed{R_{12}(0) = 0}$

11) a) state and prove parseval's theorem for Energy/power signals.

Sol ⇒ parseval's theorem defines the Energy of a Signal in terms of its Fourier transform.

$$E = \int_{-\infty}^{\infty} |x(t)|^2 dt = \frac{1}{2\pi} \int_{-\infty}^{\infty} |X(\omega)|^2 d\omega$$

(or) $$E = \int_{-\infty}^{\infty} |X(f)|^2 df$$

proof: Consider a function x(t) such that

$$x(t) \longleftrightarrow X(\omega)$$

Let $x^*(t)$ be the Conjugate of x(t) such that

The Energy of a signal x(t) is given by

$$E = \int_{-\infty}^{\infty} |x(t)|^2 dt = \int_{-\infty}^{\infty} x(t)\, x^*(t)\, dt = \int_{-\infty}^{} x^*(t)\, x(t)\, dt$$

Replacing x(t) by its inverse Fourier transform,

we have $$E = \int_{-\infty}^{\infty} x^*(t) \left[\frac{1}{2\pi} \int_{-\infty}^{\infty} X(\omega)\, e^{j\omega t} d\omega \right] dt$$

Interchanging order of integration.

$$E = \frac{1}{2\pi} \int_{-\infty}^{\infty} X(\omega) \left[\int_{-\infty}^{\infty} x^*(t)\, e^{j\omega t} dt \right] d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} X(\omega)\, X^*(-\omega)\, d\omega = \frac{1}{2\pi}\int_{-\infty}^{\infty} |X(\omega)|^2\, d\omega$$

Let $\omega = 2\pi f$ $\therefore d\omega = 2\pi\, df$

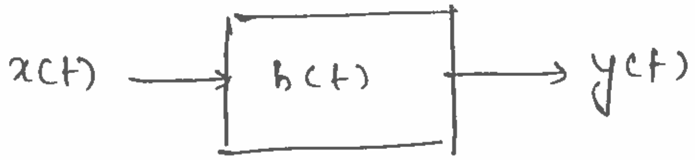
Normally $X(2\pi f)$ is written as $X(f)$, then we have

$$E = \int_{-\infty}^{\infty} |X(f)|^2\, df$$

This is called parseval's theorem for Energy signals.

11)(b) perform the convolution of $h(t) = e^{-2t}u(t)$ and $x(t) = e^{-3t}u(t)$

sol



The relation Between y(t) output response, x(t) and h(t) is given by

$$y(t) = x(t) * h(t) \quad \text{—} \ ①$$

By applying Fourier transform of Eq-①, yields

$$Y(\omega) = X(\omega)\, H(\omega) \quad \text{—} \ ②$$

Then $X[\omega] = F.T[x(t)] = FT[e^{-2t}u(t)] = \frac{1}{j\omega+2}$

$H(\omega) = F.T[h(t)] = F.T[e^{-3t}u(t)] = \frac{1}{j\omega+3}$

By substituting $X(\omega)$ & $H(\omega)$ in eq-② yields

$$Y(\omega) = \frac{1}{(j\omega+2)} \times \frac{1}{(j\omega+3)} \quad — ③$$

By Applying partial fraction on R.H.S of Eq.③ yields

$$Y(\omega) = \frac{A}{(j\omega+2)} + \frac{B}{(j\omega+3)}$$ Here $A = 1$ & $B = -1$

$$Y(\omega) = \frac{1}{(j\omega+2)} + \frac{(-1)}{(j\omega+3)} \quad — ④$$

By applying Inverse fourier transform to eq-④, yields

$$\mathcal{F}.T^{-1}[Y(\omega)] = \mathcal{F}^{-1}\left[\frac{1}{j\omega+2}\right] - \mathcal{F}^{-1}\left[\frac{1}{j\omega+3}\right]$$

$$y(t) = e^{-2t}u(t) - e^{-3t}u(t)$$

$$\boxed{y(t) = \left[e^{-2t} - e^{-3t}\right]u(t)}$$

⟹ we know that the fourier transform of $u(\omega) = \pi\delta(\omega) + \frac{1}{j\omega}$

So $$y(\omega) = \left[\pi\delta(\omega) + \frac{1}{j\omega}\right]\left[\frac{1}{1+j\frac{\omega}{\omega_c}}\right]$$

$$= \pi\delta(\omega)\frac{1}{\left(1+j\frac{\omega}{\omega_c}\right)} + \frac{1}{j\omega}\frac{1}{\left(1+j\frac{\omega}{\omega_c}\right)} \quad — ③$$

We know $X(\omega)\,\delta(\omega) = X(0)\,\delta(\omega)$ — ④

⟹ By Applying Eq-④ in eq-③ yields

$$y(\omega) = \pi\,\delta(\omega) + \frac{1}{(j\omega)}\frac{1}{\left(1+j\frac{\omega}{\omega_c}\right)} \quad — ⑤$$

By applying partial fraction for above equation

$$y(\omega) = \underbrace{\pi\,\delta(\omega) + \frac{1}{j\omega}}_{} - \frac{1}{\omega_c+j\omega} \quad — ⑥$$

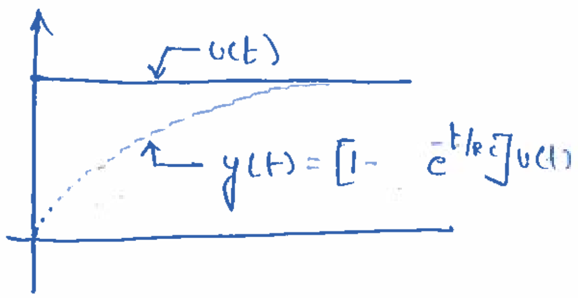
By Applying Inverse Fourier transform to above equation yields y(t)

where $\omega_c = \frac{1}{RC}$

$$y(t) = u(t) - e^{-t/RC}\,u(t)$$

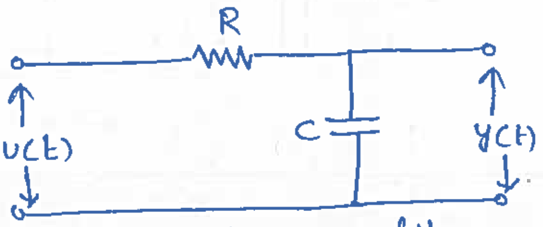
$$y(t) = \left[1 - e^{-t/RC}\right]u(t) \quad — ⑦$$

⟹ The output y(t) is increases exponentially After applying Input u(t)



⑫ Q) Relation Between Rise time and Bandwidth?

Sol



The Low pass filter is as shown in fig.

fig(a) Low pass filter

⟹ Here Input is unit step function $u(t) = \begin{cases} 1 & \text{for } t \geq 0 \\ 0 & \text{for } t < 0 \end{cases}$

⟹ The Transfer function of Low pass filter is

$$H(\omega) = \frac{1}{1 + j\omega RC} = \frac{1}{1 + j\frac{\omega}{\omega_c}} \quad \text{where } \omega_c = \frac{1}{RC}$$

Here ω_c is cutoff frequency

⟹

The Relation between Input u(t), Impulse Response h(t) and y(t) output is given by

$$\boxed{y(t) = u(t) * h(t)} \text{ — ①}$$ i.e Convolution of Two signals in time domain.

⟹ But we Know that Convolution of Two signals in time domain is equal to the product of Two Signals in frequency domain.

i.e $$\boxed{Y(\omega) = U(\omega)\, H(\omega)} \text{ — ②}$$

Rise Time: (t_r): ~~Risti~~ Rise time Indicated by t_r.

It is defined as time taken by output Response of a system $y(t)$, to reach 10% of final value to the 90% of final value.

⇒ From the definition of t_r, from the diagram of fig (b)

final value = 1

- So 10% of final value is 0.1
- and 90% of final value is 0.9.

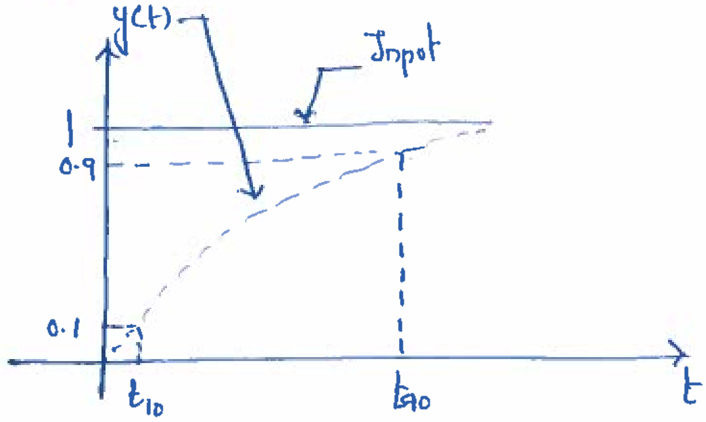


fig (c) Rise time $t_r = t_{90} - t_{10}$.

⇒ At $t = t_{10}$ $y(t_{10}) = \left[1 - e^{-\frac{t_{10}}{RC}}\right] u(t)$ —⑧

At $t = 0$ $y(t) = 0$
$t = \infty$ $y(\infty) = 1$

$$0.1 = 1 - e^{-t_{10}/RC}$$

$$e^{-t_{10}/RC} = 1 - 0.1$$

$$e^{-t_{10}/RC} = 0.9$$

By applying natural logarithm on Both sides for above equation

$$-\frac{t_{10}}{RC} = \ln(0.9)$$

$$-t = -RC(-0.105)$$

$$\boxed{t = RC(0.105)}$$

$\Rightarrow$ At $t = t_{90}$ $\quad y(t_{90}) = \left[1 - e^{-t_{90}/Rc}\right] u(t)$ —— ⑨

We know At $y(t_{90}) = 0.9$
$t = t_{90}$

$$\text{so} \quad 0.9 = 1 - e^{-t_{90}/Rc}$$

$$e^{-t_{90}/Rc} = 1 - 0.9$$

$$e^{-t_{90}/Rc} = 0.1$$

By applying natural logarithm

$$-\frac{t_{90}}{Rc} = -2.302$$

$$\boxed{t_{90} = 2.302\,Rc} \quad \text{—— ⑩}$$

Rise time: $t_r = t_{90} - t_{10} = Rc \times 2.302 - Rc \times 0.105$

$$= Rc\left[2.302 - 0.105\right]$$

$$\boxed{t_r = 2.197\,Rc} \quad \text{—— ⑪}$$

Here $\frac{1}{Rc}$ is cuttoff frequency of Low pass filter

$\boxed{\omega_c = \frac{1}{Rc}}$ which is Bandwidth of Low pass filter

$$\boxed{\omega_c = 2\pi f_{3dB} = \frac{1}{Rc}}$$

from equation - ⑪ $\quad \frac{t_r}{Rc} = 2.197$

$$t_r \cdot \omega_c = 2.197$$

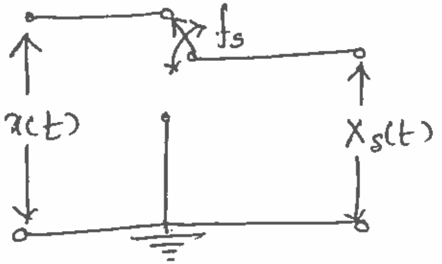
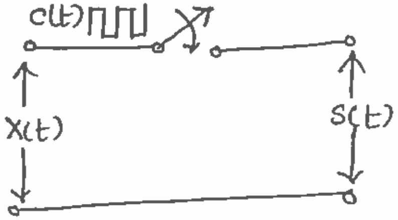
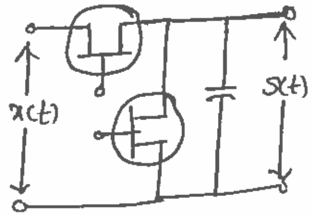
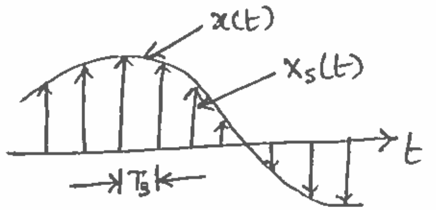
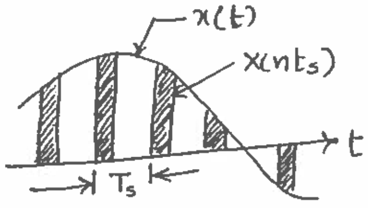
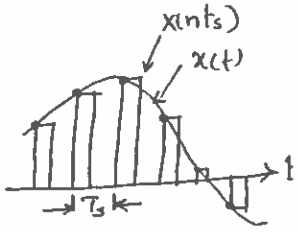
$$t_r \times 2\pi f_{3dB} = 2.197$$

$$t_r \times f_{3dB} = \frac{2.197}{2\pi}$$

$$\boxed{t_r \times f_{3dB} \approx 0.35}$$

13. Enumerate the difference between Impulse sampling, flat top Sampling and Natural Sampling ?

Sol

Sl. No	parameter of Comparison	Ideal (or) Instantaneous Sampling	Natural sampling	flat Top sampling
1	principle of Sampling	It uses multiplication by an Impulse function	It uses chopping principle	It uses Sample and hold Circuit.
2.	circuit (or) sampler			
3.	Waveforms			
4.	Realizability	This is not practically possible method	This method is used practically	This method is used practically.
5.	Sampling rate	Sampling rate tends to infinity	Sampling rate satisfies Nyquist Criteria	Sampling rate satisfies Nyquist Criteria

6.	Noise Interference	Noise Interference is maximum	Noise Interference is minimum	Noise Interference is maximum.
7.	Time domain representation	$X_\delta(t) = \sum_{n=-\infty}^{\infty} x(nT_s)\,\delta(t-nT_s)$	$s(t) = \frac{\tau A}{T_s} \sum_{n=-\infty}^{\infty} x(t)\,\mathrm{sinc}(nf_s t)\, e^{j2\pi n f_s t}$	$s(t) = \sum_{n=-\infty}^{\infty} x(nT_s)\,h(t-nT_s)$
8.	Frequency domain representation	$X_\delta(f) = f_s \sum_{n=-\infty}^{\infty} X(f-nf_s)$	$S(f) = \frac{\tau A}{T_s} \sum_{n=-\infty}^{\infty} \mathrm{sinc}(nf_s\tau)\,X(f-nf_s)$	$S(f) = f_s \sum_{n=-\infty}^{\infty} X(f-nf_s)H(f)$

⑭ . Determine the z-transform and Roc of

$x[n] = a^n u(n) - b^n u(-n-1)$ —— [12-Marks]

Sol: The given sequence is a two-sided infinite duration sequence

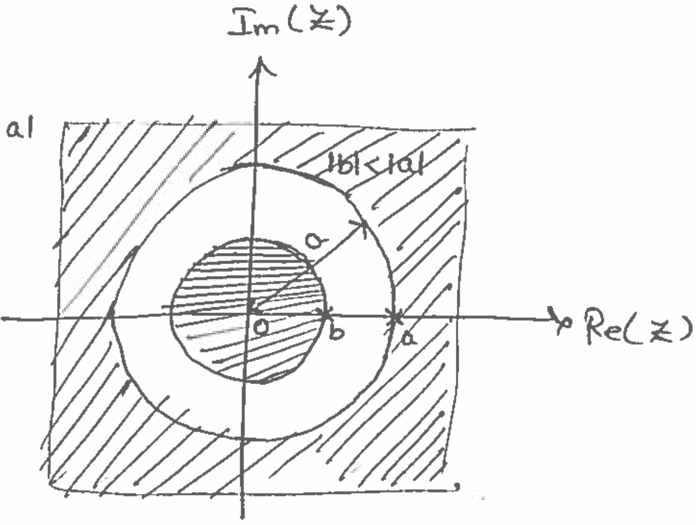
$$\therefore \quad Z[x(n)] = X(z) = \sum_{n=-\infty}^{\infty} \left[a^n u[n] - b^n u(-n-1)\right] z^{-n}$$

$$= \sum_{n=0}^{\infty} a^n z^{-n} - \sum_{n=-\infty}^{-1} b^n z^{-n}$$

$$= \sum_{n=0}^{\infty} (a z^{-1})^n - \sum_{n=1}^{\infty} [b^{-1} z]^n$$

⟹ The first series converges if $|a z^{-1}| < 1$ (or) $|z| > |a|$ and the second series converges if $|b^{-1} z| < 1$ (or) $|z| < |b|$.

⟹ If $|b| < |a|$, the two ROCs do not overlap as shown in fig (a) and the condition $|z| > |a|$ and $|z| < |b|$ cannot be satisfied simultaneously, so the z-transform $X(z)$ does not exist.

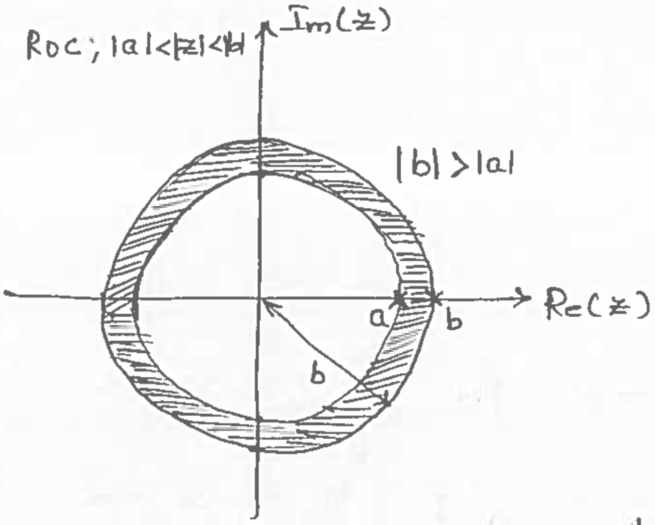


Roc of two sided sequence for (a) $|b| < |a|$

⇒ If $|b| > |a|$, the two Roc's overlap as shown in fig (b) and the conditions $|z| > |a|$ and $|z| < |b|$ can be satisfied simultaneously, so $X(z)$ exists. Therefore, the Roc of $X(z)$ is $|a| < |z| < |b|$.

This Implies that for a infinite duration two-sided signal, the Roc is a ring in the z-plane.

$$\therefore \quad X(z) = \frac{z}{z-a} + \frac{z}{z-b}$$



Roc of Two sequences for $|b| > |a|$

15) . obtain the Laplace transform of following signals

Sol : (i) Impulse function $\delta(t)$

Here $x(t) = \delta(t)$

we know that $\delta(t) = \begin{cases} 1 & \text{for } t=0 \\ 0 & \text{for } t \neq 0 \end{cases}$

$$\therefore \quad \mathcal{L}[\delta(t)] = X(s) = \int_0^\infty \delta(t)\, e^{-st}\, dt = e^{-st}\Big|_{t=0} = 1$$

(ii) unit step function $[x(t) = u(t)]$.

we know that $u(t) = \begin{cases} 1 & \text{for } t \geq 0 \\ 0 & \text{for } t < 0 \end{cases}$

$$L[u(t)] = \mathcal{L}[x(t)] = X(s) = \int_0^\infty u(t)\, e^{-st}\, dt$$

$$= \int_0^\infty 1\, e^{-st}\, dt = \frac{e^{-st}}{-s}\Bigg|_0^\infty$$

$$= -\frac{1}{s}\left[e^{-\infty} - e^{0}\right] = \frac{1}{s}.$$

(iii) $x(t) = A \sin \omega_0 t \; u(t)$

we know that $x(t) = A\left[\frac{e^{j\omega_0 t} - e^{-j\omega_0 t}}{2j}\right] u(t)$

$$\mathcal{L}[x(t)] = X(s) = \frac{A}{2j}\left\{\mathcal{L}\left[e^{j\omega_0 t} u(t)\right] - \mathcal{L}\left[e^{-j\omega_0 t} u(t)\right]\right.$$

$$X(s) = \frac{A}{2j}\left[\frac{1}{s-j\omega_0} - \frac{1}{s+j\omega_0}\right]$$

$$= \frac{A}{2j}\left[\frac{\cancel{s}+j\omega - \cancel{s}+j\omega_0}{(s-j\omega_0)(s+j\omega_0)}\right]$$

$$= \frac{A}{2\cancel{j}}\left[\frac{2\cancel{j}\omega}{s^2+\omega_0^2}\right]$$

$$X(s) = A\left(\frac{\omega_0}{s^2+\omega_0^2}\right)$$

3/1/2023

3/1/23

Semester End Regular/Supplementary Examination, Dec./Jan., 2022/2023

Degree	B. Tech. (U. G.)	**Program**	CSE, CSE (AI & ML) & CSE (DS)			**Academic Year**	2022 – 2023
Course Code	20CS303	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Database Management System**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is a Table? Give example.	20CS303.1	L1
2	What is a constraint?	20CS303.2	L1
3	List out all commands in DDL.	20CS303.3	L1
4	What is Redundancy? List-out any two problems caused by redundancy.	20CS303.4	L1
5	Define i) Lock and ii) Locking protocol.	20CS303.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	What is E-R Model? List and Explain the symbols used in E-R Diagrams with examples.	12M	20CS303.1	L2
	OR			
7	What is Data Abstraction? Illustrate and explain briefly about the 3-levels of Data abstraction.	12M	20CS303.1	L2
8	What is a Join? Explain i) Equi Join ii) Natural Join iii) Outer Join (left, right, full) with examples, in connection to SQL.	12M	20CS303.2	L2
	OR			
9	What is Relational Calculus? Explain i) Tuple Relational Calculus ii) Domain Relational Calculus with examples.	12M	20CS303.2	L2
10	Explain Nested queries and Aggregate functions with example	12M	20CS303.3	L2
	OR			
11 (a)	Apply database trigger for insertion and updating a records.	6M	20CS303.3	L3
11 (b)	What are null values? How DBMS deals with null values?	6M	20CS303.3	L2
12	What is BCNF? Explain BCNF with an example.	12M	20CS303.4	L2
	OR			
13	What is ISAM? Illustrate ISAM with an example.	12M	20CS303.4	L2
14	Explain Lock-based Concurrency Control.	12M	20CS303.5	L2
	OR			
15	Apply ARIES algorithm for system crash recovery.	12M	20CS303.5	L3

NSRIT

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Semester End Regular/Supplementary Examination, Dec./Jan., 2022/2023

Degree	B. Tech. (U. G.)	**Program**	CSE, CSE (AI & ML) & CSE (DS)			**Academic Year**	2022 – 2023
Course Code	20CS303	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Database Management System**						

PART –A SHORT ANSWER QUESTIONS 5X2=10 Marks

Questions (1 through 5)

1. **What is a table? Give example**

Ans: A table is an arrangement of information or data, typically in rows and columns, or possibly in a more complex structure.

For example, a table that contains employee data for a company might contain a row for each employee and columns representing employee information such as employee number, name, address, job title, and home telephone number.

Emp table

E.No	Emp-Nme	Designation	Dept	salary

2. **What is a constraint ?**

Ans : In DBMS, constraints are the set of rules that ensures that when an authorized user modifies the database they do not disturb the data consistency and the constraints are specified within the DDL commands like “alter” and “create” command.

Ex2

Constraints are used to limit the type of data that can go into a table. This ensures the accuracy and reliability of the data in the table. If there is any violation between the constraint and the data action, the action is aborted. Constraints can be column level or table level.

3. **List all commands in DDL**

Ans: **Data Definition Language (DDL) commands:**

- CREATE to create a new table or database.
- ALTER for alteration.
- Truncate to delete data from the table.
- DROP to drop a table.
- RENAME to rename a table.

4. **What is redundancy ? list out ant two problems caused by redundancy?**

Ans : Redundancy in DBMS is **having several copies of the same data in the database**. Redundancy in DBMS occurs when the database is not normalized.

Problems caused by Redundancy are

- insertion
- deletion
- updation

Redundancy can be avoided by normalizing the database, maintaining master data,

5. **Define i) lock ii) lock protocols**

ans: A lock is a **data variable which is associated with a data item**. This lock signifies that operations that can be performed on the data item. Locks in DBMS help synchronize access to the database items by concurrent transactions. All lock requests are made to the concurrency-control manager

Or

Database lock basically signifies the transaction about the current status of the data item i.e. whether that data is being used by other transactions or not at that point of time. Two types of Database lock are present: **Shared Lock. Exclusive Lock.**different locking protocols?

Types of Lock-Based Protocols. There are basically four lock based protocols in DBMS namely **Simplistic Lock Protocol, Pre-claiming Lock Protocol, Two-phase Locking Protocol, and Strict Two-Phase Locking Protocol**.

PART-B (LONG ANSWER QUESTIONS 5 x 12 = 60 Marks)

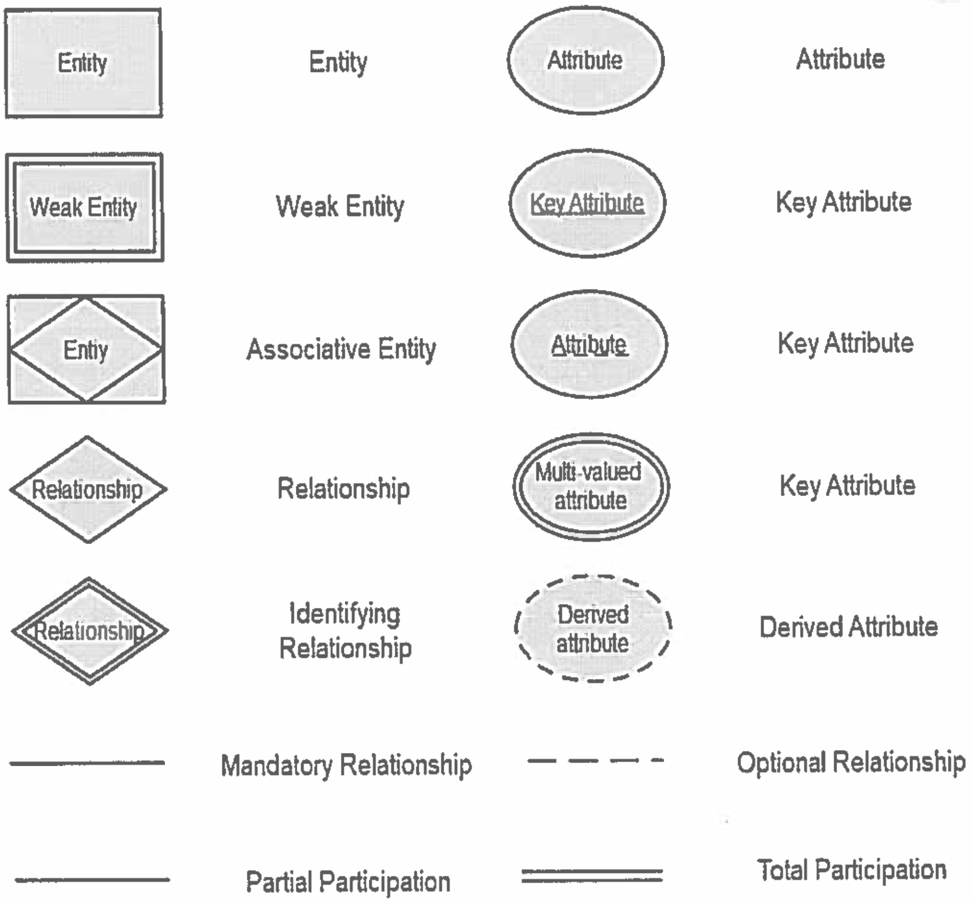
6. **What is E-R model? List and explain the symbols used in E-R diagram with examples- 12 M**

Ans :

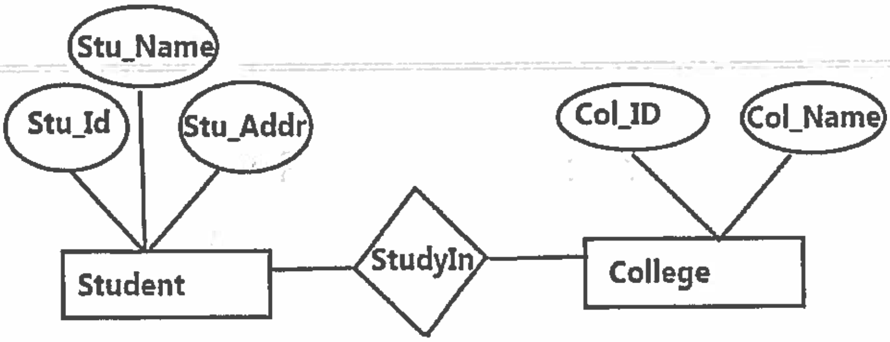
E-R model

An Entity-Relationship Model **represents the structure of the database with the help of a diagram.** ER Modelling is a systematic process to design a database as it would require you to analyze all data requirements before implementing your database

Symbols used in E-R diagram

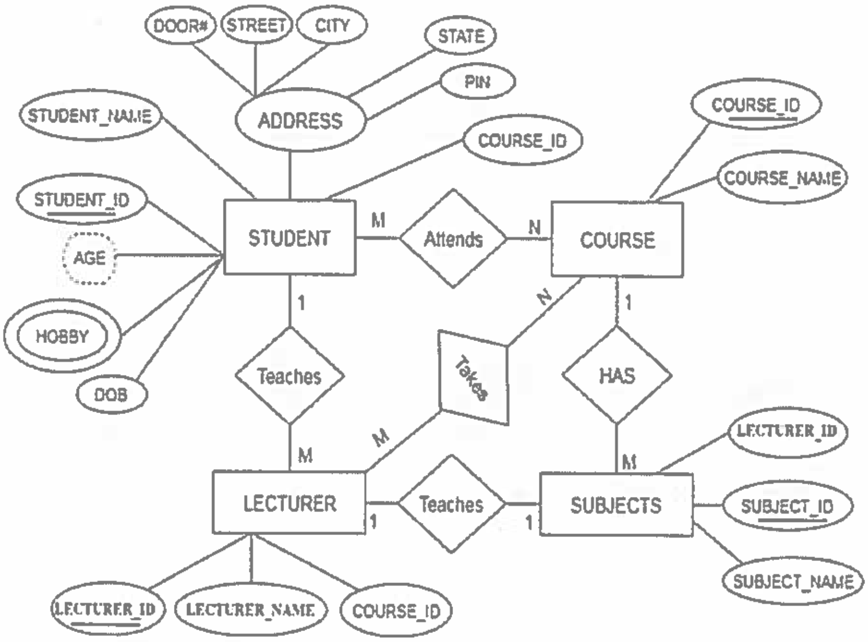


E-R MODEL examples (ANY ONE OF THE FOLLOWING)



Sample E-R Diagram

Ex2:



OR

7. What is data abstraction ? illustrate and explain 3-levels of data abstraction 12M

Ans : **Data Abstraction** is a process of hiding unwanted or irrelevant details from the end user. It provides a different view and helps in achieving data independence which is used to enhance the security of data.

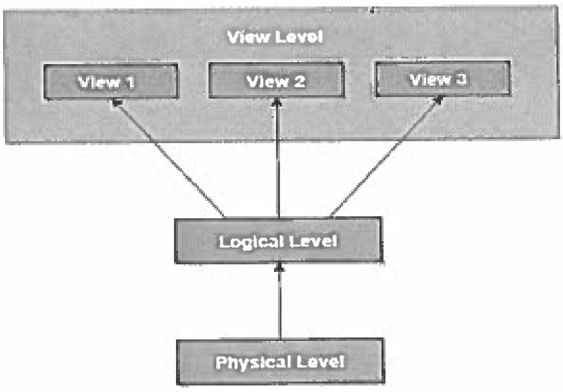
The database systems consist of complicated data structures and relations. For users to access the data easily, these complications are kept hidden, and only the relevant part of the database is made accessible to the users through data abstraction.

Levels of abstraction for DBMS

Database systems include complex data-structures. In terms of retrieval of data, reduce complexity in terms of usability of users and in order to make the system efficient, developers use levels of abstraction that hide irrelevant details from the users. Levels of abstraction simplify database design.

Mainly there are three levels of abstraction for DBMS, which are as follows –

- Physical or Internal Level
- Logical or Conceptual Level
- View or External Level



These levels are shown in the diagram

Let us discuss each level in detail.

Physical or Internal Level

It is the lowest level of abstraction for DBMS which defines how the data is actually stored, it defines data-structures to store data and access methods used by the database. Actually, it is decided by developers or database application programmers how to store the data in the database.

So, overall, the entire database is described in this level that is physical or internal level. It is a very complex level to understand. For example, customer's information is stored in tables and data is stored in the form of blocks of storage such as bytes, gigabytes etc.

Logical or Conceptual Level

Logical level is the intermediate level or next higher level. It describes what data is stored in the database and what relationship exists among those data. It tries to describe the entire or whole data because it describes what tables to be created and what are the links among those tables that are created.

It is less complex than the physical level. Logical level is used by developers or database administrators (DBA). So, overall, the logical level contains tables (fields and attributes) and relationships among table attributes.

View or External Level

It is the highest level. In view level, there are different levels of views and every view only defines a part of the entire data. It also simplifies interaction with the user and it provides many views or multiple views of the same database.

View level can be used by all users (all levels' users). This level is the least complex and easy to understand.

For example, a user can interact with a system using GUI that is view level and can enter details at GUI or screen and the user does not know how data is stored and what data is stored, this detail is hidden from the user.

8. What is a join ? explain 12 M

i) **Equi join**
ii) **Natural join**
iii) **Outer join(left, right, full) with example in connection to SQL**

Ans : **Join in DBMS** is a binary operation which allows you to combine join product and selection in one single statement. The goal of creating a join condition is that it helps you to combine the data from two or more DBMS tables. The tables in DBMS are associated using the primary key and foreign keys.

Types of Join

There are mainly two types of joins in DBMS:

1. Inner Joins: Theta, Natural, EQUI
2. Outer Join: Left, Right, Full

Let's see them in detail:

Inner Join:

Inner Join is used to return rows from both tables which satisfy the given condition. It is the most widely used join operation and can be considered as a default join-type

An Inner join or equijoin is a comparator-based join which uses equality comparisons in the join-predicate. However, if you use other comparison operators like ">" it can't be called equijoin.
Inner Join further divided into three subtypes:

- Theta join
- Natural join
- EQUI join

Theta Join

Theta Join allows you to merge two tables based on the condition represented by theta. Theta joins work for all comparison operators. It is denoted by symbol θ. The general case of JOIN operation is called a Theta join.
Syntax:

$A \bowtie_{\theta} B$
Theta join can use any conditions in the selection criteria.

Consider the following tables.

Table A			Table B	
column 1	column 2		column 1	column 2
1	1		1	1
1	2		1	3

For example:

$A \bowtie_{A.column\ 2 > B.column\ 2} (B)$

A ⋈ A.column 2 > B.column 2 (B)	
column 1	column 2
1	2

EQUI Join

EQUI Join is done when a Theta join uses only the equivalence condition. EQUI join is the most difficult operation to implement efficiently in an RDBMS, and one reason why RDBMS have essential performance problems.
For example:

$A \bowtie_{A.column\ 2 = B.column\ 2} (B)$

A ⋈ A.column 2 = B.column 2 (B)	
column 1	**column 2**
1	1

Natural Join (⋈)

Natural Join does not utilize any of the comparison operators. In this type of join, the attributes should have the same name and domain. In Natural Join, there should be at least one common attribute between two relations.
It performs selection forming equality on those attributes which appear in both relations and eliminates the duplicate attributes.

Outer join(left, right, full) with example in connection to SQL

Example:
Consider the following two tables

C	
Num	**Square**
2	4
3	9

D	
Num	**Cube**
2	8
3	18

C ⋈ D

C ⋈ D		
Num	**Square**	**Cube**
2	4	8
3	9	18

Outer Join

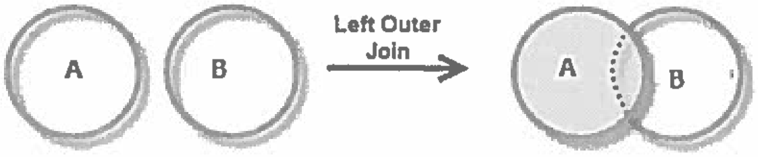
An **Outer Join** doesn't require each record in the two join tables to have a matching record. In this type of join, the table retains each record even if no other matching record exists.

Three types of Outer Joins are:

- Left Outer Join
- Right Outer Join
- Full Outer Join

Left Outer Join (A ⟕ B)

Left Outer Join returns all the rows from the table on the left even if no matching rows have been found in the table on the right. When no matching record is found in the table on the right, NULL is returned.



All rows from Left Table.

Consider the following 2 Tables

A	
Num	**Square**
2	4
3	9
4	16

B	
Num	**Cube**
2	8
3	18
5	75

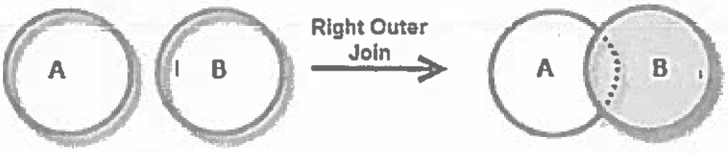
A ⟕ B

A ⋈ B		
Num	**Square**	**Cube**
2	4	8
3	9	18
4	16	–

Right Outer Join (A ⟖ B)

Right Outer Join returns all the columns from the table on the right even if no matching rows have been found in the table on the left. Where no matches have been found in the table on the left, NULL is returned. RIGHT outer JOIN is the opposite of LEFT JOIN

In our example, let's assume that you need to get the names of members and movies rented by them. Now we have a new member who has not rented any movie yet.



All rows from Right Table.

A ⟖ B

A ⋈ B		
Num	**Cube**	**Square**
2	8	4
3	18	9
5	75	–

Full Outer Join (A B)

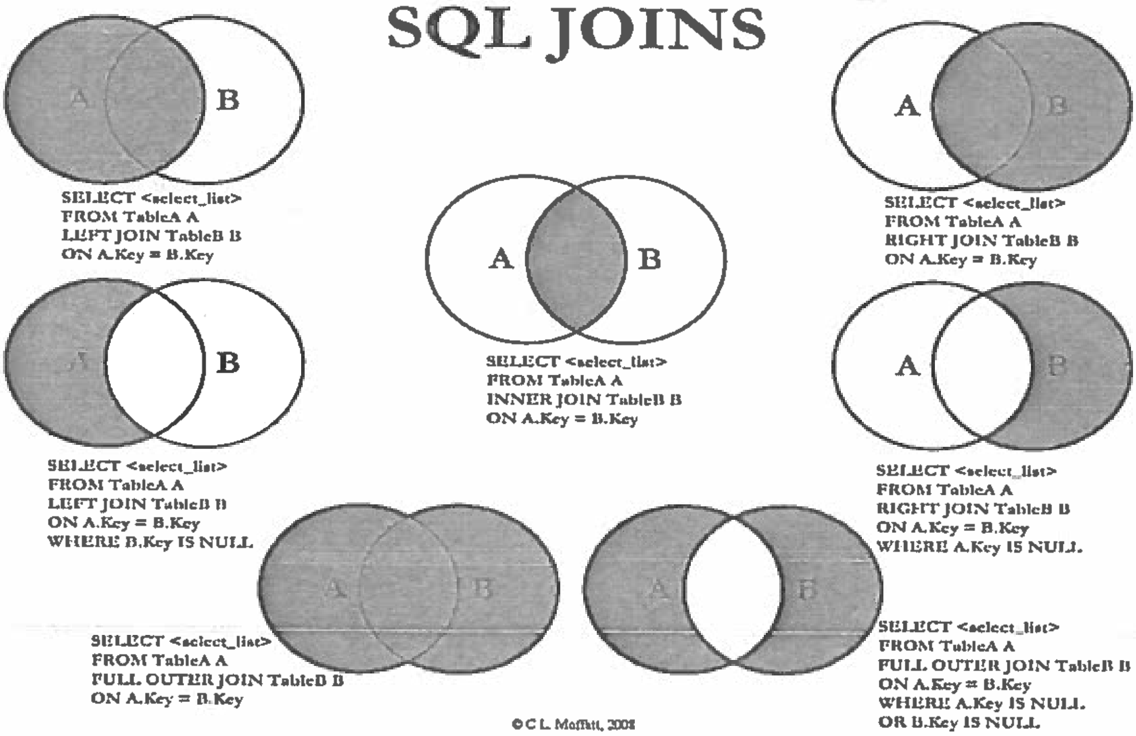
In a **Full Outer Join** , all tuples from both relations are included in the result, irrespective of the matching condition.
Example:

A ⟗ B

A ⋈ B		
Num	Square	Cube
2	4	8
3	9	18
4	16	–
5	–	75

Or

Outer join(left, right, full) with example in connection to SQL



```
SELECT <select_list>
FROM TableA A
LEFT JOIN TableB B
ON A.Key = B.Key
```

```
SELECT <select_list>
FROM TableA A
RIGHT JOIN TableB B
ON A.Key = B.Key
```

```
SELECT <select_list>
FROM TableA A
INNER JOIN TableB B
ON A.Key = B.Key
```

```
SELECT <select_list>
FROM TableA A
LEFT JOIN TableB B
ON A.Key = B.Key
WHERE B.Key IS NULL
```

```
SELECT <select_list>
FROM TableA A
RIGHT JOIN TableB B
ON A.Key = B.Key
WHERE A.Key IS NULL
```

```
SELECT <select_list>
FROM TableA A
FULL OUTER JOIN TableB B
ON A.Key = B.Key
```

```
SELECT <select_list>
FROM TableA A
FULL OUTER JOIN TableB B
ON A.Key = B.Key
WHERE A.Key IS NULL
OR B.Key IS NULL
```

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OR

9. What is Relational Calculus? Explain 12 M

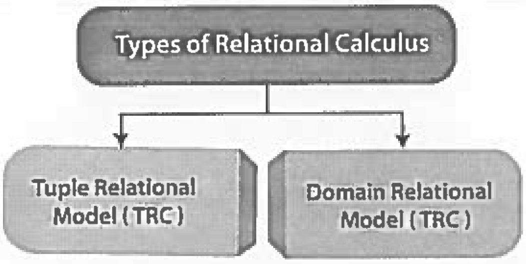
i) Tuple Relational Calculus
ii) Domain Relational Calculus with examples.

Ans :

Relational calculus is a non-procedural query language. In the non-procedural query language, the user is concerned with the details of how to obtain the end results. The relational calculus tells what to do but never explains how to do. Most commercial relational languages are based on aspects of relational calculus including SQL-QBE and QUEL.

Types of Relational calculus:

Relational Calculus



1. Tuple Relational Calculus (TRC)

It is a non-procedural query language which is based on finding a number of tuple variables also known as range variable for which predicate holds true. It describes the desired information without giving a specific procedure for obtaining that information. The tuple relational calculus is specified to select the tuples in a relation. In TRC, filtering variable uses the tuples of a relation. The result of the relation can have one or more tuples.

Notation:

A Query in the tuple relational calculus is expressed as following notation

1. {T | P (T)} or {T | Condition (T)}

Where

T is the resulting tuples

P(T) is the condition used to fetch T.

For example:

1. { T.name | Author(T) AND T.article = 'database' }

Output: This query selects the tuples from the AUTHOR relation. It returns a tuple with 'name' from Author who has written an article on 'database'.

TRC (tuple relation calculus) can be quantified. In TRC, we can use Existential (∃) and Universal Quantifiers (∀).

For example:

1. { R| ∃T ∈ Authors(T.article='database' AND R.name=T.name)}

Output: This query will yield the same result as the previous one.

2. Domain Relational Calculus (DRC)

The second form of relation is known as Domain relational calculus. In domain relational calculus, filtering variable uses the domain of attributes. Domain relational calculus uses the same operators as tuple calculus. It uses logical connectives ∧ (and), ∨ (or) and ¬ (not). It uses Existential (∃) and Universal Quantifiers (∀) to bind the variable. The QBE or Query by example is a query language related to domain relational calculus.

Notation:

1. { a1, a2, a3, ..., an | P (a1, a2, a3, ... ,an)}

Where

a1, a2 are attributes

P stands for formula built by inner attributes

For example:

1. {< article, page, subject > | ∈ javatpoint ∧ subject = 'database'}

Output: This query will yield the article, page, and subject from the relational javatpoint, where the subject is a database.

10. Explain Nested queries and Aggregate functions with example 12M

Ans :

A nested query is a query that has another query embedded within it. The embedded query is called a subquery.

A subquery typically appears within the WHERE clause of a query. It can sometimes appear in the FROM clause or HAVING clause.

Example

Let's learn about nested queries with the help of an example.

Find the names of employee who have regno=103

The query is as follows −

select E.ename from employee E where E.eid IN (select S.eid from salary S where S.regno=103);

Student table

Id	Name	classID	Marks
1	Pinky	3	2.4
2	Bob	3	1.44
3	Jam	1	3.24
4	Lucky	2	2.67
5	Ram	2	4.56

Teacher table

The teacher table is created as follows −

Id	Name	Subject	classID	Salary
1	Bhanu	Computer	3	5000
2	Rekha	Science	1	5000
3	Siri	Social	NULL	4500
4	Kittu	Maths	2	5500

Class table

The class table is created as follows −

Id	Grade	teacherID	No.ofstudents
1	8	2	20
2	9	3	40
3	10	1	38

Now let's work on nested queries

Example 1

```
Select AVG(noofstudents) from class where teacherID IN(
Select id from teacher
Where subject='science' OR subject='maths');
```

Output

You will get the following output –

```
20.0
```

Example 2

```
SELECT * FROM student
WHERE classID = (
  SELECT id
  FROM class
  WHERE noofstudents = (
    SELECT MAX(noofstudents)
    FROM class));
```

Output

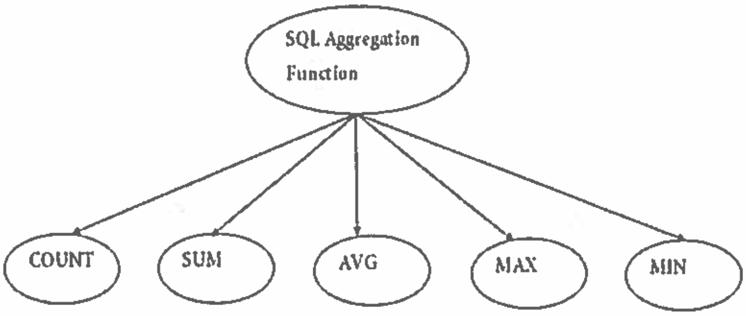
You will get the following output –

```
4|lucky |2|2.67
5|ram   |2|4.56
```

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



1. 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 )
```

Sample table:

PRODUCT_MAST

PRODUCT	COMPANY	QTY	RATE	COST
Item1	Com1	2	10	20
Item2	Com2	3	25	75
Item3	Com1	2	30	60
Item4	Com3	5	10	50
Item5	Com2	2	20	40
Item6	Cpm1	3	25	75
Item7	Com1	5	30	150
Item8	Com1	3	10	30
Item9	Com2	2	25	50
Item10	Com3	4	30	120

Example: COUNT()

```
SELECT COUNT(*)
FROM PRODUCT_MAST;
```

Output: 10

Example: COUNT with WHERE

```
SELECT COUNT(*)
FROM PRODUCT_MAST;
WHERE RATE>=20;
```

Output: 7

Example: COUNT() with DISTINCT

```
SELECT COUNT(DISTINCT COMPANY)
FROM PRODUCT_MAST;
```

Output: 3

Example: COUNT() with GROUP BY

```
SELECT COMPANY, COUNT(*)
FROM PRODUCT_MAST
GROUP BY COMPANY;
```

Output:

```
COM1 5
COM2   3
COM2  3
COM3  2
```

Example: COUNT() with HAVING

```
SELECT COMPANY, COUNT(*)
FROM PRODUCT_MAST
GROUP BY COMPANY
HAVING COUNT(*)>2;
```

Output:

```
COM1 5
COM2 3
```

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;
```

Output: 320

Example: SUM() with GROUP BY

```
SELECT SUM(COST)
FROM PRODUCT_MAST
WHERE QTY>3
```

```
GROUP BY COMPANY;
```

Output:

```
COM1 150
COM2 170
```

Example: SUM() with HAVING

```
SELECT COMPANY, SUM(COST)
FROM PRODUCT_MAST
GROUP BY COMPANY
HAVING SUM(COST)>=170;
```

Output:

```
COM1 335
COM3 170
```

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;
```

Output: 67.00

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;
```

O/P :30

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

```
1. MIN()
2. or
3. MIN( [ALL|DISTINCT] expression )
```

Example:

```
1. SELECT MIN(RATE)
2. FROM PRODUCT_MAST;
```

Output: 10

11 (a) **Apply database trigger for insertion and updating a records.** **6M**

Ans:

Triggers: Triggers are stored programs, which are automatically executed or fired when some event occurs.Trigger is invoked by Oracle engine automatically whenever a specified event occurs.Trigger is stored into database and invoked repeatedly, when specific condition match.

Syntax for creating trigger:

```
CREATE [OR REPLACE ] TRIGGER trigger_name
    {BEFORE | AFTER | INSTEAD OF }
{INSERT [OR] | UPDATE [OR] | DELETE}
[OF col_name]
    ON table_name
    [REFERENCING OLD AS o NEW AS n]
    [FOR EACH ROW]
    WHEN (condition)
    DECLARE
        Declaration-statements
    BEGIN
        Executable-statements
    EXCEPTION
        Exception-handling-statements
    END;

    /
```

PL/SQL Trigger Example

Let's take a simple example to demonstrate the trigger. In this example, we are using the following CUSTOMERS table:

Create 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

Create trigger:

Let's take a program to create a row level trigger for the CUSTOMERS table that would fire for INSERT or UPDATE or DELETE operations performed on the CUSTOMERS table.

This trigger will display the salary difference between the old values and new values:

```
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;
/
```

After the execution of the above code at SQL Prompt, it produces the following result.

```
Trigger created.
```

Updating a trigger

Check the salary difference by procedure:

Use the following code to get the old salary, new salary and salary difference after the trigger created.

```
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;
END;
/
```

Output:

```
Old salary: 20000
New salary: 25000
Salary difference: 5000
Old salary: 22000
New salary: 27000
Salary difference: 5000
Old salary: 24000
New salary: 29000
Salary difference: 5000
Old salary: 26000
New salary: 31000
Salary difference: 5000
Old salary: 28000
New salary: 33000
Salary difference: 5000
Old salary: 30000
New salary: 35000
Salary difference: 5000
6 customers updated
```

Note: As many times you executed this code, the old and new both salary is incremented by 5000 and hence the salary difference is always 5000.

After the execution of above code again, you will get the following result.

```
Old salary: 25000
New salary: 30000
Salary difference: 5000
Old salary: 27000
New salary: 32000
Salary difference: 5000
Old salary: 29000
New salary: 34000
Salary difference: 5000
Old salary: 31000
New salary: 36000
Salary difference: 5000
Old salary: 33000
New salary: 38000
Salary difference: 5000
Old salary: 35000
New salary: 40000
Salary difference: 5000
```

```
6 customers updated
```

OR

11 (b) what are null values? How DBMS deals with null values? **6M**

Ans : A field with a NULL value is a field with no value.

If a field in a table is optional, it is possible to insert a new record or update a record without adding a value to this field. Then, the field will be saved with a NULL value.

Note: A NULL value is different from a zero value or a field that contains spaces. A field with a NULL value is one that has been left blank during record creation!

It is not possible to test for NULL values with comparison operators, such as =, <, or <>.

We will have to use the IS NULL and IS NOT NULL operators instead.

IS NULL Syntax

SELECT *column_names*
FROM *table_name*
WHERE *column_name* IS NULL;

IS NOT NULL Syntax

SELECT *column_names*
FROM *table_name*
WHERE *column_name* IS NOT NULL;

Demo Database

Below is a selection from the "Customers" table in the Northwind sample database:

CustomerID	CustomerName	ContactName	Address	City	PostalCode	Country
1	Alfreds	Maria Anders	Obere Str. 57	Berlin	12209	Germany
2	Ana Trujillo	Ana Trujillo	Avda 222	México D.F.	05021	Mexico
3	Antonio Moreno	Antonio	Mataderos 2312	México D.F.	05023	Mexico
4	Around the Horn	Thomas	120 Hanover Sq.	London	WA1 1DP	UK
5	Berglunds snabbköp	Christina Berglund	Berguvsvägen 8	Luleå	S-958 22	Sweden

T

How DBMS deals with null values

The IS NULL Operator

The IS NULL operator is used to test for empty values (NULL values).

The following SQL lists all customers with a NULL value in the "Address" field:

Example

```
SELECT CustomerName, ContactName, Address
FROM Customers
WHERE Address IS NULL;
```

The IS NOT NULL Operator

The IS NOT NULL operator is used to test for non-empty values (NOT NULL values).

The following SQL lists all customers with a value in the "Address" field:

Example

```
SELECT CustomerName, ContactName, Address
FROM Customers
WHERE Address IS NOT NULL;
```

12 **What is BCNF? Explain BCNF with an example.** **12M**

Ans: BCNF (Boyce Codd Normal Form) is the advanced version of 3NF. A table is in BCNF if every functional dependency X->Y, X is the super key of the table. For BCNF, the table should be in 3NF, and for every FD. LHS is super key.

Example

Consider a relation R with attributes (student, subject, teacher).

Student	**Teacher**	**Subject**
Jhansi	P.Naresh	Database
jhansi	K.Das	C
subbu	P.Naresh	Database
subbu	R.Prasad	C

F: { (student, Teacher) -> subject

(student, subject) -> Teacher

Teacher -> subject}

Candidate keys are (student, teacher) and (student, subject).

The above relation is in 3NF [since there is no transitive dependency]. A relation R is in BCNF if for every non-trivial FD X->Y, X must be a key.

The above relation is not in BCNF, because in the FD (teacher->subject), teacher is not a key. This relation suffers with anomalies –

For example, if we try to delete the student Subbu, we will lose the information that R. Prasad teaches C. These difficulties are caused by the fact the teacher is determinant but not a candidate key.

Decomposition for BCNF

Teacher-> subject violates BCNF [since teacher is not a candidate key].

If X->Y violates BCNF then divide R into R1(X, Y) and R2(R-Y).

So R is divided into two relations R1(Teacher, subject) and R2(student, Teacher).

Ex1`

R1

Teacher	Subject
P.Naresh	database
K.DAS	C
R.Prasad	C

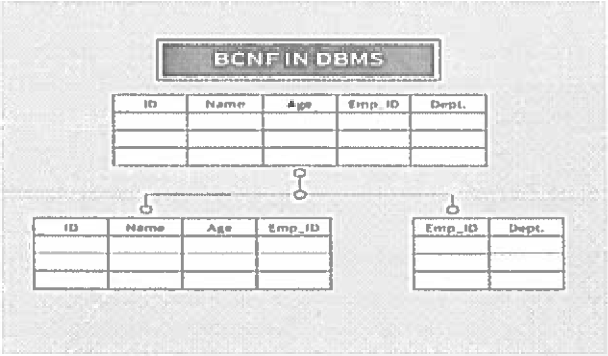
R2

Student	Teacher
Jhansi	P.Naresh
Jhansi	K.Das
Subbu	P.Naresh
Subbu	R.Prasad

All the anomalies which were present in R, now removed in the above two relations.

Or

Ex2



ID	Name	Age	Emp_ID	Dept.

ID	Name	Age	Emp_ID

Emp_ID	Dept.

Note

BCNF decomposition does not always satisfy dependency preserving property. After BCNF decomposition if dependency is not preserved then we have to decide whether we want to remain in BCNF or rollback to 3NF. This process of rollback is called denormalization.

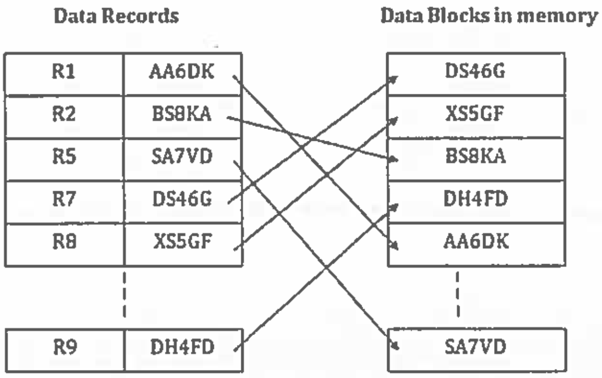
OR

13 **What is ISAM? Illustrate ISAM with an example.** **12M**

Ans:

Indexed sequential access method (ISAM)

ISAM method is an advanced sequential file organization. In this method, records are stored in the file using the primary key. An index value is generated for each primary key and mapped with the record. This index contains the address of the record in the file.



If any record has to be retrieved based on its index value, then the address of the data block is fetched and the record is retrieved from the memory.

Pros of ISAM:

- In this method, each record has the address of its data block, searching a record in a huge database is quick and easy.
- This method supports range retrieval and partial retrieval of records. Since the index is based on the primary key values, we can retrieve the data for the given range of value. In the same way, the partial value can also be easily searched, i.e., the student name starting with 'JA' can be easily searched.

Cons of ISAM

- This method requires extra space in the disk to store the index value.
- When the new records are inserted, then these files have to be reconstructed to maintain the sequence.
- When the record is deleted, then the space used by it needs to be released. Otherwise, the performance of the database will slow down.

14 Explain Lock-based Concurrency Control. 12M

Ans :

Lock-Based Protocol

In this type of protocol, any transaction cannot read or write data until it acquires an appropriate lock on it. There are two types of lock:

1. Shared lock:

- It is also known as a Read-only lock. In a shared lock, the data item can only read by the transaction.
- It can be shared between the transactions because when the transaction holds a lock, then it can't update the data on the data item.

2. Exclusive lock:

- In the exclusive lock, the data item can be both reads as well as written by the transaction.
- This lock is exclusive, and in this lock, multiple transactions do not modify the same data simultaneously.

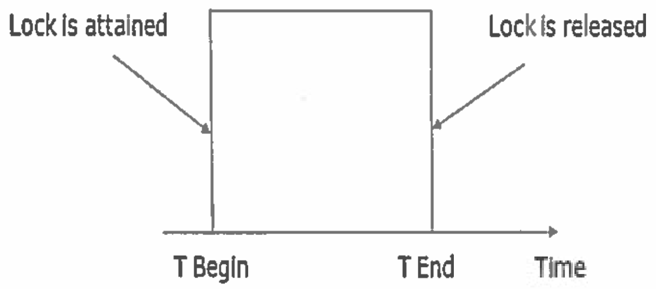
There are four types of lock protocols available:

1. Simplistic lock protocol

It is the simplest way of locking the data while transaction. Simplistic lock-based protocols allow all the transactions to get the lock on the data before insert or delete or update on it. It will unlock the data item after completing the transaction.

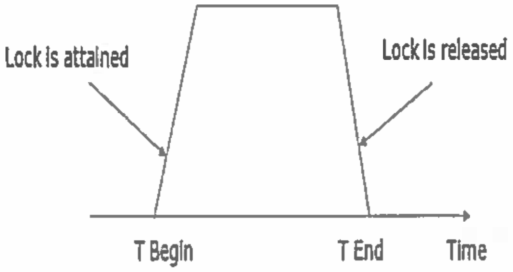
2. Pre-claiming Lock Protocol

- Pre-claiming Lock Protocols evaluate the transaction to list all the data items on which they need locks.
- Before initiating an execution of the transaction, it requests DBMS for all the lock on all those data items.
- If all the locks are granted then this protocol allows the transaction to begin. When the transaction is completed then it releases all the lock.
- If all the locks are not granted then this protocol allows the transaction to rolls back and waits until all the locks are granted.



3. Two-phase locking (2PL)

- The two-phase locking protocol divides the execution phase of the transaction into three parts.
- In the first part, when the execution of the transaction starts, it seeks permission for the lock it requires.
- In the second part, the transaction acquires all the locks. The third phase is started as soon as the transaction releases its first lock.
- In the third phase, the transaction cannot demand any new locks. It only releases the acquired locks.



There are two phases of 2PL:

Growing phase: In the growing phase, a new lock on the data item may be acquired by the transaction, but none can be released.

Shrinking phase: In the shrinking phase, existing lock held by the transaction may be released, but no new locks can be acquired.

In the below example, if lock conversion is allowed then the following phase can happen:

1. Upgrading of lock (from S(a) to X (a)) is allowed in growing phase.
2. Downgrading of lock (from X(a) to S(a)) must be done in shrinking phase.

Example:

	T1	T2
0	LOCK-S(A)	
1		LOCK-S(A)
2	LOCK-X(B)	
3	—	—
4	UNLOCK(A)	
5		LOCK-X(C)
6	UNLOCK(B)	
7		UNLOCK(A)
8		UNLOCK(C)
9	—	—

The following way shows how unlocking and locking work with 2-PL.

Transaction T1:

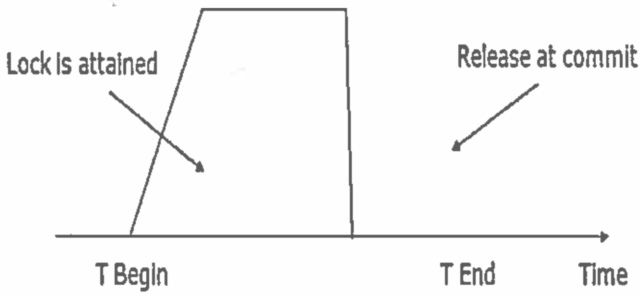
- **Growing phase:** from step 1-3
- **Shrinking phase:** from step 5-7
- **Lock point:** at 3

Transaction T2:

- **Growing phase:** from step 2-6
- **Shrinking phase:** from step 8-9
- **Lock point:** at 6

4. Strict Two-phase locking (Strict-2PL)

- The first phase of Strict-2PL is similar to 2PL. In the first phase, after acquiring all the locks, the transaction continues to execute normally.
- The only difference between 2PL and strict 2PL is that Strict-2PL does not release a lock after using it.
- Strict-2PL waits until the whole transaction to commit, and then it releases all the locks at a time.
- Strict-2PL protocol does not have shrinking phase of lock release.



It does not have cascading abort as 2PL does.

OR

15 Apply ARIES algorithm for system crash recovery. 12M

Ans :

ARIES starts by finding log records for operations that were not written to disk on crash and then replays all of them. This even includes transactions that need to rolled back. It brings the database to the same state as it was before crash. This process is called "repeating history". Note that this is same mechanism that is used for database replication in normal course of action. Once database is brought up to date, all the transactions that need to be rolled back are undone.

ARIES is a three phase algorithm:

1. Analysis phase: This phase reads the last checkpoint record in the log to figure out active transactions and dirty pages at point of crash/restart. A page is considered dirty if it was modified in memory but was not written to disk. This information is used by next two phases.

2. REDO phase: In this phase operations in log are reapplied to bring the state of the database to current.

3. UNDO phase: This phase proceeds from the end of log reverting back operations for uncommitted transactions. This has impact of rolling them back. Note that this is same procedure that database performs when rolling back a transaction in normal mode of operation for STEAL policy.

ARIES maintains two data structures and adds one more field to log record:

1. Transaction table: It contains all the transactions that are active at any point of time (i.e. are started but not committed/aborted). The table also stores the LSN of last log record written by the transaction in "lastLSN" field.

2. Dirty page table: Contains an entry for each page that has been modified but not written to disk. The table also stores the LSN of the first log record that made the associated page dirty in a field called "recoveryLSN" (also called "firstLSN"). This is the log record from which REDO need to restart for this page.

3. In addition log records are also updated to contain a field called "prevLSN" which points to previous log record for the same transaction. This creates a linked list of all log records for a transaction. When a new log record is created, "lastLSN" from transaction table is filled into its "prevLSN" field. And the LSN of current log record becomes the "lastLSN" in transaction table. Here is updated log record table with prevLSN filled in:

LSN	Prev LSN	Transaction ID	Type	Page ID
1	NIL	T1	UPDATE	P3
2	NIL	T2	UPDATE	P2
3	1	T1	COMMIT	
4	CHECKPOINT			
5	NIL	T3	UPDATE	P1
6	2	T2	UPDATE	P3
7	6	T2	COMMIT	

During checkpointing, a checkpoint log record is created. This log record contains the content of both "Transaction table" and "Dirty page table". "Analysis" phase starts by reading last checkpoint log record to get the information about active transactions and dirty pages. Here is content of "Transaction table" and "Dirty page table" at the checkpoint stage in above table at LSN 4:

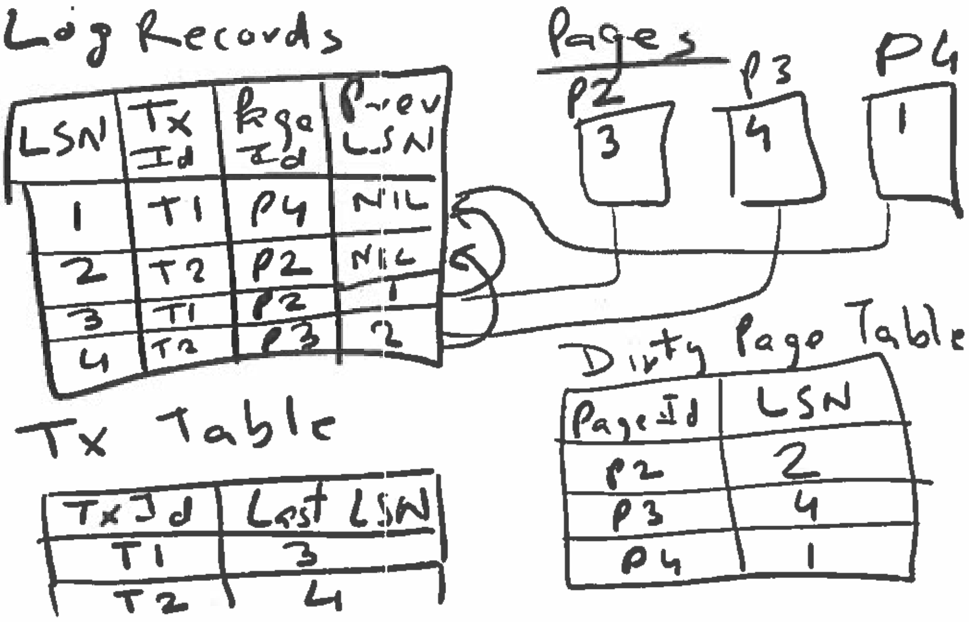
Transaction Table

Transaction ID	Last LSN	Status
T1	3	Commit
T2	2	In Progress

Dirty Page Table

Page ID	Recovery LSN
P3	1
P2	2

This whole setup can be visualized in following picture, pay attention to LSN for P2 in the "pages" list and in dirty table (dirty page table has the first LSN, whereas the P2 page has the last LSN):



Log Records

LSN	Tx Id	Page Id	Prev LSN
1	T1	P4	NIL
2	T2	P2	NIL
3	T1	P2	1
4	T2	P3	2

Tx Table

Tx Id	Last LSN
T1	3
T2	4

Dirty Page Table

Page Id	LSN
P2	2
P3	4
P4	1

ARIES Data Structures

Analysis Phase

Analysis phase reads the "checkpoint log record" for the latest checkpoint and then scans the log forward to create list of all active transactions and dirty pages. When scanning log forward:

1. If a new transaction is started, it is added to the transaction table.
2. If a commit/abort is seen for a transaction, its entry is removed from transaction table.
3. If a page is updated, a new entry is added to dirty page table and "recoveryLSN" is set to the LSN of the log record.

Hers is how the Transaction table and Dirty page table will look like after analysis is done:
Transaction Table (Rows with [x] are deleted, but are left here to show full process) Transaction ID Last LSN Status

[x] T1 3 *Commit*
[x] T2 7 *Commit*
T3 5 In ProgressDirty Page Table Page ID Recovery LSN

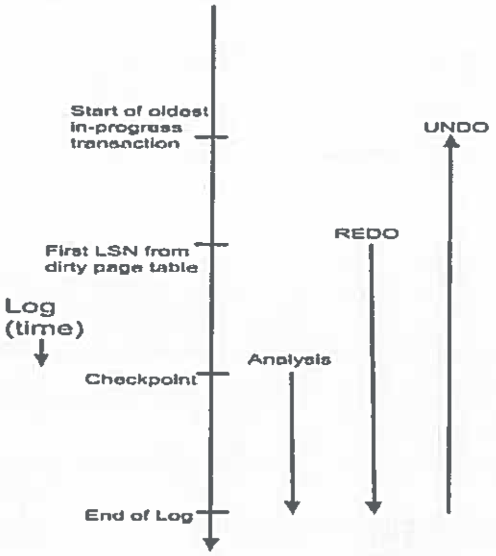
P1	5
P2	2
P3	6

REDO phase

At the end of analysis phase, the first LSN that caused a page to become dirty is known (its the lowest LSN value in dirty page table). REDO phase starts at this point in the log and processes all log entries from that point:

1. If the log record point to a page that is not dirty, ignore that log record.
2. If the log record point to a dirty page, but its "recoveryLSN" is more than current record's LSN, it means that the page was committed/aborted and a later transaction then updated the page again. So ignore this log record.
3. Otherwise check the LSN on the page (as discussed before, each page contains LSN that last updated the page). If the LSN of page is greater than this logs record's LSN, again ignore the record as this situation indicates that page made to disk successfully for a later transaction.
4. Otherwise, redo the action specified in the log record.

At the end of this phase database is in same state as it was before crash. Here is a timeline showing how three phases make a pass through the log.



UNDO phase

This phase starts from end of the log and goes backward. This phase undoes all transactions that were active at time of restart/crash. Information about these transactions are available in the transaction table. The lastLSN field in transaction table also tells the first record that need to be undone. And as we saw before, "prevLSN" field in log record creates a linked list of operations in the transaction. So all log records in this linked list need to be undone to rollback the transaction. Another good property of UNDO is that unlike REDO it doesn't need to check any state, it need to unconditionally undo log records of all active transactions.

To avoid repeating undo during repeated restart, each time a log record is undone, a "Compensation Log Record" is added to the log. In addition to the information about the undo, the record also contains the next log record for the transaction that need to be undone. And the next log record that needs to be undone is stored in the "prevLSN" of the current log being undone.

After that UNDO becomes a simple operation of going in reverse through the linked list of log records for an active transaction and undoing it. It need to do this for all active transactions in the transaction table.

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NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech. (U.G.)	**Program**	**Civil Engineering**			**Academic Year**	2022 - 2023
Course Code	20CE304	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Strength of Materials**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Write any 4 uses of mohr's circle.	20CE304.1	L1
2	Define the terms (i) Bending stress (ii) Shear stress	20CE304.2	L1
3	Calculate the maximum deflection of a simply supported beam carrying a point load of 100 KN at mid span. Span = 6 m, E= 20000 kN/m². For rectangular Cross section of the beam is 200X300 mm	20CE304.3	L2
4	Define (i) Slenderness ratio (ii) Radius of Gyration.	20CE304.4	L1
5	Distinguish between thick and thin cylinders.	20CE304.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain clearly the different types of stresses and strains.	6M	20CE304.1	L2
6 (b)	Explain in detail the behavior of mild-steel when subjected to tensile load with aid of stress strain diagram.	6M	20CE304.1	L2
	OR			
7	A steel rod of 20mm diameter passes centrally through a copper tube of 50mm external diameter and 40mm internal diameter. The tube is closed at each end by rigid plates of negligible thickness. The nuts are tightened lightly home on the projecting parts of the rod. If the temperature of the assembly is raised by 50°C, calculate the stress developed in copper and steel. Take E for steel and copper as 200 GN/m² and 100 GN/m² and α for steel and copper as 12 x 10^{-6} per °C and 18 x 10^{-6} per °C. Assume any necessary data	12M	20CE304.1	L3
8	Derive the sheer stress distribution for the following sections when it is subjected to shear force F a. Rectangular section, b X d b. Circular section, d diameter	12M	20CE304.2	L3
	OR			
9	Derive the bending equation with usual notations.	12M	20CE304.2	L2
10	A 3 m long cantilever of uniform rectangular cross–section 150 mm wide and 300 mm deep is loaded with a point load of 3 kN at the free end and a UDL of 2 kN/m over the entire length. Find the slope maximum deflection at the free end. E = 210 kN/mm². Use Macaulay's method. Assume any necessary data.	12M	20CE304.3	L2
	OR			
11	A simply supported beam of span 6 m is subjected to a udl of 2 kN/m over the entire span. Find the slope at support and maximum deflection at the mid section by method moment area method. EI is constant. Assume any necessary data.	12M	20CE304.3	L2

12 (a)	A simply supported beam of length **4 m** is subjected to uniformly distributed load of **30 kN/m** over the whole span and deflects **15 mm** at the centre. Determine the crippling loads when the beam is used as column when **one end fixed and other end hinged.**	6M	20CE304.4	L3
12 (b)	Derive an expression for crippling load when both ends of the columns are hinged.	6M	20CE304.4	L3
	OR			
13	A hollow cylindrical cast iron column is **4 m** long with both ends fixed. Determine the minimum diameter of the column if it has to carry a safe load of **250 kN** with a factor of safety. Take the internal diameter as **0.8** times the external diameter Take σ_c =550 N/mm^2 and $\alpha = \frac{1}{1600}$ in Rankine's formula.	12M	20CE304.4	L3
14	Determine the diameter of a solid shaft which will transmit **300 KN** at **250 rpm.** The maximum shear stress should not exceed **30 N/mm^2** and twist should not be more than **10** in a shaft length **2 m.** Take **modulus of rigidity = 1 x 105 N/mm^2.**	12M	20CE304.5	L3
	OR			
15 (a)	Derive an expression for thin cylinder subjected to internal fluid pressure P. determine the longitudinal stress and hoop stress.	6M	20CE304.5	L2
15 (b)	Explain various types of springs.	6M	20CE304.5	L2

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ANSWER KEY AND SCHEME OF EVALUATION

NSRIT

Course code:- 20CE304.

Strength of Materials

Dec/Jan, 2022-2023

III rd semester Regular

CIVIL ENGINEERING Department

Max marks = 70m

PART-A

1) write any four uses of Mohr's circle.

Ans: Mohr's circle is a graphical method of finding normal, tangential and resultant stress on an oblique plane. mohr's circle will be drawn for the following case.

→ A body subjected to two mutually perpendicular principal tensile stress of unequal intensities.

→ A body subjected to two mutually perpendicular principal stress which are unequal and unlike.

→ A body subjected to two mutually perpendicular principal tensile stress accompanied by simple shear stress.

2) Define the terms (i) Bending stress (ii) shear stress?

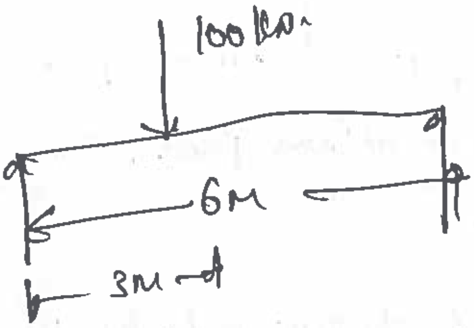
A:- Bending stress :- the beam bends or undergoes deformation the cross section of the beam will offer resistance to bending moment and these resistance are called bending stress

Shearstress: shear force produced by unbalanced vertical force develops in the beam.

- The variation of shear force in horizontal plane & vertical plane in the beam the difference in stress on both sides makes difference in resultant pull of push on the two sides of elementary length known as vertical shear.

3) Calculate the maximum deflection of a simply supported beam carrying point load of 100 kN at mid span = 6m, E = 20000 kN/m^2? for rectangular cross section of the beam is 200X300 mm.

Ans: Given data:-



length of Beam = 6M

point load (P) = 100 kN.

young's modulus (E) = 20000 kN/m^2

cross section of the beam = 200X300.

Deflection (Δ) or y = ?

Notes
Simply supported beam subjected to point load at centre,

given by deflection maximum (y) = $\frac{wl^3}{48EI}$

for slope, $\theta = \frac{wl^2}{16EI}$

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ANSWER KEY AND SCHEME OF EVALUATION

Maximum Deflection (δ) $\delta y = \frac{wl^3}{48EI}$

Deflection(y) $= \frac{100 \times 6^3 \times 4.5 \times 10^{-4}}{48 \times 20000 \times}$

$\Rightarrow$

Moment of Inertia (I) $= \frac{BD^3}{12}$

$= \frac{200 \times 300^3}{12} = \frac{0.2 \times 0.3^3}{12} = 4.5 \times 10^{-4} m^4$

maximum deflection $= \frac{100 \times 6^3}{48 \times 20000 \times 4.5 \times 10^{-4}} = 50 m$

$= 50,000 mm$

Re designate the beam as per cross section chages.

4) Define ① Slenderness ratio. ② Radius of gyration.

Ans: Slenderness : Slenderness of a column is defined as the ratio of column height to its least radius of gyration (l/k).

where l = equivalent length of column
k = least radius of gyration.

Radius of Gyration:- Radius of gyration is defined as the ratio of least second moment of Inertia to cross section Area of the column.

$$K = \sqrt{I/A}$$

where I = least moment of Inertia

A = cross section of beam.

5) Distinguish between thick and thin cylinders?

A:

thin cylinder:	thick cylinder
If the ratio of thickness to internal diameter of a cylinder is less than 1/20. the cylinder they known as thin cylinder $\frac{t}{d} < 1/20$ (or) thin cylinders.	→ If the ratio of thickness to internal diameter of a cylinder is greater than 1/20 the cylinder then is known as thick cylinder $\frac{t}{d} > 1/20$ (or) thick cylinder

③

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ANSWER KEY AND SCHEME OF EVALUATION

PART-B :

6). Explain clearly the different types of stresses and strains?

Types of stress : Normal stress: Normal stress is the stress which acts in a direction perpendicular to the area.

- The normal stress is further divided into tensile stress and compressive stress.

① Tensile stress.

② Compressive stress

③ Shear stress.

Types of strain :- ① Tensile strain

② Compressive strain

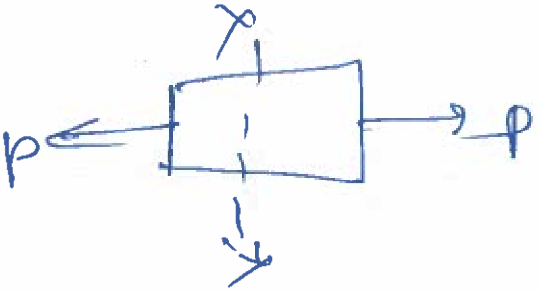
③ Volumetric strain

④ Shear strain.

Tensile stress: The stress induced in a body, when body subjected to two equal and opposite pull as shown as a result of which there is an increase in length known as tensile stress.

Tensile strain: The ratio of increase in length to original length is known as tensile strain.

Tensile stress:



Tensile stress $\sigma = \frac{\text{Resisting force}}{\text{Cross section area}} = \frac{\text{Tensile load}}{A}$ $[\because P = R]$.

Tensile strain: $\frac{\text{change in length}}{\text{original length}}$.

Compressive stress:– The stress induced in a body subjected to two equal and opposite push as shown in fig. as a result which there is decrease in length is known as compressive stress.

Compressive strain: The stress induced in a body subjected to two opposite, equal push. the corresponding strain is known as compressive strain.

$$\text{Compressive strain} = \frac{\text{Decrease in length}}{\text{original length}} = \frac{dL}{L}.$$

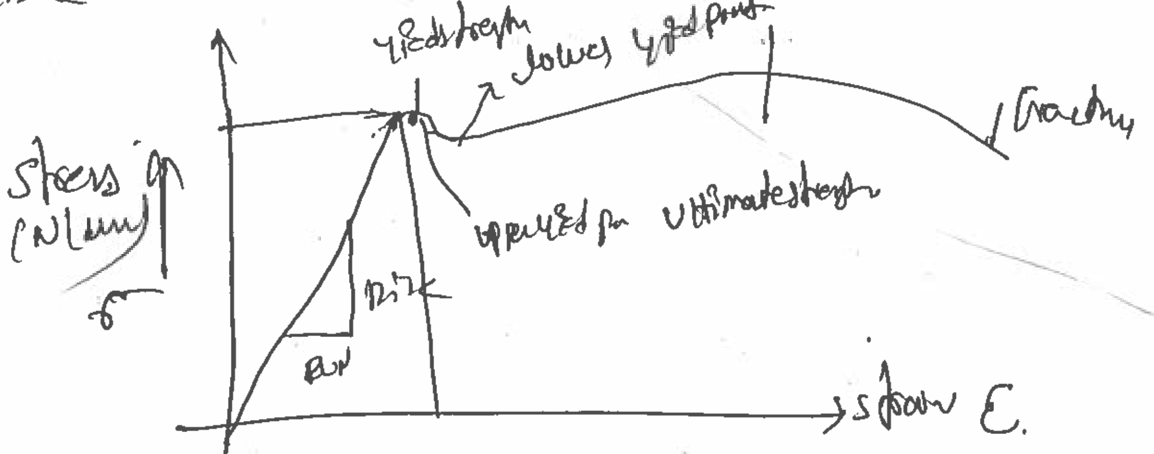
Shear stress: The stress induced in a body, when subjected to two equal and opposite force which are acting tangentially across the resisting section as shown in fig, as a result of which the body tends to shear off cross the section. the corresponding stress is known as shear stress.

$$\text{Shear stress} = \frac{\text{Shear resistance}}{\text{Shear Area}} = \frac{R}{A}.$$

$$\text{Shear strain } (\phi) = \frac{\text{Transverse displacement}}{\text{distance AD.}}$$

6b) Explain in details the behaviour of mild steel when subjected to Tensile load with aid of stress - strain diagram

A) :-



$$\text{Young's modulus} = \text{slope} = \frac{\text{Rise}}{\text{Run}}$$

Stress - Strain diagram for mild steel specimen of tensile test.

7), Given datas

Dia of steel rod $d = 20mm$.

Area of steel rod $A_s = \frac{\pi}{4} \times 20^2 = 100\pi \, mm^2$.

Area of copper tube, $A_c = \frac{\pi}{4} (50^2 - 40^2) \, mm^2$

$= 225 \pi \, mm^2$

Rise of temperature $A_c = \frac{\pi}{4} [50^2 - 40^2) \, mm^2$

$= 225 \pi \, mm^2$.

Rise of temperature, $T = 50^\circ C$

E for steel $E_s = 200 \, GN/m^2$

$= 200 \times 10^9 \, N/m^2$

$= 200 \times 10^3 \times 10^6 \, N/m^2$ [∵ $G = 10^9$, $10^6 \, N/m^2 = N/mm^2$]

$= 200 \times 10^3 \, N/mm^2$

E for copper $E_c = 100 \, GN/m^2 = 100 \times 10^9 \, N/m^2$

$= 100 \times 10^3 \times 10^6 \, N/m^2 = 100 \times 10^3 \, N/mm^2$

α for steel $\alpha_s = 12 \times 10^{-6}$ per $^\circ C$

$\alpha_c = 18 \times 10^{-6}$ Per $^\circ C$

As α for copper is more than that of steel, hence free expansion of copper will be more than that of steel when there is a rise in temperature

Let σ_s = Tensile stress in steel

σ_c = compressive stress in copper

For the equilibrium of the system,

Tensile load on steel = Compressive load on copper

$$\sigma_s A_s = \sigma_c \times A_c$$

$$\sigma_s = \frac{A_c}{A_s} \times \sigma_c$$

$$= \frac{225\pi}{100\pi} \times \sigma_c = 2.25\sigma_c$$

We know that the copper tube and steel rod will actually expand by the same amount.

Actual expansion of steel = Actual expansion of copper

But actual expansion of steel.

= Free expansion of steel + expansion due to tensile stress in steel

$$= \alpha_s T L + \frac{\sigma_s}{E_s} \cdot L.$$

Actual expansion of copper

= free expansion of copper — contraction due to compressive stress in copper

$$= \alpha_c T L - \frac{\sigma_c}{E_c} \cdot L$$

$$\alpha_s . T . L + \frac{\sigma_s}{E_s} L = \alpha_c T . L - \frac{\sigma_c}{E_c} L$$

or

$$\alpha_s . T + \frac{\sigma_s}{E_s} = \alpha_c T - \frac{\sigma_c}{E_c} \quad (\sigma_s = 2.25 \sigma_c)$$

(or)

$$12 \times 10^{-6} \times 50 + \frac{2.25 \sigma_c}{200 \times 10^3} = 18 \times 10^{-6} \times 50 - \frac{\sigma_c}{100 \times 10^3}$$

(or)

$$\frac{2.25 \sigma_c}{200 \times 10^3} + \frac{\sigma_c}{100 \times 10^3} = 18 \times 10^{-6} \times 50 - 12 \times 10^{-6} \times 50$$

$$1.125 \times 10^{-5} \sigma_c + 10^{-5} \sigma_c = 6 \times 10^{-6} \times 50$$

$$2.125 \times 10^{-5} \sigma_c = 30 \times 10^{-5}$$

$$2.125 \sigma_c = 30$$

$$\sigma_c = \frac{30}{2.125} = 14.117 \text{ N/mm}^2$$

substituting this value $\sigma_s = 14.117 \times 2.25$

$$= \underline{31.76 \text{ N/mm}^2}$$

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ANSWER KEY AND SCHEME OF EVALUATION

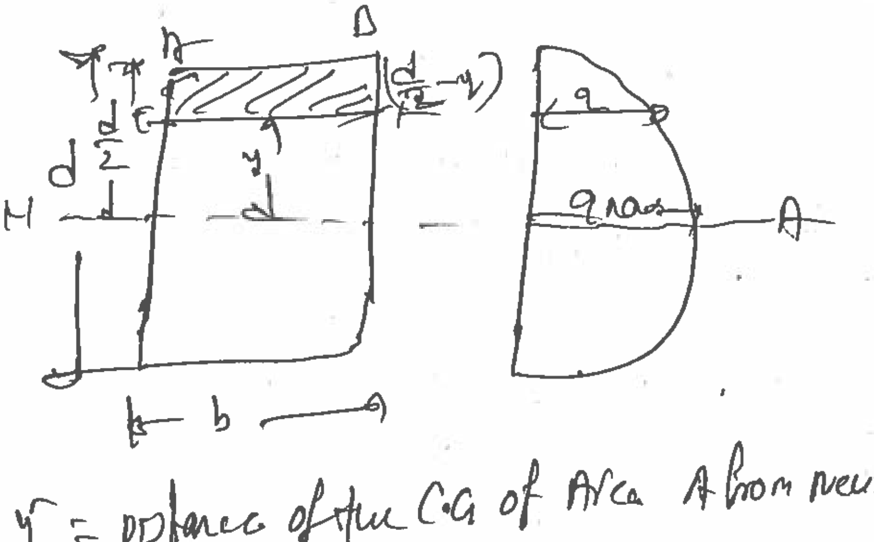
8) For Rectangular section:- A Rectangular section of beam of width b and depth d. Let F is the shear force acting at the section. Consider a level EF at a distance y from the neutral axis.

The shear stress at this level is given by equation

$$\tau = \frac{F \times A\bar{y}}{I b}$$

where A = Area of the section above y. (i.e shaded area ABFE)

$$= \left(\frac{d}{2} - y\right) \times b$$



$\bar{y}$ = Distance of the C.G of Area A from neutral axis

$$= y + \frac{1}{2}\left(\frac{d}{2} - y\right) = y + \frac{d}{4} - \frac{y}{2} = \frac{y}{2} + \frac{d}{4} = \frac{1}{2}\left(y + \frac{d}{2}\right)$$

b = Actual width of the the section at the level EF.

I = MOI of the whole section about N.A.

$$\tau = \frac{F\left(\frac{d}{2} - y\right) \times b \times \frac{1}{2}\left(y + \frac{d}{2}\right)}{b \times I}$$

$$= \frac{F}{2I}\left[\frac{d^2}{4} - y^2\right]$$

we see that τ increases as y decreases. the variation of τ with respect to y is parabola. the variation of shear stress across the section.

At the top edge $y = \frac{d}{2}$ and hence

$$\tau = \frac{F}{2I}\left[\frac{d^2}{4} - \left(\frac{d}{2}\right)^2\right] = \frac{F}{2I} \times 0 = 0.$$

At the neutral axis, $y = 0$ and hence

$$\tau = \frac{F}{2I}\left[\frac{d^2}{4} - 0\right] = \frac{F}{2I} \times \frac{d^2}{4}$$

$$= \frac{Fd^2}{8I} = \frac{Fd^2}{8 \times \frac{bd^3}{12}} = \frac{12}{8}\frac{F}{bd} = 1.5\frac{F}{bd}$$

Now average shear stress $\tau_{avg} = \frac{\text{shear force}}{\text{Area of section}} = \frac{F}{b \times d}.$

Substitute the above value in equation ① we get.

$$\boxed{\tau = 1.5 \times \tau_{avg}}$$

- The shear stress at the neutral axis where $y = 0$ this stress is also maximum shear stress.

$$\boxed{\tau_{Max} = 1.5\ \tau_{avg}}$$

$$\tau = \frac{A\bar{y}}{Ib}$$

$A\bar{y}$ = Moment of shaded area about N.A.

ANSWER KEY AND SCHEME OF EVALUATION

8) Consider a strip of thickness dy at a distance y from N.A. Let dA is the area of this strip.

$$dA = \text{Area of strip} = b \times dy.$$

Moment of the area about dA about N.A.

$$= dA \cdot y \text{ or } y \times dA$$

$$= y \times b\,dy \qquad [dA = b \times dy]$$

The moment of the shaded area about N.A is obtained by integrating the equation between the limits y to d/2.

$$\text{Moment of shaded area about N.A} = \int_y^{d/2} y \times b \times dy$$

$$= b\int_y^{d/2} y \times dy$$

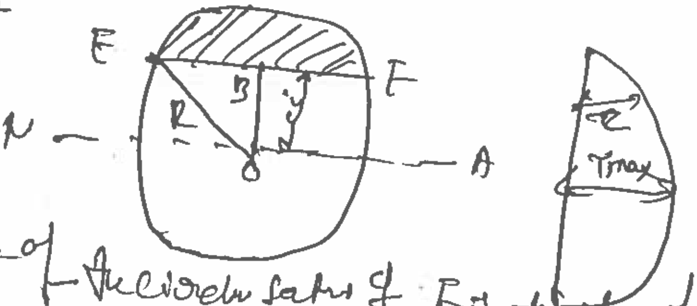
$$2b\left[\frac{y^2}{2}\right]_y^{d/2} = \frac{b}{2}\left[\left(\frac{d}{2}\right)^2 - y^2\right] = \frac{b}{2}\left(\frac{d^2}{4} - y^2\right)$$

But Moment of shaded area about N.A is equal to $A\bar{y}$.

$$A\bar{y} = \frac{b}{2}\left[\frac{d^2}{4} - y^2\right].$$

$$\tau = \frac{F \times \frac{b}{2}\left[\frac{d^2}{4} - y^2\right]}{I \times b} = \frac{F}{2I}\left[\frac{d^2}{4} - y^2\right]$$

8b) <u>Circular Section!</u>



From circular section of beam. Let R is radius of the section. Consider a level EF at a distance y from the N.A.
the circle section of FB further force acting on
the shear stress at this level is given by

$$\tau = \frac{F \times A \times \bar{y}}{I b}.$$

Consider a strip of thickness dy at a distance y from N.A. Let dA is the area of strip

$$dA = b \times dy = EF \times dy \quad [b = FF]$$

$$= 2 \times ED \times dy \quad [FF = 2 \times EB]$$

$$= 2 \times \sqrt{R^2 - y^2} \times dy.$$

$$EB = \sqrt{R^2 - y^2}$$

Moment of the area dA about N.A. = $y \times dA$

$$= y \times 2\sqrt{R^2 - y^2} \times dy \quad [\because dA = 2\sqrt{R^2 - y^2}\, dy]$$

$$= 2y \times \sqrt{R^2 - y^2}\, dy.$$

Moment of the above shaded area about the N.A is obtained by integrating the above equation between the limits y and R.

$$A\bar{y} = \int_y^R 2y \sqrt{R^2 - y^2}\, dy$$

$$= -\int_y^R (-2y) \sqrt{R^2 - y^2}\, dy.$$

Now $(-2y)$ is derivative of $R^2 - y^2$. Integral of the above equation.

$$A\bar{y} = -\left[\frac{[R^2-y^2]^{3/2}}{3/2}\right]_y^R$$

$$= -\frac{2}{3}\left[(R^2-R^2)^{3/2} - (R^2-y^2)^{3/2}\right]$$

$$= -\frac{2}{3}\left[0-(R^2-y^2)^{3/2}\right] = \frac{2}{3}(R^2-y^2)^{3/2}$$

$$q = \frac{F\times\frac{2}{3}(R^2-y^2)^{3/2}}{I\times b}$$

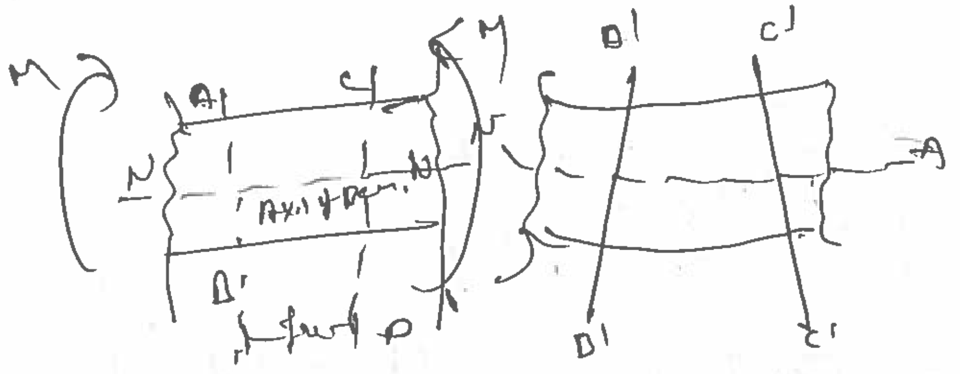
$$b = EF = 2\times EB = 2\times\sqrt{R^2-y^2}$$

$$q = \frac{\frac{2}{3}F(R^2-y^2)^{3/2}}{I\times 2\sqrt{R^2-y^2}} = \frac{F}{3I}(R^2-y^2)$$

$$\boxed{q_{max} = \frac{4}{3}\, q_{avg}}$$

9) Derive the bending equations with usual notations?

Ans: From theory of simple bending,



Before Bending After Bending

Assumptions:
- The material of the beam is homogeneous and Isotropic
- The value of Young's modulus is the same in tension and compression.
- The transverse section which were plane before bending, remain plane after the Bending.
- The beam is initially straight and all longitudinal filaments bent into circular arc with a common centre of curvature.
- The radius of curvature is large compared with dimension of the cross section
- Each layer of the beam is free to expand & contract, independently of the layer, above & below it.

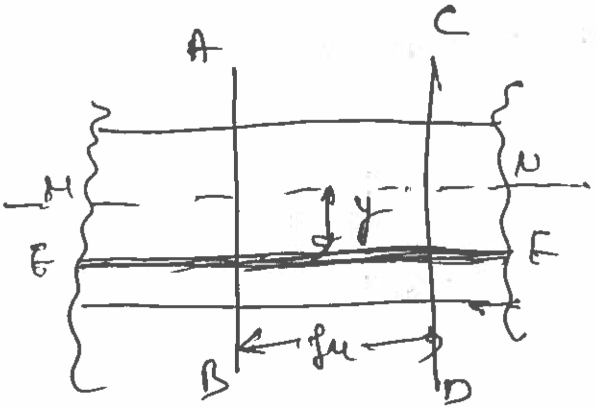
Theory of Simple Bending:

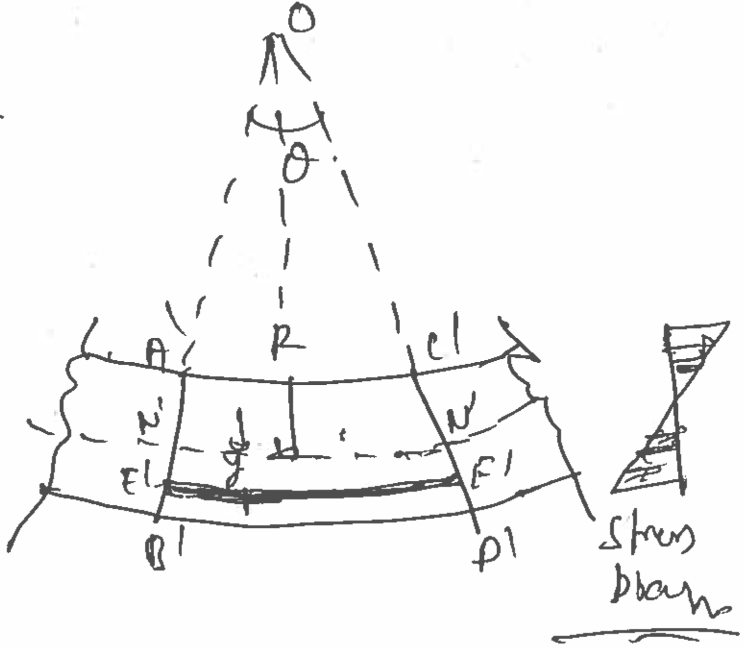
A beam is subjected to simple bending. Consider a small length δx of this part of beam. Consider two sections AB and CD which are normal to the axis of the beam N-N. Due to the action of bending moment, the part of length δx will be deformed as shown below.

→ The top layer such as AC has deformed to the shape A'C'. This layer has been shortened in its length. The bottom layer BD has deformed to the shape B'D'.

→ This layer has been elongated. or Contract at top & Bottom.

Expression for Bending stress:





From fig :-

Strain variation along the depth of Beam;

$$N'N' = R\times\theta$$

$$E'F' = (R+y)\times\theta$$

$$N'N' = NN = \delta x$$

$$\delta x = R\times\theta$$

Increase in the length of layer EF

$$= E'F' - EF = (R+y)\theta - R\times\theta$$

Strain in the layer EF

$$= \frac{\text{Increase in length}}{\text{original length}} = \frac{y\times\theta}{EF} = \frac{y\times\theta}{R\times\theta}$$

$$= y/R$$

R is constant. hence the strain in a layer is proportional to its distance from the neutral axis. it's linear.

Stress variation! let σ = stress in the layer EF

E = Young's modulus of the beam.

$$E = \frac{\text{Stress in the layer EF}}{\text{Strain in the layer EF}}$$

$$E = \frac{\sigma}{\left(\frac{y}{R}\right)}$$

$$\sigma = E\times\frac{y}{R} = \frac{E}{R}\times y.$$

$$\sigma = E \times \frac{y}{R} = \frac{E}{R} \times y.$$

Since E and R is constant. Stress in any layer is directly proportional to the distance of the layer from the neutral layer.

$$\boxed{\frac{\sigma}{y} = \frac{E}{R}}$$

Neutral axis and Moment of Resistance:

The stress at a distance y from the neutral axis $\sigma = \frac{E}{R} \times y$.

$$\boxed{\sigma = \frac{E}{R} \times y}$$

Consider a small layer at a distance y from the neutral axis.

Let dA = Area of the layer.

Now force on the layer = Stress on layer × Area on layers

$$= \sigma \times dA$$

$$= \frac{E}{R} \times y \times dA \quad \text{—} ①$$

total force on the beam section is obtained by Integrating the above equation.

total force on the beam section :

$$= \int \frac{E}{R} \times y \times dA$$

$$= \frac{E}{R} \int y \times dA$$

$\frac{E}{R} \int y \times dA = 0$ [due to No force on the section].

Moment of Resistance

Force on layer $= \frac{E}{R} \times y \times dA$

Moment of this force about N.A $=$ Force on layer $\times y$

$$= \frac{E}{R} y \times dA \times y$$

$$= \frac{E}{R} \times y^2 \times dA.$$

total moment of the forces on the section of the beam (moment of resistance)

$$= \int \frac{E}{R} \times y^2 \times dA.$$

$$= \frac{E}{R} \int y^2 \times dA.$$

Let $M =$ external moment applied on the beam section. for equilibrium the Moment of resist equal to external Bending Moment.

$$M = \frac{E}{R} \int y^2 \times dA.$$

$\left[\because \int y^2 dA = I\right)$

$$M = \frac{E \times I}{R}$$

$$\frac{M}{I} = \frac{E}{R}$$

$$\boxed{\frac{M}{I} = \frac{\sigma}{y} = \frac{E}{R}}$$

where M is expressed in N mm.

$I = mm^4$

$\sigma = N/mm^2$

$E = N/mm^2$

$R = mm$

⑪

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ANSWER KEY AND SCHEME OF EVALUATION

11) A simply supported beam of span 6m is subjected to a udl of 2kN/m over the entire span. Find the slope at support and maximum deflection at the mid section by moment area method EI is constant. Assume any necessary data.

Ans: ~~Given data~~ :

length of span = 6m

udl = 2kN/m

Max load = Wd L.

EI = constant throughout dew.

By moment area method!

By mohrs theorm I = $\frac{A}{EI}$ for slope

mohr theorem II = $\frac{A\bar{x}}{EI}$ for deflection.

B.m diagram fo given u.d.L.

Maximum bending moment at mid span $= \frac{wl^2}{8}$

$= \frac{2\times 6^2}{8}$

$= 9$ kN-m.

Area of B.M diagram $A_1 C_1 C_2$

$= \frac{2}{3}\times 3\times 9 = 18$ kN/m2.

First moment of area of B.m diagram about

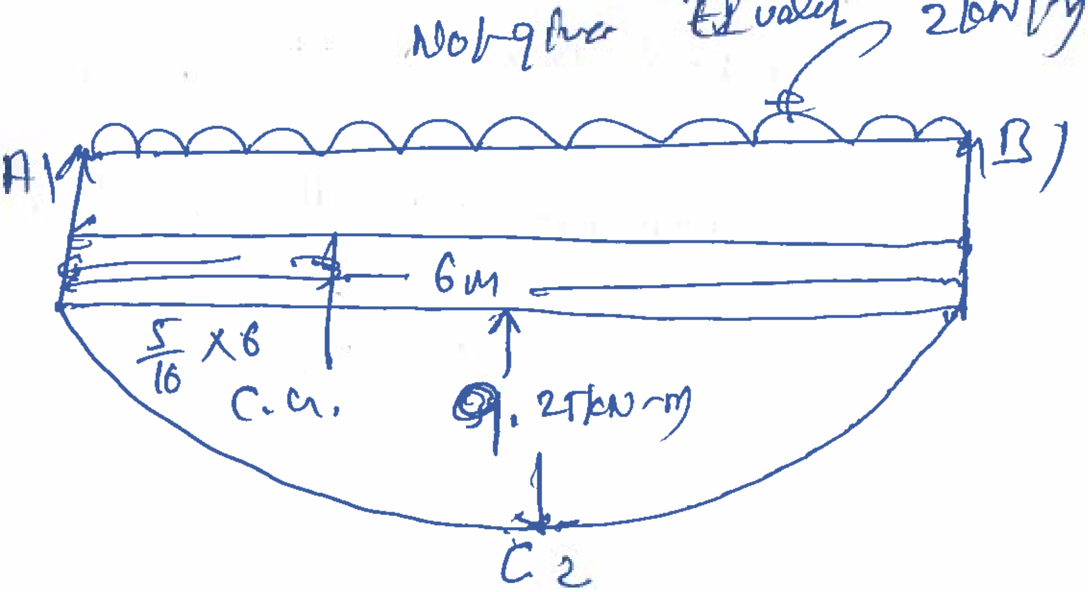
A_1 = Area x it's C.G from A_1

$A\bar{x} = 18\times\frac{5}{16}\times 6 = 33.75$ kN/m

Maximum slope at support A = $\frac{\text{Area}}{EI}$

$\theta_A = \frac{\cancel{\text{[illegible]}}\ 18}{EI}$ radian

Not given EI value



Maximum deflection at midspan C,

$$= \frac{\text{1st moment of Area } A_1C_1C_2 \text{ about } A_1}{EI}$$

$y_{max} = \frac{33.75}{EI}$ m downward

EI is constant

$\boxed{y_{max} = \frac{33.75}{EI}}$) by moment area method

10) A 3m long cantilever of uniform rectangular cross section 150mm wide and 300mm deep is loaded with a point load of 3kN at the free end and udl of 2kN/m over the entire length. Find the slope & maximum deflection at the free end. E = 210 kN/mm². Use Macaulay's method. Assume any necessary data.

Ans: Given data: Loading on cantilever as shown in fig.

Consider a section x–x at a distance x from fixed end.

A being the origin.

$$B.M \text{ at } X-X = -\left[30(3-x) + 2\frac{(3-x)^2}{2}\right]$$

i.e $$M = -[90 - 30x + 9 + x^2 - 6x]$$

$$M = -[x^2 - 36x + 99]$$

But $$EI\frac{d^2y}{dx^2} = -M$$

i.e $$EI\frac{d^2y}{dx^2} = -M$$

$$EI\frac{d^2y}{dx^2} = +[x^2 - 36x + 99]$$

Integrating the above equation successively we have.

$$EI\frac{dy}{dx} = \frac{x^3}{3} - \frac{36x^2}{2} + 99x + C_1$$

$$EI\,y = \frac{x^4}{12} - \frac{36x^3}{6} + \frac{99x^2}{2} + C_1x + C_2$$

where C_1 and C_2 are the constant of Integration

Apply the end conditions:

i.e $\frac{dy}{dx}$ = zero when x is zero.

and y = zero when x is zero.

we get $C_1 = 0$; $C_2 = 0$.

$$EI\frac{dy}{dx} = \frac{x^3}{3} - 18x^2 + 99x$$

and $EIy = \frac{x^4}{12} - 6x^3 + \frac{99x^2}{12}$

But $E = 210$ kN/mm^2

210×10^6 kN/m^2

I = moment of Inertia $= \frac{bd^3}{12}$

where $b = 150$ mm $= 0.15$ m $d = 300$ mm $= 0.30$ m

$$I = \frac{0.15 \times (0.30)^3}{12} = 3.375 \times 10^{-4} \text{ m}^4$$

a) maximum slope and deflection will occur at the free end for which $x = 3$

Substituting $x = 3$ m in the equation ①

$$210 \times 10^6 \times 3.375 \times 10^{-4} \frac{dy}{dx} = \frac{3^3}{3} - 18(3)^2 + 27(3)$$

(or)

$$\frac{dy}{dx} = \frac{144}{210 \times 3.375 \times 10^2}$$

maximum slope at the free end $= 0.00203$ radian

$= 7.21$ minutes

For maximum deflection

$$= 210 \times 10^6 \times 3.325 \times 10^{-4} \times y_{max} =$$

$$\frac{3^4}{12} - 6(3)^3 + \frac{29(3)^2}{2}$$

$$y_{max} = \frac{290.25}{210 \times 3.325 \times 10^2} = \underline{0.0040925 m}$$

Maximum Deflection = 0.0040925 m = $\underline{\underline{4.095 mm}}$

by macaulay's method.

(4)

12) a) <u>Given data:</u>

length $l = 4m = 4000mm$.

uniform distributed load (udl) = 30kN/m.

Deflection at the centre, $\delta = 15mm$.

For simply supported beam, carrying udl over the whole span. The deflection at the centre is given by,

$$\delta = \frac{5}{384} \times \frac{wl^4}{EI}$$

$$15 = \frac{5}{384} \times \frac{30 \times 4000^4}{EI}$$

$$EI = \frac{5}{384} \times \frac{30 \times 4000^4}{15}$$

$$= \frac{5}{384} \times \frac{3 \times 256}{15} \times 10^{13} = \frac{2}{3} \times 10^{13} \text{ Nmm}^2.$$

Crippling load when the beam is used as column with one end fixed and other end hinged.

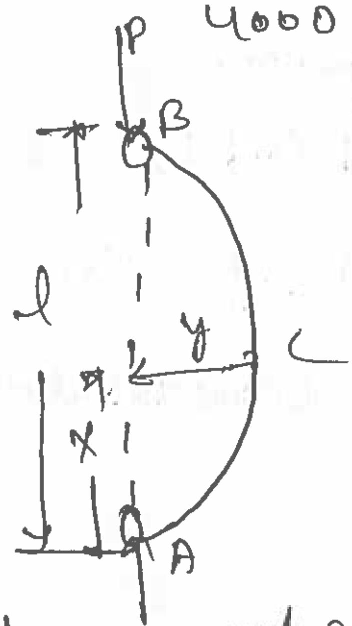
The crippling load P for this case interms of actual length is given by

$$P = \frac{2\pi^2 EI}{L_e^2}$$

when l = actual length = 4000mm

$$= \frac{2\times\pi^2\times\frac{1}{3}\times10^3}{4000^2} = 8224.5\,\text{kNs}$$

(2b)



The load at which the column just buckle (or bends) is called crippling load

→ consider a column AB of length l and uniform crosssectional area hinged at both of its ends A & B.

let deflected shape ACB

→ consider any section at a distance x from the end A.

let y = Deflection (lateral displacement) at the section.

The moment due to the crippling load at the section

$$= -Py$$

15

ANSWER KEY AND SCHEME OF EVALUATION

(-ve) sign is taken due to sign convention.

But Moment $= EI \frac{d^2y}{dx^2}$

Equating two moments, we have

$$EI \frac{d^2y}{dx^2} = -Py \text{ or } EI \frac{d^2y}{dx^2} + Py = 0$$

$$\frac{d^2y}{dx^2} + \frac{P}{EI} \cdot y = 0 \quad ;$$

The solution of the above differential equation.

$$y = C_1 \cos\left(x\sqrt{P/EI}\right) + C_2 \sin\left[x\sqrt{P/EI}\right]$$

C_1 & C_2 are constant of integration.

① At A, $x = 0$; $y = 0$

Substituting these values in equation ①.

$$0 = C_1 \cos 0 + C_2 \sin 0 \qquad [\cos 0 = 1 \text{ and } \sin 0 = 0]$$

$$= C_1 \times 1 + C_2 \times 0$$

$$= C_1$$

$$\boxed{C_1 = 0} \rightarrow (ii)$$

(ii) At B; $x = l$; $y=0$

Substituting these value in equation (i).

$$0 = C_1 \cos\left[l \times \sqrt{\frac{P}{EI}}\right] + C_2 \sin\left[l \times \sqrt{\frac{P}{EI}}\right]$$

$$= 0 + C_2 \sin\left[l \times \sqrt{P/EI}\right]$$

$$= C_2 \sin\left[l \sqrt{P/EI}\right] \rightarrow \text{(iv)}$$

From equation (iv), it is clear that eit $C_2 = 0$

$$\sin\left[l \sqrt{P/EI}\right] = 0$$

If $C_1 = 0$, then if C_2 is also equal to zero, then from equation (i) we will get $y = 0$.

$$\sin\left[l \sqrt{P/EI})\right] = 0$$

$$= \sin 0 \text{ a } \sin\pi \text{ or } \sin 2\pi \text{ or } \sin 3\pi$$

$$l \sqrt{P/EI} = 0 \text{ a } \pi \text{ a } 2\pi \text{ a } 3\pi \text{ a } 4\pi$$

Taking the least practical value

$$l \sqrt{P/EI} = \pi$$

$$\boxed{P = \frac{\pi^2 EI}{l^2}}$$

ANSWER KEY AND SCHEME OF EVALUATION

13) A hollow cylinder column length = l = 4m = 4000 mm.

End condition = both ends fixed.

effective length (l_e) = $\frac{l}{2} = \frac{4000}{2}$ = 2000mm

Safe load = 250 kN.

factor of safety = 5

let External dia = D

Internal dia = 0.8 × D

Crushing stress σ_c = 550 N/mm².

value of a = 1/1600 Rankine formulae

factor of safety = $\frac{\text{Crippling load}}{\text{Safe load}}$ or $5 = \frac{\text{Crippling load}}{250}$.

Crippling load $P = 5 \times 250 = 1250\text{kN} = 1250000$ N.

Area of column $A = \frac{\pi}{4}\left(D^2 - .64 D^2\right) = \frac{\pi}{4} \times 0.36 D^2$

$= \pi \times 0.09 D^2$

Moment of Inertia $I = \frac{\pi}{64}\left[D^4 - 0.8 D^4\right] = \frac{\pi}{64}\left[D^4 - 0.4096 D^4\right]$

$$\Rightarrow \frac{\pi}{64} \times 0.5904 \times D^4 = 0.009225 \times \pi \times D^4.$$

But $I = A k^2$ where k = radius of gyration

$$k = \sqrt{I/A} = \sqrt{\frac{0.009225 \times \pi D^4}{\pi \times 0.09 \times D^2}} = 0.32D$$

$$P = \frac{\sigma_c \times A}{1 + a\left(\frac{l_e}{k}\right)^2}$$

$$1250000 = \frac{550 \times \pi \times 0.09 D^2}{1 + \frac{1}{1600} \times \left[\frac{2000}{0.32D}\right]^2}$$

$$\frac{125000}{550 \times \pi \times 0.09} = \frac{D^2}{1 + \frac{24414}{D^2}} \text{ or } 8038 = \frac{D^2 \times D^2}{D^2 + 24414}$$

$$8038 D^2 + 8038 \times 24414 = D^4 \text{ or}$$

$$D^4 - 8038 D^2 - 8038 \times 24414 = 0$$

The above equation is quadratic.

$$D = \frac{8038 \pm \sqrt{646094 + 784958800}}{2}$$

$$D^2 = \frac{8038 \pm 29142}{2}$$

$$D = \sqrt{18592.5} = 136.3 \text{ m}$$

External diameter = 136.3 mm. Internal diameter = 0.8 × 136.3 = 109 mm

14) Given:

Power transmitted $P = 300\ kW = 300 \times 10^3 W$

speed of the shaft $N = 250 rpm$

maximum shear stress $q = 70 N/mm^2$.

Twist in shaft $\theta = 1° = \frac{\pi}{180} = 0.01745$ radians.

Length of shaft $L = 2M = 2000 m$.

Modulus of rigidity $C = 1 \times 10^5 N/mm^2$.

Let D = Diameter of the shaft.

power is given by the relation $P = \frac{2\pi NT}{60}$

$$300 \times 10^3 = \frac{2 \times \pi \times 250 \times T}{60}$$

$$T = \frac{300 \times 10^3 \times 60}{2\pi \times 250} = 11459.1 N\text{-}M \times 10^3 N\text{-}mm$$

(i) Diameter of the shaft when maximum shear stress.

$$q = 30 N/mm^2.$$

Maximum torque transmitted by solid shaft is given by relation:

$$T = \frac{\pi}{16} \times q \times D^3$$

$$11459100 = \frac{\pi}{16} \times 30 \times D^3$$

$$D = \frac{16 \times 11459100}{\pi \times 30} = \underline{124.5mm}$$

(ii) Diameter of shaft when twist should not be more than 1°.

$$\frac{T}{J} = \frac{C\theta}{L}.$$

$$J = \frac{\pi}{32} D^4$$

$$\frac{11459100}{\frac{\pi}{32} D^4} = \frac{10^5 \times 0.01745}{2000}$$

$$D^4 = \frac{32 \times 2000 \times 11459100}{10^5 \times \pi \times 0.01745} = 13377.81 \times 10^4.$$

$$D = (13372.81 \times 10^4)^{1/4} = \underline{107.5mm}$$

Diameter of the shaft = $\underline{124.5mm}$

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ANSWER KEY AND SCHEME OF EVALUATION

14) If diameter is taken smaller of the two values say 107.5mm.

$T = \frac{\pi}{16} \tau D^3$. the value of shear stress will be.

$$114591000 = \frac{\pi}{16} \times \tau \times (107.5)^3$$

$$11459100 = 243920 \tau$$

$$\tau = \frac{11459100}{243920} = 46.977 \text{ N/mm}^2$$

15) a) Expression for circumferential stress (Hoop stress).

Ans Consider a thin cylindrical vessel subjected to an internal fluid pressure. The circumferential stress will be setup in the material of the cylinder. If the bursting of the cylinder takes place as shown in fig.

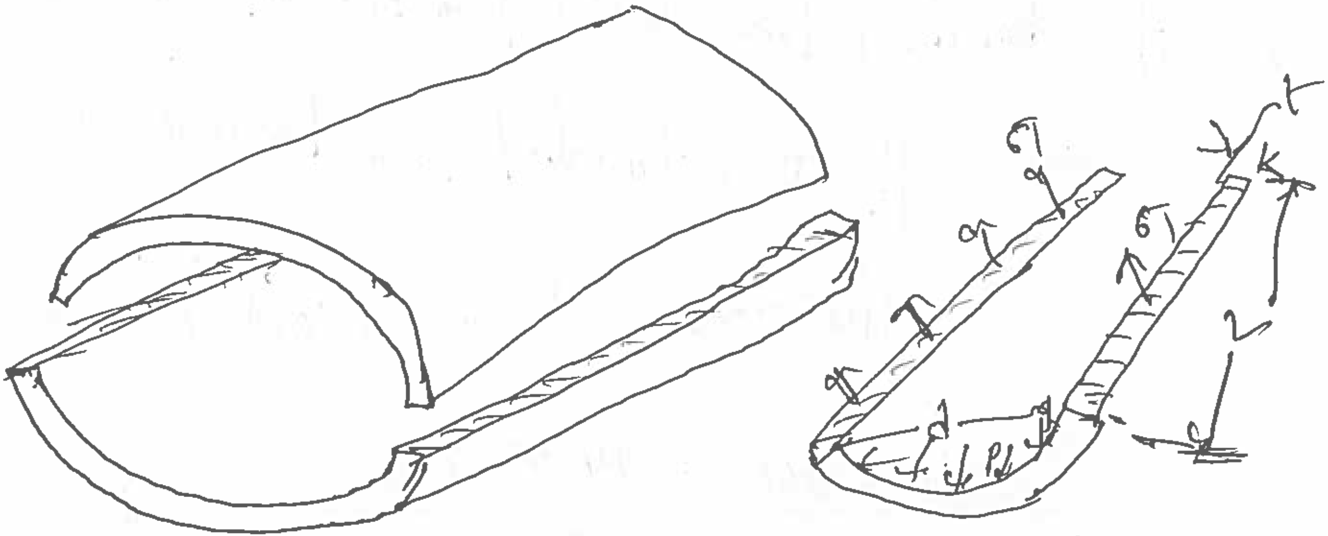
The expression for hoop stress or circumferential stress (σ_1) is obtained

Let p = Internal pressure of fluid.

d = Internal diameter of the cylinder

t = thickness of wall of the cylinder

σ_1 = circumferential or hoop stress in the material.



Expression, The bursting will take place if the force due to fluid pressure is more than the resistance force due to circumferential stress set up in the material. In the limiting case the two force should be equal

Force due to fluid pressure = $p \times$ Area on which p is acting

$= p \times (d \times L) \rightarrow$ (i)

($\because p$ is acting on projected area $d \times L$).

Force due to circumferential stress

$= \sigma_1 \times$ Area on which σ_1 is acting.

$= \sigma_1 \times (L \times t + L \times t)$.

$$= \sigma_1 \times 2Lt = 2\sigma_1 \times L \times t.$$

equating (i) and (ii) we get

$$p \times d \times L = 2\sigma_1 \times L \times t$$

$$\boxed{\sigma_1 = \frac{pd}{2t}}$$ (Corollary L)

Expression for longitudinal stress: Consider a thin cylindrical vessel subjected to internal fluid pressure, the longitudinal stress will be set up in the material of the cylinder. Let the bursting force of the cylinder takes place along the section AB of the

- The longitudinal stress (σ_2) developed in the material.

Let p = Internal pressure of fluid stored in the cylinder.

d = Internal diameter of cylinder.

t = thickness of cylinder

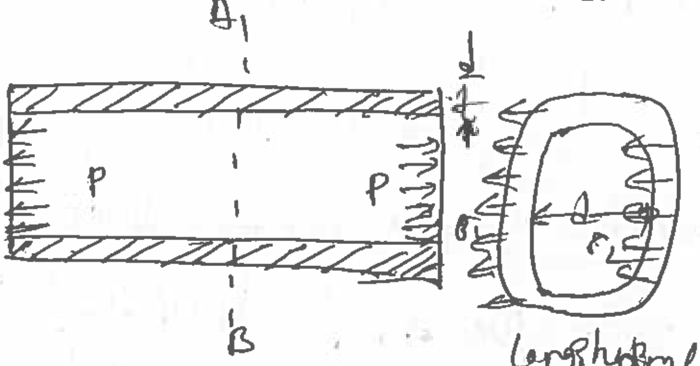
σ_2 = longitudinal stress in the material

→ The bursting will takes place if the force due to fluid pressure acting on the ends of the cylinder is more than the resisting force due to longitudinal stress σ_2 developed in the material as shown in fig.

Both force should be equal.

Force due to fluid pressure = P × Area on which P is acting

$$= P \times \frac{\pi}{4} \times d^2$$



Longitudinal stress (σ_2) develops

Resisting force = σ_2 × Area on which σ_2 is acting

$$= \sigma_2 \times \pi d \times t$$

Hence in the limiting case

Force due to fluid pressure = Resisting force

$$P \times \frac{\pi}{4} \times d^2 = \sigma_2 \times \pi d \times t$$

$$\sigma_2 = \frac{P \times \frac{\pi}{4} \times d^2}{\pi d \times t} = \frac{pd}{4t}$$

The stress σ_2 is also tensile.

$$\sigma_2 = \frac{pd}{2 \times 2t} = \frac{1}{2} \times \sigma_1$$

Longitudinal stress = Half of circumferential stress.

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ANSWER KEY AND SCHEME OF EVALUATION

15b) Explain various types of springs?

Ans) Springs are the elastic bodies which absorb energy due to resilience.

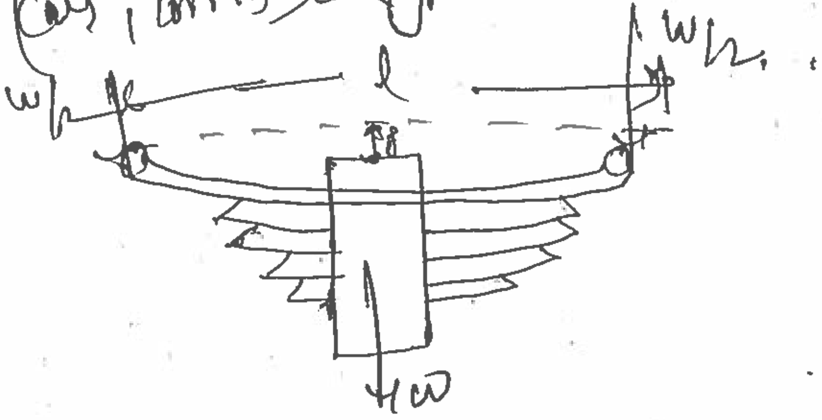
- The absorbed energy may be released as and when required. A spring which is capable of absorbing the greatest amount of energy for the given stress without getting permanently distorted is known as the best spring.

two Important types of springs:

① Laminated or leaf springs.

② Helical springs.

Laminated or leaf spring: The laminated springs are used to absorb shocks in railway wagons, coaches and road vehicles such as cars, lorries etc.



→ laminated spring which consists of a number of parallel strips of a metal having different length and same width, placed one over the other.

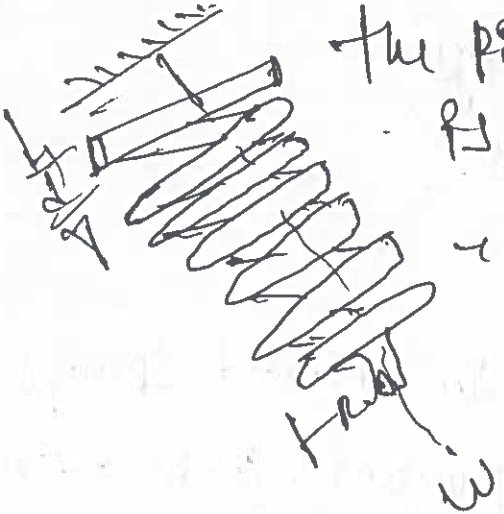
→ Initially all the plates are bent to the same radius and free to slide one over the other.

Helical Springs:- Helical springs are the thick spring wires coiled into a helix.

① closed coiled helical springs.

② open coiled helical springs.

Closed coiled helical springs:- closed - coiled helical spring are the springs in which helix angle is very small & in other words, the pitch between two adjacent turns is small.



→ A closed coiled helical spring carrying an axial load is shown in fig.

→ The helix angle in case of closed coiled helical springs are small. hence the bending effect on the spring is ignored.

→ we assume that the coils of closed coiled helical springs are stressed purely torsional stress

Course Coordinator

06/01/2023.

6/1/23.

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

Degree	B. Tech. (U.G.)	**Program**	Mechanical Engineering	**Academic Year**	2022 - 2023
Course Code	20ME304	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	III
Course	**Mechanics of Solids**				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define factor of safety.	20ME304.1	L1
2	Give the relation between S.F, B.M and rate of loading at a section of a beam.	20ME304.2	L1
3	State the formula section modulus for hollow rectangular section.	20ME304.3	L1
4	Why hollow circular shaft are preferred when compared to solid circular shaft?	20ME304.4	L1
5	Differentiate between thin cylinder and thick cylinder.	20ME304.5	L2

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw a neat sketch of stress- strain curve diagram of stainless steel and explain.	8M	20ME304.1	L2
6 (b)	The Young's modulus of a material is 210 kN/mm^2 and modulus of rigidity 75 kN/mm^2. Determine the bulk modulus.	4M	20ME304.1	L2
	OR			
7	Derive relations for normal and shear stresses acting on an inclined plane at a point in a stained material subjected to mutually perpendicular direct stresses.	12M	20ME304.1	L3
8	A cantilever beam 6 m long carries load of 30 kN, 70 kN, 40 kN and 60 kN at a distance of 1 m, 2 m, 3 m, 6 m respectively from the fixed end. Draw shear force and bending moment diagram.	12M	20ME304.2	L2
	OR			
9 (a)	A cantilever beam of length 4 m carries point loads of 1 KN, 2 KN and 3 KN at a 1, 2 and 4 m from the fixed end. Draw S.F.D and B.M.D	6M	20ME304.2	L2
9 (b)	A simply supported beam of length 8 m carries a point load of 6KN and 4 KN at a distance of 2 m and 4 m from the left end. Draw S.F.D and B.M.D	6M	20ME304.2	L2
10	State the assumptions and derive bending moment equation.	12M	20ME304.3	L2
	OR			
11	An I section with rectangular ends has the following dimensions: Flanges = 150 mm x 25 mm, Web = 300 mm x 10 mm, Total depth = 350 mm. Determine the maximum shearing stress developed in the beam for the shearing force of 25 kN.	12M	20ME304.3	L3
12	A steel cantilever 6m long carries two point loads 20 kN at the free end and 25 kN at a distance of 2 m from the free end. Find the slope and deflection at free end. Take $I = 1.3 \times 10^8$ mm^4 and $E = 2 \times 10^5$ N/mm^2.	12M	20ME304.4	L3
	OR			
13 (a)	Design a suitable diameter for a circular shaft required to transmit 80.2 kW at 200 rpm. The shear stress in the shaft is not to exceed 75 MN/m^2 and the maximum torque exceeds the mean by 40%. Also calculate the angle of twist in a length of 2 metres. Take C = 84 GN/m^2.	4M	20ME304.4	L3
13 (b)	Derive Torsional equation.	8M	20ME304.4	L3

14	A hollow cylindrical drum 750 mm in diameter and 2.6 m long has a shell thickness of 12 mm. If the drum is subjected to an internal pressure of 2.7 N/mm². Determine a. Change in diameter b. Change in length and c. Change in volume Take E =2.1 x 10^5 N/mm² and Poisson's ratio = 0.3.	12M	20ME304.5	L3
	OR			
15	Determine the ratio of strength of a solid steel column to that of a hollow column of internal diameter equal to ¾ of its external diameter. Both the columns have the same cross-sectional areas, length and end conditions.	12M	20ME304.5	L3

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ANSWER KEY AND SCHEME OF EVALUATION

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

1. Define Factor of Safety

It is the ratio of ultimate tensile stress to the working or permissible stress. It is shortly known as F.O.S

$$\text{Factor of safety} = \frac{\text{Ultimate Stress}}{\text{Working stress}}$$

2. Give the relation between S.F, B.M and rate of loading at a section of a beam

The rate of change of the bending moment with respect to x is equal to the shearing force, or the slope of the moment diagram at the given point is the shear at that point.

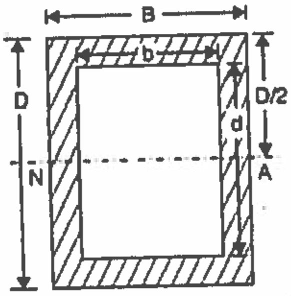
3. State the formula section modulus for hollow rectangular section.

Section modulus,

$$Z = \frac{I}{y_{\max}}$$

$$= \frac{\frac{1}{12}\left(BD^3 - bd^3\right)}{\frac{D}{2}}$$

$$= \frac{1}{6D}\left(BD^3 - bd^3\right)$$



4. Why hollow circular shaft are preferred when compared to solid circular shaft?

The stiffness of the hollow shaft is more than the solid shaft with the same weight. In the hollow shaft, the material at the centre is removed and spread at large radius.

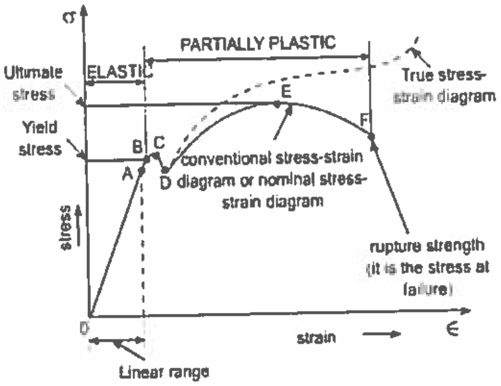
5. Differentiate between thin cylinder and thick cylinder.

Thin cylinder	Thick cylinder
The cylinder whose thickness is less than 1/10 to 1/20 of its diameter, that cylinder is called a thin cylinder.	The cylinder whose thickness is more than 1/20 of its diameter that cylinder is called a thick Cylinder.
The thin cylinder is only resisted by internal pressure.	The thick cylinder is resisted by internal as well as external pressure.
Low stress consuming capacity.	More stress consuming capacity.

Part B

(Long Answer Questions 5 x 12 = 60 Marks)

6. a. Draw a neat sketch of stress- strain curve diagram of stainless steel and explain.



Salient points of the graph:

(A) So it is evident from the graph that the strain is proportional to strain or elongation is proportional to the load giving a straight line relationship. This law of proportionality is valid upto a point A or we can say that point A is some ultimate point when the linear nature of the graph ceases or there is a deviation from the linear nature. This point is known as **the limit of proportionality or the proportionality limit.**

(B) For a short period beyond the point A, the material may still be elastic in the sense that the deformations are completely recovered when the load is removed. The limiting point B is termed as **Elastic Limit**

(C) and (D) - Beyond the elastic limit plastic deformation occurs and strains are not totally recoverable. There will be thus permanent deformation or permanent set when load is removed. These two points are termed as upper and lower yield points respectively. The stress at the yield point is called the yield strength.

(E) A further increase in the load will cause marked deformation in the whole volume of the metal. The maximum load which the specimen can with stand without failure is called the load at the ultimate strength. The highest point 'E' of the diagram corresponds to the ultimate strength of a material.

s_u = Stress which the specimen can with stand without failure & is known as Ultimate Strength or Tensile Strength. s_u is equal to load at E divided by the original cross-sectional area of the bar.

(F) Beyond point E, the bar begins to forms neck. The load falling from the maximum until fracture occurs at F.

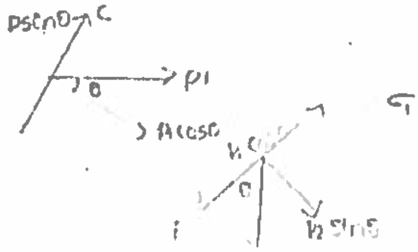
[Beyond point E, the cross-sectional area of the specimen begins to reduce rapidly over a relatively small length of bar and the bar is said to form a neck. This necking takes place whilst the load reduces, and fracture of the bar finally occurs at point F]

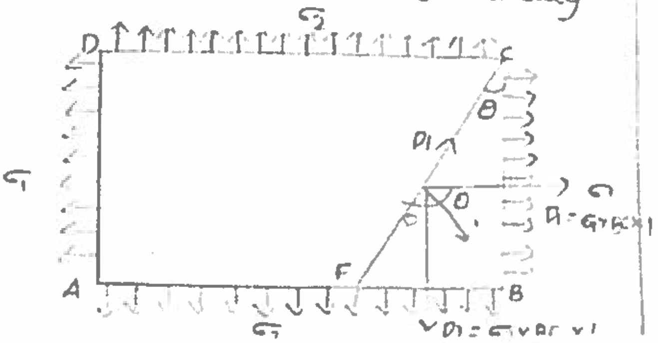
6. b. The Young's modulus of a material is 210 kN/mm2 and modulus of rigidity 75 kN/mm2. Determine the bulk modulus.

Given E= 210KN/mm^2
C= 75KN/mm^2
E=9KC/(3K+C)
210= 9*K*75/(3*K+75)
630k+15750 = 9*75*K
15750 = (675-630)K
K = 350KN/mm^2

7. Derive relations for normal and shear stresses acting on an inclined plane at a point in a stained material subjected to mutually perpendicular direct stresses.

A member subjected to like Direct stress into mutually perpendicular directions





σ_1 | σ_2 − max − AD, BC

σ_1 | σ_2 − min − CD, AB

P_1 − BC

P_2 − BF

load in BC = $\sigma_{BC} \times A = \sigma_1 \times BC \times 1$

$\boxed{P_1 = \sigma_1 \times BC}$ $\boxed{P_2 = \sigma_2 \times FB}$

$P_n = P_1 \cos\theta + P_2 \cos(90-\theta)$

$= P_1 \cos\theta + P_2 \sin\theta$

$P_t = P_1 \sin\theta - P_2 \sin(90-\theta)$

$= P_1 \sin\theta - P_2 \cos\theta$

$$\sigma_n = \frac{P_n}{A_{FC}} = \frac{P_1\cos\theta + P_2\sin\theta}{FC}$$

$$= \frac{\sigma_1 \times BC\cos\theta + \sigma_2 FB\sin\theta}{FC}$$

$$= [\sigma_1 \cos\theta] \frac{BC}{FC} + [\sigma_2 \sin\theta] \left(\frac{FB}{FC}\right)$$

Δ^{le} FBC, $\frac{BC}{FC} = \cos\theta$, $\frac{FB}{FC} = \sin\theta$

$$\boxed{\sigma_n = \sigma_1 \cos^2\theta + \sigma_2 \sin^2\theta}$$

$$\sigma_t = \frac{P_t}{A_{FC}} = \frac{P_1\sin\theta - P_2\cos\theta}{A_{FC}}$$

$$= \frac{\sigma_1 BC\sin\theta - \sigma_2 FB\cos\theta}{FC}$$

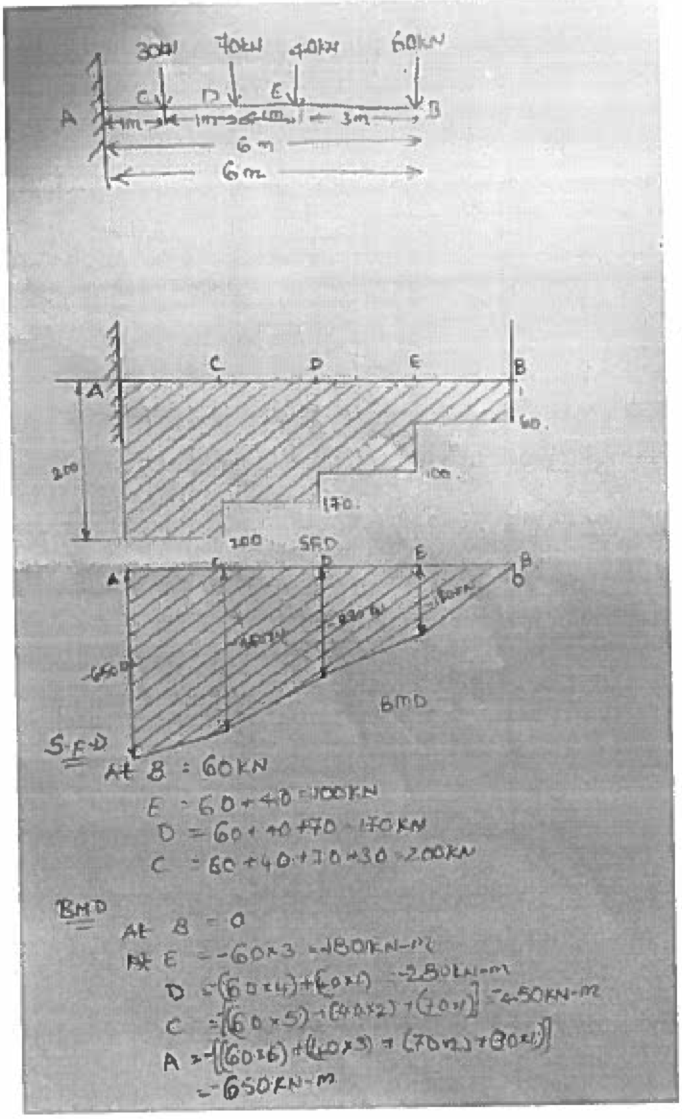
$$= [\sigma_1 \sin\theta] \frac{BC}{FC} - [\sigma_2 \cos\theta] \frac{FB}{FC}$$

$$\sigma_t = \frac{\sigma_1}{2} \sin 2\theta - \frac{\sigma_2}{2} \sin 2\theta$$

$$\sigma_t = \sin 2\theta \left[\frac{\sigma_1}{2} - \frac{\sigma_2}{2}\right]$$

$$\boxed{\sigma_t = \sin 2\theta \left[\frac{\sigma_1 - \sigma_2}{2}\right]}$$

8. A cantilever beam 6 m long carries load of 30 kN, 70 kN, 40 kN and 60 kN at a distance of 1 m, 2 m, 3 m, 6 m respectively from the fixed end. Draw shear force and bending moment diagram.



S.F.D

At B = 60KN

E = 60+40 = 100KN

D = 60+40+70 = 170KN

C = 60+40+70+30 = 200KN

BMD

At B = 0

At E = -60×3 = -180KN-m

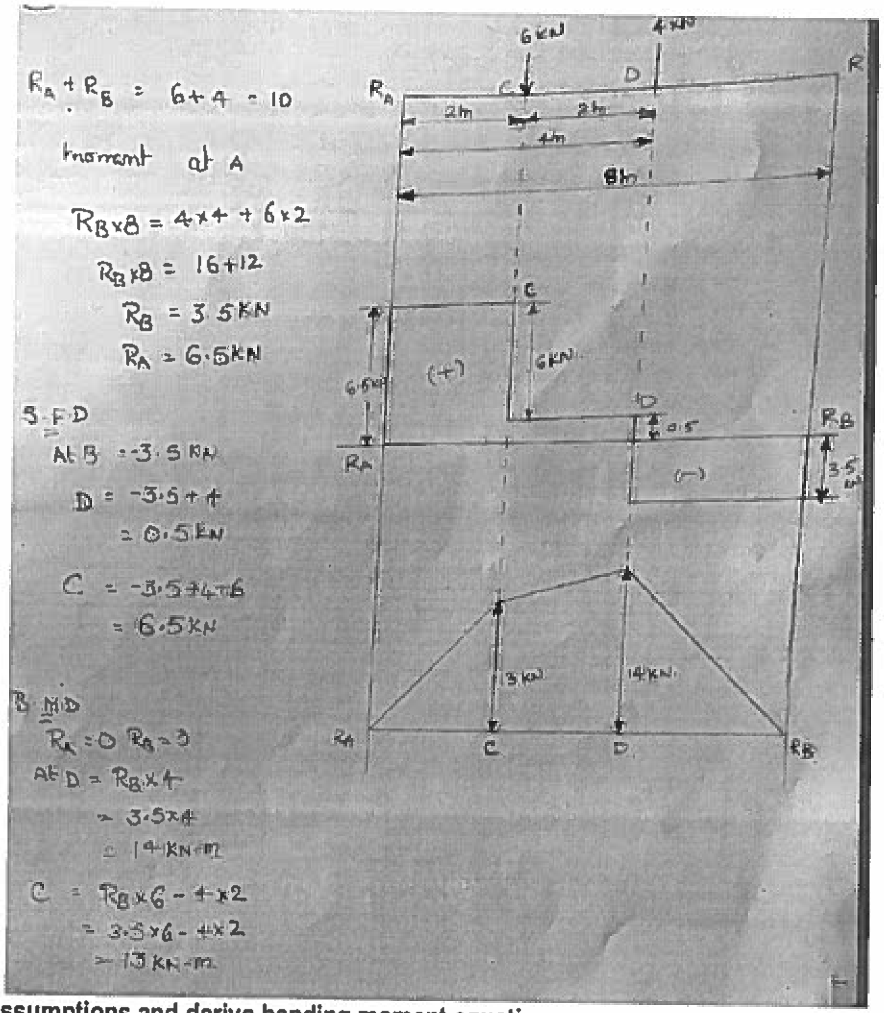
D = -(60×4)+(40×1) = -280KN-m

C = -[(60×5)+(40×2)+(70×1)] = -450KN-m

A = -[(60×6)+(40×3)+(70×2)+(30×1)]

= -650KN-m

9. a. A cantilever beam of length 4 m carries point loads of 1 KN, 2 KN and 3 KN at a 1, 2 and 4 m from the fixed end. Draw S.F.D and B.M.D



$R_A + R_B = 6+4 = 10$

moment at A

$R_B \times 8 = 4\times4 + 6\times2$

$R_B \times 8 = 16+12$

$R_B = 3.5KN$

$R_A = 6.5KN$

S.F.D

At B $= -3.5 KN$

$D = -3.5+4$

$= 0.5 KN$

$C = -3.5 \cancel{+4} + 6$

$= 6.5 KN$

B.M.D

$R_A = 0$ $R_B = 3$

At $D = R_B \times 4$

$= 3.5\times4$

$= 14 KN\text{-}m$

$C = R_B\times6 - 4\times2$

$= 3.5\times6 - 4\times2$

$= 13 KN\text{-}m$

10. State the assumptions and derive bending moment equation.

Assumptions:

1. Beam is initially straight, and has a constant cross-section.
2. Beam is made of homogeneous material and the beam has a longitudinal plane of symmetry.
3. Resultant of the applied loads lies in the plane of symmetry.
4. The geometry of the overall member is such that bending not buckling is the primary cause of failure.
5. Elastic limit is nowhere exceeded and 'E' is same in tension and compression.
6. Plane cross - sections remains plane before and after bending.

DERIVATION OF BENDING EQUATION:

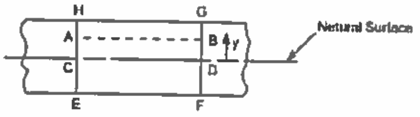


Fig 1(a)

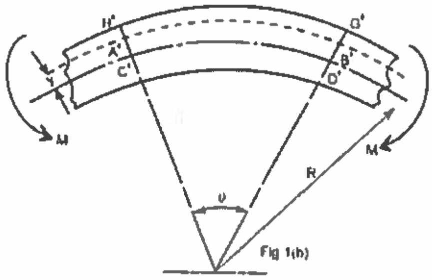
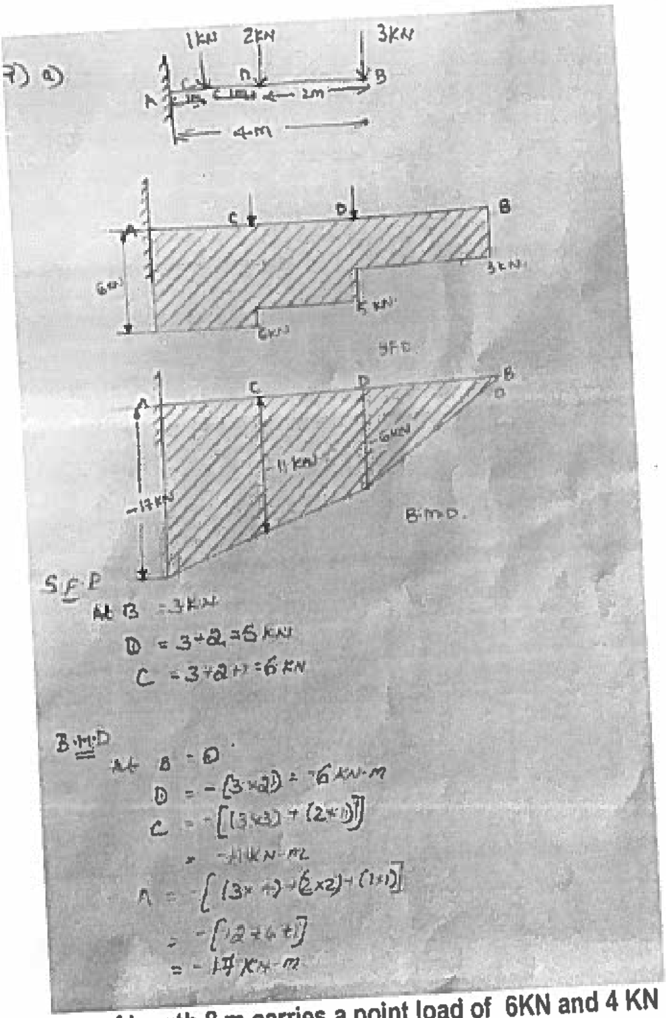


Fig 1(b)

In order to compute the value of bending stresses developed in a loaded beam, let us consider the two cross-sections of a beam **HE** and **GF**, originally parallel as shown in fig 1(a).when the beam is to bend it is assumed that these sections remain parallel i.e. **H'E'** and **G'F'**, the final position of the sections, are still straight lines, they then subtend some angle q.

Consider now fiber AB in the material, at a distance y from the N.A, when the beam bends this will stretch to A'B'. Therefore, strain in fibre AB,



S.F.D

At B = 3KN

D = 3+2 = 5 KN

C = 3+2+1 = 6 KN

B.M.D

At B = 0

D = $-(3\times2) = -6$ KN-m

C = $-[(3\times3) + (2\times1)]$

= -11 KN-m

A = $-[(3\times4) + (2\times2) + (1\times1)]$

= $-[12+4+1]$

= -17 KN-m

9.b.A simply supported beam of length 8 m carries a point load of 6KN and 4 KN at a distance of 2 m and 4 m from the left end. Draw S.F.D and B.M.D

Here, the term $\sum y^2 \delta A$ is the property of the material and is called as a second moment of area of the cross-section and is denoted by a symbol I. Therefore,

$$M = \frac{E}{R} I$$

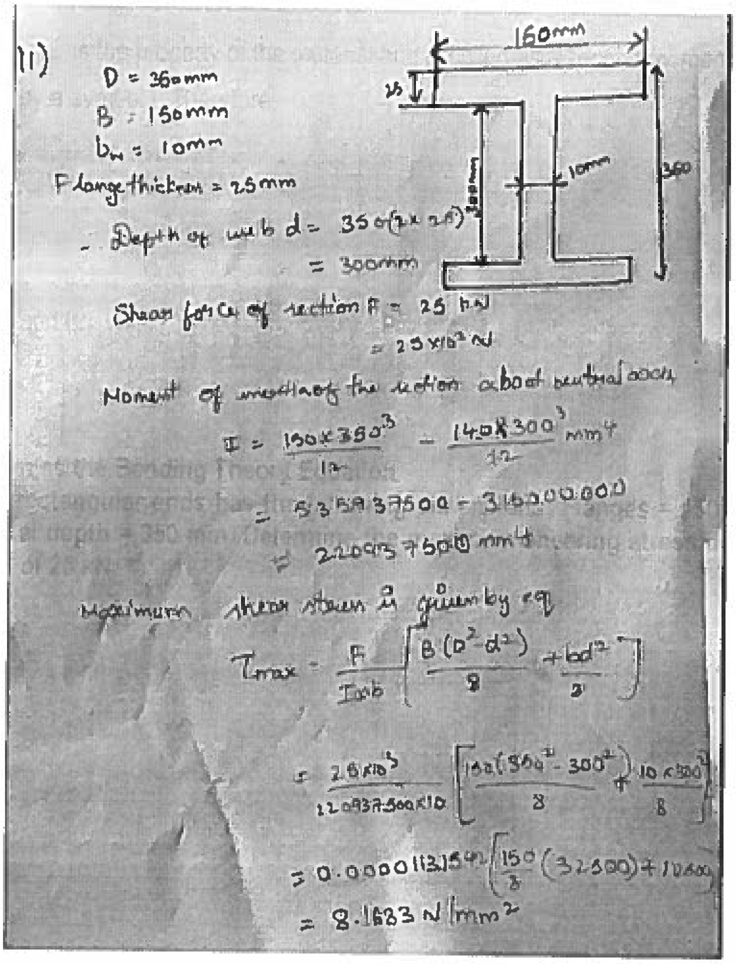
$$\frac{M}{I} = \frac{E}{R} \qquad \text{...... (2)}$$

Combine equations (1) and (2),

$$\boxed{\frac{\sigma}{y} = \frac{M}{I} = \frac{E}{R}}$$

This equation is known as the Bending Theory Equation

11. An I section with rectangular ends has the following dimensions: Flanges = 150 mm x 25 mm, Web = 300 mm x 10 mm, Total depth = 350 mm. Determine the maximum shearing stress developed in the beam for the shearing force of 25 kN.



11)

D = 360 mm

B = 150 mm

b_w = 10 mm

Flange thickness = 25 mm

∴ Depth of web d = 350 − (2 × 25)

= 300 mm

Shear force of section F = 25 kN

= 25×10^3 N

Moment of inertia of the section about neutral axis

$$I = \frac{150 \times 350^3}{12} - \frac{140 \times 300^3}{12} \text{ mm}^4$$

$$= 535937500 - 315000000$$

$$= 220937500 \text{ mm}^4$$

Maximum shear stress is given by eq

$$\tau_{max} = \frac{F}{I \times b}\left[\frac{B(D^2 - d^2)}{8} + \frac{bd^2}{8}\right]$$

$$= \frac{25 \times 10^3}{220937500 \times 10}\left[\frac{150(350^2 - 300^2)}{8} + \frac{10 \times 300^2}{8}\right]$$

$$= 0.0000113154\left(\frac{150}{8}(32500) + 10800\right)$$

$$= 8.1683 \text{ N/mm}^2$$

12. A steel cantilever 6m long carries two point loads 20 kN at the free end and 25 kN at a distance of 2 m from the free end. Find the slope and deflection at free end. Take I = 1.3 x 10^8 mm^4 and E = 2x10^5 N/mm^2.

$$\varepsilon_{AB} = \frac{\text{change in length}}{\text{original length}}$$

$$= \frac{A'B'\text{-}AB}{AB} \qquad (\text{since } AB = CD \ \& \ CD = C'D')$$

$$= \frac{A'B'\text{-}C'D'}{C'D'}$$

Since CD and C'D' are on the neutral axis and it is assumed that the Stress on the neutral axis zero. Therefore, there won't be any strain on the neutral axis.

$$\varepsilon_{AB} = \frac{A'B'\text{-}C'D'}{C'D'}$$

$$= \frac{(R+y)\theta - R\theta}{R\theta}$$

$$= \frac{y}{R}$$

Since,

$$E = \frac{\text{stress}}{\text{strain}}$$

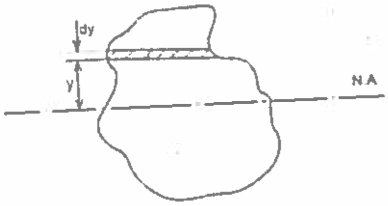
$$E = \frac{\sigma}{\frac{y}{R}}$$

$$\frac{E}{R} = \frac{\sigma}{y} \qquad \text{...... (1)}$$

Here, E is the Young's modulus of elasticity.

Consider any arbitrary a cross-section of beam, as shown below, now the stress on a fibre at a distance 'y' from the N.A, is given by the expression,

$$\sigma = \frac{E}{R} y$$



Force on the strip,

$$F = \sigma \delta A$$

$$= \left(\frac{E}{R} y \right) \delta A$$

Moment about the Neutral axis,

$$\delta M = F \times y$$

$$= \frac{E}{R} y^2 \delta A$$

Total moment,

$$M = \sum \delta M$$

$$= \sum \frac{E}{R} y^2 \delta A$$

$$= \frac{E}{R} \sum y^2 \delta A$$

13.a. Design a suitable diameter for a circular shaft required to transmit 80.2 kW at 200 rpm. The shear stress in the shaft is not to exceed 75 MN/m² and the maximum torque exceeds the mean by 40%. Also calculate the angle of twist in a length of 2 metres. Take C = 84 GN/m².

Given

$P = 80.2 kW$

$N = 200 rpm$

$\tau = 75 MPa$

$T_{max} = 1.4 T_{mean}$

$\theta = ?$ $L = 2m$

$C = 84 GN/m^2$

$$P = \frac{2\pi N T}{60}$$

$$80.2 \times 10^3 = \frac{2 \times \pi \times 200 \times T_{me}}{60}$$

$T_{mean} = 3829.2$ N-m.

$T_{max} = 1.4 \times 3829.2$

$= 5360.9$ N-m

$$T = \frac{\pi}{16} f_s d^3$$

$$5360.9 = \frac{\pi}{16} \times 75 \times 10^6 \times d^3$$

$d = 7.13$ mt

$$\frac{T}{J} = \frac{C\theta}{L}$$

$$\frac{5360.9}{\frac{\pi}{32} \times (7.13)^4} = \frac{84 \times 10^3 \times \theta}{2}$$

$\Rightarrow \theta = 0.3$ radians

13.B. Assumptions For The Derivation of Torsional Equation:

1. The material of the shaft is uniform throughout.
2. The shaft, circular in cross-section remains circular even after loading.
3. A plane section of the shaft normal to its axis before loading remains plane even after the torque has been applied.
4. The twist along the length of the shaft is uniform.
5. The distance between any two normal cross-sections remains same even after application of torque.
6. The maximum shear stress induced in the shaft due to application of torque does not exceed its elastic limit value.

DERIVATION OF TORSION FROMULA:

Let us consider,

L = length of shaft

R = radius of shaft

J = polar moment of inertia

τ_s = maximum shear stress induced

G = modulus of rigidity

γ_s = shear strain at the outer surface of the shaft

ϕ = angle of twist

1(2) Given

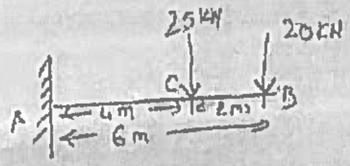
$L = 6\,m$

$W_1 = 25\,KN$

$W_2 = 20\,KN$

$I = 1.3 \times 10^{8}\,mm^4 = 1.3 \times 10^{-4}\,m^4$

$E = 2 \times 10^{5}\,N/mm^2 = 2 \times 10^{8}\,KN/m^2$



B.M.C

At B = 0

$C = -(20 \times 2)$

$= -40\,KN\text{-}m$

$A = -(20 \times 6) - (25 \times 4)$

$= -220\,KN\text{-}m$

Bending moment Area

$A_1 = \frac{1}{2}\,bh$

$= \frac{1}{2} \times 2 \times 40$

$= 40\,m^2$

$A_2 = \frac{1}{2} \times \overset{2}{\cancel{4}} \times \overset{20}{\cancel{40}}$

$= \cancel{80m} \, 60\,m^2$

$A_3 = \frac{1}{\cancel{2}} \times \overset{2}{\cancel{4}} \times 220$

$= 440\,m^2$

$A = A_1 + A_2 + A_3$

$= 40 + 60 + 440 = 540\,m^2$

According to moment Area Method

$$\text{slope } i_c = \frac{\text{Area of BMD}}{E \cdot I}$$

$$= \frac{540}{2 \times 10^{8} \times 1.3 \times 10^{-4}} = 0.020\,\text{rad}$$

Deflection At free End

$$y = \frac{A\bar{x}}{EI}$$

$\bar{x} \Rightarrow$ I $= \frac{2}{3} \times 2 = 1.33\,m$

II $= 2 + \frac{4}{2} = 4\,mt$

III $= 2 + \left(4 \times \frac{2}{3}\right) = 4.66\,m$

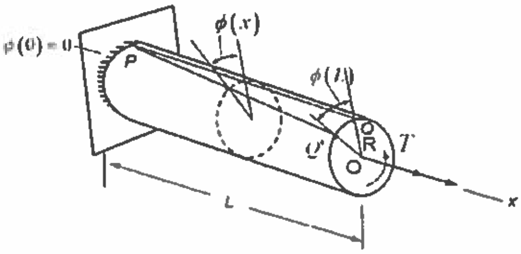
$$y = \frac{A_1 x_1 + A_2 x_2 + A_3 x_3}{EI}$$

$$= \frac{(40 \times 1.33) + (60 \times 4) + (440 \times 4.66)}{1.3 \times 10^{-4} \times 2 \times 10^{8}}$$

$$= \frac{53.2 + 240 + 2050.4}{26,000}$$

$= 0.09\,mt$

A shaft is fixed at one end and torque is being applied at the another. If a line PQ is drawn on the shaft, it will be distorted to PQ' on the application of torque. The cross-section will be twisted through an angle φ and surface by an angle of Φ .



Here,

$$\frac{QQ'}{L} = \frac{R\Phi}{L}$$

Shear strain (γ_s) =

And

$$\gamma_s = \frac{\tau_s}{G}$$

(From Hooke's law)

So, from above equations, we have;

$$\frac{\tau_s}{G} = \frac{R\Phi}{L}$$

$$\Longrightarrow \frac{\tau_s}{R} = \frac{G\Phi}{L}$$

Consider an elementary ring of thickness dx at radius x and let the shear stress at this radius be τ_x.

The turning force on the elementary ring = $2\pi x.dx.\tau_x$

Turning moment due to this force (dT) = $2\pi x^2.dx.\tau_x$

To obtain the total turning moment,

$$T = \int_0^R 2\pi x^2.dx.\tau_x$$

$$= \int_0^R 2\pi x^2.\frac{\tau_s}{R}.xdx.$$

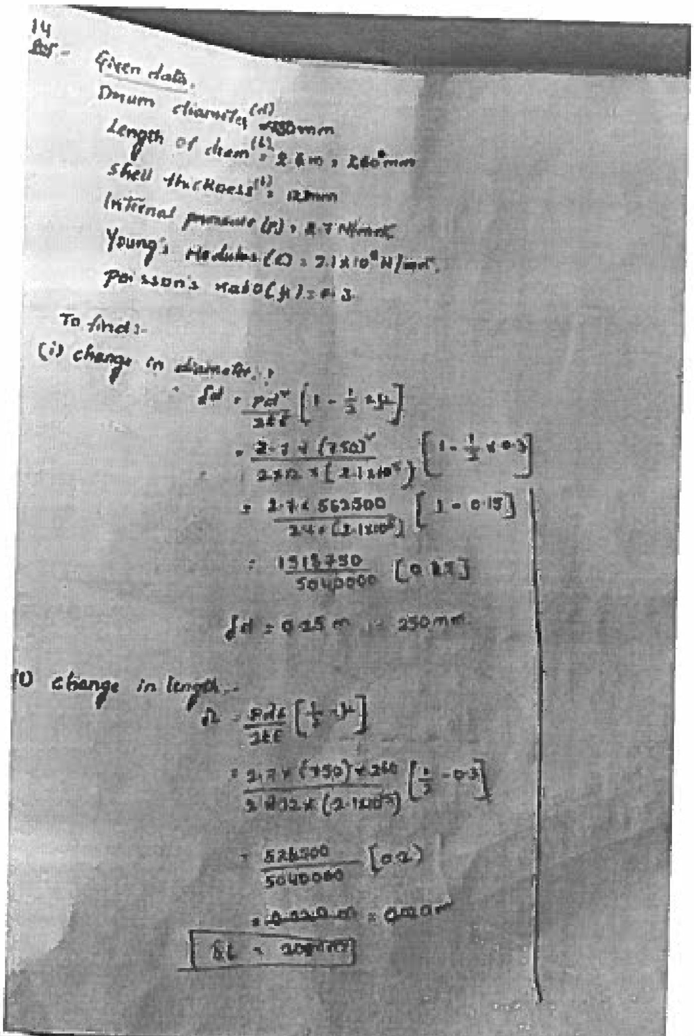
$$= 2\pi\frac{\tau_s}{R}\frac{R^4}{4} = \frac{\tau_s}{R}\frac{\pi R^4}{2} = \frac{\tau_s\tau}{R}J$$

i.e. $\frac{T}{J} = \frac{\tau_s}{R}$

Thus, we have;

$$\frac{T}{J} = \frac{\tau_s}{R} = \frac{G\Phi}{L} = \frac{\tau_x}{x}$$

14. A hollow cylindrical drum 750 mm in diameter and 2.6 m long has a shell thickness of 12 mm. If the drum is subjected to an internal pressure of 2.7 N/mm2. Determine a. Change in diameter b. Change in length and c. Change in volume Take E =2.1 x 105 N/mm2 and Poisson's ratio = 0.3.



14
Sol:- Given data.

Drum diameter (d) = 750mm

Length of drum (L) = 2.6m = 2600mm

Shell thickness (t) = 12mm

Internal pressure (p) = 2.7 N/mm²

Young's Modulus (E) = 2.1×10^5 N/mm²

Poisson's ratio (μ) = 0.3.

To find:-

(i) change in diameter:-

$$\delta d = \frac{pd^2}{2tE}\left[1-\frac{1}{2}\mu\right]$$

$$= \frac{2.7\times(750)^2}{2\times12\times(2.1\times10^5)}\left[1-\frac{1}{2}\times0.3\right]$$

$$= \frac{2.7\times562500}{24\times(2.1\times10^5)}\left[1-0.15\right]$$

$$= \frac{1518750}{5040000}\left[0.85\right]$$

$\delta d = 0.25$ m = 250mm.

(ii) change in length:-

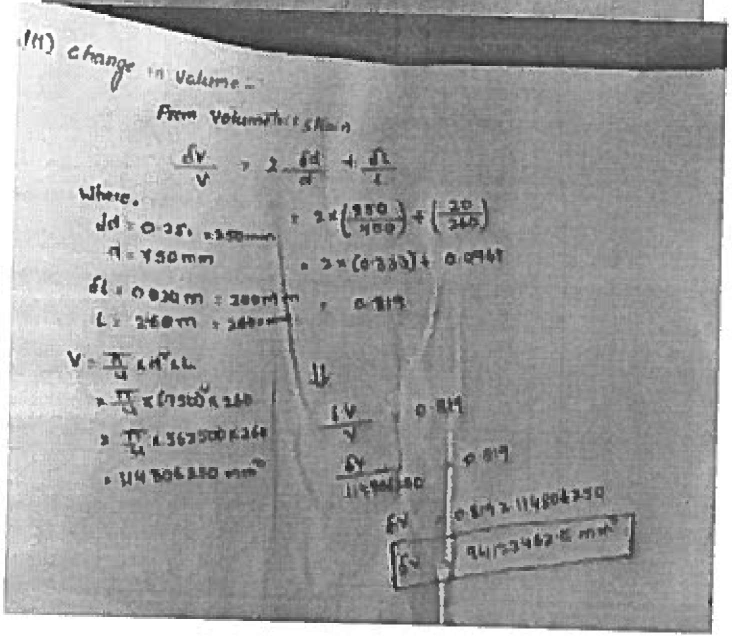
$$\delta L = \frac{pdL}{2tE}\left[\frac{1}{2}-\mu\right]$$

$$= \frac{2.7\times(750)\times260}{2\times12\times(2.1\times10^5)}\left[\frac{1}{2}-0.3\right]$$

$$= \frac{526500}{5040000}\left[0.2\right]$$

$= 0.020$ m $= 0.020$ m

$$\boxed{\delta L = 20\text{mm}}$$



(iii) change in Volume:-

From Volumetric strain

$$\frac{\delta V}{V} = 2\frac{\delta d}{d} + \frac{\delta L}{L}$$

Where,

$\delta d = 0.25 = 250$mm

$d = 750$mm

$\delta L = 0.020$m $= 20$mm

$L = 2.60$m $= 2600$mm

$$= 2\times\left(\frac{250}{750}\right)+\left(\frac{20}{260}\right)$$

$$= 2\times(0.333) + 0.0769$$

$$= 0.819$$

$$V = \frac{\pi}{4}\times d^2\times L$$

$$= \frac{\pi}{4}\times(750)^2\times260$$

$$= \frac{\pi}{4}\times562500\times260$$

$= 114806250$ mm³

⇓

$$\frac{\delta V}{V} = 0.819$$

$$\frac{\delta V}{114806250} = 0.819$$

$\delta V = 0.819\times114806250$

$$\boxed{\delta V = 94123462.5\text{ mm}^3}$$

15. Determine the ratio of strength of a solid steel column to that of a hollow column of internal diameter equal to ¾ of its external diameter. Both the columns have the same cross-sectional areas, length and end conditions.

According to Euler's column theory, the crippling or buckling load for a column of length L,

$$P_b = \frac{\pi^2 EI}{L_e^2}$$

Where L_e is the effective length of the column.

Here both the hollow and solid columns has same end conditions, so same effective length.

That implies P ∝ I

$$\therefore \frac{P_s}{P_h} = \frac{I_s}{I_h}$$

Calculation:

Given $d_i = 0.5\ d_o$ & $A_h = A_C$

$$A_h = A_c \Rightarrow \frac{\pi}{4}(d_o^2 - d_i^2) = \frac{\pi}{4}d^2$$

Substituting value of $d_i = 0.5\ d_o$ in the above expression

It comes out to be

$$\Rightarrow \frac{\pi}{4}(d_o^2 - 0.5 \times d_o^2) = \frac{\pi}{4}d^2$$

$$\Rightarrow d = 0.866\ d_o$$

$$\therefore \frac{P_s}{P_h} = \frac{I_s}{I_h} = \frac{\frac{\pi}{64}d^4}{\frac{\pi}{64}(d_o^4 - d_i^4)}$$

$$\frac{P_s}{P_h} = \frac{\frac{9}{16}}{\frac{15}{16}} = \frac{3}{5}$$

2/01/23

Signature of Faculty

Head of the Department

2/01/2023

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2022 - 2023
Course Code	20EE304	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**DC Machines & Transformers**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Distinguish between singly excited and doubly excited systems.	20EE304.1	L2
2	List any two losses occurs in DC generators.	20EE304.2	L1
3	Write the condition to get maximum efficiency in DC machine.	20EE304.3	L1
4	Define the term transformation ratio in transformers.	20EE304.4	L1
5	Mention the reason of OC test performed on LV side of a single phase transformer.	20EE304.5	L2

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Derive an expression for magnetic force developed in a doubly-excited translational magnetic system.	12M	20EE304.1	L2
	OR			
7 (a)	Derive expressions of field energy, co energy and magnetic force in a singly excited electromechanical unit.	6M	20EE304.1	L2
7 (b)	Explain the concept of energy in magnetic system with neat diagram.	6M	20EE304.1	L2
8 (a)	Explain the construction and working principle of DC generator with neat diagram	6M	20EE304.2	L2
8 (b)	A 4 pole lap wound DC shunt generator has a useful flux/pole of 0.07 Wb. The armature winding consists of 220 turns, each of 0.04 Ω resistance. Calculate the terminal voltage when running at 900 rpm, if armature current is 50 A.	6M	20EE304.1	L3
	OR			
9 (a)	Derive an expression for EMF equation of DC Generator.	6M	20EE304.2	L2
9 (b)	Explain the internal and external characteristics of DC shunt generator.	6M	20EE304.2	L2
10 (a)	Derive equation for armature torque of a dc motor. Also mention the importance of back EMF.	12M	20EE304.3	L2
	OR			
11 (a)	Explain the working principle of DC motor with neat diagram	6M	20EE304.3	L2
11 (b)	Compare the armature and field control method of speed control of dc motor.	6M	20EE304.3	L2
12 (a)	Explain the working principle of transformer with neat diagram.	6M	20EE304.4	L2
12 (b)	A 400/230 V, 50 Hz, single phase transformer has 200 turns on high voltage side. Find turns ratio, transformation ratio, and number of turns on low voltage winding. Also find the flux developed in the core.	6M	20EE304.4	L3
	OR			
13	Draw the phasor diagram of an ideal transformer on no load. Also, draw a phasor diagram of a practical transformer supplying lagging power factor load.	12M	20EE304.4	L3
14	Explain the features of OC and SC test of transformer with necessary diagrams. Also mention the advantages of these tests.	12M	20EE304.5	L2
	OR			
15 (a)	Discuss the necessity of parallel operation of transformers. Also state the conditions for satisfactory operation of three phase transformer in parallel.	6M	20EE304.5	L2
15 (b)	Explain the important features of auto transformer.	6M	20EE304.5	L2

NSRIT

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SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION
DC MACHINES & TRANSFORMERS AY 2022-2023
SEM END REGULAR/SUPPLY EXAM 02-01-2023

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

1. What are the differences between Singly excited & multiple excited magnetic field system? (2M)

Basis of Difference	Singly Excited System	Doubly Excited System
Definition	The type of excitation system used in electromechanical energy conversion which requires only one coil to produce the working magnetic flux is called singly excited system.	The excitation system used in electromechanical energy conversion which requires two independent coils excited by separate sources of power is known as doubly excited system.
Number of coils participating in energy conversion	In a singly excited system, only one coil takes active part in the electromechanical energy conversion process.	In doubly excited system, two independent coils take active part in the electromechanical energy conversion process.
Position of coil	Singly excited system has a coil on the stationary part only.	Doubly excited system has coils on stationary as well as rotating part.
Working principle	Singly excited system works on the principle of electromagnetic induction.	Doubly excited system operates on principle of synchronism or synchronous principle.
Starting torque	Singly excited system can produce the starting torque.	Doubly excited system does not produce starting torque.
Torques produced	The torque produced in a singly excited system is called reluctance torque or saliency torque.	In a doubly excited system, two torques are produced viz. reluctance torque and co-alignment torque.
Suitability	Singly excited system is suitable in variable speed applications.	Doubly excited system is suitable in constant speed applications.

2. List any two losses occurs in DC generators. (2M)

- Copper losses
  - Armature Cu loss
  - Field Cu loss
  - Loss due to brush contact resistance
- Iron Losses
  - Hysteresis loss
  - Eddy current loss

- Mechanical losses
    - Friction loss
    - Windage loss

3. Write the condition to get maximum efficiency in DC machine. (2M)

Condition for Maximum Efficiency

The efficiency of a DC generator is not constant but changes with the change in load.

Let, for a shunt generator,

I_L = load current

V = terminal voltage

Then, the output power of the DC generator is given by,

$$\text{Output Power}, = P_o = VI_L$$

$$\text{Total Input Power}, \ P_i = P_o + \text{Losses}$$

$$\Rightarrow P_i = VI_L + I_a^2R_a + W_c$$

$$\Rightarrow P_i = VI_L + (I_L + I_{sh})^2R_a + W_c$$

Where,

- $I_a^2R_a$=variable losses = Copper losses
- W_c= Constant losses = Iron losses + Mechanical losses

Practically, the shunt field current (I_{sh}) is very small as compared to load current (I_L), hence it can be neglected. Therefore,

$$P_i = VI_L + I_L^2R_a + W_c$$

Hence, the efficiency of DC generator will be,

$$\eta = \frac{P_o}{P_i} = \frac{VI_L}{VI_L + I_L^2R_a + W_c}$$

$$\eta = \frac{1}{1 + (\frac{I_LR_a}{V}) + (\frac{W_c}{VI_L})}$$

The efficiency will be maximum when the denominator of the above expression is minimum. In order to determine minimum value of denominator, differential it with respect to variable (I_L in this case) and equate it to zero, i.e.

$$\frac{d}{dI_L}[1 + (\frac{I_LR_a}{V}) + (\frac{W_c}{VI_L})] = 0$$

$$\Rightarrow 0 + \frac{R_a}{V} - \frac{W_c}{VI_L^2} = 0$$

$$\Rightarrow \frac{R_a}{V} = \frac{W_c}{VI_L^2}$$

$$\Rightarrow I_L^2R_a = W_c$$

$$\Rightarrow \text{Variable Losses} = \text{Constant Losses}$$

Hence, the efficiency of a DC generator is maximum when the load current is such that the variable losses are equal to the constant losses

4. Define the term transformation ratio in transformers (2M)

Transformation ratio: -

The transformation ratio is defined as the ratio of the number of turns in the secondary coil to the number of turns in the primary coil of the transformer.

It is defined as the ratio of output voltage to the input voltage of the transformer.

Consider the number of turns in the secondary coil is Ns and the number of turns in the primary coil is Np

The transformation ratio (r) is r=Ns/Np=Es/Ep where Es and Ep are the voltages in the secondary and primary coil respectively.

5. Mention the reason of OC test performed on LV side of a single phase transformer (2M)

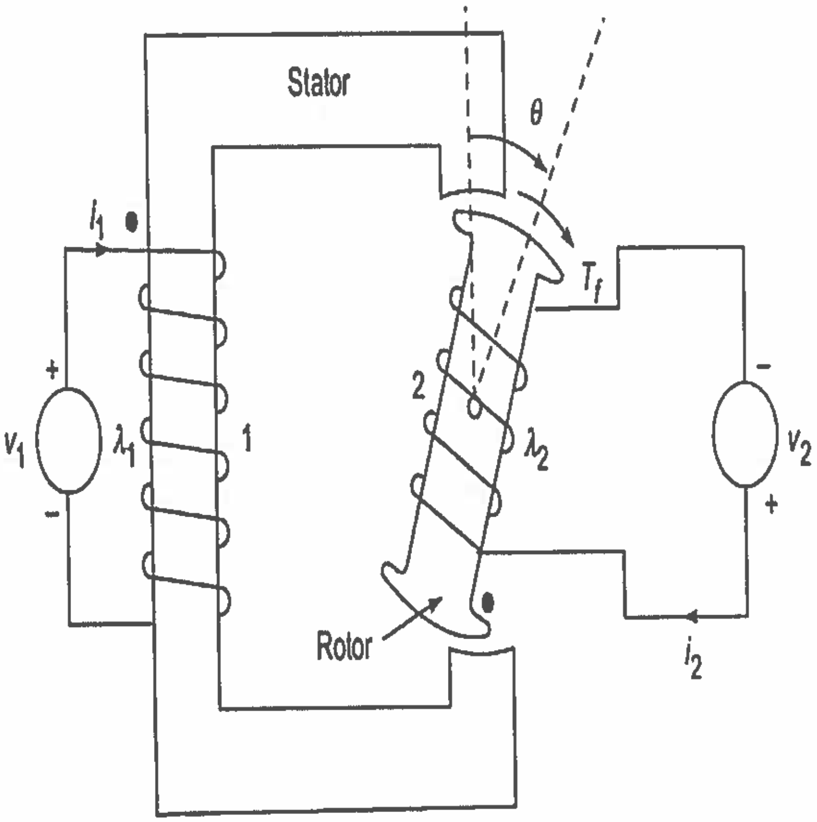
- It is done by keeping one of the windings open (without load, usually high voltage winding is open) and applying rated voltage to other winding (usually low voltage winding because it is easier to apply rated voltage).
- **No-load test (or) open-circuit test in a transform is preferred on the low voltage side because the low voltage is sufficient to obtain rated flux in the core and large on-load current for convenient reading.**
- The current drawn from this terminal is the no-load current at a low power factor corresponding to the core loss component.
- Since the no-load current is very small it doesn't contribute to the copper loss. Core loss is calculated by multiplying the applied voltage and no-load current.
- As the secondary side is open, the entire coil will be purely inductive in nature. So, the power will be **lagging due to the inductive property** of the circuit. So **LPF (Low Power Factor)** Wattmeter is used in the open circuit test of the transformer.

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

6. Derive an expression for magnetic force developed in a doubly-excited translational magnetic system. (12 M)

Below Figure shows a magnetic field system with two electrical excitations—one on stator and the other on rotor. The system can be described in either of the two sets of three independent variables; (λ_1, λ_2, q) or (i_1, i_2, q). In terms of the first set

$$T_f = -\frac{\partial W_f(\lambda_1, \lambda_2, \theta)}{\partial \theta}$$



in terms of the first set

$$T_f = \frac{\partial W_f}{\partial \theta}(\lambda_1, \lambda_2, \theta) \quad ①$$

where the field energy is given by

$$W_f(\lambda_1, \lambda_2, \theta) = \int_0^{\lambda_1} i_1 \, d\lambda_1 + \int_0^{\lambda_2} i_2 \, d\lambda_2 \quad ②$$

we know that $i = \frac{\partial W_f(\lambda, x)}{\partial \lambda}$ Like this

$$i_1 = \frac{\partial W_f}{\partial \lambda_1}(\lambda_1, \lambda_2, \theta)$$

$$i_2 = \frac{\partial W_f}{\partial \lambda_2}(\lambda_1, \lambda_2, \theta)$$

Assuming linearity

$$\lambda_1 = L_{11} i_1 + L_{12} i_2$$

$$\lambda_2 = L_{21} i_1 + L_{22} i_2$$

$$(L_{12} = L_{21})$$

∵ The flux produced by C_1 links with its own coil C_1 & secondary coil C_2

$$\Phi_1 = L_1 i_1 + M_1 i_2$$

Similarly $\Phi_2 = L_2 i_2 + M_2 i_1$

solving for i_1 & i_2 in terms of λ_1, λ_2 and substituting in eqn ②

$$W_f(\lambda_1, \lambda_2, \theta) = \frac{1}{2}\beta_{11}\lambda_1^2 + \beta_{12}\lambda_1\lambda_2 + \frac{1}{2}\beta_{22}\lambda_2^2$$

where $\beta_{11} = \frac{L_{22}}{L_{11}L_{22} - L_{12}^2}$ $\quad \beta_{12} = \beta_{21} = \frac{-L_{12}}{(L_{11}L_{22} - L_{12}^2)}$

$$\beta_{22} = \frac{L_{11}}{L_{11}L_{22} - L_{12}^2}$$

The self & mutual inductance of two exciting coils are functions of angle θ

If currents are used to describe the system state

$$T_f = \frac{\partial W_f'}{\partial \theta}(i_1, i_2\ \theta)$$

Coenergy $W_f'(i_1, i_2, \theta) = \int_0^{i_1} \lambda_1\, di_1 + \int_0^{i_2} \lambda_2\, di_2$

in linear case $W_f'(i_1, i_2, \theta) = \frac{1}{2} L_{11} i_1^2 + L_{12} i_1 i_2 + \frac{1}{2} L_{22} i_2^2$

where inductances are functions of angle θ

$$\therefore\ i_1 = \beta_{11}\lambda_{11} + \beta_{12}\lambda_2$$

$$i_2 = \beta_{21}\lambda_1 + \beta_{22}\lambda_2;\ \beta_{21} = \beta_{12}$$

$$W_f(\lambda_1, \lambda_2, \theta) = \int_0^{\lambda_1} (\beta_{11}\lambda_1 + \beta_{12}\lambda_2)\, d\lambda_1 + \int_0^{\lambda_2} (\beta_{12}\lambda_1 + \beta_{22}\lambda_2)\, d\lambda_2$$

$$= \beta_{11}\int_0^{\lambda_1} \lambda_1\, d\lambda_1 + \beta_{12}\left[\int_0^{\lambda_1} \lambda_2\, d\lambda_1 + \int_0^{\lambda_2} \lambda_1\, d\lambda_2\right] + \beta_{22}\int_0^{\lambda_2} \lambda_2\, d\lambda_2$$

$$= \left[\beta_{11}\int_0^{\lambda_1} \lambda_1\, d\lambda_1 + \beta_{12}\int_0^{\lambda_1\lambda_2} d(\lambda_1\lambda_2) + \beta_{22}\int_0^{\lambda_2} \lambda_2\, d\lambda_2\right]$$

$$= \frac{1}{2}\beta_{11}\lambda^2 + \beta_{12}\lambda_1\lambda_2 + \frac{1}{2}\beta_{22}\lambda_2^2$$

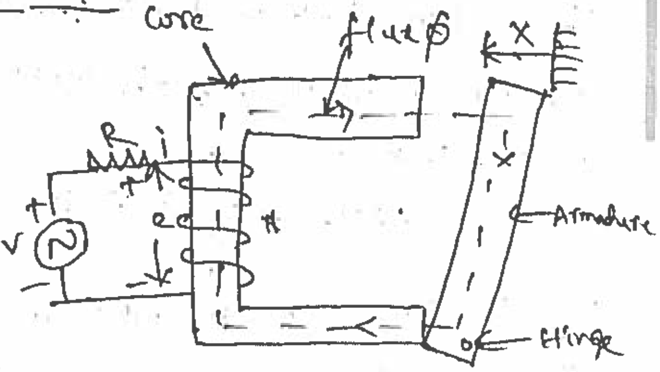
7. (a) Derive expressions of field energy, co energy and magnetic force in a singly excited electromechanical unit. (06M)

ENERGY IN MAGNETIC SYSTEM

$\lambda = N\phi$

flux linkages

emf $e = \frac{d\lambda}{dt}$



$v = iR + e$

$= iR + \frac{d\lambda}{dt}$

the electric energy I/p into the Ideal coil due to the flow of current i in time dt is

$$dW_e = ei\ dt$$

Assuming for the time being that the armature is held fixed at position 'x', all the i/p. energy stored in the magnetic field

$$dW_e = dW_f = ei\ dt\ (ei\ dt)$$

where dW_f is the change in field energy in time dt.

$$dW_e = ei\ dt = \frac{d\lambda}{dt} i\ dt = i\ d\lambda$$

$$= i\ d(N\phi)$$

$$= Ni\ d\phi$$

$$= F\ d\phi$$

$$= dW_f$$

$(F = Ni,\ \text{mmf})$

the relation i-λ or F-λ, non linear

the energy absorbed by the field for finite change in flux linkages

$$\Delta W_F = \int_{\lambda_1}^{\lambda_2} i(\lambda)\, d\lambda = \int_{\phi_1}^{\phi_2} F(\phi)\, d\phi$$

As the flux in the magnetic circuit undergoes a cycle $\phi_1 \to \phi_2 \to \phi_1$,

an irrecoverable loss in energy Takes place due to hysteresis and eddy current in the iron,

The assumption renders the ideal coil and the magnetic circuit as a conservative system with energy interchange between themselves so that the net energy is conserved.

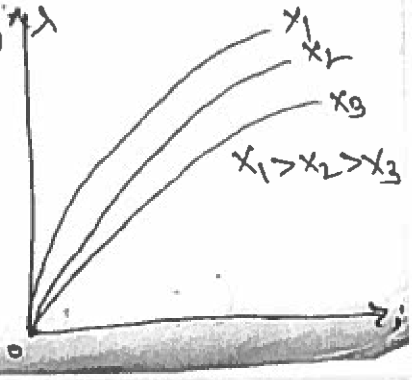
→ the energy absorbed by the magnetic system to establish flux φ (or flux linkages λ) from initial zero flux is

$$W_f = \int_0^\lambda i(\lambda)\, d\lambda = \int_0^\phi F(\phi)\, d\phi$$

→ The i-λ relationship is magnetization curve which varies with the configuration variable x

→ the air gap b/w the armature and core varies with position x of the Armature,

the total reluctance $\left(\frac{NI}{\phi}\right)$ of the magnetic path decreases as x increases



The i-λ relationship for various values of x is indicated in figure.

The relationship can be expressed as

$$i = i(\lambda, x)$$

if λ is the independent variable or as,

$$\lambda = \lambda(i, x)$$

if i is the independent variable

The field energy is in general, a function of two variables i.e. $W_f = W_f(\lambda, x)$ (or)

$$W_f = W_f(i, x)$$

→ A change in λ with fixed x (electro magnetic energy interchange

$$\left[\begin{array}{l} V = iR + \frac{d\lambda}{dt} \\ dW_e = i\,d\lambda = F\,d\varphi = dW_f \end{array}\right]$$

→ if x is allowed to change with fixed λ, energy will interchange b/w the magnetic circuit and the mechanical system. (electro + magnetic -mechanical)

→ As per $W_f = \int_0^{\lambda} i(\lambda)\, d\lambda = \int_0^{\varphi} F(\varphi)\, d\varphi$

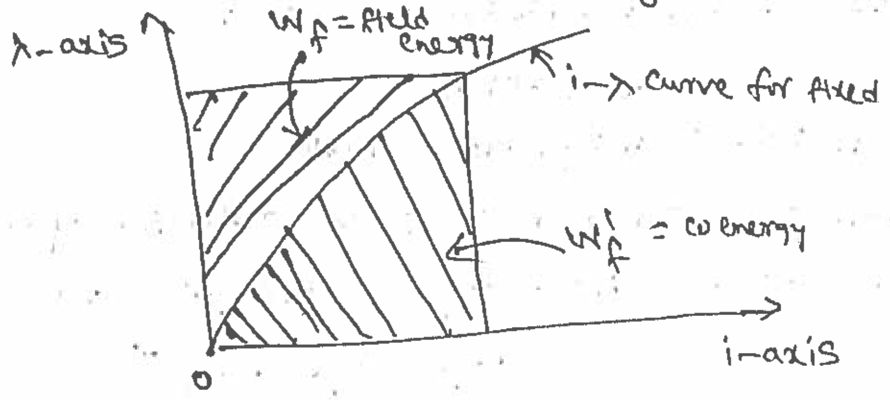
the field energy is the area b/w λ-axis and i-λ

A New term, co-energy is now defined as

$$W'_f(i, x) = i\lambda - W_f(\lambda, x)$$

wherein by expressing λ as $\lambda(i, x)$, The independent variable of W'_f becomes i and x.

The co energy on fig is shown to be the complementary area of the $i-\lambda$ curve. $W'_f = \int_0^{i} \lambda \, di$



Linear case

$$W_f = \frac{1}{2} i \lambda = \frac{1}{2} F\phi = \frac{1}{2} R\phi^2$$

$\therefore R = F/\phi$ = reluctance of the magnetic circuit.

$\therefore$ coil inductance $L = \frac{\lambda}{i}$

The field energy can be expressed as $W_f = \frac{1}{2}\frac{\lambda^2}{L}$

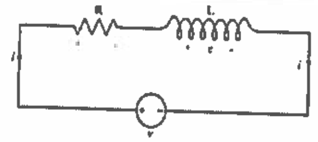
$$W_f(\lambda, x) = \frac{1}{2}\frac{\lambda^2}{L(x)}$$

The field energy is distributed throughout the space Occupied by the field. Assuming no losses const permeability

Energy density of the field $w_f = \int_0^B H\,dB = \frac{1}{2}HB = \frac{1}{2}\frac{B^2}{\mu}$

7.(b) Explain the concept of energy in magnetic system with neat diagram (6M)

Consider a coil of N turns wound around a magnetic core and is connected to voltage source (see the figure).



By applying KVL, we get,

$$V = e + iR \quad \dots(1)$$

Where,

- e is the induced EMF in the coil,
- R is the resistance of the coil circuit.

The instantaneous power input is given by,

$$p = Vi = e + i^2R \quad \dots(2)$$

Hence, the energy input to the system is,

$$W_i = \int_0^T = p\,dt = \int_0^T ei\,dt + \int_0^T i^2R\,dt \quad \dots(3)$$

The eq. (3) shows that the total input energy consists of two parts. The first part is energy stored in magnetic field and the second part is the energy dissipated in the circuit resistance in the form of heat. Therefore, the energy stored in the magnetic field is given by

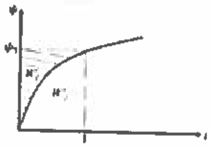
$$W_f = \int_0^T ei\,dt$$

Also, according to Faraday's law of electromagnetic induction, the induced emf is given by,

$$e = N\frac{d\phi}{dt} = \frac{d(N\phi)}{dt} = \frac{d\psi}{dt}$$

Where, $\psi = N_\phi$ is the magnetic flux linkage.

$$\therefore W_f = \int_0^T \frac{d\psi}{dt} i\,dt = \int_0^\psi i\,d\psi \quad \dots(4)$$



The eq (4) shows that the energy stored in the magnetic field is equal to the area between the ψ-i curve for the system and the flux linkage (ψ) axis.

Co-Energy and Field Energy in a Magnetically Linear System

In a magnetically linear system, the field energy is given by,

$$W_f = \int_0^\psi i\,d\psi$$

$$\because i = \frac{\psi}{L} \text{ or } \psi = Li$$

$$\therefore W_f = \int_0^\psi \frac{\psi}{L} d\psi = \frac{\psi^2}{2L}$$

$$\longrightarrow W_f = \frac{(Li)^2}{2L} = \frac{1}{2}Li^2$$

$$\longrightarrow W_f = \frac{\psi^2}{2L} = \frac{1}{2}Li^2 \quad \dots(5)$$

8. (a) Explain the construction and working principle of DC generator with neat diagram (6M)

Whether a machine is DC Generator or a Motor the construction basically remains the same.

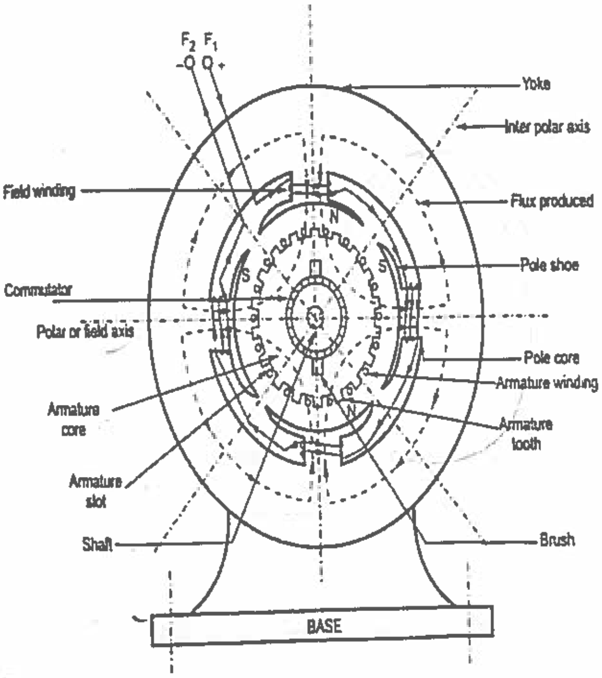
MAIN PARTS:

STATOR: That houses the field winding

ROTOR: Rotating part that rotates in magnetic field

OTHER PARTS:

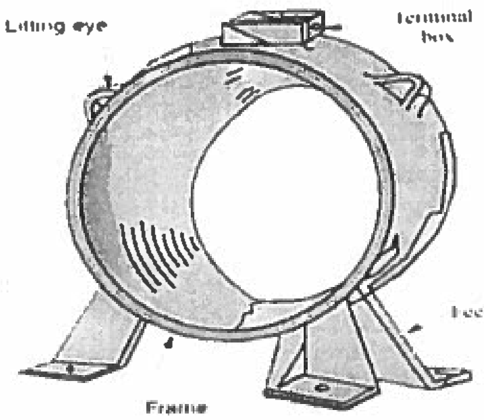
1. Yoke
2. Poles
3. Field winding
4. Armature winding
5. Commutator
6. Brushes



YOKE:

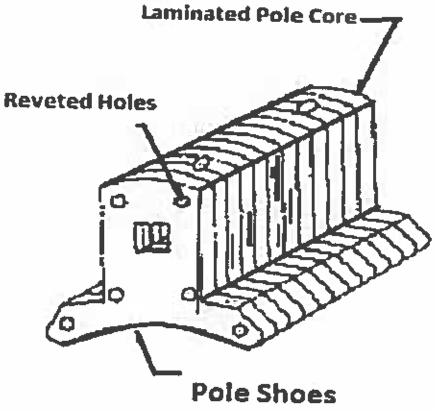
Functions:

1. It serves the purpose of outermost cover of the DC Machine . So that the insulating materials get protected from harmful atmospheric elements like moisture , dust and various gases like SO2, acidic fumes etc.
2. Made up of cast iron and steel .
3. It provides mechanical support to the poles



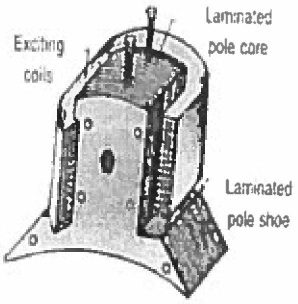
POLES:

- Each pole is divided into two parts namely 1)pole core and 2) pole shoe
- Pole core carries field winding which is necessary to produce the flux
- Pole shoe enlarges the area of armature core to come across the flux ,which is necessary to produce larger induced emf , to achieve this pole shoe has been given a particular shape
- It is made up of cast iron or cast steel.



FIELD WINDING:

- The field winding basically form an electromagnet, that produces field flux. i.e. magnetic field. exciting the pole as an electromagnet it is called field winding or exciting winding.
- Made with field coils (copper wire) wound over the slot of the pole shoe.
- These are connected in series with each other and wound in such a direction around pole cores , such that alternate "N" and "S" poles are formed.
- Total number of poles denoted as "P"



ARMATURE:

- The armature is divided into two parts namely,

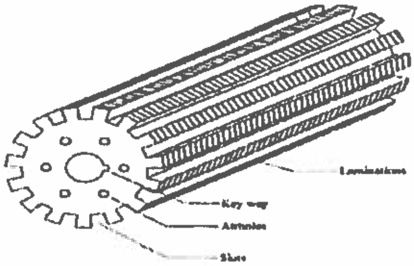
i) armature core and ii) armature winding

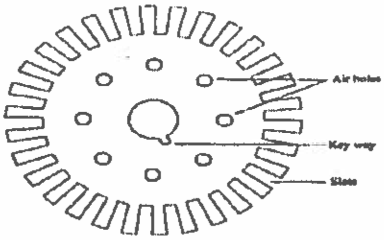
- Armature core: Armature core is cylindrical shape mounted on the shaft.it consists of slots on its periphery and the air ducts to permit the air flow through armature which serves cooling purpose.
- Functions:

1.Armature core provides house for armature winding

2.To provide a path of low reluctance to the flux.

Choice of material: copper





COMMUTATOR:

- The basic nature of EMF induced in the armature conductors is alternating. this needs rectification in case of DC generator, which is possible by commutator.
- Functions:

1. To facilitate the collection of current from the armature conductors.
2. To convert alternating emf to unidirectional dc emf

Choice of material: as it collects current from armature,it is also made up of copper segments.

BRUSHES:

- Brushes are stationary and resting on the surface of the commutator.
- Functions: to collect the current from commutator and make it available to the external circuit
- Material : to avoid the wear and tear of commutator, the brushes are made up of soft material like carbon

8.(b) A 4 pole lap wound DC shunt generator has a useful flux/pole of 0.07 Wb. The armature winding consists of 220 turns, each of 0.04 Ω resistance. Calculate the terminal voltage when running at 900 rpm, if armature current is 50 A. (6M)

⑧ b) $P = 4$ lap wound $A = P$, $N = 900\,rpm$

DC shunt generator

flux/pole $= 0.07\,wb$ $I_a = 50\,A$

NO of armature Turns $= 220$

1 Turn has 2 conductor

no of armature conductors $= 2 \times 220$

$Z = 440$

induced EMF $E = \frac{PZ\phi N}{60A} = \frac{4 \times 440 \times 0.07 \times 900}{60 \times 4}$

$= \frac{1,10,880}{240}$

$E = 462\,V$

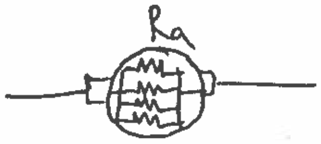
NO of Turns connected in each parallel path

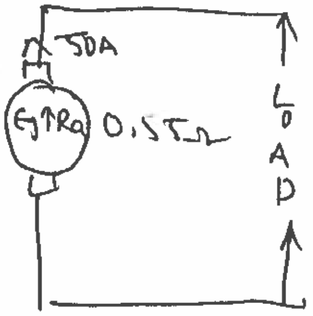
$= \frac{220}{4} = 55$

Resistance of each parallel path $= 55 \times 0.04$

$= 2.2\,\Omega$

four Resistances, each of $2.2\,\Omega$ are connected in parallel





Armature Resistance $R_a = \frac{2.2}{4} = 0.55\ \Omega$

Terminal voltage $V = E - I_a R_a$

$= 462 - (50 \times 0.55)$

$= 462 - 27.5$

$= 434.5\text{V}$

9. (a) Derive an expression for EMF equation of DC Generator (6M)

Derivation for Induced EMF of One Armature Conductor

For one revolution of the conductor,

Let,

Φ = Flux produced by each pole in weber (Wb)

and

P = number of poles in the DC generator

therefore,

Total flux produced by all the poles= Φ*P

And,

Time taken to complete one revolution =60/N

Where,

N = speed of the armature conductor in rpm.

Now, according to Faraday's law of induction, the induced emf of the armature conductor is denoted by "e" which is equal to rate of cutting the flux.

Therefore e= d Φ/dt and e= total flux/time take

Induced emf of one conductor is e= (Φ*P)/(60/N) = ΦPN/60

Derivation for Induced EMF for DC Generator

Let us suppose there are Z total numbers of conductor in a generator, and arranged in such a manner that all parallel paths are always in series.

Here,

Z = total numbers of conductor

A = number of parallel paths

Then,

Z/A = number of conductors connected in series

We know that induced emf in each path is same across the line

Therefore,

Induced emf of DC generator

E = emf of one conductor × number of conductors connected in series.

Induced emf of DC generator is

$$e = \phi P \frac{N}{60} X \frac{Z}{A}\ volts$$

9.(b) Explain the internal and external characteristics of DC shunt generator (6M)

Internal Or Total Characteristic (E/Ia)

- An internal characteristic curve shows the relation between the on-load generated emf (Eg) and the armature current (Ia).
- The on-load generated emf Eg is always less than E0 due to the armature reaction.
- Eg can be determined by subtracting the drop due to demagnetizing effect of armature reaction from no-load voltage E0.
- Therefore, internal characteristic curve lies below the O.C.C. curve.

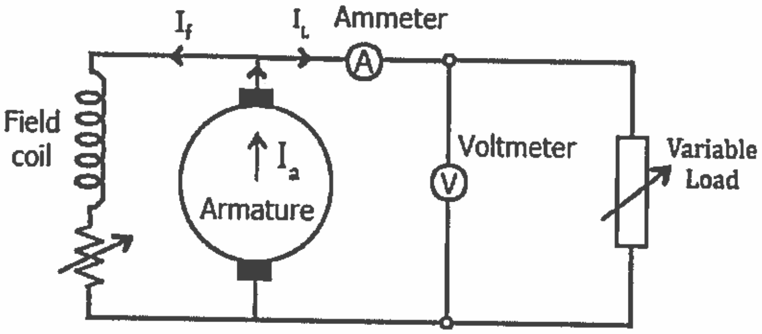
External Characteristic (V/IL)

- An external characteristic curve shows the relation between terminal voltage (V) and the load current (IL).
- Terminal voltage V is less than the generated emf Eg due to voltage drop in the armature circuit. Therefore, external characteristic curve lies below the internal characteristic curve.
- External characteristics are very important to determine the suitability of a generator for a given purpose.

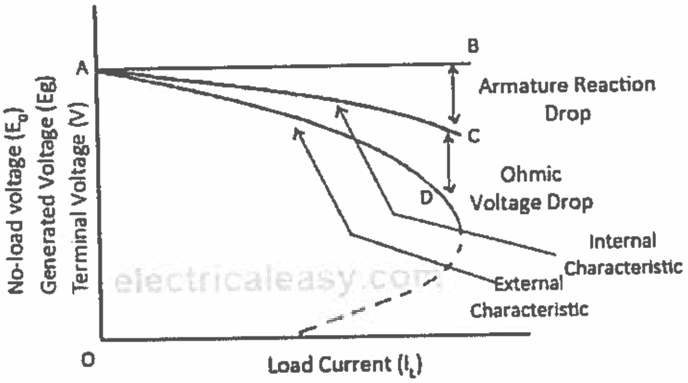
Therefore, this type of characteristic is sometimes also called as performance characteristic or load characteristic

Characteristics Of DC Shunt Generator

- To determine the internal and external load characteristics of a DC shunt generator the machine is allowed to build up its voltage before applying any external load.
- To build up voltage of a shunt generator, the generator is driven at the rated speed by a prime mover. Initial voltage is induced due to residual magnetism in the field poles.
- The generator builds up its voltage as explained by the O.C.C. curve.
- When the generator has built up the voltage, it is gradually loaded with resistive load and readings are taken at suitable intervals. Connection arrangement is as shown in the figure below.



- Unlike, separately excited DC generator, here, IL≠Ia. For a shunt generator, Ia=IL+If.
- Hence, the internal characteristic can be easily transmitted to Eg vs. IL by subtracting the correct value of If from Ia.



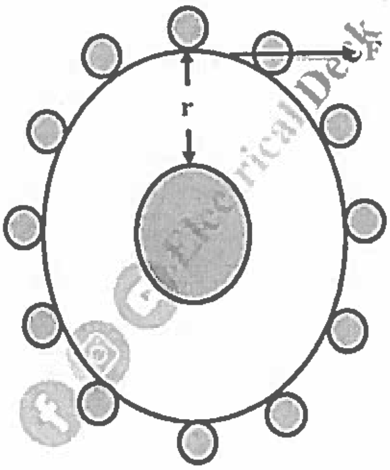
Characteristics of DC shunt generator

- During a normal running condition, when load resistance is decreased, the load current increases.
- But, as we go on decreasing the load resistance, terminal voltage also falls. So, load resistance can be decreased up to a certain limit, after which the terminal voltage drastically decreases due to excessive armature reaction at very high armature current and increased I2R losses.

Hence, beyond this limit any further decrease in load resistance results in decreasing load current. Consequently, the external characteristic curve turns back as shown by dotted line in the above figure.

10. Derive equation for armature torque of a dc motor. Also mention the importance of back EMF (12M)

Torque or moment or moment of force is the tendency of a force to rotate or move an object about an axis. A force is a push or pull, likewise, torque is a twist to an object. Mathematically, torque, T = F × r.



Let

- Tg = armature or gross torque (N-m) = Force × radius.
- r = radius of the armature in m.
- N = speed of the armature in rpm = N/60 rps.

Work done/revolution = force × distance moved per revolution

$$w.d. \;=\; F \times 2\pi r \;\; Nm$$

$$w.d. \,/\, s = F \times 2\pi r \times \frac{N}{60} \;\; Nm$$

$$w.d. \;=\; \frac{2\pi N}{60}\,(F \times r) \;\; Nm\,/\,s \;\; or\; watt$$

$$w.d. \;=\; \frac{2\pi N T_g}{60} \;\; watt \;\; .....(1)$$

The expression for voltage in dc motor is given by,

$$V \;=\; E_a + I_a\, R_a$$

$$(mutiply\; throughout\; by\; I_a\,)$$

$$V\, I_a = \; E_a\, I_a + I_a^{\;2}\, R_a$$

Electrical input = electrical power equivalent to mechanical power developed + armature copper loss

Mechanical power developed,

$$= \; E_a\, I_a \;\; watt \;\; .....(2)$$

Since equation (1) = equation (2),

$$\frac{2\pi N T_g}{60} \;=\; E_b I_a$$

$$\frac{2\pi N T_g}{60} \;=\; \frac{\phi Z N}{60} \times \frac{P}{A} \times I_a$$

$$T_g = \frac{1}{2\pi}\, \phi Z I_a \times \frac{P}{A}$$

$$T_g \;=\; 0.159\, \phi Z I_a \times \frac{P}{A} \;\; Nm$$

Hence torque of a dc motor is directly proportional to the flux/pole and armature current.

Importance of Back EMF

When the armature of the DC motor rotates under the influence of driving torque, the armature of the conductors moves through a magnetic field inducing an emf in them. The induced emf is in the opposite direction to the applied voltage and is known as the back emf.

Some advantages of back emf are listed below:

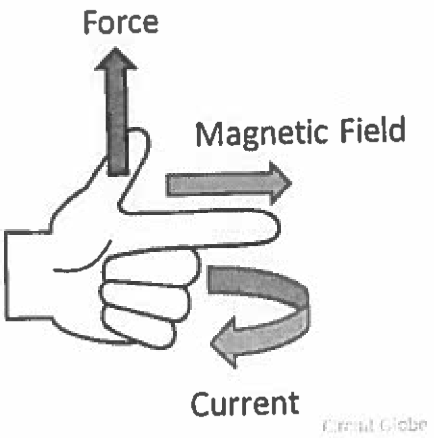
- The energy conversion in the DC motor is possible because of the back emf.
- A DC motor is made self-regulating because of back emf.

11.(a) Explain the working principle of DC motor with neat diagram (6M)

The DC motor is the device which converts the direct current into the mechanical work. It works on the principle of Lorentz Law, which states that "*the current-carrying conductor placed in a magnetic and electric field experience a force*". The experienced force is called the Lorentz force. The Fleming left-hand rule gives the direction of the force

Fleming Left Hand Rule

If the thumb, middle finger and the index finger of the left hand are displaced from each other by an angle of 90°, the middle finger represents the direction of the magnetic field. The index finger represents the direction of the current, and the thumb shows the direction of forces acting on the conductor.



The formula calculates the magnitude of the force,

$$F = BIl \quad \text{newton}$$

Before understanding the working of DC motor, first, we have to know about its construction. The armature and stator are the two main parts of the DC motor. The armature is the rotating part, and the stator is their stationary part. The armature coil is connected to the DC supply.

The armature coil consists the commutators and brushes. The commutators convert the AC induced in the armature into DC and the brushes transfer the current from rotating part of the motor to the stationary external load. The armature is placed between the north and south pole of the permanent or electromagnet.

For simplicity, consider that the armature has only one coil which is placed between the magnetic field shown below in the figure A. When the DC supply is given to the armature coil the current starts flowing through it. This current develops its own field around the coil.

Figure B shows the field induces around the coil:

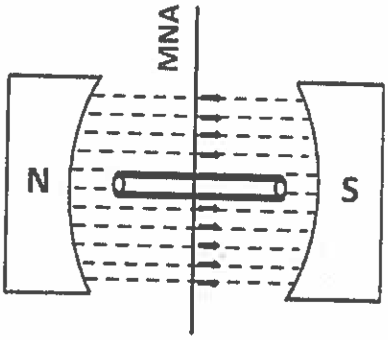


Figure A

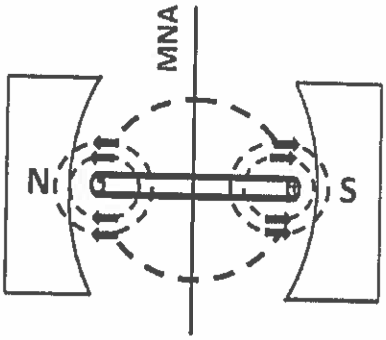


Figure B

By the interaction of the fields (produced by the coil and the magnet), the resultant field develops across the conductor. The resultant field tends to regain its original position, i.e. in the axis of the main field. The field exerts the force at the ends of the conductor, and thus the coil starts rotating.

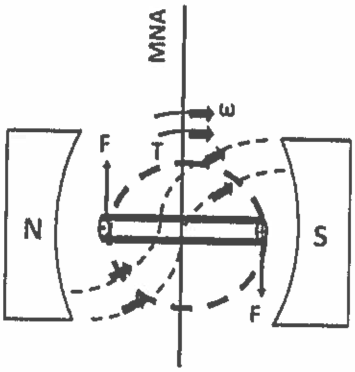
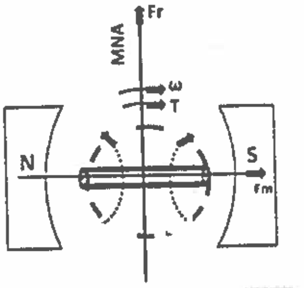


Figure C

Let the field produced by the main field be Fm, and this field rotates in the clockwise direction. When the current flows in the coil, they produce their own magnetic field say, Fr. The field Fr tries to come in the direction of the main field. Thereby, the torque act on the armature coil.



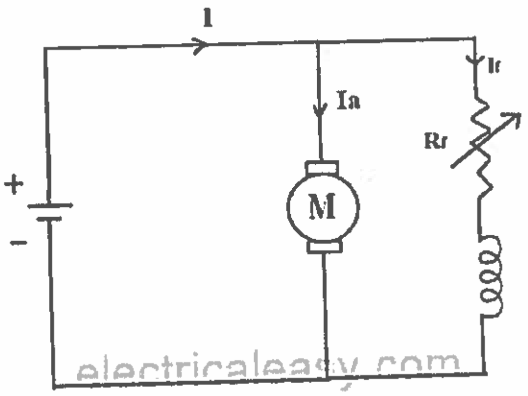
The actual DC motor consists of a large number of armature coils. *The speed of the motor is directly proportional to the number of coils used in the motor.* These coils are kept under the impact of the magnetic field.

11.(b) Compare the armature and field control method of speed control of dc motor. (6M)

Flux Control Method

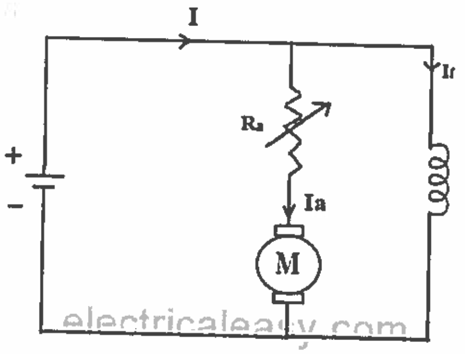
It is already explained above that the speed of a dc motor is inversely proportional to the flux per pole. Thus by decreasing the flux, speed can be increased and vice versa.

To control the flux, a rheostat is added in series with the field winding, as shown in the circuit diagram. Adding more resistance in series with the field winding will increase the speed as it decreases the flux. In shunt motors, as field current is relatively very small, Ish2R loss is small. Therefore, this method is quite efficient. Though speed can be increased above the rated value by reducing flux with this method, it puts a limit to maximum speed as weakening of field flux beyond a limit will adversely affect the commutation.



Armature Control Method

Speed of a dc motor is directly proportional to the back emf Eb and Eb = V - IaRa. That means, when supply voltage V and the armature resistance Ra are kept constant, then the speed is directly proportional to armature current Ia. Thus, if we add resistance in series with the armature, Ia decreases and, hence, the speed also decreases. Greater the resistance in series with the armature, greater the decrease in speed.



12.(a) Explain the working principle of transformer with neat diagram (6M)

A **transformer** is defined as a passive electrical device that transfers electrical energy from one circuit to another through the process of electromagnetic induction. It is most commonly used to increase ('step up') or decrease ('step down') voltage levels between circuits.

Working Principle of Transformer

The **working principle of a transformer** is very simple. Mutual induction between two or more windings (also known as coils) allows for electrical energy to be transferred between circuits. This principle is explained in further detail below.

Transformer Theory

Say you have one winding (also known as a coil) which is supplied by an alternating electrical source. The alternating current through the winding produces a continually changing and alternating flux that surrounds the winding.

If another winding is brought close to this winding, some portion of this alternating flux will link with the second winding. As this flux is continually changing in its amplitude and direction, there must be a changing flux linkage in the second winding or coil.

According to Faraday's law of electromagnetic induction, there will be an EMF induced in the second winding. If the circuit of this secondary winding is closed, then a current will flow through it. This is the basic **working principle of a transformer.**

Let us use electrical symbols to help visualize this. The winding which receives electrical power from the source is known as the 'primary winding'. In the diagram below this is the 'First Coil'.

The winding which gives the desired output voltage due to mutual induction is commonly known as the 'secondary winding'. This is the 'Second Coil' in the diagram above.

A transformer that increases voltage between the primary to secondary windings is defined as a step-up transformer. Conversely, a transformer that decreases voltage between the primary to secondary windings is defined as a step-down transformer.

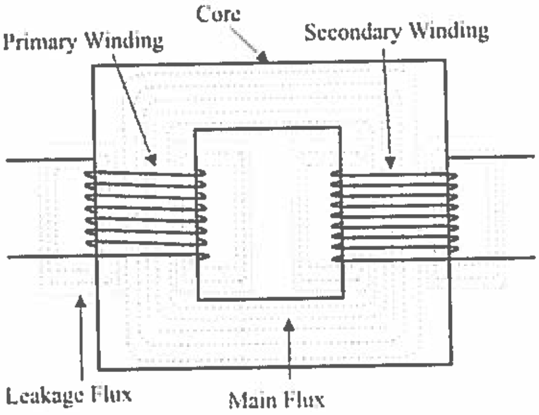
Whether the transformer increases or decreases the voltage level depends on the relative number of turns between the primary and secondary side of the transformer.

If there are more turns on the primary coil than the secondary coil than the voltage will decrease (step down).

If there are less turns on the primary coil than the secondary coil than the voltage will increase (step up).

While the diagram of the transformer above is theoretically possible in an ideal transformer – it is not very practical. This is because in the open air only a very tiny portion of the flux produced from the first coil will link with the second coil. So the current that flows through the closed circuit connected to the secondary winding will be extremely small (and difficult to measure).

The rate of change of flux linkage depends upon the amount of linked flux with the second winding. So ideally almost all of the flux of primary winding should link to the secondary winding. This is effectively and efficiently done by using a core type transformer. This provides a low reluctance path common to both of the windings.



The purpose of the transformer core is to provide a low reluctance path, through which the maximum amount of flux produced by the primary winding is passed through and linked with the secondary winding.
The current that initially passes through the transformer when it is switched on is known as the transformer inrush current.

12.(b) A 400/230 V, 50 Hz, single phase transformer has 200 turns on high voltage side. Find turns ratio, transformation ratio, and number of turns on low voltage winding. Also find the flux developed in the core. (6M)

12-b) 400/230V f = 50Hz

No of Turns on HV side = 200

Find Turns ratio, Transformation ratio,

no of Turns on low voltage side. find the flux

1) $V_1 = 400$, $V_2 = 230V$

$N_1 = 200$

$$K = \frac{V_2}{V_1} = \frac{N_2}{N_1} = \frac{E_2}{E_1} = \frac{I_1}{I_2}$$

$$\frac{V_2}{V_1} = \frac{N_2}{N_1}$$

$$\frac{230}{400} = \frac{N_2}{200}$$

$$N_2 = \frac{230}{400} \times 200$$

$$N_2 = 115$$

$$k = \frac{N_s}{N_p} = \frac{N_2}{N_1} = \frac{115}{200} = 0.575$$

emf/Turn $= 4 f \phi_m$

$$\frac{E_1}{N_1} = \frac{E_2}{N_2} = 4.44 f \phi_m$$

$$\frac{400}{200} = 4.44 \times 50 \times \phi_m$$

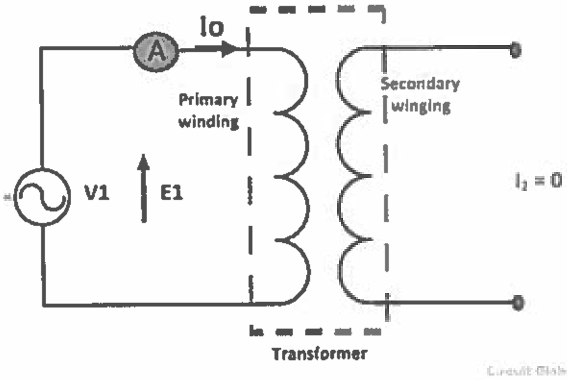
$$\frac{2}{4.44 \times 50} = \phi_m$$

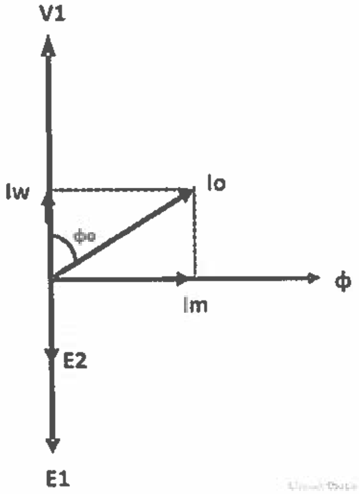
$$\phi_m = 0.009 \, Wb$$

13.(A) Draw and analyse the phasor diagram of an ideal transformer on no-load. Also, draw a phasor diagram of a practical transfomer supplying lagging power factor load (12M)

When the transformer is operating at no load, the secondary winding is open-circuited, which means there is no load on the secondary side of the transformer and, therefore, current in the secondary will be zero. While primary winding carries a small current I0 called no-load current which is **2 to 10% of the rated current.**

This current is responsible for supplying the iron losses (hysteresis and eddy current losses) in the core and a very small amount of copper losses in the primary winding. The angle of lag depends upon the losses in the transformer. The power factor is very low and varies from **0.1 to 0.15.**





The no-load current consists of two components:

- Reactive or magnetizing component Im (It is in quadrature with the applied voltage V1. It produces flux in the core and does not consume any power).
- Active or power component Iw, also know as a working component (It is in phase with the applied voltage V1. It supplies the iron losses and a small amount of primary copper loss).

The following steps are given below to draw the phasor diagram:

The function of the magnetizing component is to produce the magnetizing flux, and thus, it will be in phase with the flux.

Induced emf in the primary and the secondary winding lags the flux ϕ by 90 degrees.

The primary copper loss is neglected, and secondary current losses are zero as I2 = 0.

Therefore, the current I0 lags behind the voltage vector V1 by an angle ϕ0 called the no-load power factor angle and is shown in the phasor diagram above.

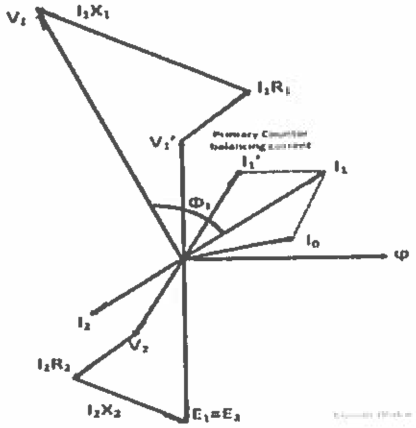
The applied voltage V1 is drawn equal and opposite to the induced emf E1 because the difference between the two, at no load, is negligible.

Active component Iw is drawn in phase with the applied voltage V1.

The phasor sum of magnetizing current Im and the working current Iw gives the no-load current I0.

Phasor Diagram of Transformer on Inductive Load

The phasor diagram of the actual transformer when it is loaded inductively is shown below:



Steps to draw the phasor diagram

- Take flux ϕ, a reference
- Induces emf E1 and E2 lags the flux by 90 degrees.
- The component of the applied voltage to the primary equal and opposite to induced emf in the primary winding. E1 is represented by V1'.
- Current I0 lags the voltage V1' by 90 degrees.

- The power factor of the load is lagging. Therefore current I2 is drawn lagging E2 by an angle ϕ2.
- The resistance and the leakage reactance of the windings result in a voltage drop, and hence secondary terminal voltage V2 is the phase difference of E2 and voltage drop.

V2 = E2 – voltage drops
I2 R2 is in phase with I2 and I2X2 is in quadrature with I2.

- The total current flowing in the primary winding is the phasor sum of I1' and I0.
- Primary applied voltage V1 is the phasor sum of V1' and the voltage drop in the primary winding.
- Current I1' is drawn equal and opposite to the current I2

V1 = V1' +voltage drop
I1R1 is in phase with I1 and I1XI is in quadrature with I1.

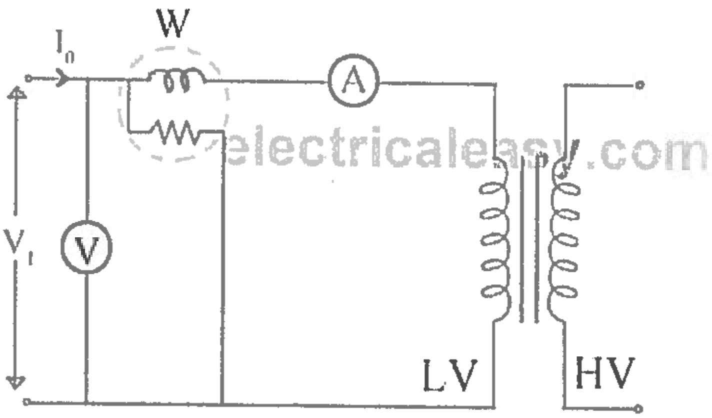
- The phasor difference between V1 and I1 gives the power factor angle ϕ1 of the primary side of the transformer.
- The power factor of the secondary side depends upon the type of load connected to the transformer.
- If the load is inductive as shown in the above phasor diagram, the power factor will be lagging, and if the load is capacitive, the power factor will be leading. Where I1R1 is the resistive drop in the primary windings I2X2 is the reactive drop in the secondary winding

14. Explain the features of OC and SC test of transformer with necessary diagrams. Also mention the advantages of these tests. (12M)

These two transformer tests are performed to find the parameters of equivalent circuit of transformer and losses of the transformer. **Open circuit test and short circuit test on transformer** are very economical and convenient because they are performed without actually loading of the transformer.

Open Circuit Or No Load Test On Transformer

Open circuit test or no load test on a transformer is performed to determine 'no load loss (core loss)' and 'no load current I0'. The **circuit diagram for open circuit test** is shown in the figure below.



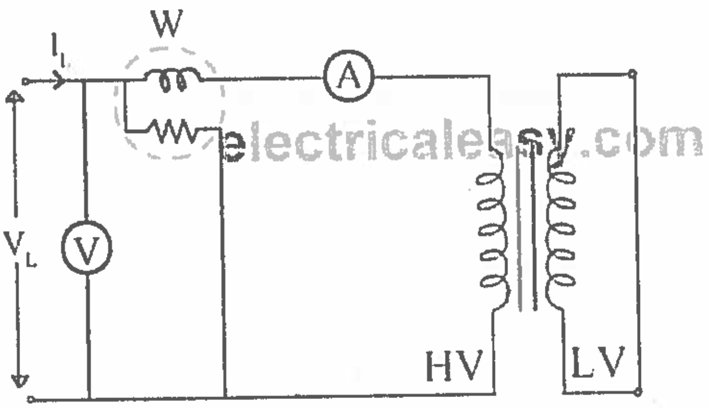
Usually high voltage (HV) winding is kept open and the low voltage (LV) winding is connected to its normal supply. A wattmeter (W), ammeter (A) and voltmeter (V) are connected to the LV winding as shown in the figure. Now, applied voltage is slowly increased from zero to normal rated value of the LV side with the help of a variac. When the applied voltage reaches to the rated value of the LV winding, readings from all the three instruments are taken.

The ammeter reading gives the no load current I0. As I0 itself is very small, the voltage drops due to this current can be neglected.

The input power is indicated by the wattmeter (W). And as the other side of transformer is open circuited, there is no output power. Hence, this input power only consists of core losses and copper losses. As described above, no-load current is so small that these copper losses can be neglected. Hence, now the input power is almost equal to the core losses. Thus, the wattmeter reading gives the core losses of the transformer.

Short Circuit Or Impedance Test On Transformer

The **connection diagram for short circuit test** or impedance test on transformer is as shown in the figure below. The LV side of transformer is short circuited and wattmeter (W), voltmere (V) and ammeter (A) are connected on the HV side of the transformer. Voltage is applied to the HV side and increased from the zero until the ammeter reading equals the rated current. All the readings are taken at this rated current.



Advantages of Tests performed on Transformers:

The above two simple **transformer tests** offer the following advantages:

(i) The power required to carry out these tests is very small as compared to the full-load output of the **transformer.** In case of **open circuit test**, the power required is equal to the iron loss whereas, for a **short-circuit test**, the power required is equal to full-load copper loss.

(ii) These tests enable us to determine the efficiency of the transformer accurately at any load and p.f. without actually loading the transformer.

(iii) The *short-circuit test* enables us to determine R01 and X01 (or R02 and X02). We can thus find the total **voltage drop in the transformer** as referred to primary or secondary.This permits us to **calculate voltage regulation of the transformer.**

15.(a) Discuss the necessity of parallel operation of transformers. Also state the conditions for satisfactory operation of three phase transformer in parallel. (6M)

Parallel operation of three phase transformer is very common in three phase power generation, transmission and distribution. It is advantageous to use two or more Transformer units in parallel instead of using a single large unit. This offers flexibility for maintenance as well as operation.

Advantage of Parallel Operation of Three Phase Transformers

- It increases the reliability of supply system. Let us try to understand how this happens. Suppose a fault occurs in any one of the Transformer unit. In such case, the faulty transformer may be taken out of service while the remaining transformers will feed the power supply. If there were only one large transformer unit is installed for supplying the load, the supply to the entire load will be interrupted during breakdown of the transformer. Thus the reliability of supply system is increased by parallel operation of transformers.
- The size of transformer increases with the increase of its rating. Therefore, a larger transformer will be bigger in size. Therefore, its transportation form manufacturer to the Site will be difficult. Whereas, transportation and installation of small sized transformers are comparatively easy.
- The maintenance opportunity in case of parallel operation is increases. One or more transformers may be taken under maintenance while the remaining transformers will supply the load at reduced power.

Condition for Parallel Operation of Three Phase Transformers

Following are the necessary conditions for parallel operation of 3 phase transformers:

- The line voltage ratio of the transformers must be same.
- The transformers should have equal per unit leakage impedance. (You may read per unit system)
- The ratio of equivalent leakage reactance to equivalent resistance should be same for all the transformers.
- The transformers should have the same polarity.

The above four conditions are also applicable for parallel operation of single phase transformers. Apart from the above four condition, there exists two more conditions which should be fulfilled for parallel operation of three phase transformers:

- The relative phase displacement between the secondary line voltages of all transformers should be zero. This means that transformers to be connected in parallel must belong to same Group number like Yy0 and Dd0 belong to same group number viz. Group 1.
- The phase sequence of secondary line voltages of all the transformers should be same.

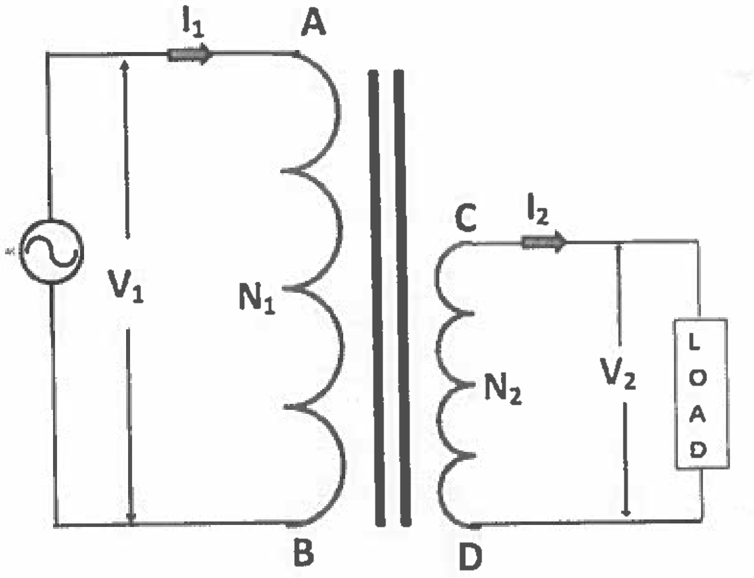
15.(B) Explain the important features of auto transformer (6M)

An **Auto Transformer** is a transformer with only one winding wound on a laminated core. An auto transformer is similar to a two winding transformer but differ in the way the primary and secondary winding are interrelated. A part of the winding is common to both primary and secondary sides.

On load condition, a part of the load current is obtained directly from the supply and the remaining part is obtained by transformer action. An Auto transformer works as a **voltage regulator.**

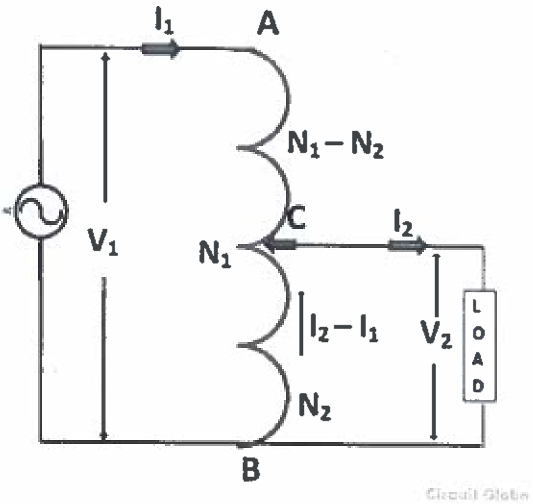
Explanation of Auto Transformer with Circuit Diagram

In an ordinary transformer, the primary and the secondary windings are electrically insulated from each other but connected magnetically as shown in the figure below. While in auto transformer the primary and the secondary windings are connected magnetically as well as electrically. In fact, a part of the single continuous winding is common to both primary and secondary.



Circuit Globe

There are two types of auto transformer based on the construction. In one type of transformer, there is continuous winding with the taps brought out at convenient points determined by the desired secondary voltage. However, in another type of auto transformer, there are two or more distinct coils which are electrically connected to form a continuous winding. The construction of Auto transformer is shown in the figure below.



The primary winding AB from which a tapping at C is taken, such that CB acts as a secondary winding. The supply voltage is applied across AB, and the load is connected across CB. The tapping may be fixed or variable. When an AC voltage V1 is applied across AB, an alternating flux is set up in the core, as a result, an emf E1 is induced in the winding AB. A part of this induced emf is taken in the secondary circuit.

Advantages of Auto transformer

- Less costly
- Better regulation
- Low losses as compared to ordinary two winding transformer of the same rating.

Disadvantages of Auto transformer

There are various advantages of the auto transformer, but then also one major disadvantage, why auto transformer is not widely used, is that

- The secondary winding is not insulated from the primary winding. If an auto transformer is used to supply low voltage from a high voltage and there is a break in the secondary winding, the full primary voltage comes across the secondary terminal which is dangerous to the operator and the equipment. So the auto transformer should not be used for interconnecting high voltage and low voltage systems.
- Used only in the limited places where a slight variation of the output voltage from input voltage is required.

03/01/23

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Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2022 - 2023
Course Code	20EC304	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Random Variables and Stochastic Processes**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Two dice are thrown. What is the probability that the sum on the dice is twelve?	20EC304.1	L1
2	Write Chehyshev's inequality.	20EC304.2	L1
3	Find the mean value of a uniform random variable.	20EC304.3	L2
4	Write $E[X^2(t)]$ in terms of the PSD of $X(t)$.	20EC304.4	L1
5	Write the expression for average noise figure of cascaded networks.	20EC304.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Discuss the significance of a Gaussian random variable using its probability density and distribution functions.	6M	20EC304.1	L2
6 (b)	State and prove Bayes theorem.	6M	20EC304.1	L2
	OR			
7 (a)	Explain the distribution and density functions of Rayleigh random variable with neat sketches.	6M	20EC304.1	L2
7 (b)	Define conditional probability distribution function and write the properties.	6M	20EC304.1	L1
8 (a)	State and prove the properties of variance of a random variable.	6M	20EC304.2	L3
8 (b)	Find the variance of a uniform random variable distributed over the interval [a, b].	6M	20EC304.2	L3
	OR			
9 (a)	A random variable X is uniformly distributed on the interval (-5, 15). Another random variable $Y=e^{x/5}$ is formed. Find E[Y].	6M	20EC304.2	L3
9 (b)	If X discrete is a random variable with probability mass function given as below table	6M	20EC304.2	L2

X	-4	-2	0	2	4
P(X)	1/5	2/5	1/10	1/15	1/5

Solve i) E[X]
ii) $E[X^2]$
iii) E[2X+3]
iv) $E[(2X+1)^2]$

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
10 (a)	If X and Y are independent, show that E[XY]=E[X] E[Y].	6M	20EC304.3	L3
10 (b)	State and prove central limit theorem for equal distributions case.	6M	20EC304.3	L3
	OR			
11 (a)	X and Y are two independent random variables related to W as W= X+Y. Find $f_W(w)$ in terms of $f_X(x)$ and $f_Y(y)$.	6M	20EC304.3	L3
11 (b)	If X and Y are two independent random variables, then $Ø_{X+Y}(\omega) = Ø_X(\omega)Ø_Y(\omega)$	6M	20EC304.3	L3

12 (a)	Explain about Poisson random processes.	6M	20EC304.4	L2
12 (b)	State any four properties of power spectral density and cross power spectral density.	6M	20EC304.4	L1
	OR			
13 (a)	A random process had the power density spectrum $S_{xx}(w)=6w^2/1+w^2$, find the average power in the process.	6M	20EC304.4	L2
13 (b)	State and prove the relationship between Power Density Spectrum and auto correlation function.	6M	20EC304.4	L3
14 (a)	Write all the properties of narrow band noise.	6M	20EC304.5	L1
14 (b)	Define convolution. List any four properties of convolution.	6M	20EC304.5	L1
	OR			
15 (a)	Explain the following i) Noise Figure ii) Noise Sources	6M	20EC304.5	L2
15 (b)	Explain the thermal noise.	6M	20EC304.5	L2

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(AUTONOMOUS)
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SCHEME OF EVALUATION

PART – A

1. Problem – 2m
2. Chebyshev's inequality statement – 2m
3. mean value of uniform r.v – 2m
4. write $E[x^2(t)]$ in terms of the PSD of $x(t)$ – 2m
5. Expression for Cascaded network – 2m

PART – B :

6. (a) Gaussian r.v derivation – 6m
   (b) State & prove Baye's theorem – 6m

7. (a) Explanation of Rayleigh distribution & density function – 6m
   (b) Properties of Condition distribution function – 6m

8. (a) Properties of Variance – 6m
   (b) Variance of Uniform r.v – 6m

9. (a) Problem – 6m
   (b) Problem – 6m

10. (a) S.T $E[XY] = E[X]\,E[Y]$ – 6m
    (b) Central limit Theorem for equal distributions – 6m

11. (a) Sum of two random variables — 6m
Explanation.

(b) S.T $\phi_{X+Y}(\omega) = \phi_X(\omega)\,\phi_Y(\omega)$.

12. (a) Poisson random Processes — 6m

(b) Properties ~~of~~ ~~[illegible]~~ — 6m

13. (a) Problem — 6m

(b) Relationship b/w PSD & autocorrelation — 6m

14. (a) Properties of Narrow Band Noise — 6m

(b) Definition of Convolution — 2m
Properties — 4m } 6m

15. (a) (i) Noise figure — 3m
(ii) Noise Sources — 3m } 6m

(b) Thermal Noise — 6m.

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(AUTONOMOUS)

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ANSWER KEY

PART – A

1. Two dice are thrown. what is the probability that the sum on the dice is twelve?

Ans:

we know that , two dice are thrown

$$n(S) = \left\{\begin{matrix} (1,1) & (1,2) & (1,3) & (1,4) & (1,5) & (1,6) \\ (2,1) & (2,2) & (2,3) & (2,4) & (2,5) & (2,6) \\ (3,1) & (3,2) & (3,3) & (3,4) & (3,5) & (3,6) \\ (4,1) & (4,2) & (4,3) & (4,4) & (4,5) & (4,6) \\ (5,1) & (5,2) & (5,3) & (5,4) & (5,5) & (5,6) \\ (6,1) & (6,2) & (6,3) & (6,4) & (6,5) & (6,6) \end{matrix}\right\} = 36$$

Total outcomes are $n(S) = 36$

Now, E be the event that the sum on the dice is twelve

$$n(E) = \{(6,6)\} = 1$$

$\therefore$ The probability of sum on the dice is twelve

$$P(E) = \frac{n(E)}{n(S)} = \frac{1}{36} \,//$$

2. Write Chebyshev's Inequality.

Ans: Chebyshev's Inequality :-

If x is a random variable with mean and variance σ^2

then (i) $P\{|x-\mu| \geq k\sigma\} \leq \frac{1}{k^2}$

(ii) $P\{|x-\mu| < k\sigma\} \geq 1-\frac{1}{k^2}$

3. Find the mean value of a uniform random variable

Ans! We know that,

Uniform random variable, $f_X(x) = \frac{1}{b-a}$ for $a \le x \le b$

$= 0$, otherwise

$$\text{Mean},\ E[x] = \int_{-\infty}^{\infty} x\, f_X(x)\, dx$$

$$= \int_{-\infty}^{a} x\, f_X(x)\, dx + \int_{a}^{b} x\, f_X(x)\, dx + \int_{b}^{\infty} x\, f_X(x)\, dx$$

$$= 0 + \int_{a}^{b} x \cdot \frac{1}{b-a}\, dx + 0$$

$$= \frac{1}{b-a} \int_{a}^{b} x\, dx$$

$$= \frac{1}{b-a}\left[\frac{x^2}{2}\right]_a^b = \frac{1}{2}\frac{1}{b-a}\left[b^2 - a^2\right]$$

$$= \frac{1}{2}\frac{1}{\cancel{b-a}}(\cancel{b-a})(b+a)$$

$$\therefore \boxed{E[x] = \frac{a+b}{2}}$$

4. Write $E[x^2(t)]$ interms of the PSD of $X(t)$.

Ans! If $S_{XX}(\omega)$ is a power spectral density (psd) of the WSS random process $x(t)$, then $\frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\, d\omega = A\{E[x^2(t)]\} = R_{XX}(0)$

(or)

the time average of the mean square value of a WSS random process equals the area under the curve of the power spectral density

5. Write the expression for average noise figure of cascaded networks.

Ans: Expression for Average noise figure of cascaded networks

$$\boxed{F = F_1 + \frac{F_2 - 1}{g_1}}$$

This is the required expression of the noise figure of cascade system having '2' stages.

ANSWER KEY AND SCHEME OF EVALUATION

PART – B

6.(a) Discuss the significance of a gaussian random variable using its probability density and distribution functions.

Ans:- Significance of Gaussian random variables.

A Gaussian random variable is one whose probability density function can be written in general form. The PDF of the Gaussian random variable has two parameters, $\underset{(or)\,a_x}{m}$ and σ, which have the interpretation of the mean and standard deviation respectively.

The Gaussian density function and distribution function of a random variable X are given by

$$f_X(x) = \frac{1}{\sqrt{2\pi\sigma_X^2}} \exp\left(\frac{-(x-a_X)^2}{2\sigma_X^2}\right) \quad \text{for all } x$$

and

$$F_X(x) = \frac{1}{\sqrt{2\pi\sigma_X^2}} \int_{-\infty}^{x} \exp\left(\frac{-(x-a_X)^2}{2\sigma_X^2}\right) dx \quad \text{for all } x$$

where $\sigma_X > 0$ and $-\infty < a_X < \infty$ are constants called standard deviation and mean value of X respectively.

The Corresponding CDF of gaussian random variable

if $$F_X(x) = \int_{-\infty}^{x} f_X(x)\,dx$$

$$= \int_{-\infty}^{x} \frac{1}{\sqrt{2\pi\sigma_X^2}}\, e^{\frac{-(x-a_X)^2}{2\sigma_X^2}}\, dx$$

The solution of this integration can be in two different forms.

(i) $$F_X(x) = P(X \le x) = 1 - P(X > x)$$

$$= 1 - \int_{x}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}}\, e^{\frac{-(x-a_X)^2}{2\sigma_X^2}}\, dx$$

$$= 1 - \int_{x}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}}\, e^{-\frac{1}{2}\left(\frac{x-a_X}{\sigma_X}\right)^2} dx$$

Let $\frac{x-a_X}{\sigma_X} = t$

$x - a_X = \sigma_X t$

$dx = \sigma_X\, dt$

Let $x = x \Rightarrow t = \frac{x-a_X}{\sigma_X}$

$x = \infty \Rightarrow t = \infty$

Substitute all the values in above eqn.

$$= 1 - \int_{\frac{x-a_X}{\sigma_X}}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}}\, e^{-\frac{1}{2}t^2} * \sigma_X\, dt$$

$$= 1 - \frac{1}{\sqrt{2\pi}} \times \frac{1}{\cancel{\sigma_X}} (\cancel{\sigma_X}) \int_{\frac{x-a_X}{\sigma_X}}^{\infty} e^{-t^2/2}\, dt$$

$$= 1 - \frac{1}{\sqrt{2\pi}} \int_{\frac{x-a_X}{\sigma_X}}^{\infty} e^{-\frac{t^2}{2}}\, dt \rightarrow ①$$

We have $Q(k) = \frac{1}{\sqrt{2\pi}} \int_{k}^{\infty} e^{-t^2/2}\, dt$ named as 'Q' function.

from ①

$$Q\left(\frac{x-a_X}{\sigma_X}\right) = \frac{1}{\sqrt{2\pi}} \int_{\frac{x-a_X}{\sigma_X}}^{\infty} e^{-t^2/2}\, dt$$

$$\therefore \boxed{F_X(x) = 1 - Q\left(\frac{x-a_X}{\sigma_X}\right)}$$

(ii) $F_X(x) = P(X \le x) = 1 - P(X > x)$

$$= 1 - \int_{x}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}} e^{-\frac{(x-a_X)^2}{2\sigma_X^2}}\, dx$$

$$= 1 - \int_{x}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}} e^{-\left(\frac{x-a_X}{\sqrt{2}\sigma_X}\right)^2} dx$$

$$= 1 - \frac{1}{\sqrt{2\pi}\sigma_X} \int_{x}^{\infty} e^{-\left(\frac{x-a_X}{\sqrt{2}\sigma_X}\right)^2} dx .$$

let $\frac{x-a_X}{\sqrt{2}\sigma_X} = t$

$x - a_X = \sqrt{2}\sigma_X t$

$dx = \sqrt{2}\sigma_X\, dt$

if $x = x \Rightarrow t = \frac{x-a_X}{\sqrt{2}\sigma_X}$

$x = \infty \Rightarrow t = \infty$

Substitute above values in eq^n ●

$$= 1 - \int_{\frac{x-a_X}{\sqrt{2}\sigma_X}}^{\infty} \frac{1}{\sqrt{2\pi\sigma_X^2}}\, e^{-t^2} \sqrt{2}\,\sigma_X\, dt$$

$$= 1 - \frac{1}{\sqrt{2\pi}\,\cancel{\sigma_X}} \sqrt{2}\,\cancel{\sigma_X} \int_{\frac{x-a_X}{\sqrt{2}\sigma_X}}^{\infty} e^{-t^2}\, dt$$

$$= 1 - \frac{1}{\cancel{\sqrt{2}}\sqrt{\pi}}\, \cancel{\sqrt{2}} \int_{\frac{x-a_X}{\sqrt{2}\sigma_X}}^{\infty} e^{-t^2}\, dt$$

$$= 1 - \frac{1}{2} \cdot \frac{2}{\sqrt{\pi}} \int_{\frac{x-a_X}{\sqrt{2}\sigma_X}}^{\infty} e^{-t^2}\, dt \rightarrow ②$$

We have the Complementory error function

$$\operatorname{erfc}(K) = \frac{2}{\sqrt{\pi}} \int_{k}^{\infty} e^{-t^2}\, dt$$

$$\boxed{F_X(x) = 1 - \frac{1}{2} \operatorname{erfc}\left(\frac{x-a_X}{\sqrt{2}\,\sigma_X}\right)}$$

(b) State and Prove Baye's theorem

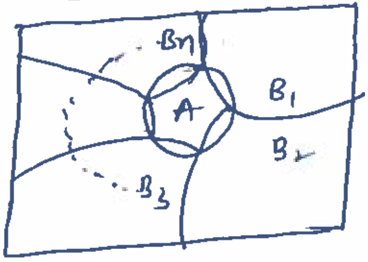
Ans: **Baye's theorem :-**

If $B_1, B_2, B_3, \ldots\ldots B_n$ are n mutually exclusive and exhaustive events such that $P(B_i) > 0$ $(i = 1, 2, 3, \ldots n)$ in a sample space S and A is any other event in S intersecting with every B_i (i.e A can only occur in combination with any one of the events $B_1, B_2, B_3, \ldots B_n$) such that $P(A) > 0$, then the Conditional probability of B_i given A is

$$P\left(\frac{B_i}{A}\right) = \frac{P(B_i)\, P(A/B_i)}{\sum_{i=1}^{n} P(B_i)\, P(A/B_i)} \quad , \quad i = 1, 2, 3, \ldots n$$

Proof :- we know that

$$A = A \cap S \quad [\because S \text{ is a Sample Space}]$$



$$A = A \cap [B_1 \cup B_2 \cup B_3 \cup \ldots \ldots \cup B_n]$$

$$A = (A \cap B_1) \cup (A \cap B_2) \cup (A \cap B_3) \cup \ldots \ldots \cup (A \cap B_n)$$

$$P(A) = P[(A \cap B_1) \cup (A \cap B_2) \cup (A \cap B_3) \cup \ldots \ldots \cup (A \cap B_n)]$$

Since $B_1, B_2, B_3, \ldots$ are mutually exclusive events

Hence $A \cap B_1, A \cap B_2, \ldots \ldots A \cap B_n$ are mutually exclusive events

$$P(A) = P(A \cap B_1) \cup P(A \cap B_2) \cup P(A \cap B_3) \cup \ldots \ldots \cup P(A \cap B_n)$$

$$P(A) = P(A \cap B_1) + P(A \cap B_2) + P(A \cap B_3) + \ldots \ldots + P(A \cap B_n)$$

$$P(A) = P(B_1)\, P(A/B_i) + P(B_2)\, P(A/B_2) + \ldots \ldots + P(B_n)\, P(A/B_n)$$

$$P(A) = \sum_{i=1}^{n} P(B_i)\, P(A/B_i)$$

Now

$$P\left(\frac{B_i}{A}\right) = \frac{P(B_i \cap A)}{P(A)}$$

$$= \frac{P(B_i)\, P(A/B_i)}{P(A)}$$

$$\boxed{P\left(\frac{B_i}{A}\right) = \frac{P(B_i)\, P(A/B_i)}{\sum_{i=1}^{n} P(B_i)\, P(A/B_i)}}$$

7. (a) Explain the distribution and density functions of Rayleigh random variable with neat sketches.

Ans:

Rayleigh density function :-

The Rayleigh density function of a random variable 'x' is given by

$$f_X(x) = \begin{cases} \frac{2}{b}(x-a)\, e^{-\frac{(x-a)^2}{b}} & , x \geq a \\ 0 & , x < a \end{cases}$$

where 'a' and 'b' are real values such that $-\infty < a < \infty, b > 0$

Rayleigh Distribution function :-

$$F_X(x) = \int_{-\infty}^{x} f_X(x)\,dx$$

$$= \int_{-\infty}^{a} f_X(x)\,dx + \int_{a}^{x} f_X(x)\,dx$$

$$= 0 + \int_{a}^{x} \frac{2}{b}(x-a)\, e^{-\frac{(x-a)^2}{b}}\,dx$$

$$= \frac{2}{b}\int_{a}^{x} (x-a)\, e^{-\frac{(x-a)^2}{b}}\,dx \rightarrow ①$$

$$\text{let } \frac{(x-a)^2}{b} = t$$

$$(x-a)^2 = bt$$

$$2(x-a)(1)\,dx = b\,dt$$

$$dx = \frac{b}{2(x-a)}\,dt$$

$$\text{if } x = a \Rightarrow t = 0$$

$$x = x \Rightarrow t = \frac{(x-a)^2}{b}$$

$$= \frac{2}{b}\int_{0}^{\frac{(x-a)^2}{b}} \cancel{(x-a)}\, e^{-t} \times \frac{b}{2\cancel{(x-a)}}\,dt$$

$$= \cancel{\frac{2}{b}} \times \cancel{\frac{b}{2}} \int_0^{\frac{(x-a)^2}{b}} e^{-t} dt$$

$$= \left[\frac{e^{-t}}{-1}\right]_0^{\frac{(x-a)^2}{b}}$$

$$= - \left[e^{-\frac{(x-a)^2}{b}} - e^0 \right]$$

$$= - \left[e^{-\frac{(x-a)^2}{b}} - 1 \right]$$

$$\therefore \boxed{F_X(x) = 1 - e^{-\frac{(x-a)^2}{b}}}$$

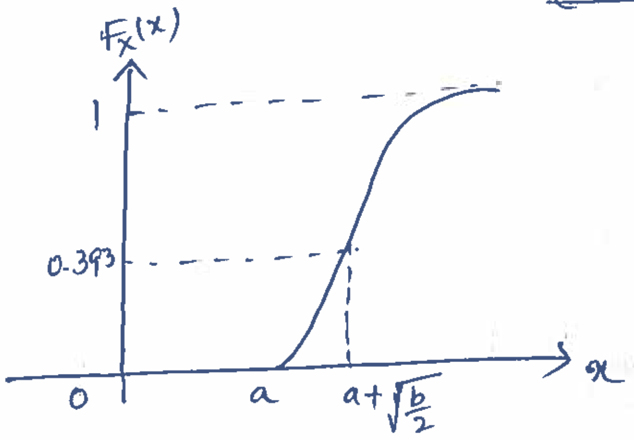
If $x = a$, $F_X(x) = 1 - e^0 = 1 - 1 = 0 \Rightarrow \boxed{F_X(x) = 0}$

If $x = a + \sqrt{\frac{b}{2}}$, $F_X(x) = 1 - e^{-\frac{\left[\cancel{a} + \sqrt{\frac{b}{2}} - \cancel{a}\right]^2}{b}}$

$$= 1 - e^{-\frac{\cancel{b}/2}{\cancel{b}}}$$

$$= 1 - e^{-1/2} = 1 - 0.607$$

$$\boxed{F_X(x) = 0.393}$$



(b) Define Conditional Probability distribution function and write the properties.

Ans: **Conditional Probability Distribution function :-**

Let A and B are two events. If A is an event $\{X \le x\}$ for random variable X, then the conditional distribution function of X, when the event B is known is denoted as $F_X(x/B)$ and is defined as

$$\boxed{F_X(x/B) = P\{X \le x/B\}}$$

we know that the conditional probability $P(A/B) = \frac{P(A \cap B)}{P(B)}$

Then

$$\boxed{F_X(x/B) = \frac{P(X \le x \cap B)}{P(B)}}$$

Properties of Conditional Distribution Functions :-

The properties of a conditional distribution function $F_X(x/B)$ are given below.

1. $F_X(-\infty/B) = 0$
2. $F_X(\infty/B) = 1$
3. $0 \le F_X(x/B) \le 1$
4. $F_X(x_1/B) \le F_X(x_2/B)$ if $x_1 < x_2$
5. $P\{x_1 < X \le x_2/B\} = F_X(x_2/B) - F_X(x_1/B)$
6. $F_X\{x^*/B\} = F_X(x/B)$.

8. (a) State and prove the properties of variance of a random variable.

Ans: Properties of Variance :-

1. The variance of a constant is zero, i.e if K is constant, then $var(K) = 0$

Proof:

$$Var(x) = E[(x-\bar{x})^2]$$

$$Var(K) = E[(K-\bar{K})^2] = E[(K-K)^2] = E[0]$$

$$\boxed{Var(K) = 0}$$

2. If K is a constant, then for a random variable X

$$Var(KX) = K^2 Var(X)$$

Proof: The variance of Kx is given by

$$\begin{aligned} Var(Kx) &= E[(Kx - K\bar{x})^2] \\ &= E[K^2 (x-\bar{x})^2] \\ &= K^2 E[(x-\bar{x})^2] \end{aligned}$$

$$\boxed{Var(Kx) = K^2 Var(x)}$$

3. For a given random variable x, the relationship between the variance and the moments is given by

$$\sigma_x^2 = m_2 - m_1^2$$

Proof: Let x be a random variable, The variance is

$$\begin{aligned} \sigma_x^2 &= E[(x-\bar{x})^2] \\ &= E[x^2 + \bar{x}^2 - 2x\bar{x}] \\ &= E[x^2] + E[\bar{x}^2] - 2E[\bar{x}x] \\ &= E[x^2] + \bar{x}^2 - 2\bar{x}E[x] \\ &= E[x^2] + \bar{x}^2 - 2\bar{x}\bar{x} \\ &= E[x^2] + \bar{x}^2 - 2\bar{x}^2 \\ \sigma_x^2 &= E[x^2] - \bar{x}^2 \Rightarrow \boxed{\sigma_x^2 = m_2 - m_1^2} \end{aligned}$$

4. If x is a random variable and a, b are real Constants, then

$$Var(ax+b) = a^2 Var(x)$$

Proof:- The variance of $(ax+b)$ is

$$\text{Variance}(ax+b) = E\left[\left((ax+b) - (\overline{ax+b})\right)^2\right]$$

W.K.T $\quad \overline{ax+b} = a\,E[x] + b$

$$Var(ax+b) = E\left[(ax+\cancel{b} - a\,E[x] - \cancel{b})^2\right]$$

$$= E\left[(ax - a\,E[x])^2\right]$$

$$= E\left[a^2(x-\bar{x})^2\right]$$

$$= a^2\,E\left[(x-\bar{x})^2\right]$$

$$\boxed{Var(ax+b) = a^2 Var(x)}$$

5. If two random variables x_1 and x_2 are independent, then $Var(x_1+x_2) = Var(x_1) + Var(x_2)$ and $Var(x_1-x_2) = Var(x_1)+Var(x_2)$.

Proof:- $$Var(x_1+x_2) = E\left[\left((x_1+x_2 - (\overline{x_1+x_2})\right)^2\right]$$

W.K.T $\cancel{\phi}$ $\overline{x_1+x_2} = \bar{x}_1 + \bar{x}_2$

So, $$Var(x_1+x_2) = E\left[(x_1+x_2 - \bar{x}_1 - \bar{x}_2)^2\right]$$

$$= E\left[\left((x_1-\bar{x}_1) + (x_2-\bar{x}_2)\right)^2\right]$$

$$= E\left[(x_1-\bar{x}_1)^2 + (x_2-\bar{x}_2)^2 + 2(x_1-\bar{x}_1)(x_2-\bar{x}_2)\right]$$

$$= E\left[(x_1-\bar{x}_1)^2\right] + E\left[(x_2-\bar{x}_2)^2\right] + 2E\left[(x_1-\bar{x}_1)(x_2-\bar{x}_2)\right]$$

Since x_1 and x_2 are independent

$$E\left[(x_1-\bar{x}_1)(x_2-\bar{x}_2)\right] = E\left[(x_1-\bar{x})\right]E\left[x_2-\bar{x}_2\right] = (\bar{x}_1 - \bar{x}_1)(\bar{x}_2-\bar{x}_2) = 0$$

Therefore $\boxed{Var(x_1+x_2) = Var(x_1) + Var(x_2)}$

Similarly, $$Var(x_1-x_2) = Var(x_1) + Var(-x_2)$$

$$= Var(x_1) + (-1)^2 Var(x_2)$$

$$\boxed{Var(x_1-x_2) = Var(x_1) + Var(x_2)}$$

(b) Find the Variance of a uniform random Variable distributed over the interval $[a, b]$.

Ans:

We Know that,

$$\text{Variance } \sigma^2 = E[x^2] - [E[x]]^2$$

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

$$E[x^2] = \int_{-\infty}^{a} x^2 f_x(x)dx + \int_{a}^{b} x^2 f_x(x)dx + \int_{b}^{\infty} x^2 f_x(x)\,dx$$

$$= 0 + \int_a^b x^2 f_x(x)dx + 0$$

$$= \int_a^b x^2 \left(\frac{1}{b-a}\right)dx$$

$$= \frac{1}{b-a}\int_a^b x^2 dx$$

$$= \frac{1}{b-a}\left[\frac{x^3}{3}\right]_a^b$$

$$= \frac{1}{3}\cdot\frac{1}{b-a}\left[b^3 - a^3\right]$$

$$= \frac{1}{3}\cdot\frac{1}{\cancel{b-a}}\,(\cancel{b-a})\,(b^2 + a^2 + ab)$$

$$\boxed{E[x^2] = \frac{a^2 + b^2 + ab}{3}}$$

$$\text{Variance } \sigma^2 = E[x^2] - [E[x]]^2$$

$$= \frac{a^2+b^2+ab}{3} - \left(\frac{a+b}{2}\right)^2 \qquad \left[\because E[x] = \frac{a+b}{2}\right]$$

$$= \frac{a^2+b^2+ab}{3} - \frac{a^2+b^2+2ab}{4}$$

$$= \frac{4(a^2+b^2+ab) - 3(a^2+b^2+2ab)}{12}$$

$$= \frac{4a^2+4b^2+4ab-3a^2-3b^2-6ab}{12}$$

$$\sigma^2 = \frac{a^2+b^2-2ab}{12} \Rightarrow \boxed{\sigma^2 = \frac{(a-b)^2}{12} \text{ (or) } \frac{(b-a)^2}{12}}$$

9.X(a). A random variable X is uniformly distributed on the interval (−5,15). Another random variable $Y = e^{X/5}$ is formed. Find E[Y].

Ans: Given X is a uniformly distributed in (−5,15).

$$f_X(x) = \begin{cases} \frac{1}{20} , & -5 < x < 15 \\ 0 , & \text{otherwise} \end{cases}$$

$$\left[\because f_X(x) = \frac{1}{b-a} , a < x < b \\ = 0, \text{ otherwise} \right]$$

Now $E[Y] = E[e^{x/5}] = \int_{-\infty}^{\infty} e^{-x/5} f_X(x) dx$

$$= \int_{-\infty}^{-5} e^{-x/5} f_X(x) dx + \int_{-5}^{15} e^{-x/5} f_X(x) dx + \int_{15}^{\infty} e^{-x/5} f_X(x) dx$$

$$= 0 + \int_{-5}^{15} e^{-x/5} \left(\frac{1}{20}\right) dx + 0$$

$$= \frac{1}{20} \int_{-5}^{15} e^{-x/5} dx$$

$$= \frac{1}{20} \left(\frac{e^{-x/5}}{-1/5} \right]_{-5}^{15}$$

$$= \frac{1}{20} \times -\frac{5}{1} \left[e^{-15/5} - e^{-5/-5} \right]$$

$$= -\frac{1}{4} \left[e^{-3} - e^{1} \right]$$

$$= \frac{1}{4} \left[e - e^{-3} \right]$$

$$\therefore \boxed{E[Y] = 0.2378}$$

9. (b) If x discrete is a random Variable with probability mass function given as below table.

X	-4	-2	0	2	4
$P(x)$	$1/5$	$2/5$	$1/10$	$1/15$	$1/5$

Solve (i) $E[x]$ (ii) $E[x^2]$ (iii) $E[2x+3]$ (iv) $E[(2x+1)^2]$.

Ans: Given data,

X	-4	-2	0	2	4
$P(x)$	$\frac{1}{5}$	$\frac{2}{5}$	$\frac{1}{10}$	$\frac{1}{15}$	$\frac{1}{5}$

Now find (i) $E[x] = \sum_{i=1}^{n} x_i\, p(x_i)$

$$= \sum_{i=1}^{5} x_i\, p(x_i)$$

$$= \cancel{-4\left(\tfrac{1}{5}\right)} - 2\left(\tfrac{2}{5}\right) + 0\left(\tfrac{1}{10}\right) + 2\left(\tfrac{1}{15}\right) + \cancel{4\left(\tfrac{1}{5}\right)}$$

$$= -\frac{4}{5} + \frac{2}{15} = \frac{-12+2}{15} = \frac{-10}{15}$$

$$\therefore \boxed{E[x] = \frac{-10}{15}} \text{ (or) } \boxed{E[x] = -0.6666}$$

(ii) $E[x^2] = \sum_{i=1}^{n} x_i^2\, p(x_i) = \sum_{i=1}^{5} x_i^2\, p(x_i)$

$$E[x^2] = (-4)^2\left(\tfrac{1}{5}\right) + (-2)^2\left(\tfrac{2}{5}\right) + (0)^2\left(\tfrac{1}{10}\right) + (2)^2\left(\tfrac{1}{15}\right) + (4)^2\left(\tfrac{1}{5}\right)$$

$$= \frac{16}{5} - \frac{8}{5} + 0 + \frac{4}{15} + \frac{16}{5}$$

$$= \frac{48 - 24 + 4 + 48}{15}$$

$$= \frac{76}{15}$$

$$\therefore \boxed{E[x^2] = 5.0666}$$

(iii) $E[2X+3] = 2E[X]+3$

$$= 2\left(\frac{-10}{15}\right)+3 = \frac{-20}{15}+3$$

$$= -1.333+3$$

$$\boxed{E[2X+3] = 1.6666}$$

(iv) $E[(2X+1)^2] = E[(2X)^2+1^2+2\cdot 2X\cdot 1]$

$$= E[4X^2+1+4X]$$

$$= 4E[X^2]+E[1]+4E[X]$$

$$= 4(5.0666)+1+4(-0.6666)$$

$$= 23.6664 - 2.6664$$

$$\boxed{E[(2X+1)^2] = 21}$$

10. (a) If X and Y are independent, Show that $E[XY] = E[X]E[Y]$

Ans: If two random variables X and Y are statistically independent, then X and Y are said to be uncorrelated.

i.e $R_{XY} = E[XY] = E[X]\,E[Y]$.

Proof :- Consider two r.v's X and Y with joint density function $f_{XY}(x,y)$ and marginal density functions $f_X(x)$ and $f_Y(y)$

If X and Y are statistically independent, then we know that $f_{XY}(x,y) = f_X(x)\cdot f_Y(y)$.

The Corelation is

$$R_{XY} = E[XY] = \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} xy\, f_{XY}(x,y)\, dx\, dy$$

$$= \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} xy \; f_X(x)\, f_Y(y)\, dx\,dy$$

$$= \int_{-\infty}^{\infty} x\, f_X(x)\, dx \int_{-\infty}^{\infty} y\, f_Y(y)\, dy$$

$$\boxed{E[XY] = E[X]\, E[Y]}$$

(b) State and prove central Limit theorem for equal distributions case.

Ans: Central Limit Theorem :

According to the Central limit theorem (CLT), the distribution of random process which is cumulative effect of a large number of independent noise sources can be assumed to be Gaussian.

For example in Communication Systems, the noise is always modelled as a random variable with gaussian distribution. This is a valid assumption since the noise in a Communication System is the cumulative effect of many random noise sources.

Equal Distribution :-

Consider N Continuous random variables $X_n, n=1,2,3 ----N$ have the same distribution and density functions

Let $Y = X_1 + X_2 + X_3 + ----+ X_n$.

Also let w be normalized random Variable

i-e $w = \frac{Y - \overline{Y}}{\sigma_Y}$

where $Y = \sum_{n=1}^{N} x_n$, $\overline{Y} = \sum_{n=1}^{N} \overline{x_n}$ and $\sigma_Y^2 = \sum_{n=1}^{N} \sigma_{x_n}^2$

$$\text{So} \quad W = \frac{\sum_{n=1}^{N} x_n - \sum_{n=1}^{N} \overline{x_n}}{\left(\sum_{n=1}^{N} \sigma_{x_n}^2\right)^{1/2}}$$

Since all random variables have the same distribution

$$\sigma_{x_n}^2 = \sigma_x^2 \;/ \left(\sum_{n=1}^{N} \sigma_x^2\right)^{1/2} = \left(N \sigma_x^2\right)^{1/2}$$

$$= \sqrt{N \sigma_X^2}$$

$$= \sqrt{N}\, \sigma_X$$

and $\overline{x_n} = \overline{x}$

$$\therefore W = \frac{1}{\sqrt{N}\, \sigma_x} \sum_{n=1}^{N} (x_n - \overline{x})$$

Then w is a Gaussian Random Variable.

11.(a) X and Y are two independent random Variables related to W as $W = X + Y$, find $f_W(w)$ interms of $f_X(x)$ and $f_Y(y)$.

Ans: Let 'x' represent the information signal voltage and Y represent the noise signal voltage at some instant of time, then sum of two independent random voltages 'x' and 'Y' available at receiver i-e

$$W = X + Y \rightarrow ①$$

By deffinition of probability distribution function,

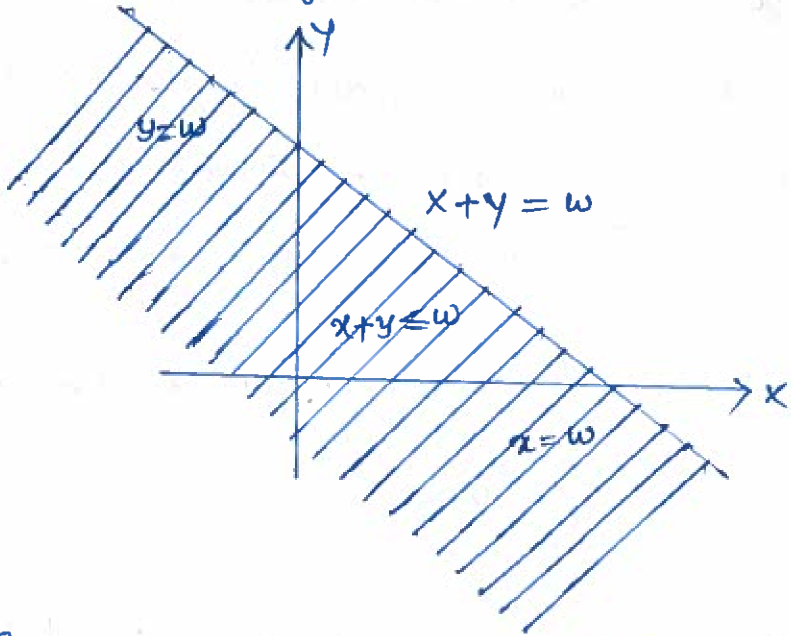
$$F_W(\omega) = P\{W \leq \omega\}$$

$$F_W(\omega) = P\{x+y \leq 3\} \quad [\because \text{from}(1)]$$

To obtain $F_W(\omega)$, all the probabilities over the region of $x+y \leq \omega$ are integrated as

$$F_W(\omega) = \int_{y=-\infty}^{\infty} \int_{x=-\infty}^{\omega-y} f_{xy}(x,y)\,dx\,dy \rightarrow ②$$

The above equation ② can be graphically represented as the shaded region below the line $x+y=\omega$



Since 'x' and 'y' are independent random variables eq ② can be written as

$$F_W(\omega) = \int_{-\infty}^{\infty} \int_{-\infty}^{\omega-y} f_X(x)\, f_Y(y)\,dx\,dy$$

$$= \int_{-\infty}^{\infty} f_Y(y)\,dy \int_{-\infty}^{\omega-y} f_X(x)\,dx$$

$$F_W(\omega) = \int_{-\infty}^{\infty} f_Y(y)\,dy\; F_X(\omega-y) \quad \left[\because \int_{-\infty}^{x} f_X(x)\,dx = F_X(x)\right]$$

derivative of distribution function is given by

$$\frac{d}{dw}\left[F_W(w)\right] = \frac{d}{dw}\int_{-\infty}^{\infty} f_Y(y)dy \; F_X(w-y)$$

$$f_W(w) = \int_{-\infty}^{\infty} f_Y(y)dy \; \frac{d}{dw} F_X(w-y)$$

$$f_W(w) = \int_{-\infty}^{\infty} f_Y(y)dy \; f_X(w-y)$$

$$\therefore f_W(w) = \int_{-\infty}^{\infty} f_Y(y)\, f_X(w-y)\, dy$$

∴ The above equation represents density function of Sum of two independent random variables.

It can be concluded that density function of Sum of two Independent random variables is the convolution of their Independent density functions.

i.e $f_W(w) = \int_{-\infty}^{\infty} f_Y(y)\, f_X(w-y)\, dy = f_Y(y) * f_X(x)$.

(b) If X and Y are two independent random variables then $\phi_{X+Y}(w) = \phi_X(w)\,\phi_Y(w)$.

Ans:- Consider two random Variables X and Y. The Characteristic function of the sum of the random Variables is

$$\phi_{X+Y}(w) = E\left[e^{j(X+Y)w}\right]$$

$$= \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} e^{j(x+y)w} f_{XY}(x,y)\, dx\, dy$$

Since X and Y are independent, $f_{XY}(x,y) = f_X(x) f_Y(y)$

Then $\phi_{X+Y}(\omega) = \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} e^{jx\omega + jy\omega} f_X(x) f_Y(y) dx dy$

$$= \int_{-\infty}^{\infty}\int_{-\infty}^{\infty} e^{jx\omega} \cdot e^{jy\omega} f_X(x) f_Y(y) dx dy$$

$$= \int_{-\infty}^{\infty} e^{jx\omega} f_X(x) dx \int_{-\infty}^{\infty} e^{jy\omega} f_Y(y) dy$$

$$= \phi_X(\omega) \phi_Y(\omega)$$

$$\therefore \boxed{\phi_{X+Y}(\omega) = \phi_X(\omega) \phi_Y(\omega)}$$

~~Given~~

12.(a) Explain about Poisson random Processes.

Ans:

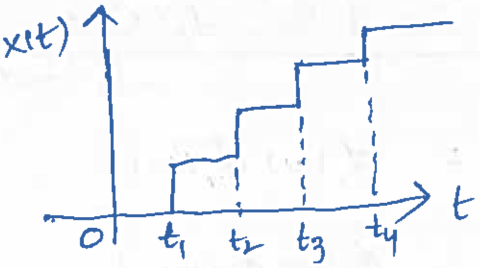
Poisson random Processes:

The poisson process X(t) is a discrete random process which represents the number of times that some event has occured as a function of time. X(t) has integer valued, non-decreasing sample functions, such as check in registers, arrival of a customer, arrival of vehicles at a particular point etc.

In these functions, a single event occurs at a random time. counting the number of occurances with time is a poisson process. It is, therefore also called a counting process.

The Conditions for a poisson process $x(t)$ are

1. $x(0) = 0$
2. Only one event occurs in any instant of time, i.e in an infinitesimal time interval.
3. The number of events that occur in any given time interval is independent increments.



12(b) State any four properties of power Spectral density and Cross power spectral density.

Ans:- <u>Properties of power Spectral density;</u> ~~Spect~~

1. $S_{XX}(\omega) \geq 0$
2. The power Spectral density at zero frequency is equal to the area under the Curve of the autocorrelation $R_{XX}(\tau)$ i.e $S_{XX}(0) = \int_{-\infty}^{\infty} R_{XX}(\tau)\, d\tau$.
3. The power density Spectrum of a real process $x(t)$ is an even function.
   i.e $S_{XX}(-\omega) = S_{XX}(\omega)$, $x(t)$ is real
4. $S_{XX}(\omega)$ is always a real function
5. If $S_{XX}(\omega)$ is a power spectral density of the WSS random process $x(t)$, then
   $\frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\, d\omega = A\{E[x^2(t)]\} = R_{XX}(0)$.

Properties of Cross power density Spectrum:

1. $S_{xy}(\omega) = S_{yx}(-\omega) = S^{*}_{yx}(\omega)$

2. The real part of $S_{xy}(\omega)$ and the real part of $S_{yx}(\omega)$ are even functions of ω, i.e $Re[S_{xy}(\omega)]$ and $Re[S_{yx}(\omega)]$. are even functions.

3. The imaginary part of $S_{xy}(\omega)$ and the imaginary part of $S_{yx}(\omega)$ are odd functions of ω.

   i.e $Im[S_{xy}(\omega)]$ and $Im[S_{yx}(\omega)]$ are odd functions.

4. If $x(t)$ and $y(t)$ are uncorrelated and have constant mean values $\bar{x}$ and $\bar{y}$ then

$$S_{xy}(\omega) = 2\pi \bar{x}\bar{y}\,\delta(\omega)$$

5. $S_{xy}(\omega) = 0$ and $S_{yx}(\omega) = 0$ if $x(t)$ and $y(t)$ are orthogonal.

13 (a) A random process had the power density spectrum $S_{yx}(\omega) = \frac{6\omega^2}{1+\omega^2}$, find the average power in the process.

Ans:

Given $S_{xx}(\omega) = \frac{6\omega^2}{1+\omega^2}$

Average power of $x(t)$ is

$$P_{XX} = E[x^2(t)] = \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\,d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} \frac{6\omega^2}{1+\omega^2}\,d\omega$$

$$= \frac{6}{2\pi}\left[\int_{-\infty}^{\infty} \frac{\omega^2}{1+\omega^2}\,d\omega\right]$$

$$= \frac{\overset{3}{\cancel{6}}}{\cancel{2\pi}}\left[\frac{\cancel{\pi}}{2\sqrt{2}}\right]$$

$$P_{xx} = \frac{3}{2\sqrt{2}}$$

$$\boxed{P_{xx} = 1.06 \text{ watts}}.$$

13.(b) State and prove the relationship between power Density spectrum and auto Correlation function.

<u>Ans:</u> <u>Relationship between power Spectral density and auto correlation</u> :

Let $x(t)$ be an ensemble of random process $x(t)$

Let us define $x_T(t)$ as

$$x_T(t) = \begin{cases} x(t) & \text{for } -T \le t \le T \\ 0, & \text{otherwise} \end{cases}$$

The fourier transform of $x_T(t)$ is given by

$$X_T(j\omega) = \int_{-T}^{T} x_T(t)\, e^{-j\omega t}\, dt$$

Using parseval's theorem, we can write

$$\int_{-T}^{T} x^2(t)\,dt = \frac{1}{2\pi}\int_{-\infty}^{\infty} |X_T(j\omega)|^2\, d\omega$$

The average power is given by

$$P(T') = \frac{1}{2T}\int_{-T}^{T} x^2(t)\,dt = \frac{1}{2\pi}\int_{-\infty}^{\infty}\frac{|X_T(j\omega)|^2}{2T}\,d\omega$$

To find the average power of the random process, we take the expected value with T tending to infinity in the above equation

$$P_{XX} = \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T} E[x^2(t)]\,dt$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty}\underset{T\to\infty}{Lt}\ \frac{E[|X_T(j\omega)|^2]}{2T}\,d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\,d\omega$$

Here $$S_{XX}(\omega) = \underset{T\to\infty}{Lt}\ \frac{E[|X_T(j\omega)|^2]}{2T}$$

We know that $$X_T(j\omega) = \int_{-T}^{T} x_T(t)\,e^{-j\omega t}\,dt$$

Using the above equation, we can obtain

$$S_{XX}(\omega) = \underset{T\to\infty}{Lt}\ E\left[\frac{1}{2T}\int_{-T}^{T} x(t_1)\,e^{j\omega t}dt_1 \int_{-T}^{T} x(t_2)\,e^{-j\omega t_2}dt_2\right]$$

$$= \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} E[x(t_1)x(t_2)]\,e^{-j\omega(t_2-t_1)}\,dt_2\,dt_1$$

We know $E[x(t_1)x(t_2)] = R_{xx}(t_1,t_2)$

$$S_{XX}(\omega) = \underset{T\to\infty}{Lt}\ \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\,e^{-j\omega(t_2-t_1)}\,dt_2\,dt_1$$

Now the inverse fourier transform of $S_{XX}(\omega)$ is

$$F^{-1}[S_{XX}(\omega)] = \frac{1}{2\pi}\int_{-\infty}^{\infty} S_{XX}(\omega)\, e^{j\omega\tau}\, d\omega$$

$$= \frac{1}{2\pi}\int_{-\infty}^{\infty} \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, e^{-j\omega(t_2-t_1)} e^{j\omega\tau}\, dt_2\, dt_1\, d\omega$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2) \int_{-\infty}^{\infty} e^{-j\omega(t_2-t_1-\tau)}\, d\omega\, dt_2\, dt_1$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, \frac{1}{2\pi}\left[2\pi\, \delta(t_2-t_1-\tau)\right] d\omega\, dt_2\, dt_1$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T}\int_{-T}^{T} R_{XX}(t_1,t_2)\, \delta(t_2-t_1-\tau)\, dt_2\, dt_1$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T} R_{XX}(t_1, t_1+\tau)\, dt_1$$

$$= \underset{T\to\infty}{Lt} \frac{1}{2T}\int_{-T}^{T} R_{XX}(t, t+\tau)\, d\tau$$

$$= A\left[R_{XX}(t, t+\tau)\right].$$

which is the time average of the process autocorrelation function. The above equation says that the inverse fourier transform of the power density spectrum is the time average of the auto correlation function of the process. Therefore, we can write

$$S_{XX}(\omega) = \int_{-\infty}^{\infty} A\left[R_{XX}(t,t+\tau)\right] e^{-j\omega\tau}\, d\tau$$

for a wide sense stationary process

$$A\left[R_{XX}(t,t+\tau)\right] = R_{XX}(\tau)$$

Therefore $S_{XX}(\omega) = \int_{-\infty}^{\infty} R_{XX}(\tau) e^{j\omega\tau} d\tau \rightarrow ①$

and $R_{XX}(\tau) = \frac{1}{2\pi} \int_{-\infty}^{\infty} S_{XX}(\omega) e^{j\omega\tau} d\omega \rightarrow ②$

i.e power Spectral density is fourier transform of autocorrelation function. and eq^n ① & ② is also known as weiner - Khintchen relations.

14. (a) Write all the properties of Narrow band Noise.

Ans:

Properties of Narrow Band Noise :-

1. The in phase component $N_i(t)$ and the quadrature phase component $N_Q(t)$ of a narrow band Noise $N(t)$ have zero mean, since $N(t)$ has zero mean.

2. If the narrow band noise $N(t)$ is gaussian, then its in phase Component $N_i(t)$ and quadrature Component, $N_Q(t)$ are jointly gaussian.

3. If the narrow band noise $N(t)$ is wide sense Stationary then its in phase Component $N_i(t)$ and quadrature Component $N_Q(t)$ are jointly wide sense stationary.

4. Both the in phase component $N_i(t)$ and the quadrature Component $N_Q(t)$ have the same power spectral density, which is related to the power Spectral density $S_N(\omega)$ of the narrow band noise $N(t)$ as follows.

$$S_{N_i}(\omega) = S_{N_Q}(\omega) = S_N(\omega-\omega_c) + S_N(\omega+\omega_c) \quad -\omega_0 \leq \omega \leq \omega_0.$$

where it is assumed that $S_N(\omega)$ occupies the frequency interval $\omega_c - \omega_0 \le |\omega| \le \omega_0 + \omega_0$ and $\omega_c > \omega_0$.

5. Quadrature Components $N_Q(t)$ and $N_I(t)$ have the same variance as the narrow band noise $N(t)$.

6. The Cross-spectral densities of the quadrature components of a narrow band noise are purely imaginary.

i.e $S_{N_I N_Q}(\omega) = -S_{N_Q N_I}(\omega)$

and
$$S_{N_I N_Q}(\omega) = \begin{cases} \frac{-j}{2}[S_N(\omega+\omega_c) - S_N(\omega-\omega_c) & -\omega_0 \le \omega \le \omega_0 \\ 0 \quad , & \text{elsewhere} \end{cases}$$

7. If the narrow band noise $N(t)$ is gaussian with zero mean and a power spectral density $S_N(\omega)$ that is locally symmetric about the mid-band frequency $\pm\omega_0$, then the inphase noise $N_I(t)$ and the quadrature noise $N_Q(t)$ are statistically independent.

i.e $R_{N_I N_Q}(0) = 0$.

14)(b) Define Convolution. List any four properties of convolution.

Ans: Convolution:

The relationship between the input to a linear shift invariant system, $x(t)$ and the output $y(t)$ is given by the convolution sum.

$$y(t) = x(t) * h(t) = \int_{-\infty}^{\infty} x(\tau)\, h(t-\tau)\, d\tau .$$

(or)

$$y(t) = \int_{-\infty}^{\infty} h(\tau)\, x(t-\tau)\, d\tau .$$

Properties of Convolution :-

Convolution is a linear operator and therefore, has a number of important properties including the Commutative, associative, and distributive properties. The definitions and interpretations of these properties are summerized below.

1. Commutative Property :

The commutative property states that the order in which two sequences are convolved is not important. Mathematically the commutative property

$$x(t) * h(t) = h(t) * x(t).$$

2. Associative Property :

The convolution operator satisfies the associative property, which is

$$[x(t) * h_1(t)] * h_2(t) = x(t) * [h_1(t) * h_2(t)]$$

3. Distributive Property :

The distributive property of the convolution operator states that.

$$x(t) * [h_1(t) + h_2(t)] = x(t) * h_1(t) + x(t) * h_2(t)$$

15) (a) Explain the following (i) Noise Figure (ii) Noise Sources.

Ans: **Noise Figure :-**

Noise figure gives the amount of noise internally generated by the system. It is the ratio of the power density of the total noise available at the output of the network to the power density available at the output only due to the input noise source. Noise figure gives a measure of the system performance of the noise.

It is mathematically expressed as

$$F = \frac{S_{no}(\omega)}{S'_{no}(\omega)} = \frac{S'_{no}(\omega) + S''_{no}(\omega)}{S'_{no}(\omega)} = 1 + \frac{S''_{no}(\omega)}{S'_{no}(\omega)}$$

Where $S_{no}(\omega)$ = the total noise power spectral density at the output $S'_{no}(\omega)$ is the noise power spectral density at the output due to input noise and $S''_{no}(\omega)$ = noise power spectral density at the output due to the noise generated internally by the system. and $S_{no}(\omega) = S'_{no}(\omega)$ then $F = 1$

Note: If $F > 1$, the system is said to be a noisy system. The range of F is $1 < F < \infty$; As F increases, the system becomes noisy.

(ii) **Noise Sources :-** There are two types of noise sources

1. External noise
2. Internal noise.

External Noise :- Noise whose sources are external to the receiver is called external noise. Most external noise is added into the desired signal is communication channels.

In externoise having
(i) Atmosphere Noise
(ii) ~~Extraterest~~
(ii) Extraterestrial Noise
(iii) Industrial Noise.

Internal Noise :- The Noise created with in a device (or) a system is called internal noise

→ Internal noise generated by any of the active (or) passive devices found in systems. This noise is also called function noise.

Internal noise having →
(i) Shot noise
(ii) Transit time noise
(iii) Flicker Noise
(iv) Thermal Noise

(b) Explain the Thermal Noise

Ans: Thermal Noise :-

In any conducting material, electrons move randomly the noise produced is called thermal noise.

→ Each free electron inside a conducting medium is in motion due to temparature. When the temparature increases random motion of electrons increases, and hence noise increases. At 0°K, there is no random motion of electrons and hence noise is zero.

→ The thermal noise amplitude mainly depends on resistance so it is called resistor noise

→ The thermal noise power is propotional to the temparature in degree Kelvin and the band width of the system.

$$\boxed{P_n \propto T_B}$$

where T = temparature in degree Kelvin
B = Band width in Hz

Noise power is $P_n = KTB$ watts

K = Boltzmann's constant
$= 1.38 \times 10^{-23}$ J / °K

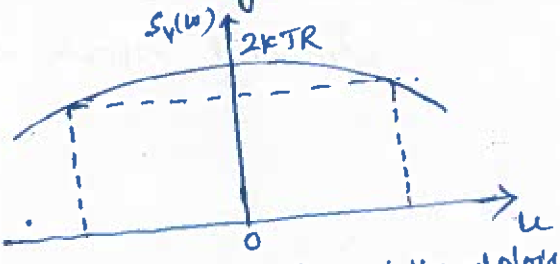
The power density spectrum of the noise voltage contributing to the thermal noise is given by

$$\boxed{S_V(\omega) = \frac{2KTR}{1+\left(\frac{\omega}{2}\right)^2}}$$

where R is the resistance and α is the average collisions per second per electron

→ The thermal noise distribution may be approximated to a gaussian distribution with zero mean value

i-e $\boxed{S_V(\omega) = 2KTR}$



Power density Spectrum of thermal Noise

The power density spectrum for noise current is

$$S_I(\omega) = S_V(\omega)/R^2$$

$$= \frac{2KT\cancel{R}}{R^{\cancel{2}}} \qquad [\because S_V(\omega) = 2KTR]$$

$$= \frac{2KT}{R}$$

$$S_I(\omega) = 2KT\left(\frac{1}{R}\right)$$

$$\boxed{S_I(\omega) = 2KTG} \qquad \left[\because G = \frac{1}{R}\right]$$

where G is a conductance of the resistor.

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 – 2023

Degree	B. Tech. (U. G.)	**Program**	CSE, CSE (AI & ML) & CSE (DS)			**Academic Year**	2022 - 2023
Course Code	20CS304	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Object Oriented Programming through C++**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the use of Scope resolution operator?	20CS304.1	L1
2	Define Inline function.	20CS304.2	L2
3	Develop the class X includes a routine to overload the – operator, Write a statement that subtracts an object of class X, x1 from another such object x2 and places the result in x3.	20CS304.3	L3
4	List out the two types of exceptions provided by C++.	20CS304.4	L1
5	Define iterators.	20CS304.5	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 principles of object oriented programming with illustrative examples.	6M	20CS304.1	L2
6 (b)	Develop a C++ program to explain default arguments with an example.	6M	20CS304.1	L2
	OR			
7 (a)	Build a C++ program to find the volume of shapes using function overloading.	6M	20CS304.1	L3
7 (b)	Construct a program in C++ to convert the binary number to octal. Eg: (000100)2- (04)8.	6M	20CS304.1	L3
8 (a)	Define a class to represent employee database and calculate the net salary.	6M	20CS304.2	L2
8 (b)	Explain friend functions with an example.	6M	20CS304.2	L2
	OR			
9 (a)	Define a class called complex. Include function for reading and displaying complex objects. Write a function to overload +operator to add two complex objects.	6M	20CS304.2	L1
9 (b)	Construct a C++ program to convert Fahrenheit object to Celsius object.	6M	20CS304.2	L2
10 (a)	Demonstrate the C++ program to generate a Fibonacci series.	6M	20CS304.3	L2
10 (b)	Construct a C++ program to add two distance objects by overloading the addition operator.	6M	20CS304.3	L3
	OR			
11 (a)	Develop a C++ program to implement single inheritance with parent class is teacher and inherit the teacher's name into the child class student.	6M	20CS304.3	L3
11 (b)	Explain about the multiple inheritance with an example.	6M	20CS304.3	L2
12 (a)	Explain the virtual function with an example.	6M	20CS304.4	L2
12 (b)	Compare static binding and dynamic binding. Explain it with neat example code.	6M	20CS304.4	L2
	OR			
13 (a)	Define an exception "Division by Zero" that is thrown when any number is divided by zero. Write a program that uses this exception.	6M	20CS304.4	L2

13 (b)	What is standard template library? Explain vector class implementation with an example.	6M	20CS304.5	L2
14	Illustrate a function template to sort arrays of float and int using bubble sort.	12M	20CS304.5	L2
	OR			
15	Extend a class template to generate a class matrix automatically for a matrix of any particular type Using the class template definition, the program should handle the arithmetic operations (+,-,*,/)For any particular type. (Such as int, float, double, char).	12M	20CS304.5	L2

Object Oriented Programming through C++ KEY

Short Answer questions:

1.What is the use of scope resolution operator?

Scope Resolution Operator in C++

This section will discuss the scope resolution operator and its various uses in the C++ programming language. The scope resolution operator is used to reference the global variable or member function that is out of scope. Therefore, we use the scope resolution operator to access the hidden variable or function of a program. The operator is represented as the double colon (::) symbol.

For example, when the global and local variable or function has the same name in a program, and when we call the variable, by default it only accesses the inner or local variable without calling the global variable. In this way, it hides the global variable or function. To overcome this situation, we use the scope resolution operator to fetch a program's hidden variable or function.

Uses of the scope resolution Operator

1. It is used to access the hidden variables or member functions of a program.
2. It defines the member function outside of the class using the scope resolution.
3. It is used to access the static variable and static function of a class.
4. The scope resolution operator is used to override function in the Inheritance.

Program to access the hidden value using the scope resolution (::) operator

Program1.cpp

```
#include <iostream>
using namespace std;
// declare global variable
int num = 50;
int main ()
{
// declare local variable
int num = 100;

// print the value of the variables
cout << " The value of the local variable num: " << num;

// use scope resolution operator (::) to access the global variable
cout << "\n The value of the global variable num: " << ::num;
return 0;
}
```

2.Define Inline function.

A) If make a function as inline, then the compiler replaces the function calling location with the definition of the inline function at compile time.

Any changes made to an inline function will require the inline function to be recompiled again because the compiler would need to replace all the code with a new code; otherwise, it will execute the old functionality.

Syntax for an inline function:

```
inline return_type function_name(parameters)
{
   // function code?
}
```

Let's understand the difference between the normal function and the inline function.

Inside the **main()** method, when the function fun1() is called, the control is transferred to the definition of the called function. The addresses from where the function is called and the definition of the function are different. This control transfer takes a lot of time and increases the overhead.

When the inline function is encountered, then the definition of the function is copied to it. In this case, there is no control transfer which saves a lot of time and also decreases the overhead.

Let's understand through an example.

```
#include <iostream>
using namespace std;
inline int add(int a, int b)
{
   return(a+b);
}
int main()
{
   cout<<"Addition of 'a' and 'b' is:"<<add(2,3);
   return 0;

}
```

The main use of the inline function in C++ is to save memory space. Whenever the function is called, then it takes a lot of time to execute the tasks, such as moving to the calling function. If the length of the function is small, then the substantial amount of execution time is spent in such overheads, and sometimes time taken required for moving to the calling function will be greater than the time taken required to execute that function.

The solution to this problem is to use macro definitions known as macros. The preprocessor macros are widely used in C, but the major drawback with the macros is that these are not normal functions which means the error checking process will not be done during the compilation.

C++ has provided one solution to this problem. In the case of function calling, the time for calling such small functions is huge, so to overcome such a problem, a new concept was introduced known as an inline function. When the function is encountered inside the main() method, it is expanded with its definition thus saving time.

We cannot provide the inlining to the functions in the following circumstances:

- If a function is recursive.
- If a function contains a loop like for, while, do-while loop.
- If a function contains static variables.
- If a function contains a switch or go to statemen

3. Develop the class X includes a routine to overload the – operator, Write a statement that subtracts an object of class X, x1 from another such object x2 and places the result in x3.

```
#include<iostream>

using namespace std;



class x {

private:

  int real;

public:

Complex(int r ) {real = r;}

  x operator - (x const&obj) {

     x res;

res.real = real + obj.real;

     return res;

  }
```

```
  void print() { cout<< real << '\n'; }

};


int main()

{

   x c1(10), c2(2);

   x c3 = c1 -c2;

   c3.print();

}
```

4.List out the two types of exceptions provided by C++.

Exceptions: Exceptions are runtime anomalies or unusual conditions that a program may encounter whileexecuting.Anomaliesmightincludeconditionssuchassdivisionbyzero,accessinganarrayoutsideofitsbounds orrunningoutofmemoryordiskspace.Whenaprogramencountersanexceptioncondition,itmustbe identified andhandled.

Exceptionsprovideawaytotransfercontrolfromonepartofaprogramtoanother.C++exceptionhandling isbuiltuponthree keywords: try,catch, andthrow.

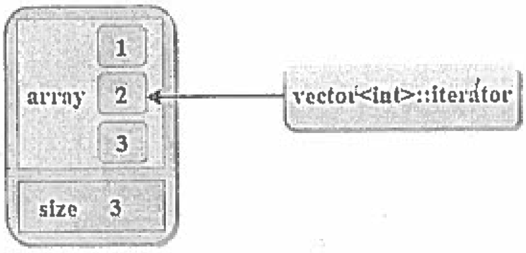
Typesofexceptions:Therearetwokindsofexceptions1.Synchronousexceptions
1.Asynchronousexceptions

1.Synchronousexceptions:Errorssuchas"Out-of-rangeindex"and"overflow"aresynchronousexceptions
2.Asynchronous exceptions:Theerrors thataregeneratedby any eventbeyondthecontrolof theprogramarecalled asynchronousexceptions

Thepurposeof exceptionhandlingistoprovide ameanstodetectandreportanexceptionalcircumstance.

5.Define iterators.

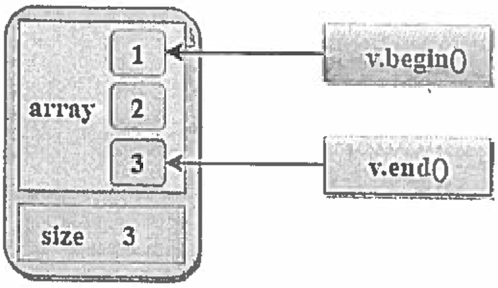
- Iterators are pointer-like entities used to access the individual elements in a container.
- Iterators are moved sequentially from one element to another element. This process is known as iterating through a container.



- Iterator contains mainly two functions:

begin(): The member function begin() returns an iterator to the first element of the vector

end(): The member function end() returns an iterator to the past-the-last element of a container.



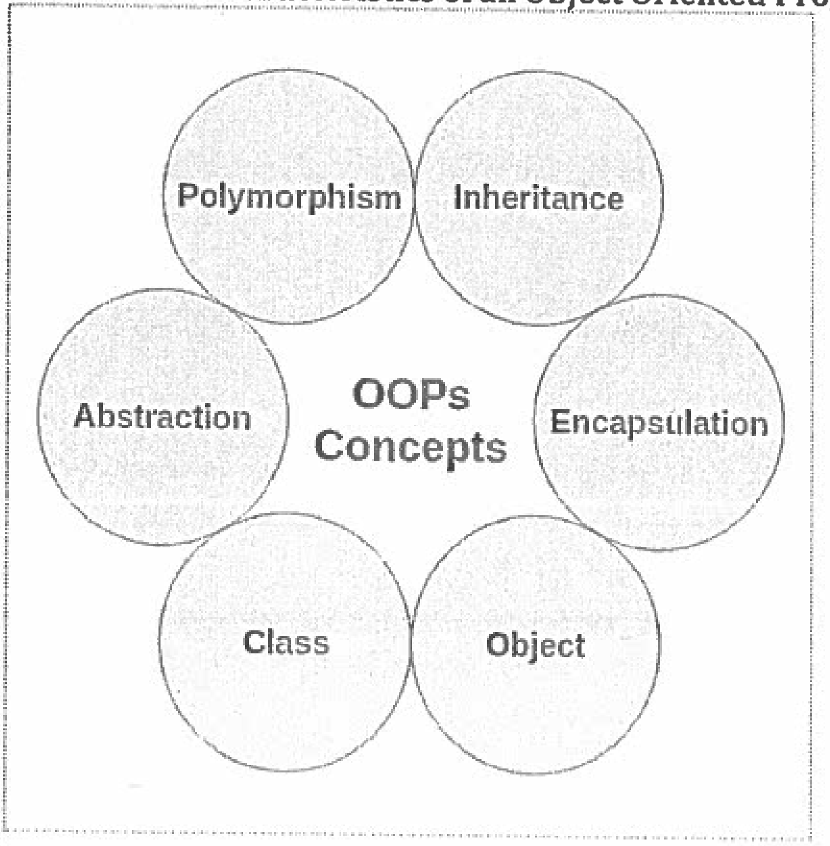
LONG ANSWER QUESTIONS:

6)a)Explain the principles of object oriented programming with illustrative examples.

1. *Introduction*
2. *Class*
3. *Objects*
4. *Encapsulation*
5. *Abstraction*
6. *Polymorphism*
7. *Inheritance*
8. *Dynamic Binding*
9. *Message Passing*

Object-oriented programming – As the name suggests uses objects in programming. Object-oriented programming aims to implement real-world entities like inheritance, hiding, polymorphism, etc in programming. The main aim of OOP is to bind together the data and the functions that operate on them so that no other part of the code can access this data except that function.

Characteristics of an Object Oriented Programming language



Class: The building block of C++ that leads to Object-Oriented programming is a Class. It is a user-defined data type, which holds its own data members and member functions, which can be accessed and used by creating an instance of that class. A class is like a blueprint for an object.
For Example: Consider the Class of Cars. There may be many cars with different names and brand but all of them will share some common properties like all of them will have 4 wheels, Speed Limit, Mileage range etc. So here, Car is the class and wheels, speed limits, mileage are their properties.

- A Class is a user-defined data-type which has data members and member functions.
- Data members are the data variables and member functions are the functions used to manipulate these variables and together these data members and member functions define the properties and behaviour of the objects in a Class.
- In the above example of class Car, the data member will be speed limit, mileage etc and member functions can apply brakes, increase speed etc.

We can say that a **Class in C++** is a blue-print representing a group of objects which shares some common properties and behaviours.

Object: An Object is an identifiable entity with some characteristics and behaviour. An Object is an instance of a Class. When a class is defined, no memory is allocated but when it is instantiated (i.e. an object is created) memory is allocated.

```
classperson
{
    charname[20];
    intid;
public:
    voidgetdetails(){}
};

intmain()
{
   person p1; // p1 is a object
}
```

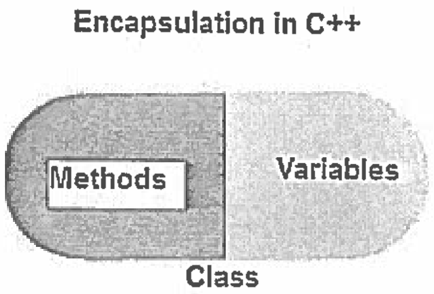
Object take up space in memory and have an associated address like a record in pascal or structure or union in C.

When a program is executed the objects interact by sending messages to one another.

Each object contains data and code to manipulate the data. Objects can interact without having to know details of each other's data or code, it is sufficient to know the type of message accepted and type of response returned by the objects.

Encapsulation: In normal terms, Encapsulation is defined as wrapping up of data and information under a single unit. In Object-Oriented Programming, Encapsulation is defined as binding together the data and the functions that manipulate them.
Consider a real-life example of encapsulation, in a company, there are different sections like the accounts section, finance section, sales section etc. The finance section handles all the financial transactions and keeps records of all the data related to finance. Similarly, the sales section handles all the sales-related activities and keeps records of all the sales. Now there

may arise a situation when for some reason an official from the finance section needs all the data about sales in a particular month. In this case, he is not allowed to directly access the data of the sales section. He will first have to contact some other officer in the sales section and then request him to give the particular data. This is what encapsulation is. Here the data of the sales section and the employees that can manipulate them are wrapped under a single name "sales section".



Encapsulation in C++

Encapsulation also leads to *data abstraction or hiding*. As using encapsulation also hides the data. In the above example, the data of any of the section like sales, finance or accounts are hidden from any other section.
Abstraction: Data abstraction is one of the most essential and important features of object-oriented programming in C++. Abstraction means displaying only essential information and hiding the details. Data abstraction refers to providing only essential information about the data to the outside world, hiding the background details or implementation.
Consider a real-life example of a man driving a car. The man only knows that pressing the accelerators will increase the speed of the car or applying brakes will stop the car but he does not know about how on pressing accelerator the speed is actually increasing, he does not know about the inner mechanism of the car or the implementation of accelerator, brakes etc in the car. This is what abstraction is.

- *Abstraction using Classes*: We can implement Abstraction in C++ using classes. The class helps us to group data members and member functions using available access specifiers. A Class can decide which data member will be visible to the outside world and which is not.
- *Abstraction in Header files*: One more type of abstraction in C++ can be header files. For example, consider the pow() method present in math.h header file. Whenever we need to calculate the power of a number, we simply call the function pow() present in the math.h header file and pass the numbers as arguments without knowing the underlying algorithm according to which the function is actually calculating the power of numbers.

Polymorphism: The word polymorphism means having many forms. In simple words, we can define polymorphism as the ability of a message to be displayed in more than one form.
A person at the same time can have different characteristic. Like a man at the same time is a father, a husband, an employee. So the same person posses different behaviour in different situations. This is called polymorphism.

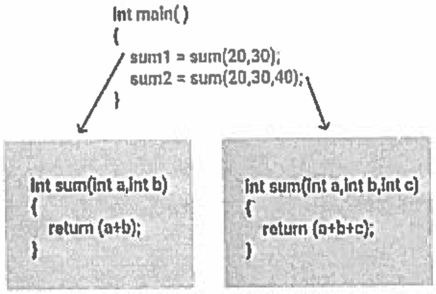
An operation may exhibit different behaviours in different instances. The behaviour depends upon the types of data used in the operation.

C++ supports operator overloading and function overloading.

- *Operator Overloading*: The process of making an operator to exhibit different behaviours in different instances is known as operator overloading.
- *Function Overloading*: Function overloading is using a single function name to perform different types of tasks.

  Polymorphism is extensively used in implementing inheritance.

Example: Suppose we have to write a function to add some integers, some times there are 2 integers, some times there are 3 integers. We can write the Addition Method with the same name having different parameters, the concerned method will be called according to parameters.



```
int main( )
{
 sum1 = sum(20,30);
 sum2 = sum(20,30,40);
}
```

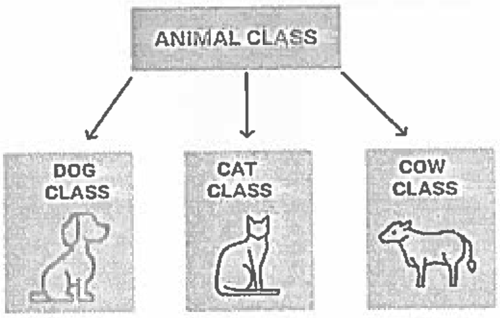
```
int sum(int a,int b)
{
  return (a+b);
}
```

```
int sum(int a,int b,int c)
{
  return (a+b+c);
}
```

Inheritance: The capability of a class to derive properties and characteristics from another class is called Inheritance. Inheritance is one of the most important features of Object-Oriented Programming.

- **Sub Class**: The class that inherits properties from another class is called Sub class or Derived Class.
- **Super Class**:The class whose properties are inherited by sub class is called Base Class or Super class.
- **Reusability**: Inheritance supports the concept of "reusability", i.e. when we want to create a new class and there is already a class that includes some of the code that we want, we can derive our new class from the existing class. By doing this, we are reusing the fields and methods of the existing class.

Example: Dog, Cat, Cow can be Derived Class of Animal Base Class.



Dynamic Binding: In dynamic binding, the code to be executed in response to function call is decided at runtime. C++ has virtual functions to support this.

Message Passing: Objects communicate with one another by sending and receiving information to each other. A message for an object is a request for execution of a procedure and therefore will invoke a function in the receiving object that generates the desired results. Message passing involves specifying the name of the object, the name of the function and the information to be sent.

Advantage of OOPs over Procedure-oriented programming language

1. OOPs makes development and maintenance easier where as in Procedure-oriented programming language it is not easy to manage if code grows as project size grows.
2. OOPs provide data hiding whereas in Procedure-oriented programming language a global data can be accessed from anywhere.
3. OOPs provide ability to simulate real-world event much more effectively. We can provide the solution of real word problem if we are using the Object-Oriented Programming language.

6)b)Develop a C++ program to explain default arguments with an example.

In a function, arguments are defined as the values passed when a function is called. Values passed are the source, and the receiving function is the destination.

Now let us understand the concept of default arguments in detail.

Definition

A default argument is a value in the function declaration automatically assigned by the compiler if the calling function does not pass any value to that argument.

Characteristics for defining the default arguments

Following are the rules of declaring default arguments -

- The values passed in the default arguments are not constant. These values can be overwritten if the value is passed to the function. If not, the previously declared value retains.
- During the calling of function, the values are copied from left to right.
- All the values that will be given default value will be on the right.

Example

- void function(intx,int y, int z = 0)
  Explanation - The above function is valid. Here z is the value that is predefined as a part of the default argument.
- Void function(int x, int z = 0, int y)
  Explanation - The above function is invalid. Here z is the value defined in between, and it is not accepted.

Code

```
#include<iostream>
using namespace std;
int sum(int x, int y, int z=0, int w=0) // Here there are two values in the default arguments
{ // Both z and w are initialised to zero
```

```
    return (x + y + z + w); // return sum of all parameter values
}
int main()
{
    cout << sum(10, 15) << endl; // x = 10, y = 15, z = 0, w = 0
    cout << sum(10, 15, 25) << endl; // x = 10, y = 15, z = 25, w = 0
    cout << sum(10, 15, 25, 30) << endl; // x = 10, y = 15, z = 25, w = 30
    return 0;
}
```

Explanation

In the above program, we have called the sum function three times.

- Sum(10,15)
  When this function is called, it reaches the definition of the sum. There it initializes x to 10 y to 15, and the rest values are zero by default as no value is passed. And all the values after sum give 25 as output.
- Sum(10,15, 25)
  When this function is called, x remains 10, y remains 15, the third parameter z that is passed is initialized to 25 instead of zero. And the last value remains 0. The sum of x, y, z, w, is 50 which is returned as output.
- Sum(10,15,25,30)
  In this function call, there are four parameter values passed into the function with x as 10, y as 15, z is 25, and w as 30. All the values are then summed up to give 80 as the output. 7.a.

7)a)Build a C++ program to find the volume of shapes using function overloading.

```
#include<iostream>

using namespace std;
#define PI 3.1416

// function prototypes
float volume(float length, float breadth, float height);
float volume(float radius);
float volume(float radius, float height);

int main(){
        float cube_l = 40.0, cube_b = 30.0, cube_h = 10.0;
```

```
o           float sphere_r = 2.5;
o           float cylinder_r = 2.5, cylinder_h = 10.0;
o           cout<<"Volume of Cube ="<<volume(cube_l, cube_b, cube_h)<<endl;
o           cout<<"Volume of Sphere ="<<volume(sphere_r)<<endl;
o           cout<<"Volume of Cylinder ="<<volume(cylinder_r, cylinder_h)<<endl;
o           return 0;
o   }
o   // function defination
o   float volume(float length, float breadth, float height){
o           return length * breadth * height;
o   }
o   float volume(float radius){
o           return (4.0/3.0) * PI * radius * radius *radius;
o   }
o   float volume(float radius, float height){

    return PI * radius *radius * height;

}

    )
```

7)B) Construct a program in C++ to convert the binary number to octal.

Eg: (000100)2- (04)8.

Program to convert binary to octal conversion.

```
#include <iostream>
#include <cmath>


using namespace std;


int convertBinarytoOctal(long long);
int main()
{
   long longbinaryNumber;
```

```
cout<< "Enter a binary number: ";
cin>>binaryNumber;

cout<<binaryNumber<< " in binary = " <<convertBinarytoOctal(binaryNumber) << " in octal ";

   return 0;
}

int convertBinarytoOctal(long long binaryNumber)
{
   int octalNumber = 0, decimalNumber = 0, i = 0;

while(binaryNumber != 0)
   {
decimalNumber += (binaryNumber%10) * pow(2,i);
      ++i;
binaryNumber/=10;
   }

i = 1;

   while (decimalNumber != 0)
   {
octalNumber += (decimalNumber % 8) * i;
decimalNumber /= 8;
i *= 10;
   }

   return octalNumber;
}
```

8)a)Define a class to represent employee database and calculate the net salary.

```
#include<iostream>
using namespace std;

class Employee
{
```

```
        int eid;
        char ename[100];
        float basic_salary, hra, da, i_tax, net_salary;

        public:
        void accept_details()
        {
cout<<"\n Enter  Employee Id : ";
cin>>eid;
cout<<"\n Enter Employee Name : ";
cin>>ename;
cout<<"\n Enter Basic Salary : ";
cin>>basic_salary;

hra = 800;
                da = 0.25 * basic_salary;
i_tax = 0.15 * basic_salary;
net_salary = basic_salary + da + hra - i_tax;
        }
        void display_details()
        {
cout<<"\n ---------------------- ";
cout<<"\n Employee Id       : "<<eid;
cout<<"\n Employee Name  : "<<ename;
cout<<"\n Basic Salary       : "<<basic_salary;
cout<<"\n HRA                    : "<<hra;
cout<<"\n DA                      : "<<da;
cout<<"\n I-Tax                    : "<<i_tax;
cout<<"\n Net Salary           : "<<net_salary;
        }
};
int main()
{
        Employee e[3];
        int i;
        for(i=0;i<3;i++)
        {
        e[i].accept_details();
        }
        for(i=0;i<3;i++)
{        e[i].display_details();
}
        return 0;
}
```

8)b)Explain friend functions with an example.

```
#include <iostream>

using namespace std;
```

```
class Distance {
 private:
     int meter;

     // friend function
     friend int addFive(Distance);

   public:
 Distance() : meter(0) {}

 };

 // friend function definition
 int addFive(Distance d) {

    //accessing private members from the friend function
  d.meter += 5;
    return d.meter;
  }

  int main() {
     Distance D;
  cout<< "Distance: " <<addFive(D);
     return 0;
  }
```

9)a)Define a class called complex. Include function for reading and displaying complex objects. Write a function to overload +operator to add two complex objects.#include <iostream>

```
#include <sstream>
#include <cmath>
using namespace std;
class Complex {
   private:
      int real, imag;
   public:
Complex(){
      real = imag = 0;
   }
   Complex (int r, int i){
```

```
        real = r;
    imag = i;
      }
      string to_string(){
    stringstream ss;
    if(imag>= 0)
          ss << "(" << real << " + " <<imag<< "i)";
        else
          ss << "(" << real << " - " << abs(imag) << "i)";
        return ss.str();
      }
      Complex operator+(Complex c2){
        Complex ret;
    ret.real = real + c2.real;
    ret.imag = imag + c2.imag;
        return ret;
      }
    };
    int main(){
      Complex c1(8,-5), c2(2,3);
      Complex res = c1 + c2;
    cout<<res.to_string();
}
```

9)b) Construct a C++ program to convert Fahrenheit object to Celsius object.

```
#include<iostream>
using namespacestd;
intmain()
{
floatfahrenheit, celsius;
cout<<"Enter the Temperature in Fahrenheit: ";
cin>>fahrenheit;
celsius = (fahrenheit-32)/1.8;
cout<<"\nEquivalent Temperature in Celsius: "<<celsius;
cout<<endl;
return 0;
}
```

10)a)Demonstrate the C++ program to generate a Fibonacci series.

```
o  #include <iostream>
o  using namespace std;
o  int main() {
o   int n1=0,n2=1,n3,i,number;
o   cout<<"Enter the number of elements: ";
o   cin>>number;
o   cout<<n1<<" "<<n2<<" "; //printing 0 and 1
o   for(i=2;i<number;++i) //loop starts from 2 because 0 and 1 are already printed
o   {
```

```
o    n3=n1+n2;
o    cout<<n3<<" ";
o    n1=n2;
o    n2=n3;
o   }
o    return 0;
  }
```

10)b)Construct a C++ program to add two distance objects by overloading the addition operator.

```
#include <iostream>
usingnamespace std;

classDistance {
private:
int feet, inches;

public:
   // function to read distance
voidreadDistance(void)
   {
cout<< "Enter feet: ";
cin>> feet;
cout<< "Enter inches: ";
cin>> inches;
   }

   // function to display distance
voiddispDistance(void)
   {
cout<< "Feet:" << feet << "\t"
<< "Inches:" << inches <<endl;
   }

   // add two Distance using + operator overloading
   Distance operator+(Distance& dist1)
   {
      Distance tempD; // to add two distances
tempD.inches = inches + dist1.inches;
tempD.feet = feet + dist1.feet + (tempD.inches / 12);
tempD.inches = tempD.inches % 12;
returntempD;
   }
};

intmain()
{
   Distance D1, D2, D3;

cout<< "Enter first  distance:" <<endl;
```

```
    D1.readDistance();
cout<<endl;

cout<< "Enter second distance:" <<endl;
    D2.readDistance();
cout<<endl;

    // add two distances
    D3 = D1 + D2;

cout<< "Total Distance:" <<endl;
    D3.dispDistance();

cout<<endl;

return0;
}
```

11)a)Explain single inheritance with example.

C++ Inheritance

In C++, inheritance is a process in which one object acquires all the properties and behaviors of its parent object automatically. In such way, you can reuse, extend or modify the attributes and behaviors which are defined in other class.

In C++, the class which inherits the members of another class is called derived class and the class whose members are inherited is called base class. The derived class is the specialized class for the base class.

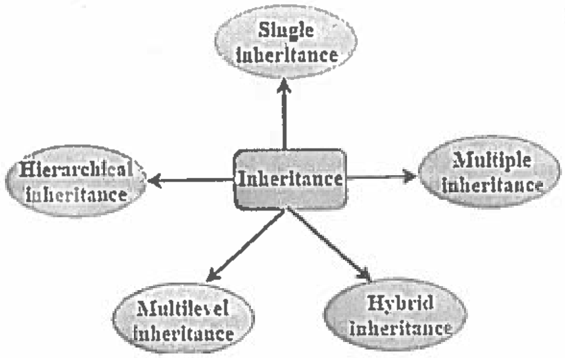
Advantage of C++ Inheritance

Code reusability: Now you can reuse the members of your parent class. So, there is no need to define the member again. So less code is required in the class.

Types Of Inheritance

C++ supports five types of inheritance:

- Single inheritance
- Multiple inheritance
- Hierarchical inheritance
- Multilevel inheritance
- Hybrid inheritance



Derived Classes

A Derived class is defined as the class derived from the base class.

The Syntax of Derived class:

1. class derived_class_name :: visibility-mode base_class_name
2. {
3. // body of the derived class.
4. }

Where,

derived_class_name: It is the name of the derived class.

visibility mode: The visibility mode specifies whether the features of the base class are publicly inherited or privately inherited. It can be public or private.

base_class_name: It is the name of the base class.

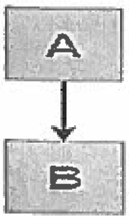
- When the base class is privately inherited by the derived class, public members of the base class becomes the private members of the derived class. Therefore, the public members of the base class are not accessible by the objects of the derived class only by the member functions of the derived class.
- When the base class is publicly inherited by the derived class, public members of the base class also become the public members of the derived class. Therefore, the public members of the base class are accessible by the objects of the derived class as well as by the member functions of the base class.

Note:

- In C++, the default mode of visibility is private.
- The private members of the base class are never inherited.

C++ Single Inheritance

Single inheritance is defined as the inheritance in which a derived class is inherited from the only one base class.



Where 'A' is the base class, and 'B' is the derived class.

C++ Single Level Inheritance Example: Inheriting Fields

When one class inherits another class, it is known as single level inheritance. Let's see the example of single level inheritance which inherits the fields only.

```
#include <iostream>
using namespace std;
 class Account {
   public:
   float salary = 60000;
 };
   class Programmer: public Account {
   public:
   float bonus = 5000;
   };
int main(void) {
     Programmer p1;
     cout<<"Salary: "<<p1.salary<<endl;
     cout<<"Bonus: "<<p1.bonus<<endl;
    return 0;
}
```

11)b)ExplainMultiple inheritance with example.

Multiple Inheritance in C++

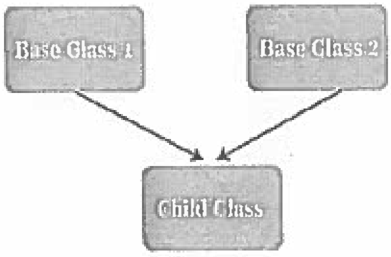
This section will discuss the Multiple Inheritances in the C++ programming language. When we acquire the features and functionalities of one class to another class, the process is called **Inheritance**. In this way, we can reuse, extend or modify all the attributes and behaviour of the parent class using the derived class object. It is the most important feature of object-oriented programming that reduces the length of the program.

A class that inherits all member functions and functionality from another or parent class is called the **derived class**. And the class from which derive class acquires some features is called the **base** or **parent class.**

Multiple Inheritance is the concept of the Inheritance in C++ that allows a child class to inherit properties or behaviour from multiple **base** classes. Therefore, we can say it is the process that enables a derived class to acquire member functions, properties, characteristics from more than one base class.

Diagram of the Multiple Inheritance

Following is the diagram of the Multiple Inheritances in the C++ programming language



In the above diagram, there are two-parent classes: **Base Class 1** and **Base Class 2,** whereas there is only one **Child Class.** The Child Class acquires all features from both Base class 1 and Base class 2. Therefore, we termed the type of Inheritance as Multiple Inheritance.

Syntax of the Multiple Inheritance

```
class A
{
// code of class A
}
class B
{
// code of class B
}
class C: public A, public B (access modifier class_name)
{
// code of the derived class
}
```

In the above syntax, class A and class B are two base classes, and class C is the child class that inherits some features of the parent classes.

Let's write the various program of Multiple Inheritance to inherit the member functions and functionality from more than one base class using the derived class.

Example 1: Program to use the Multiple Inheritance

Program1.cpp

```
#include <iostream>
using namespace std;

// create a base class1
class Base_class
{
  // access specifier
  public:
  // It is a member function
  void display()
  {
    cout << " It is the first function of the Base class " << endl;
  }
};

// create a base class2
class Base_class2
{
  // access specifier
  public:
  // It is a member function
  void display2()
  {
    cout << " It is the second function of the Base class " << endl;
  }
};

/* create a child_class to inherit features of Base_class and Base_class2 with access specifier. */
class child_class: public Base_class, public Base_class2
{

  // access specifier
  public:
  void display3() // It is a member function of derive class
  {
    cout << " It is the function of the derived class " << endl;
  }

```

```
};

int main ()
{
   // create an object for derived class
   child_class ch;
   ch.display(); // call member function of Base_class1
   ch.display2(); // call member function of Base_class2
   ch.display3(); // call member function of child_class
}
```

12)a)Explain the virtual function with an example.

```
#include<iostream>
using namespace std;

class base {
public:
   virtual void print()
   {
cout<< "print base class\n";
   }

   void show()
   {
cout<< "show base class\n";
   }
};

class derived : public base {
public:
   void print()
   {
cout<< "print derived class\n";
   }

   void show()
   {
cout<< "show derived class\n";
   }
};

int main()
```

```
{
   base *bptr;
   derived d;
bptr = &d;

   // Virtual function, binded at runtime
bptr->print();

   // Non-virtual function, binded at compile time
bptr->show();

   return 0;
}
```

12)b)Compare static binding and dynamic binding. Explain it with neat example code.

Binding generally refers to a mapping of one thing to another. In the context of compiled languages, binding is the link between a function call and the function definition. When a function is called in C++, the program control binds to the memory address where that function is defined.

There are two types of binding in C++: static (or early) binding and dynamic (or late) binding. This post provides an overview of the differences between static and dynamic binding in C++.

The static binding happens at the compile-time, and dynamic binding happens at the runtime. Hence, they are also called early and late binding, respectively.

In static binding, the function definition and the function call are linked during the compile-time, whereas in dynamic binding, the function calls are not resolved until runtime. So, they are not bound until runtime.

Static binding happens when all information needed to call a function is available at the compile-time. Dynamic binding happens when the compiler cannot determine all information needed for a function call at compile-time.

Static binding can be achieved using the normal function calls, function overloading, and operator overloading, while dynamic binding can be achieved using the virtual functions.

Since all information needed to call a function is available before runtime, static binding results in faster execution of a program. Unlike static binding, a function call is not resolved until runtime for later binding, resulting in somewhat slower execution of code.

The major advantage of dynamic binding is that it is flexible since a single function can handle different types of objects at runtime. This significantly reduces the size of the codebase and also makes the source code more readable.

Example of Static Binding in C++:

Consider the following code, where the sum() function is overloaded to accept two and three integer arguments. Even though there are two functions with the same name inside the ComputeSum class, the function call sum() binds to the correct function depending on the parameters passed to those functions. This binding is done statically during compile time.

```
// C++ program to illustrate the concept of static binding
#include <iostream>
using namespace std;

class ComputeSum
{
    public:

    int sum(int x, int y) {
        return x + y;
    }

    int sum(int x, int y, int z) {
        return x + y + z;
    }
};

int main()
{
ComputeSumobj;
cout<< "Sum is " <<obj.sum(10, 20) <<endl;
```

```
cout<< "Sum is " <<obj.sum(10, 20, 30) <<endl;


   return 0;

}
```

Example of Dynamic Binding in C++:

Consider the following code, where we have a base class B, and a derived class D. Base class B has a virtual function f(), which is overridden by a function in the derived class D, i.e., D::f() overrides B::f().

Now consider lines 30-34, where the decision as to which class's function will be invoked depends on the dynamic type of the object pointed to by basePtr. This information can only be available at the runtime, and hence f() is subject to the dynamic binding.

```
// C++ program to illustrate the concept of dynamic binding

#include <iostream>

using namespace std;


class B

{

   public:


   // Virtual function

   virtual void f() {

cout<< "The base class function is called.\n";

   }

};


class D: public B

{

   public:
```

```
   void f() {
cout<< "The derived class function is called.\n";
   }
};

int main()
{
    B base;
    D derived;

    B *basePtr = &base;
 basePtr->f();

 basePtr = &derived;
 basePtr->f();

     return 0;
 }
```

13)a) Define an exception "Division by Zero" that is thrown when any number is divided by zero. Write a program that uses this exception.

Exceptionhandling

Exceptions: Exceptions are runtime anomalies or unusual conditions that a program may encounter whileexecuting.Anomaliesmightincludeconditionssuchassdivisionbyzero,accessinganarrayoutsideofitsboundsorrunningoutofmemoryordiskspace.Whenaprogramencountersanexceptioncondition,itmustbe identified andhandled.

Exceptionsprovideawaytotransfercontrolfromonepartofaprogramtoanother.C++exceptionhandlingisbuiltuponthree keywords: try,catch, andthrow.

Typesofexceptions:Therearetwokindsofexceptions1.Synchronousexceptions
2.Asynchronousexceptions

3.Synchronousexceptions:Errorssuchas"Out-of-

rangeindex"and"overflow"aresynchronousexceptions
4.Asynchronous exceptions:Theerrors thataregeneratedby any eventbeyondthecontrolof theprogramarecalled asynchronousexceptions

Thepurposeof exceptionhandlingistoprovide ameanstodetectandreportanexceptionalcircumstance

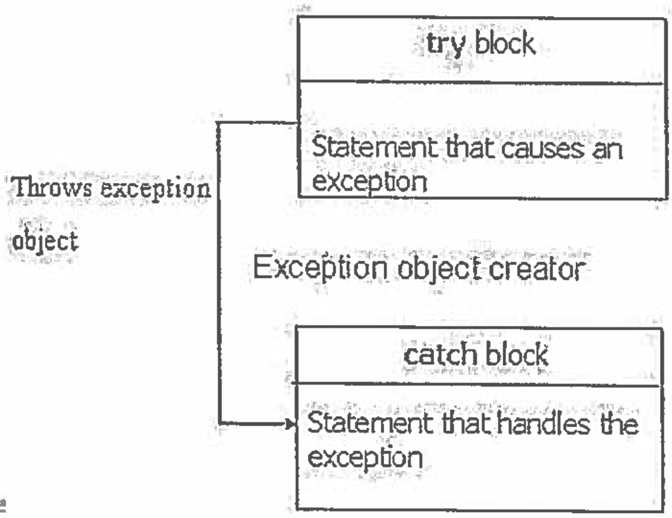
ExceptionHandlingMechanism:

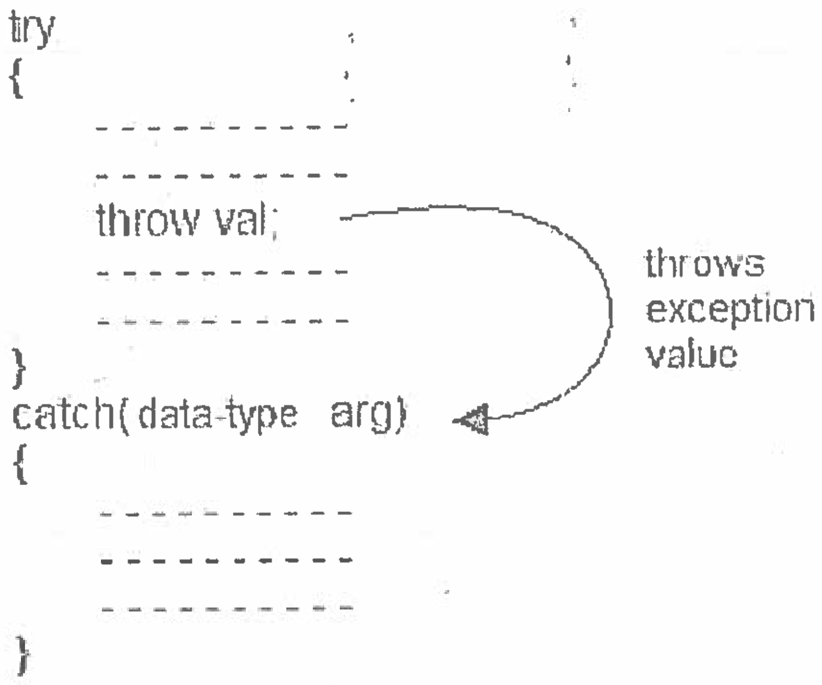
An exception is said to be thrown at the place where some error or abnormal condition is detected. Thethrowing will cause the normal program flow to be aborted, in a raised exception. An exception is thrownprogrammatic,theprogrammer specifiestheconditionsofa throw.

In handled exceptions, execution of the program will resume at a designated block of code, calleda catch block, which encloses the point of throwing in terms of program execution. The catch block canbe,and usuallyis,locatedin adifferent function than thepoint ofthrowing.

C++exceptionhandlingisbuiltuponthreekeywords:try,catch,andthrow.

Try is used to preface a block of statements which may generate exceptions. This block of statements is knownas try block. When an exception is detected it is thrown by using throw statement in the try block. Catch blockcatchesthe exception thrownbythrowstatementinthe tryblock andhandlesitappropriately.





```
try
{
        - - - - - - - - - - -
        - - - - - - - - - - -
        throw val;
        - - - - - - - - - - -
        - - - - - - - - - - -
}
catch( data-type   arg)
{
        - - - - - - - - - - -
        - - - - - - - - - - -
        - - - - - - - - - - -
}
```

```
#include<iostream>usin
gnamespacestd;intmain()
    {
    inta,b;
    cout<<"Enter any two integer
    values";cin>>a>>b;
     intx=a-b;

          try
          {
                if(x!=0)
                {
                cout<<"Result(a/x)="<<a/x<<endl;
                }
                else
                {
                throw x;
                }
      }
          catch(intex)
          {
          cout<<"Exceptioncaught:DivideByZero\n";
          }
    }
```

13)b) What is standard template library? Explain vector class implementation with an example.

- VECTOR Implementation:
- Vectors have the ability to resize itself automatically like dynamic arrays when an element is inserted or deleted, the container handle their storage automatically. Vector elements are placed in contiguous storage so that they can be accessed and traversed using iterators. Data can be inserted or erased at the begin, middle or end of the vector.
    - **Functions and descriptions:**
    - List of functions used here:
    - v.size() = Returns the size of vector.
    - v.push_back() = It is used to insert elements to the vector from end.
    - v.pop_back() = To pop out the value from the vector from back.
    - v.capacity() = Returns the size of the storage space currently allocated to the vector as number of elements.
    - v.clear() = Clears the vector.
    - Example Code

```
#include <iostream>
#include <vector>
using namespace std;
int main() {
  vector<int> v;
  vector<int>::iterator it;
  int c, i;
```

```
while (1) {
 cout<<"1.Size of the Vector"<<endl;
 cout<<"2.Insert Element into the Vector"<<endl;
 cout<<"3.Delete Last Element of the Vector"<<endl;
 cout<<"4.Display the capacity of vector"<<endl;
 cout<<"5.Display by Iterator"<<endl;
 cout<<"6.Clear the Vector"<<endl;
 cout<<"7.Exit"<<endl;
 cout<<"Enter your Choice: ";
 cin>>c;
 switch(c) {
   case 1:
     cout<<"Size of Vector: ";
     cout<<v.size()<<endl;
   break;
   case 2:
     cout<<"Enter value to be inserted: ";
     cin>>i;
     v.push_back(i);
   break;
   case 3:
     cout<<"Delete Last Element Inserted:"<<endl;
     v.pop_back();
   break;
   case 4:
     cout<<"Displaying capacity of vector: ";
     cout<<v.capacity()<<endl;
   break;
   case 5:
     cout<<"Displaying Vector by Iterator: ";
     for (it = v.begin(); it != v.end(); it++) {
       cout<<*it<<" ";
     }
     cout<<endl;
   break;
   case 6:
     v.clear();
     cout<<"Vector Cleared"<<endl;
   break;
   case 7:
     exit(1);
   break;
   default:
     cout<<"Wrong Choice"<<endl;
  }
 }
 return 0;
}
```

14)Illustrate a function template to sort arrays of float and int using bubble sort.

BUBBLE Sort Using Function Templates:

```
#include<iostream>
#include<vector>
using namespace std;
// template function to perform bubble sort on array, arr
// n: size of arr
template<typename T>
void BubbleSort(T arr[], int n)
{
        for(int i=0;i<n-1;++i){
                for(int j=0;j<n-i-1;++j){
                        if(arr[j]>arr[j+1]){
                                T temp = arr[j+1];
                                arr[j+1] = arr[j];
                                arr[j] = temp;
                        }
                }
        }
}

// Template function to print array
// n: size of arr[]
template<typename T>
void PrintArray(T arr[], int n)
{
  for (int i = 0; i< n; ++i)
cout<<arr[i] << " ";
cout<< "\n\n";
}

int main()
```

```
{

   int arr[] = { 1, 10, 90, 100, -1, 11, 9, 14, 3, 2, 20, 19 };

   int n = sizeof(arr) / sizeof(int);


cout<< "Array Before Sorting: " <<endl;

PrintArray(arr, n);


BubbleSort(arr, n);


cout<< "Array After Sorting: " <<endl;

PrintArray(arr, n);


}
```

15.Extend a class template to generate a class matrix automatically for a matrix of any particular type Using the class template definition, the program should handle the arithmetic operations (+,-,*,/)For any particular type. (Such as int, float, double, char).

```
#include<iostream>

using namespace std;

const int r=5,c=5;

template<class T>
class matrix
{
 T m[r][c];
public:
 void get_value()
 {
  for(int i=0;i<r;i++)
  {
   for(int j=0;j<c;j++)
   {
 cout<<"\n M["<<i<<"]["<<j<<"] = ";
 cin>>m[i][j];
   }
  }
 }
```

```
void operator +(matrix ob)
{
 T p[r][c];

 for(int i=0;i<r;i++)
 {
  for(int j=0;j<c;j++)
  {
   p[i][j]=m[i][j]+ob.m[i][j];
cout<<" "<<p[i][j]<<" ";
  }
cout<<"\n";
 }
}

void operator -(matrix ob)
{
 T p[r][c];

 for(int i=0;i<r;i++)
 {
  for(int j=0;j<c;j++)
  {
   p[i][j]=m[i][j]-ob.m[i][j];
cout<<" "<<p[i][j]<<" ";
  }
cout<<"\n";
 }
}

void operator *(matrix ob)
{
 T p[r][c];

 for(int i=0;i<r;i++)
 {
  for(int j=0;j<c;j++)
  {
   p[i][j]=0;
   for(int k=0;k<c;k++)
   {
    p[i][j]+=(m[i][k] * ob.m[k][j]);
   }
  }
 }
```

```
 for(int i=0;i<r;i++)
 {
  for(int j=0;j<c;j++)
  {
cout<<" "<<p[i][j]<<" ";
  }
cout<<"\n";
 }
}
}

 void display()
 {
  for(int i=0;i<r;i++)
  {
   for(int j=0;j<c;j++)
   {
 cout<<" "<<m[i][j]<<" ";
   }
 cout<<"\n";
  }
 cout<<"\n\n";
 }
 };

 void main()
 {
  matrix<int> m1,m2;
  int ch;

 cout<<"\n Enter Elements of Matrix A\n";
  m1.get_value();
 cout<<"\n Enter Elements of Matrix B\n";
  m2.get_value();

  while(1)
  {
   system("cls");
  cout<<"\n----------MATRIX OPERATIONS-----------\n\n";
  cout<<"\n 1. Sum";
  cout<<"\n 2. Difference";
  cout<<"\n 3. Product";
  cout<<"\n 5. Display";
  cout<<"\n 0. EXIT\n";
  cout<<"\n Enter your choice: ";
  cin>>ch;
```

```
 switch(ch)
 {
 case 1: cout<<"\n\n Matrices Sum \n\n";
   m1 + m2;
  break;
 case 2: cout<<"\n\n Matrices Subtraction \n\n";
   m1-m2;
  break;
 case 3: cout<<"\n\n Matrices Product \n\n";
   m1*m2;
  break;
 case 4: cout<<"\n\n MATRIX A\n";
   m1.display();
cout<<"\n\n Transposed Matrix\n";
   m1.transpose();
cout<<"\n\n MATRIX B\n";
   m2.display();
cout<<"\n\n Transposed Matrix\n";
   m2.transpose();
  break;
 case 5: cout<<"\n\n MATRIX A\n";
   m1.display();
cout<<"\n\n MATRIX B\n";
   m2.display();
  break;
 case 0: exit(0);
 default: cout<<"\n\n Invalid choice";
 }
 system("pause");
 }
}
```

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	Civil Engineering			**Academic Year**	2022 - 2023
Course Code	20CE305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Fluid Mechanics**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define vapour pressure and surface tension of a fluid.	20CE305.1	L1
2	List any two applications of flownet.	20CE305.2	L1
3	State Hagen-Poiseuille Law.	20CE305.3	L1
4	Define draft tube.	20CE305.4	L1
5	What is boundary layer?	20CE305.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Explain the difference between Piezometer and U-Tube Manometer with derivations and diagrams.	12M	20CE305.1	2
	OR			
7 (a)	A block of material of specific gravity 0.45 floats in water. Determine the meta-centric height of the block if its size is 3 m * 2 m* 0.8 m.	6M	20CE305.1	L3
7 (b)	A Rectangular plane surface 1 m wide and 3 m deep lies in water in such a way that its plane surface makes an angle of 300 with the free surface of water. Determine the total pressure and the depth of centre of pressure when upper edge of the plate is 2 m below the free surface.	6M	20CE305.1	L3
8 (a)	Explain various types of fluid flows.	4M	20CE305.2	L2
8 (b)	The velocity potential function is given by φ = y 2 – x 2 – xy 3 /2 + x3 y/2. Find the velocity components in x and y directions. Show that φ represents a possible case of flow.	8M	20CE305.2	L3
	OR			
9	Derive Bernoulli's equation for a compressible frictionless fluid.	12M	20CE305.2	L2
10	Derive Darcy Weisbach equation for head loss in case of pipe flow.	12M	20CE305.3	L2
	OR			
11	Describe Reynolds experiment for characterization of flows in pipe.	12M	20CE305.3	L2
12	Draw inlet and outlet triangles of Pelton turbine and explain the working of the turbine with neat sketches.	12M	20CE305.4	L2
	OR			
13 (a)	What are unit quantities? Define unit quantities for a turbine. Why they are important?	6M	20CE305.4	L1
13 (b)	Explain the factors which influence the selection of turbine.	6M	20CE305.4	L2
14	How will you find the drag on a flat plate due to laminar and turbulent boundary layers?	12M	20CE404.5	L3
	OR			
15	Derive the expression for displacement thickness and discuss the characteristics of laminar and turbulent boundary layer.	12M	20CE404.5	L2

NSRIT

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(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Degree: B.Tech (U.G)   Program: Civil Engineering   Academic year: 2022-23

Course code: 20CE305   Test Duration: 3hrs   Max Marks: 70   Sem: III

Course: fluid Mechanics

PART-A-

① Surface tension & Vapour pressure: Defined as tensile force acting on the surface of a liquid in contact with gas (or) on the surface between two immiscible liquids such that the contact surface behaves like membrane under tension.

Vapour pressure: When liquid vapourizes at 100°c vaporisation takes place, the molecules escapes from free surface of the liquid. The vapour molecules get accumulated at free surface of liquid this pressure known as vapour pressure.

② Two- Applications flow net:

① Rate of Seepage loss
② Seepage pressure
③ uplift pressure
④ Exit gradient

③ **Hagen - Poise uille Law**

$$\frac{P_1 - P_2}{\rho g} = h_f = \frac{32 \mu \bar{u} L}{\rho g D^2}$$

④ **Draft tube**

Draft tube is a pipe of gradualy increasing area which connects the outlet of the runner to the tail race.

⑤ **Boundary layer**

variation of velocity from zero to free stream velocity in the direction normal to boundary takes place in narrow region in the vicinity of solid boundary. This narrow region is called boundary layer

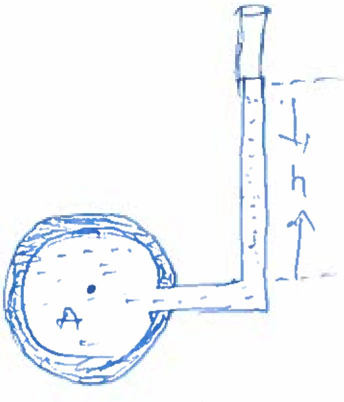
PART-B :

⑥ **Piezometer:** → Simplest form used to measure guage pressure

→ One end connected to point where pressure is to be measured and other open to atmosphere

→ Rise of liquid gives pressure head

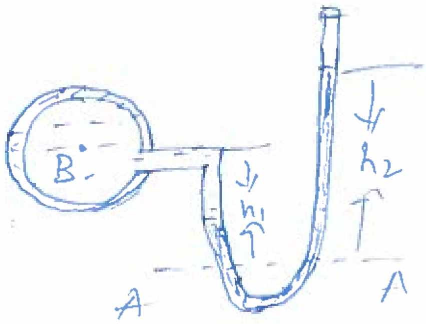
→ If at 'A' the height of liquid 'h' in Piezometer tube, then pressure at A $= \rho \times g \times h \frac{N}{m^2}$



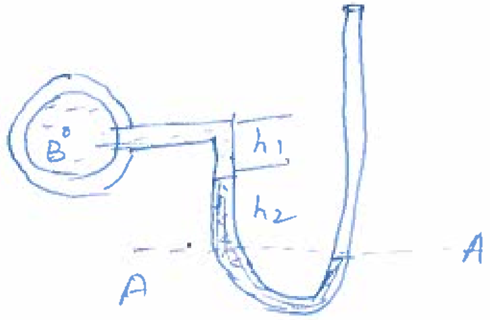
Piezometer

ANSWER KEY AND SCHEME OF EVALUATION

U-tube Manometers: It consists of glass tube bent in U-shape, one end of which is connected to a point at which pressure is to be measured and other end remains open to the atmosphere as shown. The tube generally contains mercury (or) any other liquid whose specific gravity is greater than the specific gravity of the liquid whose pressure is to be measured



for Guage Pressure



for vaccum pressure

(a) for Guage Pressure

Pressure above A-A in left column $= P + S_1 * g \times h_1 \rightarrow ①$

Pressure above A-A in right column $= S_2 \times g \times h_2 \rightarrow ②$

Equating ① & ②

$$P + S_1 g h_1 = P_2 g h_2$$

$$\boxed{P = (S_2 g h_2 - S_1 \times g \times h_1)}$$

ⓑ for vacum Pressure

Pressure above A-A in the left column = $\rho_2 g h_2 + \rho_1 g h_1 + P$

Pressure head in the right colum above A-A = 0

$$\rho_2 g h_2 + \rho_1 g h_1 + P = 0$$

$$\boxed{P = -(\rho_2 g h_2 + \rho_1 g h_1)}$$

(7a) Dimension of block = 3m × 2m × 0·8m , depth of immersion = h m

Spr gr wood = 0·45

Weight of wood piece = weigh density × volume

$= 0·45 \times 1000 \times 9·81 \times 3 \times 2 \times 0·8$

weight of water displaced = weight density of water × volume of wood submerged in water

$= 1000 \times 9·81 \times 2 \times 3 \times \cancel{0·8}$ 'h' m

for equilibrium weight of wood = weight of water displace

$\therefore 0·45 \times 1000 \times 9·81 \times 3 \times 2 \times 8 = 1000 \times 9·81 \times 2 \times 3 \times h$

$\Rightarrow h = 3·6$ ~~m~~

$AB = \frac{h}{2} = \frac{3·6}{2} = 1·8\,m$, $AG = \frac{0·8}{2} = 0·4\,m$

$BG = AG - AB = 0·4 - 1·8 = 1·4\,m$

$GM = \frac{I}{\forall} - B·G$, $I = \frac{1}{12} \times 3 \times (2)^3 = 2\,m^4$

$\forall$ = volume of wood in water = $3 \times 2 \times 3·6 = 21·6\,m^3$

$GM = \frac{2}{21·6} - 1·4 = 1·3\,m$

No plate velocity 'U' at 1-1

Mass of fluid per second through strip

$= \rho \times \text{velocity} \times \text{Area} = \rho \times U \times b \times dy \rightarrow (ii)$

Reduction in mass = mass/se (ii) - mass/sec (i)

$= \rho U b dy - \rho u b dy = \rho b (U-u) dy$

∴ Total reduction in mass of fluid, flowing B C

$= \int_0^\delta \rho b (U-u) dy = \rho b \int_0^\delta (U-u) dy$

plate displaced by δ^*, velocity (U)

$\rho \times V \times A = \rho \times U \times \delta^* \times b$

=

Eq (iii) & (iv) $\rho b \int_0^\delta (U-u) dy = \rho \times U \times \delta^* b$

Cancelling ρb both sides

$$\int_0^\delta (U-u)\, dy = U \times \delta^*$$

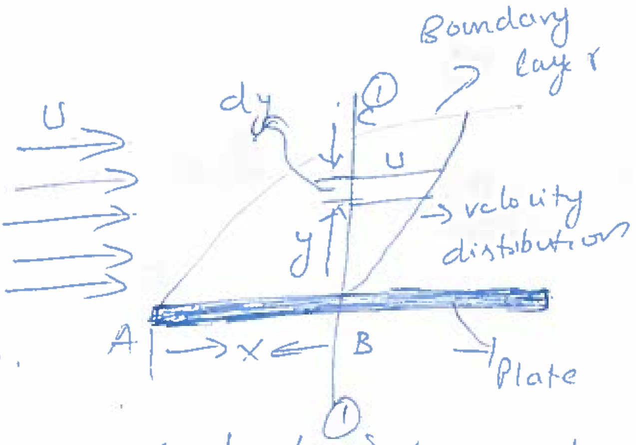
$$\delta^* = \frac{1}{U} \int_0^\delta (U-u) dy = \int_0^\delta \frac{(U-u) dy}{U}$$

$$\boxed{\delta^* = \int_0^\delta \left(1 - \frac{u}{U}\right) dy}$$

ANSWER KEY AND SCHEME OF EVALUATION

⑮ Displacement thickness (δ^*): Distance measured perpendicular to boundary layer by which the free stream is displaced due to the formation of boundary layer.

Consider flow of fluid having free-stream velocity equal to U over a thin smooth plate. At distance x from leading edge consider a section 1-1.



The velocity of fluid at B is zero and at 'C', which lies on the boundary layer is 'U'. Thus velocities varies from zero at B to U at 'C'. where BC is equal to thickness of boundary layer i.e

Distance BC = δ

At section 1-1, consider elemental strip

Then area of elemental strip
$dA = b \times dy$

y = distance of elemental strip from plate

dy = thickness of elemental strip

u = velocity of fluid at elemental strip

b = width of plate

Mass of fluid per second flowing through strip

$= \rho \times$ velocity $\times$ Area

$= \rho u \times dA = \rho u b\, dy \rightarrow ①$

frictional resistance = frictional resistance per unit wetted area per unit velocity × wetted area × (velocity)^2

$$F_1 = f' \times \pi d L \times V^2 \quad \left(\because \text{ wetted area} = \pi d \times L,\ \text{velocity} = V = V_1 = V_2\right)$$

$$= f' \times P \times L \times V^2 \quad \left(\because \pi d = \text{Perimeter} = P\right)$$

force acting on fluid b/w 1-1 & 2-2

① . Pressure force at section 1-1 = $P_1 \times A$ (A = Area of Pipe)

② Pressure force Sect 2-2 = $P_2 \times A$

③ frictional force F_1

Resolving all forces

$$P_1 A - P_2 A - F_1 = 0$$

$$(P_1 - P_2) A = F_1 = f' \times P \times L \times v^2$$

$$P_1 - P_2 = \frac{f' \times P \times L \times V^2}{A} \rightarrow ②$$

from eq ①

$$P_1 - P_2 = \rho g h_f$$

Equating ① & ②

$$\rho g h_f = \frac{f' \times P \times L \times V^2}{A}$$

$$h_f = \frac{f'}{\rho g} \times \frac{P}{A} \times L \times V^2 \rightarrow (iii)$$

eq (iii) $\frac{P}{A} = \frac{\pi d}{\pi/4\, d^2} = \frac{4}{d}$

$$h_f = \frac{f'}{\rho g} \times \frac{4}{d} \times L \times v^2 = \frac{f'}{\rho g} \times \frac{4 L V^2}{d} \rightarrow (iv)$$

Putting $\frac{f'}{\rho g} = \frac{f}{2}$,

where 'f' is know as coeff of Friction

Equation (iv) becomes as

$$\boxed{h_f = \frac{4 \cdot f}{2g} \cdot \frac{L V^2}{d} = \frac{4 f L V^2}{d \times 2g}}$$

$\downarrow$

Equation is Darcy - Weisbach eq

⑩ Darcy - Weisbach equation for head loss in case of Pipe flow :

Consider a uniform horizontal pipe, having steady flow as show ~~in~~ ~~fig~~. let 1-1 & 2-2 are two sections of Pipe. let P_1 = Pressure intensity at 1-1

V_1 = velocity of flow at 1-1

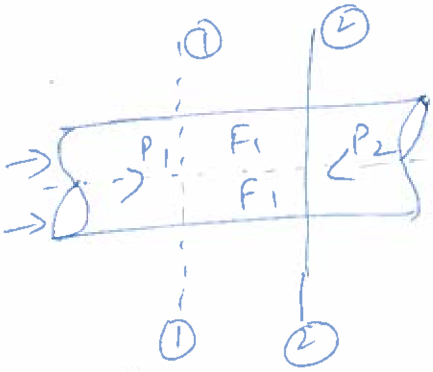
L = length of Pipe 1-1 & 2-2 Section

d - diameter

f' = frictional resistance per unit wetted area per unit velocity

h_f = loss of head due to friction

P_2, V_2 = are value of pressure intensity & velocity at sec 2-2



Apply Bernoullis 1-1 & 2-2

Total head 1-1 = Total head 2-2 + loss of head b/w 1-1 & 2-2

$$\frac{P_1}{\rho g} + \frac{V_1^2}{2g} + z_1 = \frac{P_2}{\rho g} + \frac{V_2^2}{2g} + z_2 + h_f$$

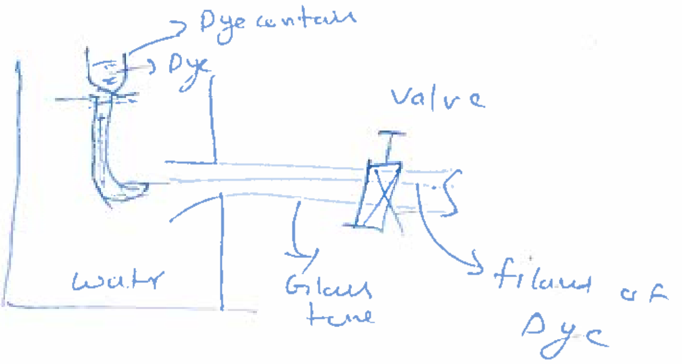
$z_1 = z_2 \Rightarrow$ Pipe horizontal

$V_1 = V_2$ as dia same 1-1 & 2-2

$$\frac{P_1}{\rho g} = \frac{P_2}{\rho g} + h_f \quad \text{(or)} \quad h_f = \frac{P_1}{\rho g} - \frac{P_2}{\rho g} \rightarrow (i)$$

⑪ Reynold's Experiment

$$\frac{\rho V \times d}{\mu}$$



Apparatus has

(i) A tank containing water at constant head

(ii) A glass tube having a bell-mouthed entrance at one end & regulating valve at other end

(a)

Dye filament → LAMINAR FLOW

(b)

wavy filament → TRANSITION

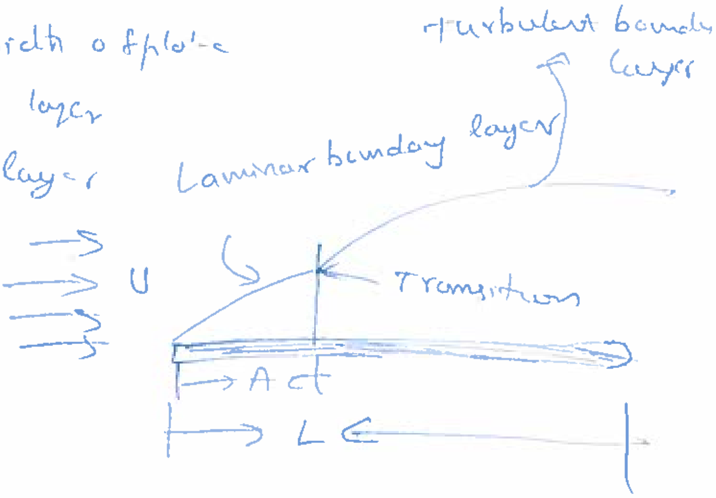
(c)

Diffused filament → TURBULENT FLOW

⑭ L = Total length of plate, b = width of plate

A = length of laminar boundary layer

L − A = length of turbulent boundary layer



① find the length from the leading edge up to which laminar boundary layer ends.

Done by $5 \times 10^5 = \frac{Ux}{\nu}$

② Drag using Blasius solution of laminar boundary layer

③ Drag Due to turbulent

④ drag due to turbulent boundary layer for a length A

$$= \boxed{② + ③ - ④ = \text{Drag due laminar for length } A + L{-}A}$$

ANSWER KEY AND SCHEME OF EVALUATION

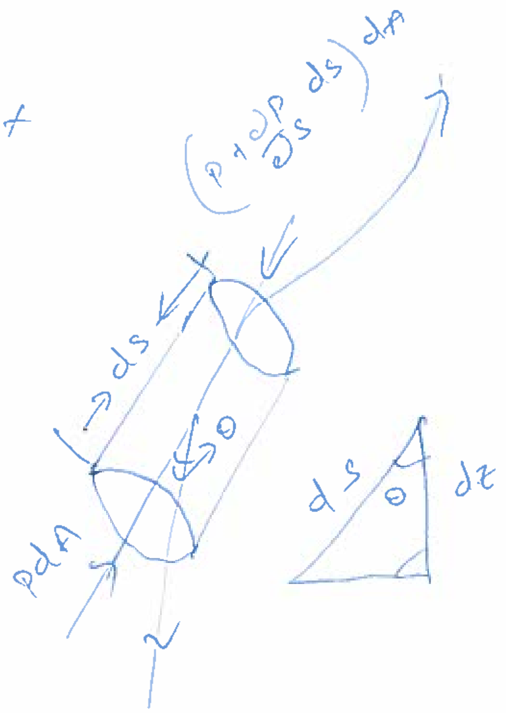
$\Rightarrow \frac{P}{\rho g} + \frac{v^2}{2g} + z = \text{constant}$

$\frac{P}{\rho g}$ = Pressure energy

$\frac{v^2}{2g}$ = kinetic energy

z = Potential energy

$a_s = \frac{dv}{dt}$, where 'v' is a function of s & t



$$= \frac{\partial v}{\partial s}\frac{ds}{dt} + \frac{\partial v}{\partial t} = \frac{v\partial v}{\partial s} + \frac{\partial v}{\partial t}$$

$$\left\{\because \frac{ds}{dt} = v\right\}$$

flow stead, $\frac{\partial v}{\partial t} = 0$

$$a_s = \frac{v\partial v}{\partial s}$$

Substituting value of a_s in eq (6.2)

$$-\frac{\partial p}{\partial s} ds\, dA - \rho g\, dA\, ds \cos\theta = \rho\, dA\, ds \times \frac{v\partial v}{\partial s}$$

Dividing by $\rho\, ds\, dA$, $-\frac{\partial p}{\rho \partial s} - g\cos\theta = \frac{v\partial v}{\partial s}$

(or) $$\frac{\partial p}{\rho\partial s} + g\cos\theta + v\frac{v\partial v}{\partial s} = 0$$

but $\cos\theta = \frac{dz}{\partial s}$

$$\therefore \frac{1}{\rho}\frac{\partial p}{\partial s} + g\frac{dz}{\partial s} + \frac{v\partial v}{\partial s} = 0 \text{ (or) } \frac{\partial p}{\rho} + g\,dz + v\,dv = 0$$

(or) $\frac{\partial p}{\rho} + g\,dz + v\,dv = 0$ → Euler's eq

Bernoulli's eq obtained by integrating Euler's

$$\int \frac{dp}{\rho} + \int g\,dz + \int v\,dv = \text{Constant}$$

If flow incompressible, ρ is constant

$$\frac{p}{\rho} + gz + \frac{v^2}{2} = \text{Constant}$$

⑨ consider stream line in which the force due to gravity and pressure are taken into consideration flow takes place in s-direction. Consider cylindrical element of cross-section dA & length ds. forces acting are

① Pressure force PdA in direction of flow

② Pressure force $\left(P + \frac{\partial P}{\partial s} ds\right) dA$ opposite to the direction of flow

③ weight of element $\rho g\, dA\, ds$.

let 'θ' is the angle b/w the direction of flow and the line of action of the weight of element.

Resultant force on the fluid element in the direction 's' must be equal to mass of fluid element × acceleration in direction 's'

$$= PdA - \left(P + \frac{\partial P}{\partial s} ds\right) dA - \rho g\, dA\, ds \cos\theta \times a_s$$

$$= \rho\, dA\, ds \times a_s$$

where 'as' is the acceleration in direction of s'

(13) (a) unit quantities

(i) unit speed (2) unit power (3) unit discharge

In order to predict the behaviour of turbine working under varying condition of head, speed, output & gate opening result expressed in unit quantities

(1) unit speed = Defined as speed of turbine working under unit head (under head 1m)

N_u = Exp for unit speed

N = Speed of turbine under H

H = Head under which turbine works

u = Tangential velocity

$u \propto V$ & $V \propto \sqrt{H}$ & $u = \frac{\pi D N}{60}$

$u \propto N$ (or) $N \propto u$ (or) $N \propto \sqrt{H}$ & $N = K_1\sqrt{H}$

$H = 1, N = N_u$

$N_u = K_1\sqrt{1.0} = K_1$ & $N = N_u\sqrt{H}$ (or)

$$\boxed{N_u = \frac{N}{\sqrt{H}}}$$

(2) unit discharge $Q_u = \frac{Q}{\sqrt{H}}$

(3) unit power = $P_u = \frac{P}{H^{3/2}}$

β angle ,

a = Area of Jet $= \frac{\pi}{4} d^2$

work done by jet to runner =

$$F_x \times u = \rho a V_1 (V_{w1} + V_{w2}) \times u \ \text{N m/s}$$

Power given to runner by jet $= \frac{\rho a V_1 (V_{w1} + V_{w2}) \times u}{1000}$ kW

Work done $= \frac{\rho a V_1 \left[V_{w1} + V_{w2}\right] \times u}{1000}$ kW

Work done per unit wt. $= \frac{\rho a V_1 (V_{w1} + V_{w2}) \times u}{\rho a V_1 \times g}$

K.E $= \frac{1}{2}(\rho a V_1 \times \underline{V_1^2}$

H.E $= \eta_h = \frac{\text{work done}}{\text{K.E}}$

$$\eta_h = \frac{\rho a V_1 \left[V_{w1} + V_{w2}\right] \times u}{\frac{1}{2}\left[\rho a V_1\right] \times V_1^2} = \frac{2\left[V_{w1} + V_{w2}\right] \times u}{V_1^2}$$

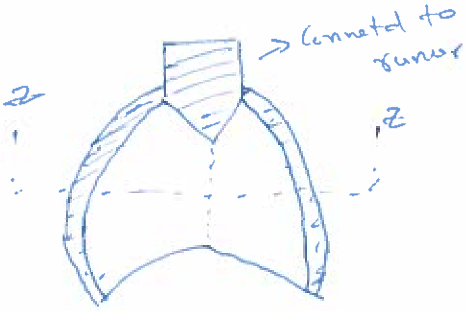
$V_{w1} = V_1$, $V_{r1} = V_1 - u_1 = (V_1 - u)$, $V_{r2} = (V_1 - u)$

$V_{w2} = V_{2r} \cos\phi - u_2 = V_{r2} \cos\phi - u = (V_1 - u)\cos\phi - u$

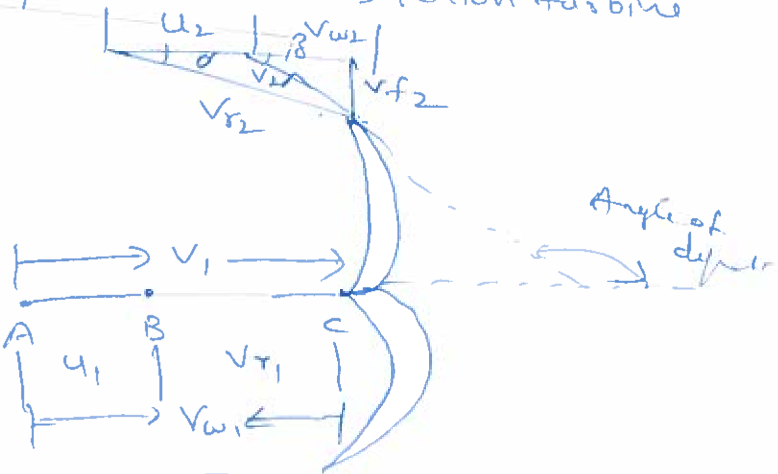
Sub V_{w1} & V_{w2}

$$\boxed{\eta_h = \frac{2(V_1 - u)\left[1 + \cos\phi\right] u}{V_1^2}}$$

⑫ Tangential flow Impulse turbine → Pelton turbine



ⓐ Splitter



ⓑ velocity triangles

H = Net head acting on Pelton wheel

N = Hg − hf

Hg = Gross head & $h_f = \frac{4fLV^2}{D\times 2g}$

D^* = Dia of Penstock

D = Dia of wheel

N = Speed of wheel in r.p.m

d = Dia of jet

V_1 = velocity of jet at inlet = $\sqrt{2gH}$

$$u = u_1 = u_2 = \frac{\pi DN}{60}$$

The velocity Δle at inlet ab will be strt whe

$$V_{r1} = V_1 - u_1 = V_1 - u$$

$V_{w1} = V_1$, $\alpha = 0$ & $\theta = 0$

velocity Δle outlet

$V_{r2} = V_{r1}$ & $V_{w2} = V_{r2}\cos\phi - u_2$

force exerted by jet of water

$$F_x = \rho a V_1 (V_{w1} + V_{w2})$$

⑧ ⓐ Types of fluid flows

(i) Stead & unsteady flow

(ii) uniform & non uniform flows

(iii) Laminar & turbulent

(iv) compressible & incompressible flows

(v) Rotational & Irrotational flows

(vi) one, two & three dimensional flows

8ⓑ $\phi = y^2 - x^2 - xy\frac{3}{2} + x^3\frac{y}{2}$

finding $\frac{\partial\phi}{\partial x} = -2x - \frac{3}{2}y + 3x^2.\frac{y}{2}$ &

$\frac{\partial\phi}{\partial y}$

ⓐ7b

→ Dimension 1m × 3m,

→ Plane surface angle of 300

→ Total pressure =

→ Depth of pressure when plate is 2m below the free surface

Signature of faculty

Signature of HoD (CE)

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022-2023

Degree	B. Tech. (U. G.)	**Program**	Mechanical Engineering			**Academic Year**	2022 - 2023
Course Code	20ME305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Manufacturing Process**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four properties required for a moulding sand.	20ME305.1	L1
2	What are the functions of risers?	20ME305.2	L1
3	State the principle of liquid penetrant testing of welds.	20ME305.3	L1
4	List any four forging defects.	20ME305.4	L1
5	Recall spring back and it remedies.	20ME305.5	L2

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 advantages and disadvantages of different pattern materials?	6M	20ME305.1	L1
6 (b)	List out the steps involved in making a casting. Explain with the help of neat sketches.	6M	20ME305.1	L2
	OR			
7 (a)	Why is it desirable to make the pattern allowances as small as possible?	6M	20ME305.1	L1
7 (b)	Explain the different types of gating systems. Mention advantages and demerits of each.	6M	20ME305.1	L2
8 (a)	Make a list of safety considerations and precautions that should be taken concerning all aspects of melting and casting of metals, including equipment involved.	6M	20ME305.2	L1
8 (b)	Sketch and explain the construction and operation of hot chamber die casting process.	6M	20ME305.2	L2
	OR			
9 (a)	Is there any difference in the tendency for shrinkage void formation for metals with short and long freezing ranges, respectively? Explain.	6M	20ME305.2	L2
9 (b)	Explain the different zones in cupola.	6M	20ME305.2	L2
10 (a)	Describe the difference between brazing and soldering.	6M	20ME305.3	L2
10 (b)	What are the parameters that control the weld quality in manual metal-arc welding?	6M	20ME305.3	L1
	OR			
11 (a)	Explain the difference between TIG welding and MIG welding.	6M	20ME305.3	L2
11 (b)	Describe with the help of a neat sketch the principle of spot welding.	6M	20ME305.3	L2
12 (a)	Explain the features of different types of rolling mills.	6M	20ME305.4	L2
12 (b)	How are seamless tubes produced? Explain with neat diagram.	6M	20ME305.4	L2
	OR			
13	Distinguish between forward extrusion and backward extrusion processes.	12M	20ME305.4	L2
14	Explain blanking and piercing operations.	12M	20ME305.5	L2
	OR			
15	Explain the working principle of electromagnetic forming. Discuss the various advantages and applications.	12M	20ME305.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY
(AUTONOMOUS)
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Degree	B.Tech (U.G.)			Year	II	**Academic Year**	2022 - 2023
Course Code	20ME305	**Test Duration**	3 Hrs	Max. Marks	70	**Semester**	III
Course	**Manufacturing Process**						

Part A

No.	Answers	Marks
1	Strength, Permeability, Moisture Content, Flowability, Grain Size, Grain Shape, Collapsibility. Refractory Strength.	Any 4 properties - each - 1M
2	Riser is provided to act as a reservoir of molten metal to compensate for shrinkage of casting.	Function of riser -2M
3	The basic principle of liquid penetrant testing (PT) is capillary action, which allows the penetrant to enter in the opening of the defect, remain there when the liquid is removed from the material surface, and then re-emerge on the surface on application of a developer, which has a capillary action similar to blotting paper. Liquid penetrant testing can be performed on any solid, non-porous material, such as metals, ceramics or plastics. It is commonly used to detect defects in castings, forgings and welded parts, for example, to discover flaws in corrosion tests on materials	principle of liquid penetrant testing – 2M
4	Unfilled Section, Cold Shut, Scale Pits, Die Shift, Flakes, Surface Cracking	Any 4 Defects - each - 1M
5	Springback is the geometric change made to a part at the end of the forming process when the part has been released from the forces of the forming tool. Upon completion of sheet metal forming, deep-drawn and stretch-drawn parts spring back and thereby affect the dimensional accuracy of a finished part. Remedies are Overbending, Re-bending, Increased forming force	Spring back and its remedies -2M

Part B

No.	Answers	Marks
6a	**Pattern:** Pattern is the shape of object, which is made by different material to get a required output from it. To prepare a pattern there are different materials available which is explained below. The materials that used in pattern making has different properties and advantages. **Pattern Materials:**	Any 4 properties - each - 1M

Pattern materials are different types according to the output required, pattern materials are used in making patterns. According to the required dimensions and surface finish pattern materials are selected.

Different types of materials are:

Wood Patterns:

Wood is the most common material used in making a pattern, this type of material is easy in making patterns. to make wood into a pattern,it should dried properly.

Advantages of Wood Patterns:

- Wood is highly available in nature and less cost.
- Making patterns is easy.
- We can obtain to a high surface finish easily with less effort.
- Wood can construct into different forms easily.
- They have comparatively more strength to weight ratio.
- This type of pattern material is easy to machine.

Disadvantages of Wood Patterns:

- Wood material is not high accuracy in dimensions we required.
- This material should handle with care and smooth.
- It has less wear resistance.
- Wood has low abrasion resistance.

Metal Patterns:

Metal patterns are one of the strong and hard patterns and these patterns are more costly than different types pattern materials.

Advantages of Metal Patterns:

- Metal patterns are comparatively more strong.
- We can obtain high surface finish.
- Metals can obtain to high dimensional accuracy.
- They are more resistant to wear, corrosion etc..
- Metals can be stored for long periods than wood.

Disadvantages of Metal Patterns:

- Metals are more weight, it is not easy to handle.
- They are high in cost compared to many pattern materials.
- Pattern should take care to avoid any damages like corrosion, deformation etc.

Plastic Patterns:

A plastic material is high in use, but they are not safe and not used at high temperatures.

Advantages of Plastic Patterns:

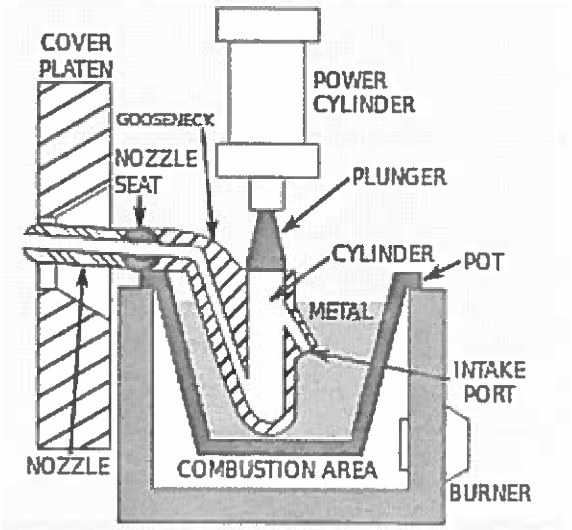
- Pattern making with plastic is low in weight.
- Plastic can obtain to high accuracy in dimensions.
- This type of patterns are strong and smooth.
- Plastic patterns have not to effect from nature like corrosion, rust, moisture etc..
- This type of patterns has a long life without protection.
- Handling is easy.

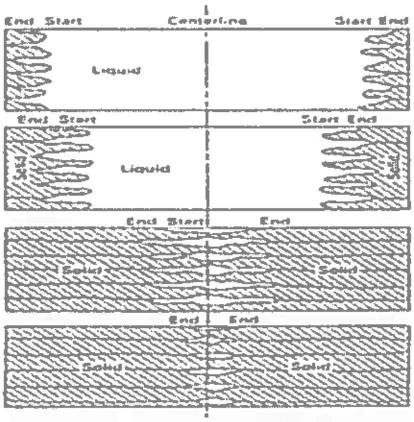
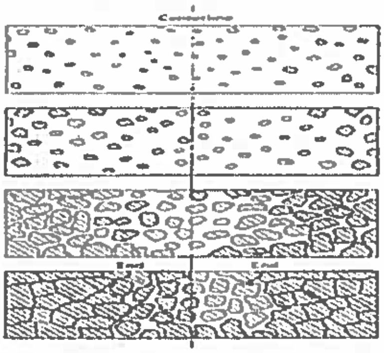
Disadvantages of plastic Patterns:

- Plastic patterns are not useful at higher temperatures.
- Stronger plastic materials are more costly to use.

6b	**Basic Steps in Casting Process** Patternmaking, Coremaking, Molding, Melting and pouring, Finishing, Patternmaking. To create a casting mold, a manufacturer must first design a physical model. The process of fabricating this model is called patternmaking. Using computer-assisted design (CAD) systems, the manufacturer designs dimensions and geometry of a mold, and then packs an aggregate material, such as sand, concrete or plastic, around the pattern. Once the pattern is removed, the mold cavity in the sand can be filled. **Coremaking** Many part designs require the inclusion of cores in the casting mold. Cores are solid materials placed inside the mold cavity to create interior surfaces of a casting. For example, a metal pipefitting will require a cylindrical core inside the mold cavity to create the hollow construction of the component's interior. **Molding** At this point, the manufacturer can create the casting mold. A material such as sand, plaster or wax is used in expendable mold casting, whereas metal and other durable materials are used in non-expendable mold casting techniques. The material fills the casting mold model and is allowed to harden, at which point the manufacturer removes it from the cavity and the casting of the component can now begin. **Melting and Pouring** Metal must be properly melted prior to being placed in the mold. Typically, this is done by using what is known as a crucible. Crucibles are containers made of porcelain or another melt-resistance substance in which a manufacturer can heat a metal beyond its melting point. Once properly melted, the molten metal is poured into the casting mold to cool and harden. **Finishing** Because metal can sometimes fill in cracks in a casting mold or sprues, the pouring channel for the mold, manufacturers must often finish the metal following casting. This can be accomplished through a variety of finishing techniques, including sanding, grinding and buffing. Once proper appearance and surface texture has been achieved, further post-treatment processes such as painting or electroplating may be necessary for some applications.	6
7 a	There are different types of pattern allowances: (i) Draft allowance:- The pattern needs to be removed from each mold it shapes without breaking or distorting it. The draft is a taper that facilitates pattern removal. The exact angle of the taper depends on the complexity of the pattern, the mold type, and surface type. (ii) Shrinkage allowance:- Shrinkage allowance compensates for the amount that a metal will shrink during Cooling. The precise allowance depends on the metal being cast. (iii) Distortion allowance:- Patterns may be intentionally distorted to compensate for expected cooling distortion. (iv) Machining allowance:- Some castings are finished by machining. The patterns for machine-finished castings intentionally include excess material to compensate for material that will be lost in the finishing stage. It is desirable to make pattern allowances as small as possible for easy removal of the product and a good surface finish.	6

7 b	**Types of Gating Systems:** The gating systems are of two types: • Pressurized gating system • Un-pressurized gating system **Pressurized Gating System** • The total cross sectional area decreases towards the mold cavity • Back pressure is maintained by the restrictions in the metal flow • Flow of liquid (volume) is almost equal from all gates • Back pressure helps in reducing the aspiration as the sprue always runs full • Because of the restrictions the metal flows at high velocity leading to more turbulence and chances of mold erosion **Advantages** 1. Good liquidity. 2. Simple construction. 3. Wide range of applicable resins. 4. Material filling is good. 5. Surface of molded products shrinks less. 6. Omit processing of flow path. 7. Less pressure loss. 8. It can form large or deeper molded products. **Disadvantages** 1. Only one molded product can be formed at a time, and it is not possible to take several multi-point gates unless a multi-nozzle molding machine is used. 2. There are traces of gate residue that affect appearance and increase post-processing. 3. Flat and shallow molded articles are easy to warp and twist. 4. Gate cycle must be determined. 5. Residual stress near gate is large, which may cause cracking or deformation. **Un-Pressurized Gating System** • The total cross sectional area increases towards the mold cavity • Restriction only at the bottom of sprue • Flow of liquid (volume) is different from all gates • aspiration in the gating system as the system never runs full • Less turbulence **Advantages** 1. Low residual stress. 2. Gate size is correct (rectangular section). 3. Gate is easy to separate from the molded product. 4. Prevent material from flowing backwards. 5. Part of gate generates frictional heat, which can raise temperature of material again and promote filling. **Disadvantages** 1. Flow resistance is large. 2. Pressure loss is large. 3. Materials with poor fluidity can easily cause insufficient filling or halfway curing. 4. A flat or large-sized molded products is likely to cause bubbles or flow marks due to small size of gate	6

8 a	*Personal Protection Equipment* When coming in close proximity to these molten metals it doesn't take much to get an extremely serious burn. Even if the metal doesn't actually touch your skin, it can cause some serious problems. This is why heat resistant personal protection equipment is essential for proper casting safety. The following are some key elements of this equipment: • **Gloves** – You must have thick gloves that can shield your hands from the heat. They should also provide protection in the event that any of the molten metal comes in contact with them. • **Boots** – When metal bubbles up or splatters it most often falls down near your feet. If it lands on normal shoes or boots, it can burn right through them to your feet. Even if it hits the floor, you could accidently step on it while it is still extremely hot. This is why heat resistant boots are essential. • **Long Pants & Sleeves** – Wearing long thick pants and long sleeves that can help protect you from the heat is important. While it may seem like the long clothes would keep you hot, they can actually help to shield you from the scorching heat used in the casting process. • **Face Protection** – If you're working directly with the metal, having a face shield to ensure it doesn't splatter up onto your face is very important. • **Protective Apron** – In almost all cases it is best to have a heavy apron that will help protect your body.	6
8 b	Hot-chamber die casting, also known gooseneck machines, rely upon a pool of molten metal to feed the die. At the beginning of the cycle the piston of the machine is retracted, which allows the molten metal to fill the "gooseneck". The pneumatic- or hydraulic-powered piston then forces this metal out of the gooseneck into the die. The advantages of this system include fast cycle times (approximately 15 cycles a minute) and the convenience of melting the metal in the casting machine. The disadvantages of this system are that it is limited to use with low-melting point metals and that aluminium cannot be used because it picks up some of the iron while in the molten pool. Therefore, hot-chamber machines are primarily used with zinc-, tin-, and lead-based alloys. 	6
9 a	Short freezing range alloys are closely similar but in these instances the front is not perfectly plain but is serrated as illustrated in figure, which shows a strong tendency toward skin formation, and the front of the crystals solidifying inward (start of freeze) will not advance much	6

	faster than their bases (end of freeze). Such relative, short crystalline growth helps keep liquid feed metal in contact with all solidifying surfaces. Such strong progressive solidification in these short freezing range alloys promotes the development of directional solidification along any temperature gradients in the solidifying casting figure.  For long freezing range alloys, the development of directional solidification is difficult. Although a thin skin may initially form on the mold walls, almost as soon as the crystallites are formed, their growth becomes drastically inhibited. The first-formed crystals are much lower in alloying elements than the liquid metal from which they are formed. Hence many of the atoms of alloying elements present in the liquid metal from which crystallites are formed are expelled into the liquid surrounding the crystallites. These surrounding liquids thus become enriched in alloying element and this enrichment considerably depresses the freezing point of the liquid. If freezing were very slow, so that ample time is available for diffusion of the excess atoms from this enriched layer of liquid into remainder of the liquid metal, the crystallites would continue to grow and a columnar structure would result. However, in most practical situations there is not nearly sufficient time for these concentrations in the liquid to be dissipated by diffusion so that the liquid metal surrounding the crystallites is unable to freeze till a much later stage during freezing of the casting as a whole. As illustrated in figure, the material at first being fluid, then mushy and finally rigid, It has been estimated that in many alloys, rigidity is not established until the casting is about 60-70% solid. This mushy or pasty mode of solidification results in the development of numerous small channels of liquid metal late in solidification. Feeding through these channels is restricted, and dispersed shrinkage porosity occurs throughout the casting. 	
9 b	**Different zones in Cupola furnace are:-** **Legs:** Legs are provided for supporting purposes. **Slag Hole or Slag spout:** The slag hole is used for removing or extracting the slag from the melting iron. **Sand Bed:** This is in taper form and from this, the melted iron comes out easily. **Tuyeres:** By tuyeres, we enter the gas to the proper burn of fuel.	6

<table>
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<td>

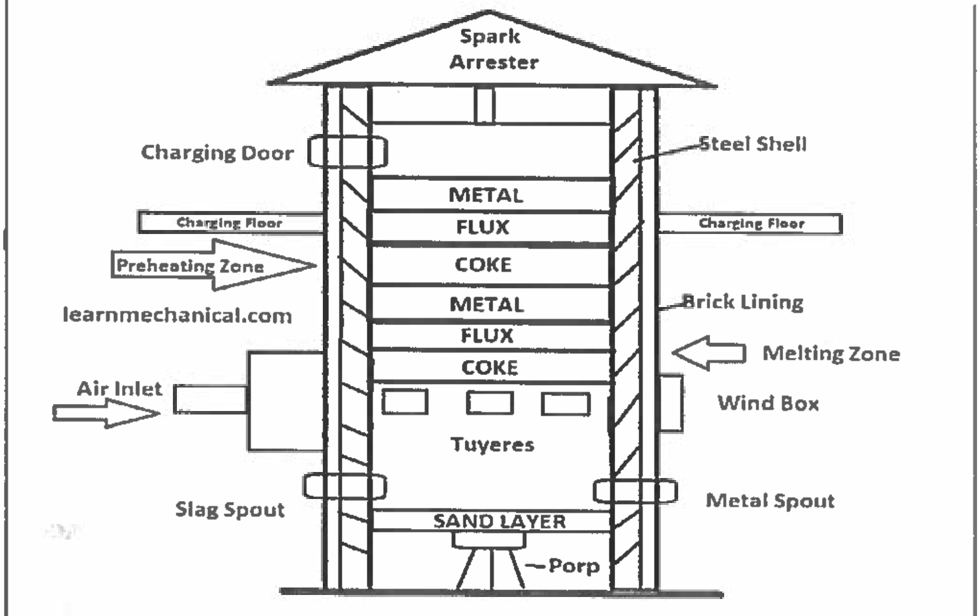
Preheating Zone: In the Preheating zone, the heating process started and heats the metal charge about 1090 degrees Celsius.

Melting Zone: In the melting zone, we do not provide much heat to melt the metal charge because it's already melted in the preheating zone with a temperature of about 1090 degrees Celsius.

Charging door: From here we supply the charge to the furnace. The various charges are for the cupola furnace are Pig Iron, Coke and limestone.

Brick lining and Steel shell: The shell of the cupola furnace is being usually made of steel and it's called a steel shell.

Spark Arrester: This device used in the system for preventing the emission from the fireplace.



CUPOLA FURNACE

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<tr>
<td>10 a</td>
<td>
<table>
<tr><th>Soldering</th><th>Brazing</th></tr>
<tr><td>Soldering is a metal-joining process where the melting temperature of the filler metal is relatively low.</td><td>Brazing is a metal-joining process where the melting temperature of the filler metal is usually above 450°C.</td></tr>
<tr><td>The most common type of filler metal used in soldering is a 60:40 tin: lead alloy.</td><td>Brass alloys are commonly used in brazing.</td></tr>
<tr><td>Soldering occurs at temperatures around 400°C.</td><td>Brazing occurs at temperatures above 450°C.</td></tr>
<tr><td>When the gap between the metals parts is not a fine gap, soldering can be used.</td><td>Brazing is used to cover a fine gap between the metal parts to be joined.</td></tr>
<tr><td>Soldering is a softer metal-joining process where the metals parts are not held very tightly.</td><td>Brazing creates a tight fit between the metal parts joined.</td></tr>
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<td>6</td>
</tr>
<tr>
<td>10 b</td>
<td>The parameters that control the weld quality in manual metal arc welding are welding current, arc voltage, welding speed, heat input rate. Increase in current and voltage is responsible for the amount of heat generation which in turn affects the weld quality. The welding speed is responsible for the joint formation. More speed leads to insufficient local heating whereas low speed leads to high heat. Both the conditions are undesirable. Hence. an optimum value of the parameters should be maintained in order to attain a sound joint.</td>
<td>6</td>
</tr>
<tr>
<td>11 a</td>
<td>MIG is generally seen as being easier to learn and perform as well as being faster and better for welding thick materials. However, TIG welding offers greater control and precision, is better for</td>
<td>6</td>
</tr>
</table>

thinner materials and offers neater welds with little finishing required.

Aside from these general differences, there are a number of key differences that can be categorised according to different properties, cost, ease of learning, and more:

1. Weld Strength

TIG welded joins are typically stronger than those produced by MIG welding. This is because the narrow, focused arc created by TIG welders offers better penetration of the metal. In addition, the TIG weld beads, when applied correctly, contain few holes and other defects that can weaken the weld. Despite this generalisation, MIG welds can still produce strong welds with good penetration by grinding or cutting a V-shaped groove into the joint before starting to weld to increase penetration. Good travel speed and torch positioning will also improve the weld strength of MIG welds.

2. Weld Speed

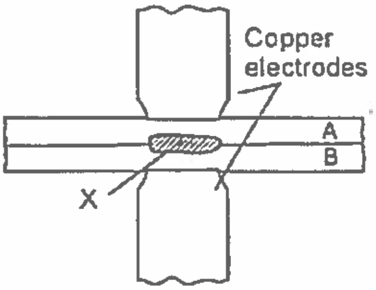
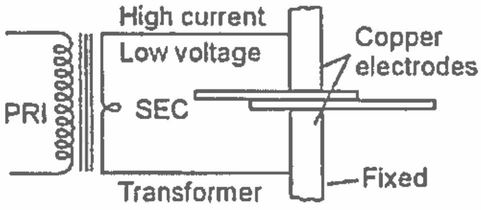
MIG welders typically provide faster welding speeds in a production setting.This is because air-cooled MIG welders auomatically feed filler material into the weld pool and have a rounder and broader arc that dissipates heat better. This allows welders to move the weld puddle faster and make longer runs without overheating.By contrast, TIG welders cannot move the weld puddle as fast or supply enough filler rod to compete with MIG welding speeds. In addition, the air-cooled torches used in TIG welding get too hot during lengthy welding runs, meaning they need to cool or be swapped for more expensive water-cooled torches.

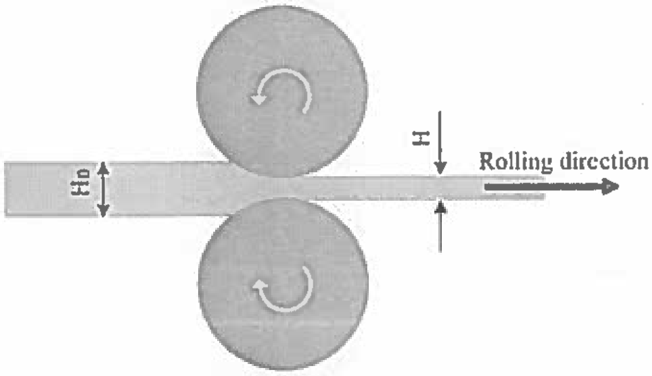
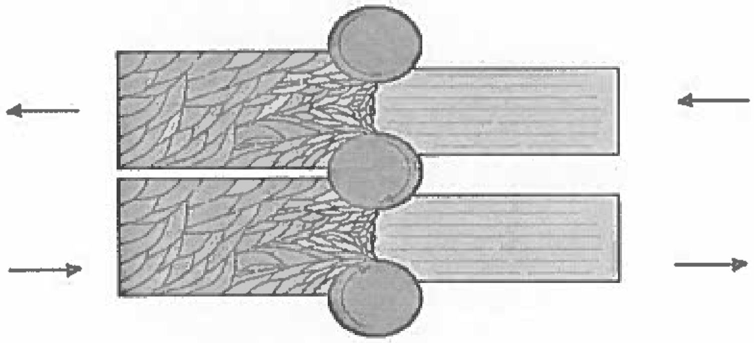
3. Shielding Gas

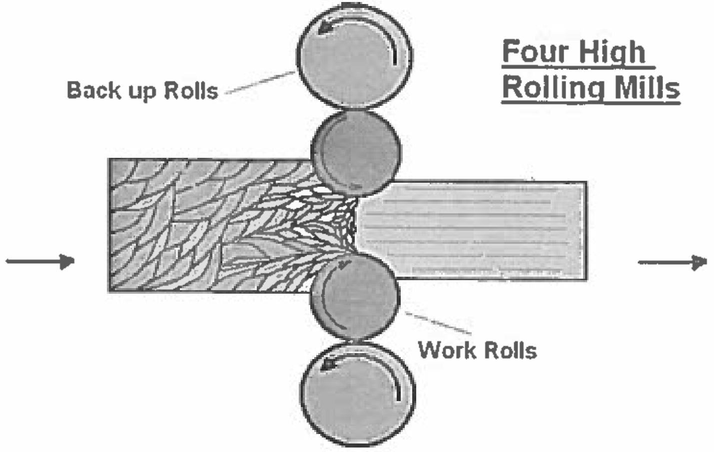
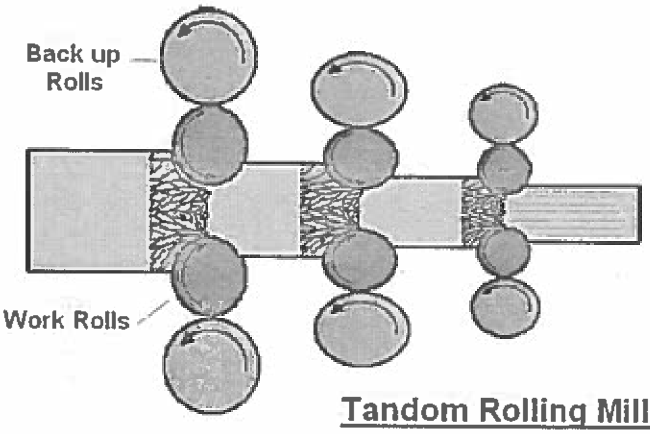
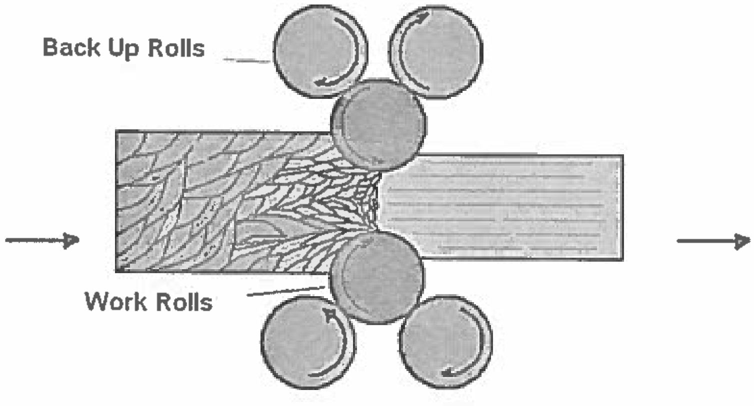
Both MIG and TIG welding use a shielding gas ensure quality welds. The shielding gas protects the weld puddle from reactive gases found in the air that can cause impurities in the weld. TIG welds typically use pure argon gas to protect the weld as the tungsten electrode is more sensitive to reactive gases like oxygen or CO_2. MIG welds are typically performed with a blend of argon and CO_2 (typically 75/25%), as the small addition of carbon dioxide stabilises the arc and aids penetration. There are exceptions to these general shielding gas rules, depending on the application. TIG welding may sometimes use a blend of argon with helium, hydrogen or nitrogen, while MIG welding is performed with 100% pure argon when welding aluminium and can be performed with pure CO_2 to save costs and increase weld penetration. The two techniques also use different shielding gas flow rates, with MIG typically using 35 to 50 cubic feet per hour and TIG welding operating a shielding gas flow of 15 to 25 cubic feet per hour.

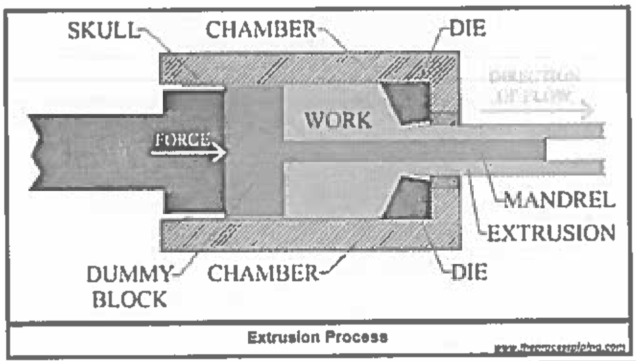
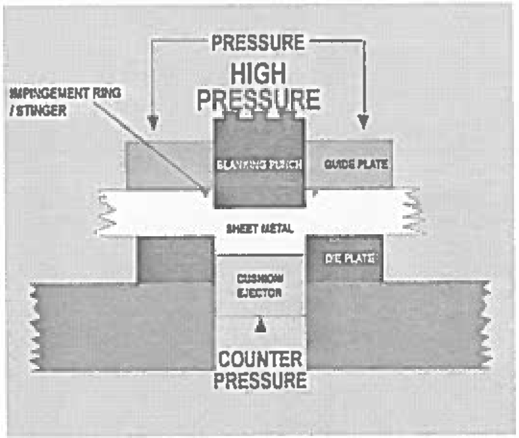
4. Weld Aesthetics

TIG welds tend to show better aesthetic qualities than MIG welds. With very little or no spatter, TIG welds usually only require light polishing to finish and remove any discolouration. TIG welded stacks of 'coins' are often seen as the most aesthetically-pleasing weld and can serve to make unpainted welds (in steel or aluminium, for example) look their best. By contrast, MIG welds have a less desirable appearance, even though an experienced welder can still create good-looking MIG weld beads. Despite the aesthetic differences, MIG welds are often fine for applications where appearance is less important or where the welds are to be coated, covering

	the appearance of the joint. **5. Process Difficulty** MIG welding is much easier to learn and master than TIG welding. TIG welding requires the use of two hands, one to move the welding torch and the other to feed the filler rod into the weld pool. Plus, there is often a foot pedal with which you can control the amperage. While these various movements allow for greater control, they can be difficult to master. The metals to be joined must also be cleaned and prepared well for TIG welds and welders tend to consider TIG as a more advanced process. MIG welding, by contrast, is much easier to learn. There is no foot pedal to master and the filler material is automatically fed through the welding gun, meaning that you only need to use one hand to complete the weld. **6. Cost** TIG welding costs more per foot of bead than MIG welding. This is due to the lower deposition rates associated with TIG as well as the need for more experienced, and thereby more expensive, welders. In addition, TIG welding requires more prep work, which also adds to the cost. Finally, MIG welding supplies and machines tend to be less expensive than TIG.All of these factors combine to make TIG welding more expensive than MIG welding.	
11 b	In this method the heating effect occurs when a current flows through a resistance. A transformer supplies A.C. at low voltage and high current to the electrodes. The spot weld is made by overlapping the parts and gripping the overlapping sections between two electrode points, through which the current is passed for localised heating. After the current is switched-off, the molten metal solidifies, leaving a welded spot  Fig. 24.1  Two copper electrodes are pressed against the steel sheet plates squeezing them together. The electrical resistance is greatest at the interface where the sheets or plates are in contract when a large current at low voltage is passed between the electrodes through the sheets. For any given joint between two sheets a suitable time is selected and the current varied until a sound weld is obtained. If welds are made near each other some of the current is shunted through the adjacent weld. Once current and time are set, other welds will be of consistent	6

	quality.	
12 a	As per the requirement of the process and arrangement of the rolls, the rolling mill rolls can be divided into the following categories: **1.Two High Rolling Mills** It consists of two rollers, which rotate in the opposite direction for the desired movement of the workpiece. The workpiece is fed between the rollers, which apply a full force, and tends to deform a workpiece and convert it into the desired shape. If you want a robust and quality two high rolling mill, you can look for mill rolls manufacturers to know which one suits your purpose. The two high rolling mills are further divided into two more categories namely high non-reversible machine in which rollers rotate in only one direction, and the workpiece can be fed in only one direction. On the other hand, the second one high reversible machine, which both rollers rotate in both directions.  Two high rolling mill **2.Three High Rolling Mills** The Three High Rolling Mills comprises of a roll stand with three parallel rolls one above another. The adjacent rolls rotate in the opposite direction to pass the material between the top and middle roll in single direction, and the bottom and central roll in the opposite direction. The workpiece is rolled on both forward and return passes. The workpiece passes through the bottom and intermediate rolls, and returns between the middle and top rolls. Various steel roll manufacturers provide top-quality roll mills to meet every type of industrial requirement.  Three High Rolling Mills **3.Four High Rolling Mills** The Four High Rolling Mills have a roll stand with four parallel rolls placed one above another. The top and bottom rolls work in the opposite direction. The two in the middle are smaller than the top and bottom rolls, which are also known as backup rolls.	6

	 Four High Rolling Mills **4.Tandem Rolling Mills** The Tandem Rolling Mills comprises of a set of two or three strands of roll set in parallel alignment. A continuous pass might be possible through each one with the change in the direction of the material. Many mill rolls manufacturers provide quality tandem rolling mills to various industries.  Tandom Rolling Mills **5.Cluster Rolling Mills** Cluster Rolling Mills is a first four high rolling mills, where each of the working rolls is backed up by two or more larger rolls for rolling hard material. At times, one might need to employ work rolls of minimum diameter.  Cluster Rolling Mills	
12 b	Seamless tubes are as defined – they do not have a welded seam. The tubing is manufactured through an extrusion process where the tube is drawn from a solid stainless steel billet and extruded into a hollow form. The billets are first heated and then formed into oblong circular molds that are hollowed in a piercing mill. While hot, the molds are drawn through a mandrel rod and elongated. The mandrel milling process increases the molds length by twenty times to form a seamless tube shape. Tubing is further shaped through pilgering, a cold rolling process, or	6

<table>
<tr>
<td></td>
<td>cold drawing.

Extrusion Process</td>
<td></td>
</tr>
<tr>
<td>13</td>
<td>
<table>
<tr><th>Forward Extrusion</th><th>Backward Extrusion</th></tr>
<tr><td>1.Simple, but the material must slide along the chamber wall.</td><td>1. In this case, material does not move but die moves.</td></tr>
<tr><td>3. High extrusion forces required but mechanically simple and uncomplicated.</td><td>3. 25–30% less extruding force required as compared to direct extrusion. But hollow rarequired limited application.</td></tr>
<tr><td>4. High scrap or material waste—18–20% on an average.</td><td>4. Low scrap or material waste only 5–6% of billet weight.</td></tr>
<tr><td>2. High friction forces must be overcome.</td><td>2. Low friction forces are generated as the mass of material does not move</td></tr>
</table>
</td>
<td>12</td>
</tr>
<tr>
<td>14</td>
<td>**Blanking and piercing:**
Blanking and piercing are shearing processes in which a punch and die are used to produce parts from coil or sheet stock. Blanking produces the outside features of the component, while piercing produces internal holes or shapes. The terms "piercing" and "punching" can be used interchangeably.
 </td>
<td>12</td>
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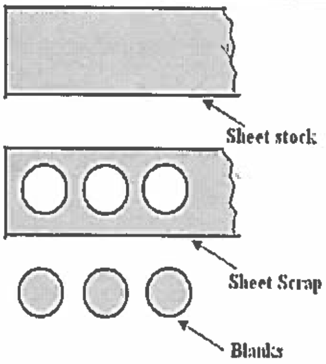
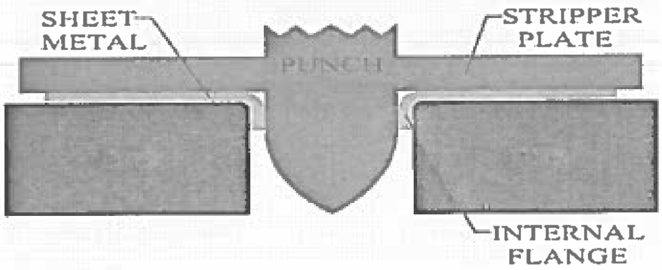
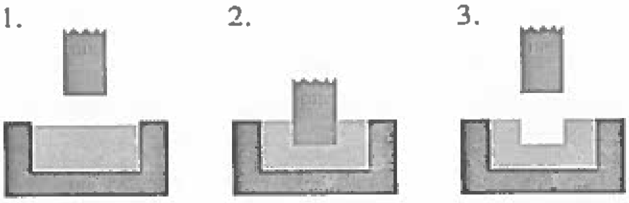


BLANKING

1. Blanking and piercing are manufacturing processes by which certain geometrical shapes are sheared off a sheet metal.
2. If the sheared off part is the one required, the processes referred to as blanking and if the remaining part in the sheet is the one required, the process is referred to as piercing.
3. these processes are reviewed and discussed The main parameters affecting these processes are presented and discussed.
4. These include: the radial clearance percentage, punch and die geometrical parameters, for example punch and die profile radii.
5. The abovementioned parameters on the force and energy required to effect
6. blanking together with their effect on the quality of the products are also presented and discussed.
7. Recent experimental results together with photo macrographs and photomicrographs are also included and discussed.
8. Finally, the effect of punch and die wear on the quality of the blanks is also given and discussed.
9. It focuses on blanking and piercing operations in a press tool to form and shape the final part geometry.
10. The types of piercing operations include conventional piercing, piercing with a pointed punch, piece-and-extrude operations, slotting, countersinking, and cutting and lancing of tabs.
11. The article provides information on the punch assembly, the die assembly, and the stripper and discusses the factors considered during piercing operations.
12. It reviews the applications of the four types of blanks used in sheet-forming operations, namely, rectangular blank, rough blank, partially developed blank, and fully developed blank.
13. It concludes with a discussion on the process capabilities, applications, and limitations of fine-edge blanking and piercing.

	PIERCING  PIERCING  	
15	**Electromagnetic forming:** The electrical energy stored in a capacitor bank is used to produce opposing magnetic fields around a tubular work piece, surrounded by current carrying coils. The coil is firmly held and hence the work piece collapses into the die cavity due to magnetic repelling force, thus assuming die shape. **Process details/ Steps:** i) The electrical energy is stored in the capacitor bank ii) The tubular work piece is mounted on a mandrel having the die cavity to produce shape on the tube. iii) A primary coil is placed around the tube and mandrel assembly. iv) When the switch is closed, the energy is discharged through the coil v) The coil produces a varying magnetic field around it. vi) In the tube a secondary current is induced, which creates its own magnetic field in the opposite direction. vii) The directions of these two magnetic fields oppose one another and hence the rigidly held coil repels the work into the die cavity. viii) The work tube collapses into the die, assuming its shape. **Process parameters:** i) Work piece size ii) ii) Electrical conductivity of the work material. iii) iii) Size of the capacitor bank iv) iv) The strength of the current, which decides the strength of the magnetic field and the force applied. v) v) Insulation on the coil. vi) vi) Rigidity of the coil. **Advantages:** i) Suitable for small tubes ii) ii) Operations like collapsing, bending and crimping can be easily done. iii) Electrical energy applied can be precisely controlled and hence the process is iv) accurately controlled.	12

v) The process is safer compared to explosive forming.

vi) Wide range of applications.

Limitations:

i) Applicable only for electrically conducting materials.

ii) Not suitable for large work pieces.

iii) Rigid clamping of primary coil is critical.

iv) Shorter life of the coil due to large forces acting on it.

Applications:

i) Crimping of coils, tubes, wires

ii) Bending of tubes into complex shapes

iii) Bulging of thin tubes.

Nadimpalli Satyanarayana Raju Institute of Technology (Autonomous). IQAC: Quality Management System (QMS)

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2022 - 2023
Course Code	20EE305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Power Generation and Transmission**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any two disadvantages of nuclear power plant.	20EE305.1	L1
2	Define the diversity factor.	20EE305.2	L1
3	Define GMR and GMD.	20EE305.3	L1
4	Recall surge impedance loading (SIL).	20EE305.4	L1
5	Mention any two methods of improving string efficiency.	20EE305.5	L1

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

<table>
<tr><th>No.</th><th>Questions (6 through 15)</th><th>Marks</th><th>Learning Outcome (s)</th><th>DoK</th></tr>
<tr><td>6</td><td>Explain the construction and working principle of thermal power plant.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>7</td><td>Explain the layout, classification, and operation of hydro power plant.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td>8</td><td>A generating station has the following daily load cycle:
<table>
<tr><td>Time (Hrs)</td><td>0-6</td><td>6-10</td><td>10-12</td><td>12-16</td><td>16 -20</td><td>20 -24</td></tr>
<tr><td>Load (MW)</td><td>40</td><td>50</td><td>60</td><td>50</td><td>70</td><td>40</td></tr>
</table>
70MW 1200 MWh
Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.
50MW 71.4</td><td>12M</td><td>20EE305.2</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>9</td><td>Explain the types of Tariff methods.</td><td>12M</td><td>20EE305.2</td><td>L2</td></tr>
<tr><td>10</td><td>Derive the equation for inductance of a three phase over head line.</td><td>12M</td><td>20EE305.3</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>11</td><td>Derive the equation for capacitance of a two-wire over head line.</td><td>12M</td><td>20EE305.3</td><td>L3</td></tr>
<tr><td>12</td><td>Classify the types of transmission lines with model representations.</td><td>12M</td><td>20EE305.4</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>13 (a)</td><td>Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.</td><td>6M</td><td>20EE305.4</td><td>L3</td></tr>
<tr><td>13 (b)</td><td>Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.</td><td>6M</td><td>20EE305.4</td><td>L3</td></tr>
<tr><td>14</td><td>Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.</td><td>12M</td><td>20EE305.5</td><td>L2</td></tr>
</table>

	OR			
15 (a)	Derive an expression for sag of a line supported between two supports of the same tower height.	8M	20EE305.5	L3
15 (b)	A 132 kV transmission line has the following data: Wt. of conductor = 680 kg/km; Length of span = 260 m Ultimate strength = 3100 kg; Safety factor = 2. Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.	4M	20EE305.5	L3

Height = 13.7M

Tension, T = 1550

d = 3.7 M

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

Part-A:

Q1: Two ~~ad~~ disadvantages – 2M. (1M for each).

Q2: Definition – 1M
Formula – 1M.

Q3: GMR – 1M
GMD – 1M

04: SIL Definition – 2M.

05: Two methods – 1M for each.

Part-B

Q6: Figure : 4M
Components List : 4M
Explanation : 4M.

Q7: Figure : 4M
Component List : 4M
Explanation : 4M.

Q8: (i) 3M ; (ii) 3M ; (iii) 3M ; (iv) 3M
Curve : 3M.

Q9: 6 Types 2M for each types.

Q10: Basic explanation : 4M.

Unsymmetrical : 4M.
Derivation

Transposition : 4M.
Derivation.

Q11: Explanation of Law : 4M.

Derivation : 8M.

Q12: Short T/m Line : 3M

Medium T/m Line : 6M.

Long T/m Line : 3M.
DC T/m line : 2M

Q13a: Derivation : 6M.

b: Figures : 3M.
Derivation : 3M.

Q14: Each Method : 3M.

Q15 a: Figure : 4M
Derivation : 4M.

b: Complete Solution : 4M.

1. List any two disadvantages of nuclear power plant.
   Any two of the below (2m)
   a) Strong pressure vessel is required due to the use of high-pressure water system
   b) Formation of low temperature steam
   c) Use of expensive cladding material for prevention of corrosion
   d) 4.High losses from heat exchanger
   e) 5.High power consumption by auxiliaries
   f) 6.Requires more elaborate safety devices

2. Define the diversity factor.
   It shows the diversity of load connected to a power station (2m)
   Diversity factor = Sum of individual maximum demands / Maximum demand on power system

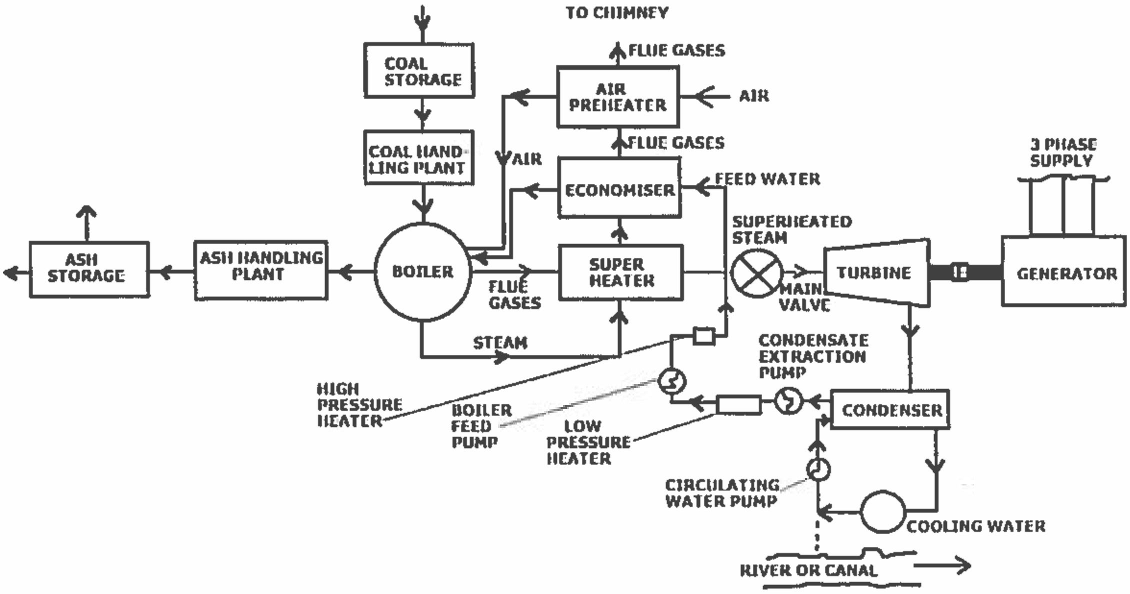
3. Define GMR and GMD.
   GMR stands for Geometrical Mean Radius. It is also called the self GMD (Geometrical Mean Distance)
   GMR = 0.7788R
   Where R is the radius of the conductor (1m)

   GMD stands for Geometrical Mean Distance GMD represents the geometrical mean distance from one conductor to the other. GMD for a different arrangement of conductors has different values. (1m)

4. Recall surge impedance loading (SIL).
   SIL is defined as the maximum load (at unity power factor) that can be delivered by the transmission line when the loads terminate with a value equal to surge impedance (Zs) of the line. (2m)

5. Mention any two methods of improving string efficiency.
   Any of the following (2m)
   By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap
   By Using Longer Cross-Arms
   By Capacitance Grading
   By Static Shielding

6. Explain the construction and working principle of thermal power plant.



A Thermal Power plant converts the heat energy of coal into electrical energy. Coal is burnt in boiler which converts water into steam.

The expansion of steam in turbine produces mechanical power which drives the Alternator coupled to the turbine. Thermal power plants contribute maximum to the generation of power for any country.

In thermal generating stations coal, oil, natural gas etc. are employed as primary courses of energy.

Main Components

- Coal handling plant
- Pulverizing plant
- Boiler
- Turbine
- Condenser
- Cooling towers and ponds
- Feed water heater
- Economizer
- Air preheater

Coal Handling Plant

- Coal is transported to power station by rail or road and stored in coal storage plant and then pulverized.
- The function of coal handling plant is automatic feeding of coal to the boiler furnace.
- A thermal power plant burns enormous amounts of coal.
- A 200MW plant may require around 2000 tons of coal daily.

Pulverizing Plant

- In modern thermal power plant, coal is pulverized i.e. ground to dust like size and carried to the furnace in a stream of hot air. Pulverizing is a means of exposing a large surface area tot the action of oxygen and consequently helping combustion.
- Pulverizing process consists 3 stages classified as:
  1.Feeding
  2.Drying.
  3.Grinding

Boiler

- The function of boiler is to generate steam at desired pressure and temperature by transferring heat produced by burning of fuel in a furnace to change water into steam.

Turbine

- In thermal power plants generally 3 turbines are used to increase the efficiency.
- High pressure turbine
- Intermediate pressure turbine
- Low pressure turbine

Condenser

- The surface condenser is a shell and tube heat exchanger where cooling water flows through tubes and exhaust steam fed into the shell surrounds the tubes, as a result, steam condense outside the tubes.

Cooling Towers and Ponds

- A condenser needs huge quantity of water to condense the steam.
- Most plants use cooled cooling system where warm water coming from condenser is cooled and reused.
- Cooling tower is a steel or concrete hyperbolic structure with the height of 150m.

Feed water heater

- Feed water heating improves overall plant efficiency
- Thermal stresses due to cold water entering the boiler drum are avoided.
- Quality of steam produced by the boiler is increased.

Economizer

- Flue gases coming out of the boiler carry lot of heat. An economizer extracts a part of this heat from flue gases and uses it for heating fee water.

- Saving coal consumption and higher boiler efficiency.

Air Preheater

- The function of air preheaters is to preheat the air before entering to the furnace by utilizing some of the energy left in the flue gases before exhausting them to the atmosphere.
- After flue gases leave economizer, some further heat can be extracted from them and used to heat incoming heat. Cooling of flue gases by 20-degree centigrade increases the plant efficiency by 1%.

Ash Handling plant

- The ash from the boiler is collected in two forms
- Bottom ash (Slurry): It's a waste which is dumped into ash pond
- Fly ash: Fly ash is separated from flue gases in ESP.

Water Handling Plant

- Water in a power plant is used for
- Production of steam - for rotating turbine
- Cooling purpose - For cooling of various equipment.
- Water is recycled and used for various purpose:
  Raw water - For cooling purposes - Steam - Condenser - Raw water
- About 4 cubic meter water is lost/day/MW.

Electrostatic precipitator (ESP)

- An ESP electrically charges the ash particles and imparts a strong electric field in the flue gas to collect and remove them. ESP is comprised of a series of parallel, vertical metallic plates (collecting electrodes) forming lanes through which the flu gases pass.

7. Explain the layout, classification, and operation of hydro power plant.

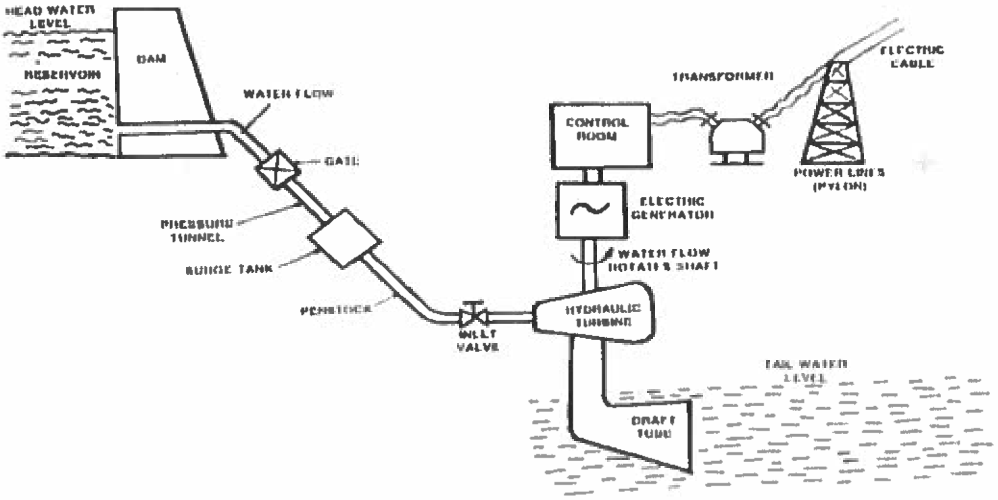


Fig.Layout of Hydro electric Power plant

(4m)

In hydroelectric power station potential and kinetic energy of stored water is converted into electrical energy.

Working Principle

- Hydroelectric power is the power obtained from the energy falling water whereas hydroelectric power plant is the power utilizing the potential energy of water at a high level for the generation of the electrical energy.

Main Components

- Reservoir
- Dam
- Trace rack
- Turbine
- Forebay

- Surge tank
- Penstock
- Spillway
- Turbine
- Powerhouse (4m)

Dam
- Develops a reservoir to store water
- Builds up head for power generation

Spillway
- To safeguard the dam when water level in the reservoir rises

Intake
- Contains trash racks to filter out debris which may damage the turbine

Forebay
- The forebay is used as "regulating reservoir" storing water temporarily during light load and providing the same for initial increases on account of increasing load during which water in canal is being accelerated. A forebay is an enlarged body of water just above the intake which is used to store water temporarily to meet the hourly load fluctuations. A forebay is not required if plant is located just at the base of the dam but, if the plants are situated away from the storage reservoir, a forebay is must.

Surge tank
- Surge tank is a small reservoir in which the water level rises or falls to reduce the pressure swings so that they are not transmitted to the penstock.
- When the load demand is reduced on the power station then, it causes rise in water level in the surge tank which produces a retarding head and reduces the velocity of water in the penstock and hence avoiding the undesirable phenomenon called "water hammer"
- When the load on the plant is increased, the governor causes the turbine to open the gates in order to allow more water to flow through the penstock to supply the increased load and there is a tendency to cause a vacuum or a negative pressure in the penstock.

Penstock
- Penstock is a closed conduit which connects the forebay or surge tank to the scroll case of the turbine. In case of high head plants, a single penstock is provided.

Valves and Gates
- Gates are used in low head plants at the entrance to the turbine casing to shut-off the flow and provide for unwatering the turbine for inspection and repairs. Valves are used at the entrance to the turbine casing if a long or medium length penstocks is used in the hydro power plant.

Trash Racks
- Trash racks are used to prevent the ingress of floating and other material to the turbine. These are built up from long, flat bars set vertically or nearly so and spaced in accordance with the minimum width of water passage through the turbine.

Tail race
- After the useful work is done by water, it is discharged to the tail race.

Draft tube
- It is an airtight pipe of suitable diameter attached to the runner outlet and conducting water down from the wheel and discharging it under the surface of the water in the tail race. With the help of draft tube operating head on the turbine is increased resulting in increase in output and efficiency

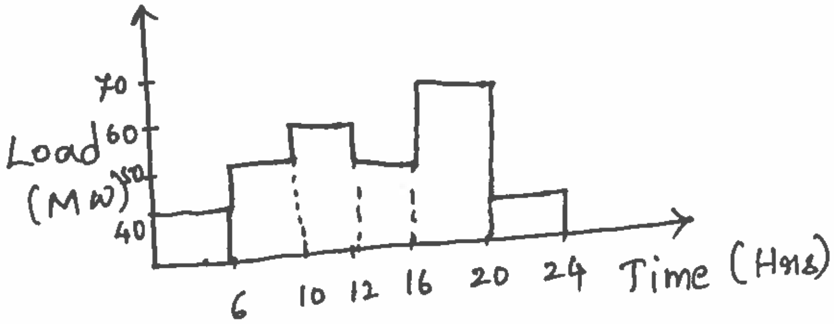
Prime Movers or water turbines
- In hydroelectric plants water turbines serve the purpose of prime mover which converts the kinetic energy of water into mechanical energy which is further utilized to drive the alternators generating electrical energy. (4m)

---

8. A generating station has the following daily load cycle:

Time (Hrs)	0-6	6-10	10-12	12-16	16 -20	20 -24
Load (MW)	40	50	60	50	70	40

Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.



(i) Maximum Demand = 70MW.

(ii) Units Generated per day = (6×40) + (4×50) + (2×60) + (4×50) + (4×70) + (4×40)

= 1200 MWh.

(iii) Average Load = $\frac{\text{Units generated/day}}{\text{Hrs in a day}}$

$= \frac{12{\cdot}00 \text{ MWh}}{24} = \frac{1200 \times 10^6}{24}$

$= 50 \times 10^6 = 50\text{MW}.$

(iv) Load Factor $= \frac{\text{Average Demand} \times 100}{\text{Maximum Demand}} = \frac{50\text{MW}}{70\text{MW}} \times 100$

$= 0.714 \times 100$

$= 71.4.$

9. Explain the types of Tariff methods.

Simple Tariff: When there is a fixed rate per unit of energy consumed, it is the Simple Tariff. In this type of tariff, the price charged per unit is constant. It does not vary with increase or decrease with the number of Units consumed.

Flat Rate Tariff: When different types of Consumers are charged at different uniform per unit rates, it is the Flat Rate Tariff. The rate for each type of consumer is arrived at by taking its load factor, diversity factor into consideration. The Bill will be Total units Consumed x Rate/Unit.

Block Rate Tariff: When a given block of energy is charged at a specific rate and the succeeding blocks of energy are charged progressively at reduced rates. Then the Tariff is called the Block Rate Tariff.

If the number of units generated increases, then the cost of generation per-unit-automatically decreases. For the first 30 units may be charged at the rate of 60paisa per unit, the next 25 units at the rate of 55paise per unit and the remaining additional units may be charged at the rate of 30 paise per unit. This type of tariff is being majorly used for residential and small commercial consumers.

Two Part Tariff: When the rate of electric energy is charged based on maximum demand of the consumer and the units consumed, it is called the Two-part Tariff.

In two-part tariff, the total charge to be made from the consumer is spilt into two components, fixed charges and running charges. The fixed charges depend upon the maximum demand of the consumer, while the running charges depend upon the no. of units consumed by the consumer.

Total Charges = Rs. (B kW+ C kWh)

B = Charges per kW of maximum demand

C = Charges per kWh of energy consumed

This type of Tariff is mostly applicable to Industrial Consumers.

Maximum Demand Tariff: It is similar to Two Part Tariff with the only difference that the maximum demand is actually measured by installing maximum demand meter in the premises of the consumer.

This type of tariff is mostly applied to big consumer, as a separate maximum demand meter is required.

Three-part tariff: When the total charge to be made from the consumer split into three parts, fixed charges, Semi fixed charges and running Charges.

Total Charge = Rs. (A + B kW + C kWh)

A = Fixed Charges during each billing period

B = Charge per kW of Maximum demand

C = Charge per kWh of energy consumed

Power Factor Tariff:

The Tariff in which Power factor of the Consumer's load is taken into consideration is known as Power Factor Tariff.

The following are the important types of Power Factor tariff

(i) kVA maximum demand tariff: It is a modified form of two-part tariff. The fixed charges are made on the basis of maximum demand in kVA and not in kW. As kVA is inversely proportional to power factor, a consumer having low power factor has to contribute more towards the fixed charge.

(ii) Sliding Scale Tariff: This is also known as average PF tariff. In this case an average power factor say 0.8 lagging is taken as reference. If the pf of the consumer falls below this factor, suitable additional charges are made. If the PF is above the reference, a discount is allowed to the consumer.

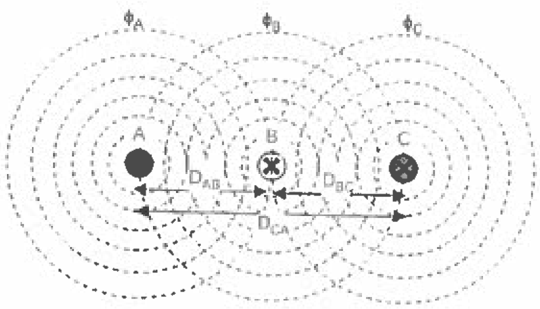
(iii) kW and kVAR Tariff: In this type both active power (kW) and reactive power (kVAR) supplied are charged separately. A consumer having low PF will draw more reactive Power and hence shall have to pay more charges.

10. Derive the equation for inductance of a three-phase overhead line.

Inductance in Three Phase Transmission Line:

In the three-phase transmission line, three conductors are parallel to each other. The direction of the current is same through each of the conductors.

Let us consider conductor A produces magnetic flux φ_A, Conductor B produces magnetic flux φ_B, And conductor C produces magnetic flux φ_C. When they carry the current of the same magnitude "I", they are in flux linkage with each other.

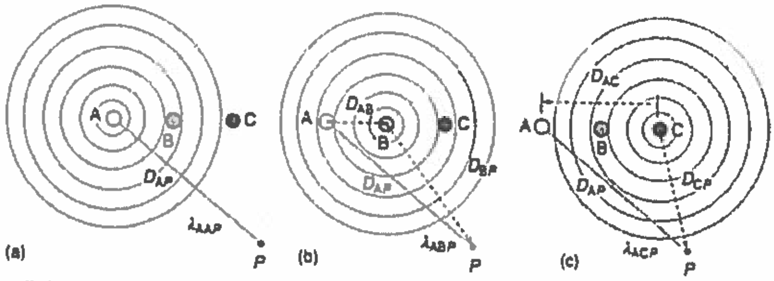


Now, let us consider a point P near three conductors. So, flux linkage at point P due to current through conductor A is,

$$\lambda_{AP} = \lambda_{AAP} + \lambda_{ABP} + \lambda_{ACP}$$

Flux linkage at point P for conductor A due to current through conductor A =

$$\lambda_{AAP} = \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D_{AP}}{GMR_A}\right) \quad Wb/m$$



Flux linkage at point P for conductor A due to current through conductor B =

$$\lambda_{ABP} = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D_{BP}}{D_{AB}}\right) \quad Wb/m$$

Flux linkage at point P for conductor A due to current through conductor C =

$$\lambda_{ACP} = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D_{CP}}{D_{AC}}\right) \quad Wb/m$$

Therefore, flux linkage at point P for conductor A,

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right]$$
$$+ \frac{\mu_0}{2\pi}\left[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})\right] \quad Wb/m$$

(2m)

As, $D_{AP} = D_{BP} = D_{CP}$ and $I_A + I_B + I_C = 0$ in balanced system, then we can write that $I_A = -I_B - I_C$

$$\therefore \frac{\mu_0}{2\pi}[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})]$$

$$= \frac{\mu_0}{2\pi}[-I_B \times ln(D_{AP}) - I_C \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})] = 0$$

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right] + 0 = \lambda_A (say)$$

$$So, \ \lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right]$$

$$Similarly, \ \lambda_B = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{BA}}\right) + I_B \times ln\left(\frac{1}{GMR_B}\right) + I_C \times ln\left(\frac{1}{D_{BC}}\right)\right]$$

$$and,\ \lambda_C = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{CA}}\right) + I_B \times ln\left(\frac{1}{D_{CB}}\right) + I_C \times ln\left(\frac{1}{GMR_C}\right)\right]$$

(2m)

For a balanced system,

$$D_{AB} = D_{BC} = D_{CA} = D$$

$$I_A + I_B + I_C = 0$$

In balanced system, then we can write that, $I_A = -I_B - I_C$

$$\lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D}\right) + I_C \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (I_B + I_C) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (-I_A) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D}{GMR_A}\right) \quad Wb/m$$

$$\lambda_B = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D}{GMR_B}\right) \quad Wb/m$$

$$\lambda_C = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D}{GMR_C}\right) \quad Wb/m$$

(2m)

$$So,\ inductance\ per\ metre\ per\ phase,\ \ L_{phase} = \frac{\mu_0}{2\pi} \times ln\left(\frac{D}{GMR_{phase}}\right) \quad H/m$$

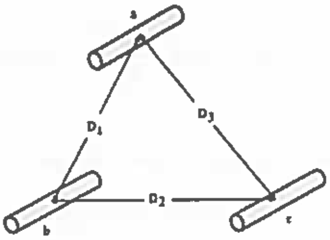
Consider a 3-phase overhead transmission line with phase conductors a, b, and c of radius r each spaced unsymmetrically such that the distance between the three conductors is D_1, D_2, and D_3 as shown in the figure below. Let I_a, I_b, and I_c be the currents flowing in the conductors a, b, and c respectively.

Assume that the whole system is balanced which leads to the flow of equal currents in the conductors i.e., $I_a = I_b = I_c = I$(say). The currents I_a, I_b, and I_c are displaced at 120° apart from each other. If I_a is taken as the reference phasor then the currents are given by,

$$I_a = I(1 - j0)$$

$$I_b = I(-0.5 - j0.866)\ and$$

$$I_c = I(-0.5 + j0.866)$$



Unsymmetrically Spaced 3-Phase Conductors

We know that the flux linkage of any conductor, in a group of conductors, is due to its own current and currents in the other conductors. Therefore, flux linkage of the conductor 'a' is due to its own current and currents in the conductor's 'b' and 'c', and it is given by,

$$\lambda_a = 2 \times 10^{-7}\left[I_a \ln\frac{1}{r'} + I_b \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_3}\right]$$

Similarly, flux linkages of conductors 'b' and 'c' is given by,

$$\lambda_b = 2 \times 10^{-7}\left[I_b \ln\frac{1}{r'} + I_a \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_2}\right]$$

$$\lambda_c = 2 \times 10^{-7}\left[I_c \ln\frac{1}{r'} + I_a \ln\frac{1}{D_3} + I_b \ln\frac{1}{D_2}\right]$$

(2m)

Where, r' = Geometric mean radius (GMR) of conductor = 0.07788 × r. Substituting the values of I_a, I_b, and I_c in the above equation we get,

$$\lambda_a = 2\times10^{-7}\left[I\ \ln\frac{1}{r'} + I(-0.5 - j0.866)\ln\frac{1}{D_1} + I(-0.5 + j0.866)\ln\frac{1}{D_3}\right]$$

$$\lambda_a = 2\times10^{-7}\,I\left[\ln\frac{1}{r'} + \ln\sqrt{D_1} + \ln\sqrt{D_3} + j0.866\ln D_1 + j0.866\ln\frac{1}{D_3}\right]$$

$$\lambda_a = 2\times10^{-7}\,I\left[\ln\frac{1}{r'} + \ln\sqrt{D_1D_3} + j\frac{\sqrt{3}}{2}\ln D_1 + j\frac{\sqrt{3}}{2}\ln\frac{1}{D_3}\right]$$

$$\lambda_a = 2\times10^{-7}\,I\left[\ln\frac{1}{r'} + \ln\sqrt{D_1D_3} + j\sqrt{3}\ln\sqrt{\left(\frac{D_1}{D_3}\right)}\right]$$

We know that the inductance L_a is given by,

$$L_a = \frac{\lambda_a}{I_a}$$

$$L_a = \frac{2\times10^{-7}}{I}\,I\left[\ln\frac{1}{r'} + \ln\sqrt{D_1D_3} + j\sqrt{3}\ln\sqrt{\left(\frac{D_1}{D_3}\right)}\right]$$

$$\therefore L_a = 2\times10^{-7}\left[\ln\frac{1}{r'} + \ln\sqrt{D_1D_3} + j\sqrt{3}\ln\sqrt{\left(\frac{D_1}{D_3}\right)}\right]H/m$$

Similarly, the inductance due to the conductor's 'b' and 'c' can be calculated and they are given by,

$$L_b = 2\times10^{-7}\left[\ln\frac{1}{r'} + \ln\sqrt{D_1D_2} + j\sqrt{3}\ln\sqrt{\left(\frac{D_2}{D_1}\right)}\right]H/m$$

$$L_c = 2\times10^{-7}\left[\ln\frac{1}{r'} + \ln\sqrt{D_2D_3} + j\sqrt{3}\ln\sqrt{\left(\frac{D_3}{D_2}\right)}\right]H/m$$

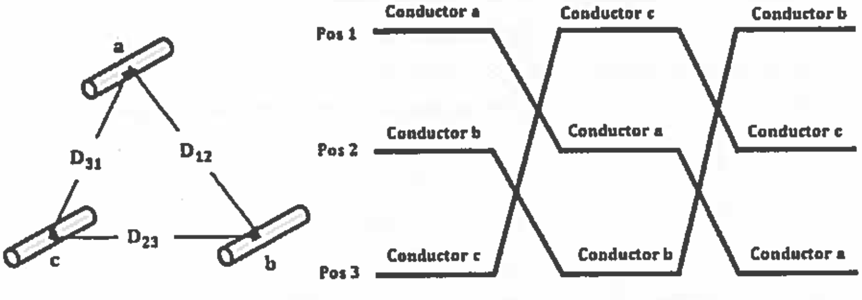
When conductors are unsymmetrically spaced in a 3-phase line, the flux linkage and inductance per phase are not identical, and knowing the inductance of each phase becomes complicated and it results in an unbalanced circuit.

The equilibrium in the circuit can be retained by shifting the places of the conductor at every period such that the conductors take the initial place of every conductor at the same distances.

Such an arrangement of the conductor's position obtained by shifting their places is called transposition. By transposition, we can obtain almost the same inductance between the conductors. Let us see the expression for inductance when the 3-phase line is transposed.

Inductance of Unsymmetrically Spaced 3-Phase Transmission Line When Transposed: (2m)

The simple configuration of a 3-phase conductor is shown in the figure. As the conductors are transposed, so their position in the transposition cycle would be as shown in the figure below.



Unsymmetrically Spaced 3-Phase Conductors with Transposition Cycle

When the conductors are connected in parallel, it results in low inductance keeping the distance between phases as small as possible. For deriving the inductance per phase of 3-phase conductors placed unsymmetrically, we need to determine the flux linkage of each conductor for every position it occupies in a transposition cycle. After that, we can determine the average flux linkages. So, the flux linkage of phase 'a' in position 1 is given as,

$$\lambda_{a1} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}} + I_c \ln\frac{1}{D_{31}}\right]$$

When 'a' is in position 2, 'b' in position 3 and 'c' in position 1,

$$\lambda_{a3} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{31}} + I_c \ln\frac{1}{D_{23}}\right]$$

The average value of flux linkage of a single-phase 'a' is,

$$\lambda_a = \frac{\lambda_{a1} + \lambda_{a2} + \lambda_{a3}}{3}$$

$$\lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}D_{23}D_{31}} + I_c \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

We know that

$$I_a + I_b + I_c = 0$$
$$I_b + I_c = -I_a$$

$$\therefore \lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} - I_a \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

$$\lambda_a = 2 \times 10^{-7} I_a \ln\sqrt{\frac{\sqrt[3]{D_{12}D_{23}D_{31}}}{D_s}}$$

Therefore, the average inductance per phase is,

$$L_a = 2 \times 10^{-7} \ln\frac{D_{eq}}{D_s}\ H/m$$

$$Where, D_{eq} = \sqrt[3]{D_{12}D_{23}D_{31}}$$

Where,

D_{eq} = Geometric mean of three distances of unsymmetrical line

D_s = GMR (geometric mean radius) of the conductor.

The above expression is the inductance per phase of a 3-phase transmission line with unsymmetrical spacing but lines are transposed. Nowadays the transposition of conductors is made at switching stations to balance inductance.

If the conductors are equispaced,

D1 = D2 = D3 = D

$$L = 2\,x\,10^{-7} \ln\frac{\sqrt[3]{d^3}}{r'} \Rightarrow L = 2\,x\,10^{-7} \ln\frac{d}{r'}$$

(2m)

11. Derive the equation for capacitance of a two-wire overhead line.

Electric Field Intensity due infinite line charge:

Consider a long wire having q coulomb/m as shown in figure

Using Gauss's law, Field Intensity (E) at a point P, which is r metre from the conductor, can be calculated as

$$\oint \overline{D} \cdot \overline{ds} = Q$$

The Flux density at point P, considering the cylindrical shell of radius r and length l can be calculated using Gauss's Law

$$D \cdot 2\Pi r \cdot l = ql \Rightarrow D = \frac{q}{2\pi r}\ coulomb/m^2$$

And Electric Field intensity E is given by

$$E = \frac{D}{\varepsilon_0}$$

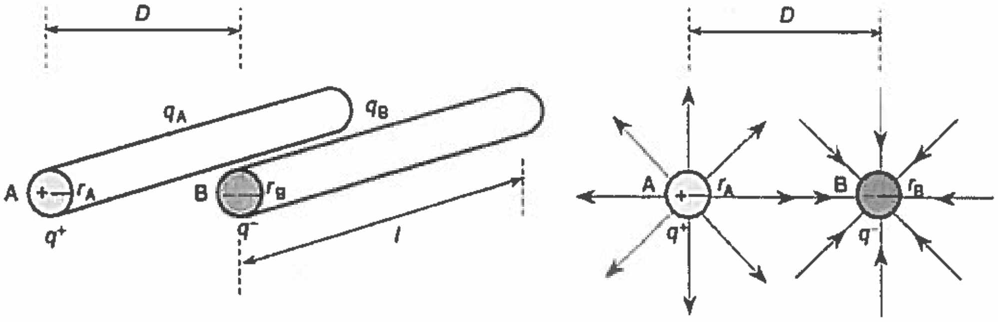
Capacitance and Capacitive Reactance:

Capacitance exists among transmission line conductors due to their potential difference. To evaluate the capacitance between conductors in a surrounding medium with permittivity ε, it is necessary to determine the voltage between the conductors, and the electric field strength of the surrounding. (4m)

Capacitance of a Single-Phase Line with Two Wires:

Consider a two-wire single-phase line with conductors A and B with the same radius r, separated by a distance D > r_A and r_B. The conductors are energized by a voltage source such that conductor A has a charge q+ and conductor B a charge q- as shown in Fig.

The charge on each conductor generates independent electric fields. Charge q+ on conductor A generates a voltage V_{AB-A} between both conductors. Similarly, charge q- on conductor B generates a voltage V_{AB-B} between conductors.



(4m)

Electric field produced from a two-wire single-phase system.

V_{AB-A} is calculated by integrating the electric field intensity, due to the charge on conductor A, on conductor B from r_A to D

$$V_{AB-A} = \int_{r_A}^{D} E_A \, dx = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D}{r_A}\right]$$

V_{AB-B} is calculated by integrating the electric field intensity due to the charge on conductor B from D to r_B

$$V_{AB-B} = \int_{D}^{r_B} E_B \, dx = \frac{-q}{2\pi\varepsilon_0} \ln\left[\frac{r_B}{D}\right]$$

The total voltage is the sum of the generated voltages V_{AB-A} and V_{AB-B}

$$V_{AB} = V_{AB-A} + V_{AB-B} = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D}{r_A}\right] - \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{r_B}{D}\right] = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D^2}{r_A r_B}\right]$$

If the conductors have the same radius, $r_A=r_B=r$, then the voltage between conductors V_{AB}, and the capacitance between conductors C_{AB}, for a 1-m line length are

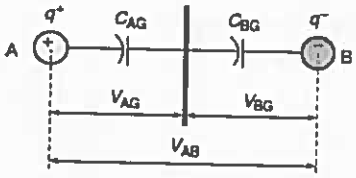
$$V_{AB} = \frac{q}{\pi\varepsilon_0} \ln\left[\frac{D}{r}\right] \quad (\mathrm{V})$$

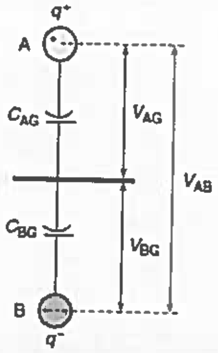
$$C_{AB} = \frac{\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \quad (\mathrm{F/m})$$

The voltage between each conductor and ground (G) is one-half of the voltage between the two conductors. Therefore, the capacitance from either line to ground is twice the capacitance between lines (4m)

$$V_{AG} = V_{BG} = \frac{V_{AB}}{2} \quad (\mathrm{V})$$

$$C_{AG} = \frac{q}{V_{AG}} = \frac{2\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \quad (\mathrm{F/m})$$





12. Classify the types of transmission lines with model representations.

Classification of Transmission Lines - Short, Medium & Long Transmission Lines:

A. AC transmission line, and
B. DC transmission line

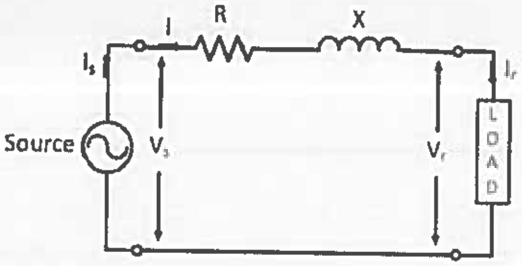
Depending upon the operating voltage and length, the overhead ac transmission lines are classified as,

1. Short transmission lines
2. Medium transmission lines
3. Long transmission lines

Short Transmission Line:

- ✓ Overhead transmission line is less than 50km
- ✓ Operating voltages of less than 20kV.

In these lines, the effect of capacitance is neglected due to smaller length and low operating voltage. Hence, the resistance and inductance effects of the line are considered while determining the performance of the short transmission line as shown below.



Equivalent circuit model of a short line

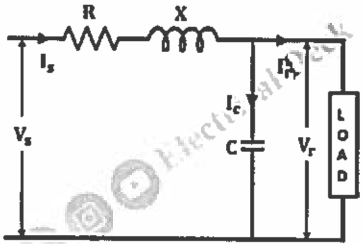
Circuit Globe

Medium Transmission Line:

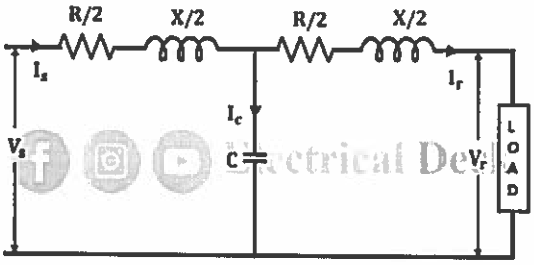
- ✓ Length of the overhead transmission line is in the range of 50-150km
- ✓ the operating voltage is greater than 20kV.

Based on the location of the capacitance at different places, the medium transmission lines have different configurations. These configurations show the different ways in which the effect of capacitance is taken into consideration. The three configurations based on the location of capacitance are,

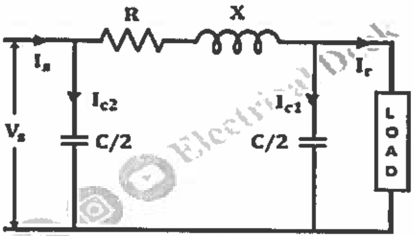
i. End condenser representation of medium transmission lines
ii. Nominal-T representation of medium transmission lines
iii. Nominal-π representation of medium transmission lines.



End Condenser Method of Medium Transmission Line



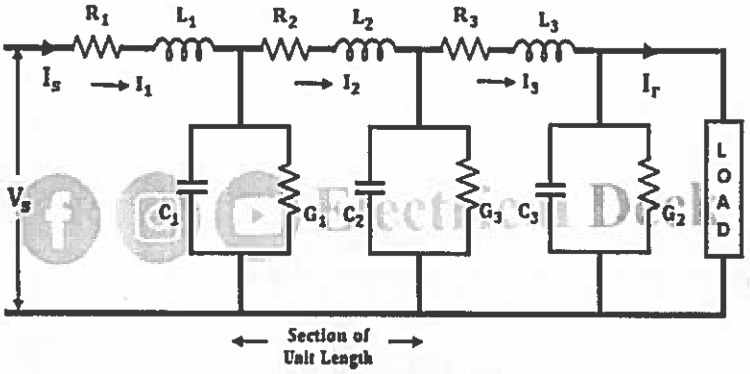
Nominal-T Method of Medium Transmission Line



Nominal-π Method of Medium Transmission Line

Long Transmission Line:

- ✓ Overhead transmission lines whose length is more than 150km
- ✓ Operating voltage of these lines is more than 100kV



Long Transmission Line

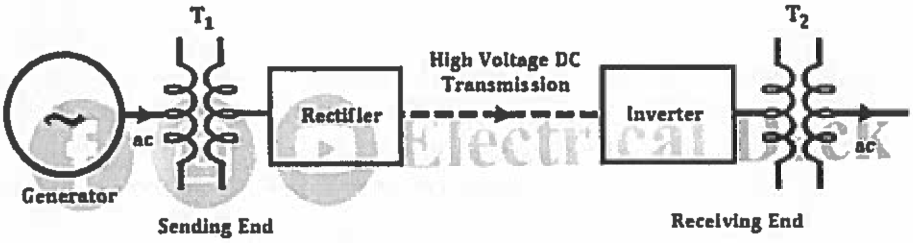
DC Transmission Line

In AC transmission lines, for transmission of power over long distances at higher voltages, the cost of transmission line and loss increases.

Also, the AC long transmission line suffers from problems like stability limits, voltage control, line compensation, interconnection of lines, ground impedance, etc due to an increase in voltage levels and distance.

The various problems associated with long-distance AC transmission have led to the development of HVDC (high voltage direct current) transmission nothing but a dc transmission line.

The use of DC power for long transmission lines has various advantages like no stability problem, absence of charging current, no skin effect, need for reactive compensation, bulk power transfer, economic power transmission, etc.



HVDC Transmission System

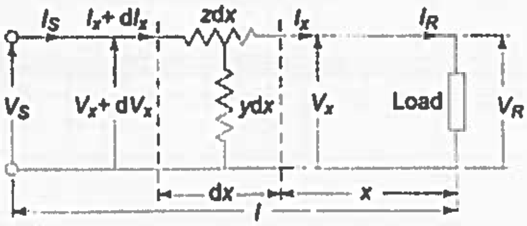
It is inefficient to use a dc transmission system for shorter and medium distances.

13. (a) Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.

ABCD parameters in Long Transmission Line-rigorous Solution:

For rigorous solution, consider the Line Impedance and admittance uniformly distributed and not lumped.

Consider a small element of length dx situated at a distance x from receiving end. Let Z and Y denoted respectively the Series Impedance and shunt Admittance of the line per unit length.



Schematic diagram of a long line

The voltage at two ends of the element are denotes as V (towards receiving end) and V + dV (towards sending end) respectively.

Let dx be an elemental section of the line at a distance x from the receiving-end having a series impedance zdx and a shunt admittance ydx. The rise in voltage to neutral over the elemental section in the direction of increasing x is dV_x.

We can write the following differential relationships across the elemental section:

$$dVx = I_x z dx \text{ or } \frac{dV_x}{dx} = zI_x \qquad (5.14)$$

$$dIx = V_x y dx \text{ or } \frac{dI_x}{dx} = yVx \qquad (5.15)$$

It may be noticed that the kind of connection (e.g. T or π) assumed for the elemental section, does not affect these first order differential relations. Differentiating Eq. (5.14) with respect to x, we obtain

$$\frac{d^2V_x}{dx^2} = \frac{dI_x}{dx}z$$

Substituting the value of dI_x/dx from Eq. (5.15), we get

$$\frac{d^2V_x}{dx^2} = yzV_x \qquad (5.16)$$

This is a linear differential equation whose general solution can be written as follows:

$$V_x = C_1 e^{\gamma x} + C_2 e^{-\gamma x} \qquad (5.17)$$

where

$$\gamma = \sqrt{yx} \qquad (5.18)$$

and C_1 and C_2 are arbitrary constants to be evaluated.

Differentiating Eq. (5.17) with respect to x:

$$\frac{dV_x}{dx} = C_1 \gamma e^{\gamma x} - C_2 \gamma e^{-\gamma x} = zI_x$$

$$I_x = \frac{C_1}{Z_r} e^{\gamma x} - \frac{C_2}{Z_c} e^{-\gamma x} \qquad (5.19)$$

Where

$$Z_c = \left(\frac{z}{y}\right)^{1/2} \qquad (5.20)$$

The constants C_1 and C_2 may be evaluated by using the end conditions, i.e. when x = 0, $V_x = V_R$ and $I_x = I_R$. Substituting these values in Eqs. (5.17) and (5.19) gives

$$V_R = C_1 + C_2$$

$$I_R = \frac{1}{Z_c}(C_1 - C_2)$$

which upon solving yield

$$C_1 = \frac{1}{2}(V_R + Z_c I_R)$$

$$C_2 = \frac{1}{2}(V_R - Z_c I_R)$$

With C_1 and C_2 as determined above, Eqs. (5.17) and (5.19) yield the solution for V_x and I_x as

$$V_x = \left(\frac{V_R + Z_c I_R}{2}\right)e^{\gamma x} + \left(\frac{V_R - Z_c I_R}{2}\right)e^{-\gamma x}$$

$$I_x = \left(\frac{V_R/Z_c + I_R}{2}\right)e^{\gamma x} - \left(\frac{V_R/Z_c - I_R}{2}\right)e^{-\gamma x} \qquad (5.21)$$

Here Z_c is called the **characteristic impedance** of the Long Transmission Line and γ is called the **propagation constant.**

Knowing V_R, I_R and the parameters of the line, using Eq. (5.21) complex number rms values of V_x and I_x at any distance x along the line can be easily found out.

A more convenient form of expression for voltage and current is obtained by introducing hyperbolic functions. Rearranging Eq. (5.21), we get

$$V_x = V_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right) + I_R Z_c\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right)$$

$$I_x = V_R \frac{1}{Z_c}\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right) + I_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right)$$

These can be rewritten after introducing hyperbolic functions, as

$$V_x = V_R \cosh \gamma x + I_R Z_c \sinh \gamma x \qquad (5.22)$$

$$I_x = I_R \cosh \gamma x + V_R \frac{1}{Z_c} \sinh \gamma x$$

where x = *l*, $V_x = V_s$, $I_x = I_s$

$$\therefore \quad \begin{bmatrix} V_S \\ I_S \end{bmatrix} = \begin{bmatrix} \cosh \gamma l & Z_c \sinh \gamma l \\ \frac{1}{Z_c} \sinh \gamma l & \cosh \gamma l \end{bmatrix} \begin{bmatrix} V_R \\ I_R \end{bmatrix} \qquad (5.23)$$

Here

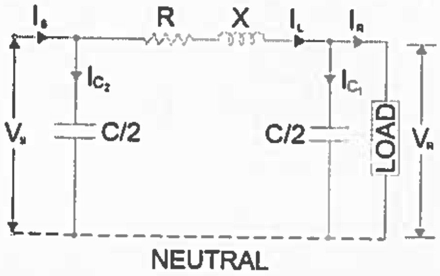
$$A = D = \cosh \gamma l$$

$$B = Z_c \sinh \gamma l \qquad (5.24)$$

$$C = \frac{1}{Z_c} \sinh \gamma l$$

13. (b) Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.

In Nominal π Method, the shunt capacitance of each line i.e. phase to neutral is divided into two equal parts. One part is lumped at the sending end while the other is lumped at receiving end as shown in figure below.



Notice that, in this method there is no effect of shunt capacitance at sending end on the line voltage drop and hence on voltage regulation but this accounts for the charging current in sending end.

Let

IR = Load Current per phase

R = Resistance per phase

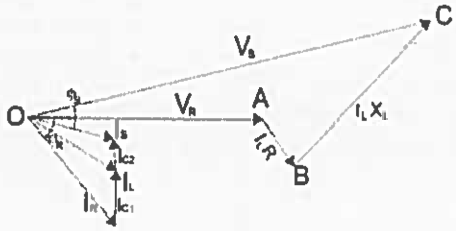
X = Reactance per phase

C = Capacitance per phase

Cos Ø R = Receiving end power factor (lagging)

Vs = Sending end voltage

Let us now draw the phasor. Assume receiving end voltage VR as reference and load current IR lagging this voltage by ØR.



Therefore,

VR= VR+ j0

and IR= IR<-ØR = IR (CosØR – jSinØR)

Charging Current at load end IC1 = j(ωC/2) VR = jπfCVR

Line Current **IL = IR + IC1** (phasor sum)

Sending end voltage Vs = VR + IL (R+jX)

Now,

Charging current at sending end IC2 = j(ωC/2)Vs = jπfCVs

Hence, Sending end current **Is = IL + IC2** (phasor sum)

Thus, sending end current and voltage is calculated as above and from these parameters the performance of line is evaluated.

14. Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.

Methods of Improving String Efficiency

Method # 1. By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap:

It is clear from the expression of string efficiency that the string efficiency increases with the decrease in value of K (i.e. the ratio of shunt capacitance to mutual capacitance). One method is to design the units such that the mutual capacitance (capacitance of each unit) is much greater than the shunt capacitance (capacitance to earth). This can be achieved by using insulators with larger discs or providing each insulator unit with a metal cap. The ratio K can be made 1/6 to 1/10 by this method.

Method # 2. By Using Longer Cross-Arms:

The ratio of shunt capacitance to mutual capacitance, K can alternatively be reduced by using longer cross-arms so that the horizontal distance from line support (pole or tower) is increased thereby decreasing the shunt capacitance. But the limitations of cost and mechanical strength of line supports do not allow the cross-arms to be too long and it has been found that in practice it is not possible to obtain the value of K less than 0.1.

Method # 3. By Capacitance Grading:

It is seen that non-uniform distribution of voltage across an insulator string is due to leakage current from the insulator pin to the supporting structure, which cannot be eliminated. However, it is possible that discs of different capacities are used such that the product of their capacitive reactance and the current flowing through the respective unit is same. This can be achieved by grading the mutual capacitance of the insulator units i.e., by having lower units of more capacitance—maximum at the line unit and minimum at the top unit, nearest to the cross-arm. By this method complete equality of voltage across the units of an insulator string can be obtained but this method needs a large number of different-sized insulator units and maintaining spares of all varieties of insulator discs. So this method is not used in practice below 200 kV.

Consider a 4-unit string. Let C be the capacitance of the top unit and let the capacitances of others units are C_2, C_3 and C_4, as shown in Fig.

Assume $C_1 = k\ C$

Applying Kirchhoff's first law to node A we get, $I_2 = I_1 + i_1$

$\Rightarrow \omega\ C_2 v = \omega\ C\ v + \omega\ C_1 v$ or $C_2 = C + K\ C = C\ (1 + K)$ ... (9.11)

Applying Kirchhoff's first law to node B we get, $I_3 = I_2 + i_2$

or $\omega\ C_3 v = \omega\ C_2 v + \omega\ C_1 \times 2\ v$

(or) $C_3 = C_2 + 2\ K\ C = C\ (1 + K) + 2\ K\ C = C\ (1 + 3\ K)$ ... (9.12)

Applying Kirchhoff's first law to node C we get,

$I_4 = I_3 + i_3$

or $\omega\ C_4 v = \omega\ C_3 v + \omega\ C_1 \times 3\ v$

or $C_4 = C_3 + 3\ K\ C = C\ (1 + 3\ K) + 3\ K\ C = C\ (1 + 6\ K)$ ...(9.13)

Thus, it will be possible to equalize the potential across all the units, if their capacitances are in the ratio of 1: (1 + K): (1 + 3 K): (1 + 6 K) and so on.

But in practice it is impossible to obtain such units which will have their capacitances in above ratio, although nearby results can be obtained by employing standard insulators for most of the units and employing larger units adjacent to the line.

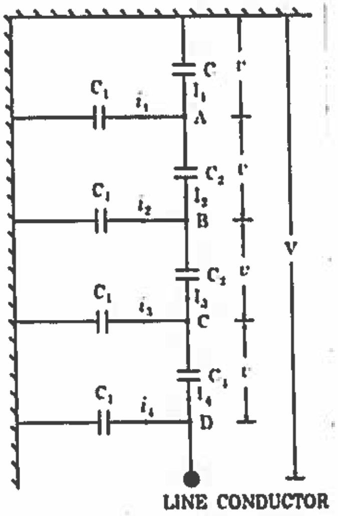
Method # 4. By Static Shielding:

In case of capacitance grading, insulator units of different capacitances are used so that the flow of different currents through the respective units produce equal voltage drop. In static shielding, pin to supporting structure charging currents are exactly cancelled so that the same current flows through the identical insulator units and produce equal voltage drops across each insulator unit.

This arrangement is given is Fig. 9.22. In this method a guard or grading ring, which usually takes the form of a large metal ring surrounding the bottom unit and electrically connected to the metal work at the bottom of this unit, and therefore to the line conductor.

The guard ring screens the lower units, reduces their earth capacitance C_1 and introduces a number of capacitances between the line conductor and the various insulator unit caps. These capacitances are greater for lower units and thus the voltages across them are reduced. With this method also it is impossible to obtain in practice an equal distribution of voltage but considerable improvements are possible.

Let the capacitances between the links and the shield be C_x, C_y and C_z respectively as shown in Fig. 9.22, and let v be the potential across each unit.



Since the capacitance of each unit is same, therefore, their charging currents I_1, I_2, I_3, and I_4 would be same, let it be I. If $C_1 = K C$

Applying Kirchhoff s first law to node A we get,

$I + i_x = I + i_1$ or $i_x = i_1$... (9.14)

Similarly, $i_y = i_2$... (9.15)

and $i_z = i_3$... (9.16)

Also, $i_1 = \omega C_1 v = \omega K C v$ ... (9.17)

$i_2 = 2 \omega C_1 v = 2 \omega K C v$ ... (9.18)

$i_3 = 3 \omega C_1 v = 3 \omega K C v$ ... (9.19)

The potential causing current i_x is 3 v (voltage across three units leaving the top one).

So, $i_x = \omega C_x \times 3 v = 3 \omega C_x v$ ... (9.20)

Comparing Eqs. (9.14), (9.17) and (9.20), we have,

$3 \omega C_x v = \omega K C v$ or $C_x = K C/3$ ... (9.21)

The potential causing current y is 2 v and therefore,

$I_y = 2 \omega C_y v$ ... (9.22)

Comparing Eqs. (9.15), (9.18) and (9.22) we have,

$2 \omega C_y v = 2 \omega K C v$ or $C_y = K C$ ... (9.23)

The potential causing current i_z is v and therefore,

$i_z = \omega C_z v$ ... (9.24)

Comparing Eqs. (9.16), (9.19) and (9.24), we have,

$\omega C_z v = 3 \omega K C v$

or $C_z = 3 K C$

In general, if there are n units

$i_1 = \omega K C v$ and $i_x = (n - 1) \omega C_x v$

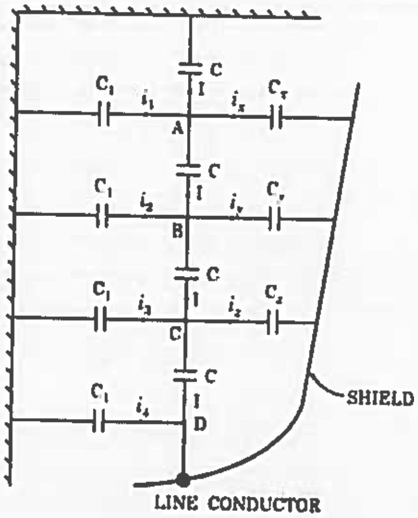
or $C_x = KC/ (n-1)$

Similarly, $C_y = 2KC/ (n-2)$

and $C_z = 3KC/ (n-3)$

or The capacitance of the pth metal link to the line is given as:

$C_p = pKC/ (n-p)$

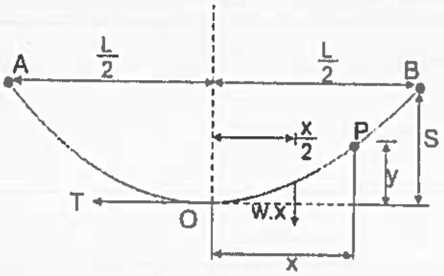


15. (a) Derive an expression for sag of a line supported between two supports of the same tower height.

Calculation of Sag in a Transmission Line:

Sag calculation for supports is at equal levels

Suppose, AOB is the conductor. A and B are points of supports. Point O is the lowest point and the midpoint.



Let, L = length of the span, i.e. AB

w = weight per unit length of the conductor

T = tension in the conductor.

We have chosen any point on the conductor, say point P.

The distance of point P from the Lowest point O is x.

y is the height from point O to point P.

Equating two moments of two forces about point O as per the figure above we get,

$$Ty = wx \times \frac{x}{2}$$

$$Now, y = \frac{wx^2}{2T},$$

The maximum dip (sag) is represented by the value of y at either of the supports A and B. At support A, x = l/2 and y = S.

$$Then\ S = \frac{wL^2}{8T}$$

15. (b) A 132 kV transmission line has the following data:
Wt. of conductor = 680 kg/km;
Length of span = 260 m
Ultimate strength = 3100 kg;
Safety factor = 2.
Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.

Sol: Weight of conductor/mt run, $w = \frac{680}{1000} = 0.68$ kg.

Working Tension, $T = \frac{\text{Ultimate Strength}}{\text{Safety factor}}$

$= \frac{3100}{2} = 1550$kg.

Span length, $l = 260$m.

$\therefore$ Sag $= \frac{wl^2}{8T} = \frac{0.68 \times 260^2}{8 \times 1550}$

$= 3.7$m.

Height = GC + Sag = 10 + 3.7

= 13.7m.

UshaRani
(Faculty:)

(HOD) 7/1/23.

Nadimpalli Satyanarayana Raju Institute of Technology (Autonomous). IQAC: Quality Management System (QMS)

NSRIT

Semester End Regular/Supplementary Examination, Dec./Jan., 2022 - 2023

Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2022 - 2023
Course Code	20EE305	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	III
Course	**Power Generation and Transmission**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any two disadvantages of nuclear power plant.	20EE305.1	L1
2	Define the diversity factor.	20EE305.2	L1
3	Define GMR and GMD.	20EE305.3	L1
4	Recall surge impedance loading (SIL).	20EE305.4	L1
5	Mention any two methods of improving string efficiency.	20EE305.5	L1

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

<table>
<tr><th>No.</th><th>Questions (6 through 15)</th><th>Marks</th><th>Learning Outcome (s)</th><th>DoK</th></tr>
<tr><td>6</td><td>Explain the construction and working principle of thermal power plant.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>7</td><td>Explain the layout, classification, and operation of hydro power plant.</td><td>12M</td><td>20EE305.1</td><td>L2</td></tr>
<tr><td>8</td><td>A generating station has the following daily load cycle:
<table>
<tr><td>Time (Hrs)</td><td>0-6</td><td>6-10</td><td>10-12</td><td>12-16</td><td>16 -20</td><td>20 -24</td></tr>
<tr><td>Load (MW)</td><td>40</td><td>50</td><td>60</td><td>50</td><td>70</td><td>40</td></tr>
</table>
70MW 1200 MWh
Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.
50MW 71.4</td><td>12M</td><td>20EE305.2</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>9</td><td>Explain the types of Tariff methods.</td><td>12M</td><td>20EE305.2</td><td>L2</td></tr>
<tr><td>10</td><td>Derive the equation for inductance of a three phase over head line.</td><td>12M</td><td>20EE305.3</td><td>L3</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>11</td><td>Derive the equation for capacitance of a two-wire over head line.</td><td>12M</td><td>20EE305.3</td><td>L3</td></tr>
<tr><td>12</td><td>Classify the types of transmission lines with model representations.</td><td>12M</td><td>20EE305.4</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>13 (a)</td><td>Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.</td><td>6M</td><td>20EE305.4</td><td>L3</td></tr>
<tr><td>13 (b)</td><td>Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.</td><td>6M</td><td>20EE305.4</td><td>L3</td></tr>
<tr><td>14</td><td>Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.</td><td>12M</td><td>20EE305.5</td><td>L2</td></tr>
</table>

	OR			
15 (a)	Derive an expression for sag of a line supported between two supports of the same tower height.	8M	20EE305.5	L3
15 (b)	A 132 kV transmission line has the following data: Wt. of conductor = 680 kg/km; Length of span = 260 m Ultimate strength = 3100 kg; Safety factor = 2. Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.	4M	20EE305.5	L3

Height = 13.7M

Tension, T = 1550

d = 3.7 M

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ANSWER KEY AND SCHEME OF EVALUATION

Part- A:

Q1: Two ~~ad~~ disadvantages — 2M. (1M for each).

Q2: Definition — 1M
Formula — 1M.

Q3: GMR – 1M
GMD – 1M

Q4: SIL Definition — 2M.

Q5: Two methods — 1M for each.

Part- B

Q6: Figure : 4M
Components List : 4M
Explanation : 4M.

Q7: Figure : 4M
Component List : 4M
Explanation : 4M.

Q8: (i) 2M ; (ii) 3M ; (iii) 2M ; (iv) 2M
Curve: 3M.

Q9: 6 Types 2M for each types.

Q10: Basic explanation : 4M.

Unsymmetrical Derivation : 4M.

Transposition Derivation : 4M.

Q 11: Explanation of Law: 4M.

Derivation : 8M.

Q 12: Short T/m Line : 2M

Medium T/m Line : 6M.

Long T/m Line : 2M.
DC T/m line : 2M

Q 13a: Derivation : 6M.

b: Figures : 3M.
Derivation : 3M.

Q14 : Each Method : 3M.

Q 15 a: Figure : 4M
Derivation : 4M.

b: Complete Solution : 4M.

1. List any two disadvantages of nuclear power plant.
   Any two of the below (2m)
   a) Strong pressure vessel is required due to the use of high-pressure water system
   b) Formation of low temperature steam
   c) Use of expensive cladding material for prevention of corrosion
   d) 4.High losses from heat exchanger
   e) 5.High power consumption by auxiliaries
   f) 6.Requires more elaborate safety devices

2. Define the diversity factor.
   It shows the diversity of load connected to a power station (2m)
   Diversity factor = Sum of individual maximum demands / Maximum demand on power system

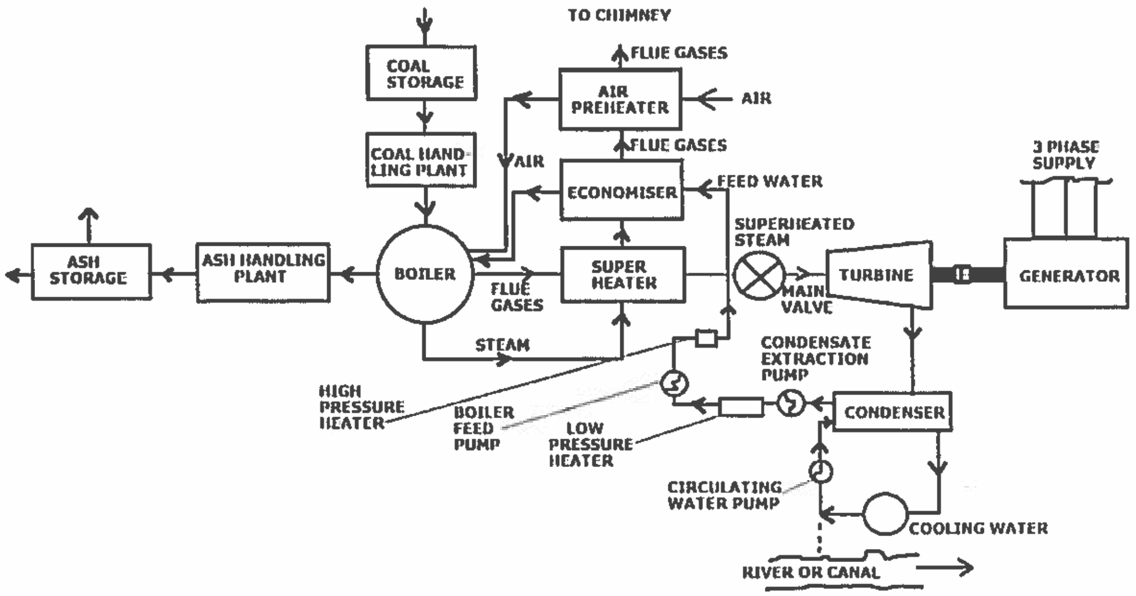
3. Define GMR and GMD.
   GMR stands for Geometrical Mean Radius. It is also called the self GMD (Geometrical Mean Distance)
   GMR = 0.7788R
   Where R is the radius of the conductor (1m)

   GMD stands for Geometrical Mean Distance GMD represents the geometrical mean distance from one conductor to the other. GMD for a different arrangement of conductors has different values. (1m)

4. Recall surge impedance loading (SIL).
   SIL is defined as the maximum load (at unity power factor) that can be delivered by the transmission line when the loads terminate with a value equal to surge impedance (Zs) of the line. (2m)

5. Mention any two methods of improving string efficiency.
   Any of the following (2m)
   By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap
   By Using Longer Cross-Arms
   By Capacitance Grading
   By Static Shielding

6. Explain the construction and working principle of thermal power plant.



A Thermal Power plant converts the heat energy of coal into electrical energy. Coal is burnt in boiler which converts water into steam.

The expansion of steam in turbine produces mechanical power which drives the Alternator coupled to the turbine. Thermal power plants contribute maximum to the generation of power for any country.

In thermal generating stations coal, oil, natural gas etc. are employed as primary courses of energy.

Main Components

- Coal handling plant
- Pulverizing plant
- Boiler
- Turbine
- Condenser
- Cooling towers and ponds
- Feed water heater
- Economizer
- Air preheater

Coal Handling Plant

- Coal is transported to power station by rail or road and stored in coal storage plant and then pulverized.
- The function of coal handling plant is automatic feeding of coal to the boiler furnace.
- A thermal power plant burns enormous amounts of coal.
- A 200MW plant may require around 2000 tons of coal daily.

Pulverizing Plant

- In modern thermal power plant, coal is pulverized i.e. ground to dust like size and carried to the furnace in a stream of hot air. Pulverizing is a means of exposing a large surface area tot the action of oxygen and consequently helping combustion.
- Pulverizing process consists 3 stages classified as:
  1.Feeding
  2.Drying.
  3.Grinding

Boiler

- The function of boiler is to generate steam at desired pressure and temperature by transferring heat produced by burning of fuel in a furnace to change water into steam.

Turbine

- In thermal power plants generally 3 turbines are used to increase the efficiency.
- High pressure turbine
- Intermediate pressure turbine
- Low pressure turbine

Condenser

- The surface condenser is a shell and tube heat exchanger where cooling water flows through tubes and exhaust steam fed into the shell surrounds the tubes, as a result, steam condense outside the tubes.

Cooling Towers and Ponds

- A condenser needs huge quantity of water to condense the steam.
- Most plants use cooled cooling system where warm water coming from condenser is cooled and reused.
- Cooling tower is a steel or concrete hyperbolic structure with the height of 150m.

Feed water heater

- Feed water heating improves overall plant efficiency
- Thermal stresses due to cold water entering the boiler drum are avoided.
- Quality of steam produced by the boiler is increased.

Economizer

- Flue gases coming out of the boiler carry lot of heat. An economizer extracts a part of this heat from flue gases and uses it for heating fee water.

- Saving coal consumption and higher boiler efficiency.

Air Preheater

- The function of air preheaters is to preheat the air before entering to the furnace by utilizing some of the energy left in the flue gases before exhausting them to the atmosphere.
- After flue gases leave economizer, some further heat can be extracted from them and used to heat incoming heat. Cooling of flue gases by 20-degree centigrade increases the plant efficiency by 1%.

Ash Handling plant

- The ash from the boiler is collected in two forms
- Bottom ash (Slurry): It's a waste which is dumped into ash pond
- Fly ash: Fly ash is separated from flue gases in ESP.

Water Handling Plant

- Water in a power plant is used for
- Production of steam - for rotating turbine
- Cooling purpose - For cooling of various equipment.
- Water is recycled and used for various purpose:
  Raw water - For cooling purposes - Steam - Condenser - Raw water
- About 4 cubic meter water is lost/day/MW.

Electrostatic precipitator (ESP)

- An ESP electrically charges the ash particles and imparts a strong electric field in the flue gas to collect and remove them. ESP is comprised of a series of parallel, vertical metallic plates (collecting electrodes) forming lanes through which the flu gases pass.

7. Explain the layout, classification, and operation of hydro power plant.

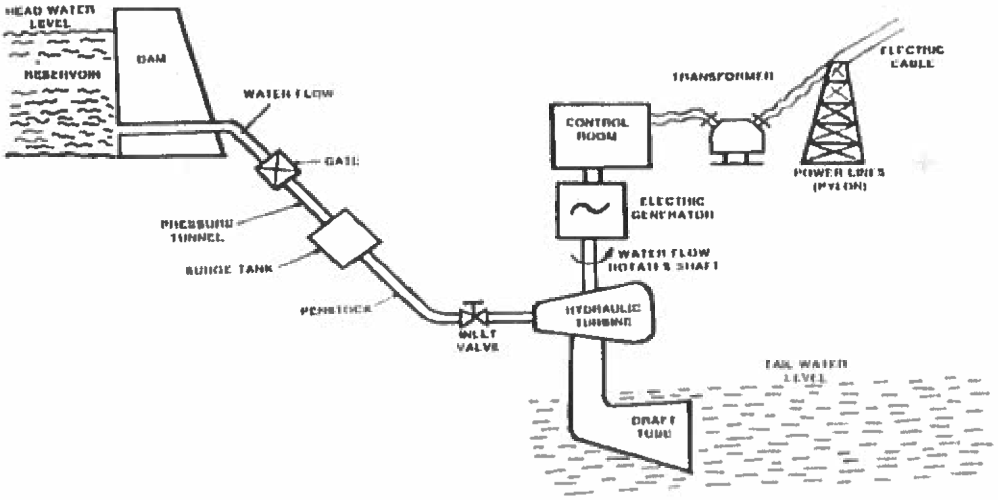


Fig.Layout of Hydro electric Power plant

(4m)

In hydroelectric power station potential and kinetic energy of stored water is converted into electrical energy.

Working Principle

- Hydroelectric power is the power obtained from the energy falling water whereas hydroelectric power plant is the power utilizing the potential energy of water at a high level for the generation of the electrical energy.

Main Components

- Reservoir
- Dam
- Trace rack
- Turbine
- Forebay

- Surge tank
- Penstock
- Spillway
- Turbine
- Powerhouse (4m)

Dam

- Develops a reservoir to store water
- Builds up head for power generation

Spillway

- To safeguard the dam when water level in the reservoir rises

Intake

- Contains trash racks to filter out debris which may damage the turbine

Forebay

- The forebay is used as "regulating reservoir" storing water temporarily during light load and providing the same for initial increases on account of increasing load during which water in canal is being accelerated. A forebay is an enlarged body of water just above the intake which is used to store water temporarily to meet the hourly load fluctuations. A forebay is not required if plant is located just at the base of the dam but, if the plants are situated away from the storage reservoir, a forebay is must.

Surge tank

- Surge tank is a small reservoir in which the water level rises or falls to reduce the pressure swings so that they are not transmitted to the penstock.
- When the load demand is reduced on the power station then, it causes rise in water level in the surge tank which produces a retarding head and reduces the velocity of water in the penstock and hence avoiding the undesirable phenomenon called "water hammer"
- When the load on the plant is increased, the governor causes the turbine to open the gates in order to allow more water to flow through the penstock to supply the increased load and there is a tendency to cause a vacuum or a negative pressure in the penstock.

Penstock

- Penstock is a closed conduit which connects the forebay or surge tank to the scroll case of the turbine. In case of high head plants, a single penstock is provided.

Valves and Gates

- Gates are used in low head plants at the entrance to the turbine casing to shut-off the flow and provide for unwatering the turbine for inspection and repairs. Valves are used at the entrance to the turbine casing if a long or medium length penstocks is used in the hydro power plant.

Trash Racks

- Trash racks are used to prevent the ingress of floating and other material to the turbine. These are built up from long, flat bars set vertically or nearly so and spaced in accordance with the minimum width of water passage through the turbine.

Tail race

- After the useful work is done by water, it is discharged to the tail race.

Draft tube

- It is an airtight pipe of suitable diameter attached to the runner outlet and conducting water down from the wheel and discharging it under the surface of the water in the tail race. With the help of draft tube operating head on the turbine is increased resulting in increase in output and efficiency

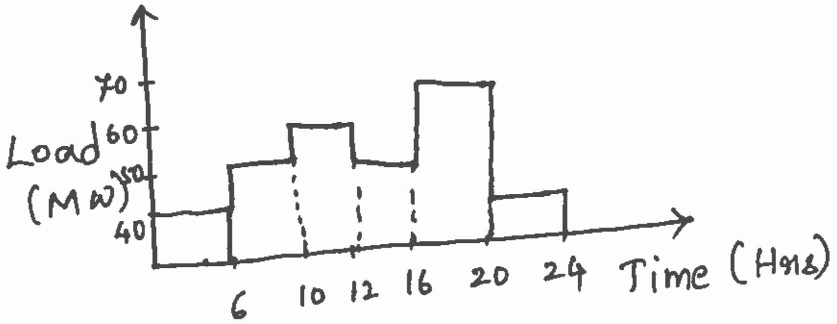
Prime Movers or water turbines

- In hydroelectric plants water turbines serve the purpose of prime mover which converts the kinetic energy of water into mechanical energy which is further utilized to drive the alternators generating electrical energy. (4m)

8. A generating station has the following daily load cycle:

Time (Hrs)	0-6	6-10	10-12	12-16	16 -20	20 -24
Load (MW)	40	50	60	50	70	40

Draw the load curve and find (i) maximum demand (ii) units generated per day (iii) average load and (iv) load factor.



(i) Maximum Demand = 70 MW.

(ii) Units Generated per day $= (6\times40)+(4\times50)+(2\times60)$
$+(4\times50)+(4\times70)+(4\times40)$

$= 1200$ MWh.

(iii) Average Load $= \dfrac{\text{Units generated/day}}{\text{Hrs in a day}}$

$= \dfrac{1200\text{ MWh}}{24} = \dfrac{1200\times10^6}{24}$

$= 50\times10^6 = 50$ MW.

(iv) Load Factor $= \dfrac{\text{Average Demand}\times100}{\text{Maximum Demand}} = \dfrac{50\text{MW}}{70\text{MW}}\times100$

$= 0.714\times100$

$= 71.4.$

9. Explain the types of Tariff methods.

Simple Tariff: When there is a fixed rate per unit of energy consumed, it is the Simple Tariff. In this type of tariff, the price charged per unit is constant. It does not vary with increase or decrease with the number of Units consumed.

Flat Rate Tariff: When different types of Consumers are charged at different uniform per unit rates, it is the Flat Rate Tariff. The rate for each type of consumer is arrived at by taking its load factor, diversity factor into consideration. The Bill will be Total units Consumed x Rate/Unit.

Block Rate Tariff: When a given block of energy is charged at a specific rate and the succeeding blocks of energy are charged progressively at reduced rates. Then the Tariff is called the Block Rate Tariff.

If the number of units generated increases, then the cost of generation per unit automatically decreases. For the first 30 units may be charged at the rate of 60paisa per unit, the next 25 units at the rate of 55paise per unit and the remaining additional units may be charged at the rate of 30 paise per unit. This type of tariff is being majorly used for residential and small commercial consumers.

Two Part Tariff: When the rate of electric energy is charged based on maximum demand of the consumer and the units consumed, it is called the Two-part Tariff.

In two-part tariff, the total charge to be made from the consumer is spilt into two components, fixed charges and running charges. The fixed charges depend upon the maximum demand of the consumer, while the running charges depend upon the no. of units consumed by the consumer.

Total Charges = Rs. (B kW+ C kWh)

B = Charges per kW of maximum demand

C = Charges per kWh of energy consumed

This type of Tariff is mostly applicable to Industrial Consumers.

Maximum Demand Tariff: It is similar to Two Part Tariff with the only difference that the maximum demand is actually measured by installing maximum demand meter in the premises of the consumer.

This type of tariff is mostly applied to big consumer, as a separate maximum demand meter is required.

Three-part tariff: When the total charge to be made from the consumer split into three parts, fixed charges, Semi fixed charges and running Charges.

Total Charge = Rs. (A + B kW + C kWh)

A = Fixed Charges during each billing period

B = Charge per kW of Maximum demand

C = Charge per kWh of energy consumed

Power Factor Tariff:

The Tariff in which Power factor of the Consumer's load is taken into consideration is known as Power Factor Tariff.

The following are the important types of Power Factor tariff

(i) kVA maximum demand tariff: It is a modified form of two-part tariff. The fixed charges are made on the basis of maximum demand in kVA and not in kW. As kVA is inversely proportional to power factor, a consumer having low power factor has to contribute more towards the fixed charge.

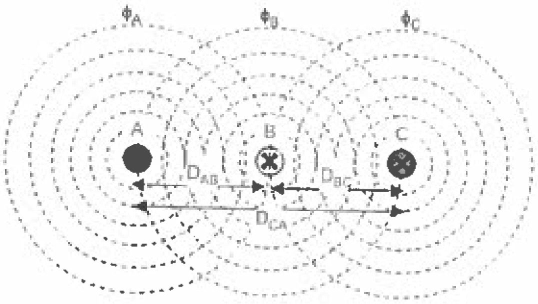
(ii) Sliding Scale Tariff: This is also known as average PF tariff. In this case an average power factor say 0.8 lagging is taken as reference. If the pf of the consumer falls below this factor, suitable additional charges are made. If the PF is above the reference, a discount is allowed to the consumer.

(iii) kW and kVAR Tariff: In this type both active power (kW) and reactive power (kVAR) supplied are charged separately. A consumer having low PF will draw more reactive Power and hence shall have to pay more charges.

10. Derive the equation for inductance of a three-phase overhead line.

Inductance in Three Phase Transmission Line:

In the three-phase transmission line, three conductors are parallel to each other. The direction of the current is same through each of the conductors.



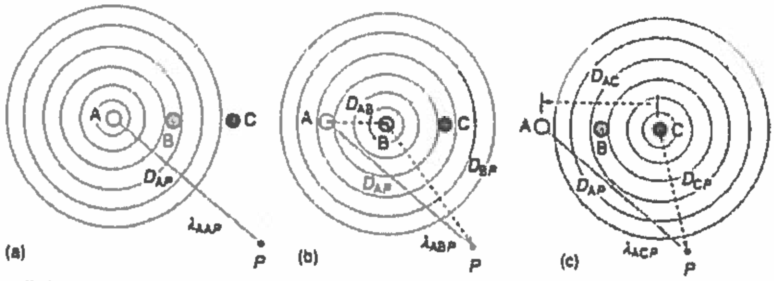
Let us consider conductor A produces magnetic flux φA, Conductor B produces magnetic flux φB, And conductor C produces magnetic flux φC. When they carry the current of the same magnitude "I", they are in flux linkage with each other.

Now, let us consider a point P near three conductors. So, flux linkage at point P due to current through conductor A is,

$$\lambda_{AP} = \lambda_{AAP} + \lambda_{ABP} + \lambda_{ACP}$$

Flux linkage at point P for conductor A due to current through conductor A =

$$\lambda_{AAP} = \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D_{AP}}{GMR_A}\right) \quad Wb/m$$



Flux linkage at point P for conductor A due to current through conductor B =

$$\lambda_{ABP} = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D_{BP}}{D_{AB}}\right) \quad Wb/m$$

Flux linkage at point P for conductor A due to current through conductor C =

$$\lambda_{ACP} = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D_{CP}}{D_{AC}}\right) \quad Wb/m$$

Therefore, flux linkage at point P for conductor A,

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right]$$
$$+ \frac{\mu_0}{2\pi}\left[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})\right] \quad Wb/m$$

(2m)

As, $D_{AP} = D_{BP} = D_{CP}$ and $I_A + I_B + I_C = 0$ in balanced system, then we can write that $I_A = -I_B - I_C$

$$\therefore \frac{\mu_0}{2\pi}[I_A \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})]$$

$$= \frac{\mu_0}{2\pi}[-I_B \times ln(D_{AP}) - I_C \times ln(D_{AP}) + I_B \times ln(D_{BP}) + I_C \times ln(D_{CP})] = 0$$

$$\Rightarrow \lambda_{AP} = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right] + 0 = \lambda_A(say)$$

$$So, \ \lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D_{AB}}\right) + I_C \times ln\left(\frac{1}{D_{AC}}\right)\right]$$

$$Similarly, \ \lambda_B = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{BA}}\right) + I_B \times ln\left(\frac{1}{GMR_B}\right) + I_C \times ln\left(\frac{1}{D_{BC}}\right)\right]$$

$$and,\ \lambda_C = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{D_{CA}}\right) + I_B \times ln\left(\frac{1}{D_{CB}}\right) + I_C \times ln\left(\frac{1}{GMR_C}\right)\right]$$

(2m)

For a balanced system,

$$D_{AB} = D_{BC} = D_{CA} = D$$

$$I_A + I_B + I_C = 0$$

In balanced system, then we can write that, $I_A = -I_B - I_C$

$$\lambda_A = \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + I_B \times ln\left(\frac{1}{D}\right) + I_C \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (I_B + I_C) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi}\left[I_A \times ln\left(\frac{1}{GMR_A}\right) + (-I_A) \times ln\left(\frac{1}{D}\right)\right]$$

$$= \frac{\mu_0}{2\pi} I_A \times ln\left(\frac{D}{GMR_A}\right)\ \ Wb/m$$

$$\lambda_B = \frac{\mu_0}{2\pi} I_B \times ln\left(\frac{D}{GMR_B}\right)\ \ Wb/m$$

$$\lambda_C = \frac{\mu_0}{2\pi} I_C \times ln\left(\frac{D}{GMR_C}\right)\ \ Wb/m$$

(2m)

$$So,\ inductance\ per\ metre\ per\ phase,\ \ L_{phase} = \frac{\mu_0}{2\pi} \times ln\left(\frac{D}{GMR_{phase}}\right)\ \ H/m$$

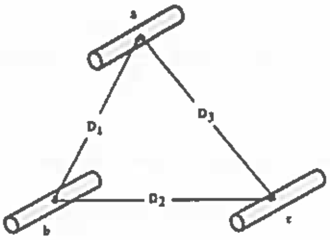
Consider a 3-phase overhead transmission line with phase conductors a, b, and c of radius r each spaced unsymmetrically such that the distance between the three conductors is D_1, D_2, and D_3 as shown in the figure below. Let I_a, I_b, and I_c be the currents flowing in the conductors a, b, and c respectively.

Assume that the whole system is balanced which leads to the flow of equal currents in the conductors i.e., $I_a = I_b = I_c = I$(say). The currents I_a, I_b, and I_c are displaced at 120° apart from each other. If I_a is taken as the reference phasor then the currents are given by,

$$I_a = I(1 - j0)$$

$$I_b = I(-0.5 - j0.866)\ and$$

$$I_c = I(-0.5 + j0.866)$$



Unsymmetrically Spaced 3-Phase Conductors

We know that the flux linkage of any conductor, in a group of conductors, is due to its own current and currents in the other conductors. Therefore, flux linkage of the conductor 'a' is due to its own current and currents in the conductor's 'b' and 'c', and it is given by,

$$\lambda_a = 2 \times 10^{-7}\left[I_a \ln\frac{1}{r'} + I_b \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_3}\right]$$

Similarly, flux linkages of conductors 'b' and 'c' is given by,

$$\lambda_b = 2 \times 10^{-7}\left[I_b \ln\frac{1}{r'} + I_a \ln\frac{1}{D_1} + I_c \ln\frac{1}{D_2}\right]$$

$$\lambda_c = 2 \times 10^{-7}\left[I_c \ln\frac{1}{r'} + I_a \ln\frac{1}{D_3} + I_b \ln\frac{1}{D_2}\right]$$

(2m)

Where, r' = Geometric mean radius (GMR) of conductor = 0.07788 × r. Substituting the values of I_a, I_b, and I_c in the above equation we get,

$$\lambda_a = 2 \times 10^{-7} \left[I \ln\frac{1}{r'} + I(-0.5 - j0.866) \ln\frac{1}{D_1} + I(-0.5 + j0.866) \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1} + \ln\sqrt{D_3} + j0.866 \ln D_1 + j0.866 \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\frac{\sqrt{3}}{2} \ln D_1 + j\frac{\sqrt{3}}{2} \ln\frac{1}{D_3} \right]$$

$$\lambda_a = 2 \times 10^{-7}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln\sqrt{\left(\frac{D_1}{D_3}\right)} \right]$$

We know that the inductance L$_a$ is given by,

$$L_a = \frac{\lambda_a}{I_a}$$

$$L_a = \frac{2 \times 10^{-7}}{I}\, I \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln\sqrt{\left(\frac{D_1}{D_3}\right)} \right]$$

$$\therefore L_a = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_3} + j\sqrt{3} \ln\sqrt{\left(\frac{D_1}{D_3}\right)} \right] H/m$$

Similarly, the inductance due to the conductor's 'b' and 'c' can be calculated and they are given by,

$$L_b = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_1 D_2} + j\sqrt{3} \ln\sqrt{\left(\frac{D_2}{D_1}\right)} \right] H/m$$

$$L_c = 2 \times 10^{-7} \left[\ln\frac{1}{r'} + \ln\sqrt{D_2 D_3} + j\sqrt{3} \ln\sqrt{\left(\frac{D_3}{D_2}\right)} \right] H/m$$

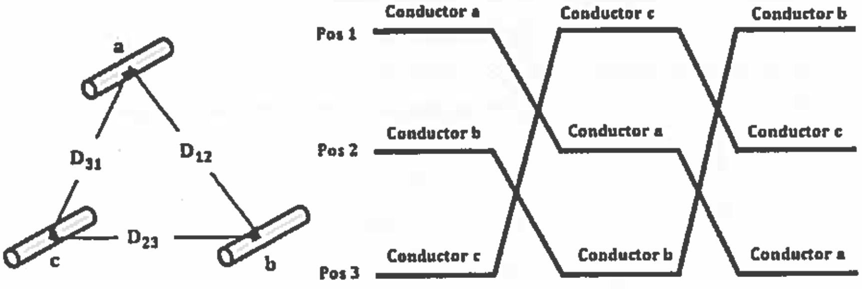
When conductors are unsymmetrically spaced in a 3-phase line, the flux linkage and inductance per phase are not identical, and knowing the inductance of each phase becomes complicated and it results in an unbalanced circuit.

The equilibrium in the circuit can be retained by shifting the places of the conductor at every period such that the conductors take the initial place of every conductor at the same distances.

Such an arrangement of the conductor's position obtained by shifting their places is called transposition. By transposition, we can obtain almost the same inductance between the conductors. Let us see the expression for inductance when the 3-phase line is transposed.

Inductance of Unsymmetrically Spaced 3-Phase Transmission Line When Transposed: (2m)

The simple configuration of a 3-phase conductor is shown in the figure. As the conductors are transposed, so their position in the transposition cycle would be as shown in the figure below.



Unsymmetrically Spaced 3-Phase Conductors with Transposition Cycle

When the conductors are connected in parallel, it results in low inductance keeping the distance between phases as small as possible. For deriving the inductance per phase of 3-phase conductors placed unsymmetrically, we need to determine the flux linkage of each conductor for every position it occupies in a transposition cycle. After that, we can determine the average flux linkages. So, the flux linkage of phase 'a' in position 1 is given as,

$$\lambda_{a1} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}} + I_c \ln\frac{1}{D_{31}}\right]$$

When 'a' is in position 2, 'b' in position 3 and 'c' in position 1,

$$\lambda_{a3} = 2 \times 10^{-7}\left[I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{31}} + I_c \ln\frac{1}{D_{23}}\right]$$

The average value of flux linkage of a single-phase 'a' is,

$$\lambda_a = \frac{\lambda_{a1} + \lambda_{a2} + \lambda_{a3}}{3}$$

$$\lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} + I_b \ln\frac{1}{D_{12}D_{23}D_{31}} + I_c \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

We know that

$$I_a + I_b + I_c = 0$$

$$I_b + I_c = -I_a$$

$$\therefore \lambda_a = \frac{2 \times 10^{-7}}{3}\left[3I_a \ln\frac{1}{D_s} - I_a \ln\frac{1}{D_{12}D_{23}D_{31}}\right]$$

$$\lambda_a = 2 \times 10^{-7} I_a \ln \sqrt{\frac{\sqrt[3]{D_{12}D_{23}D_{31}}}{D_s}}$$

Therefore, the average inductance per phase is,

$$L_a = 2 \times 10^{-7} \ln\frac{D_{eq}}{D_s}\ H/m$$

$$Where, D_{eq} = \sqrt[3]{D_{12}D_{23}D_{31}}$$

Where,

D_{eq} = Geometric mean of three distances of unsymmetrical line

D_s = GMR (geometric mean radius) of the conductor.

The above expression is the inductance per phase of a 3-phase transmission line with unsymmetrical spacing but lines are transposed. Nowadays the transposition of conductors is made at switching stations to balance inductance.

If the conductors are equispaced,

D1 = D2 = D3 = D

$$L = 2\,x\,10^{-7} \ln\frac{\sqrt[3]{d^3}}{r'} \Rightarrow L = 2\,x\,10^{-7} \ln\frac{d}{r'}$$

(2m)

11. Derive the equation for capacitance of a two-wire overhead line.

Electric Field Intensity due infinite line charge:

Consider a long wire having q coulomb/m as shown in figure

Using Gauss's law, Field Intensity (E) at a point P, which is r metre from the conductor, can be calculated as

$$\oint \overline{D} \cdot \overline{ds} = Q$$

The Flux density at point P, considering the cylindrical shell of radius r and length l can be calculated using Gauss's Law

$$D \cdot 2\Pi r \cdot l = ql \Rightarrow D = \frac{q}{2\pi r}\ coulomb/m^2$$

And Electric Field intensity E is given by

$$E = \frac{D}{\varepsilon_0}$$

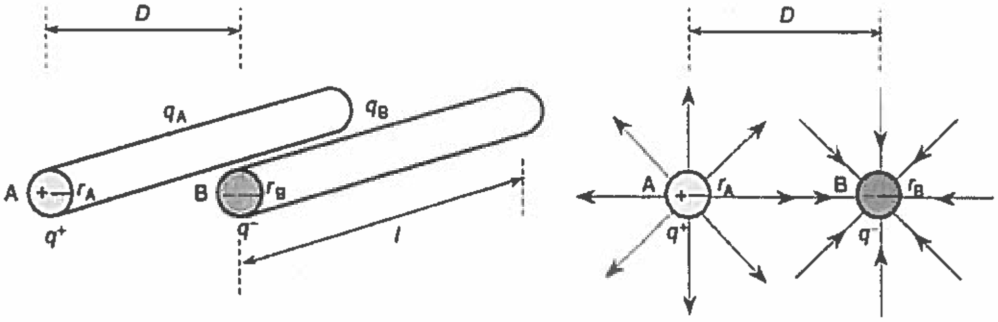
Capacitance and Capacitive Reactance:

Capacitance exists among transmission line conductors due to their potential difference. To evaluate the capacitance between conductors in a surrounding medium with permittivity ε, it is necessary to determine the voltage between the conductors, and the electric field strength of the surrounding. (4m)

Capacitance of a Single-Phase Line with Two Wires:

Consider a two-wire single-phase line with conductors A and B with the same radius r, separated by a distance D > r_A and r_B. The conductors are energized by a voltage source such that conductor A has a charge q+ and conductor B a charge q- as shown in Fig.

The charge on each conductor generates independent electric fields. Charge q+ on conductor A generates a voltage V_{AB-A} between both conductors. Similarly, charge q- on conductor B generates a voltage V_{AB-B} between conductors.



(4m)

Electric field produced from a two-wire single-phase system.

V_{AB-A} is calculated by integrating the electric field intensity, due to the charge on conductor A, on conductor B from r_A to D

$$V_{AB-A} = \int_{r_A}^{D} E_A \, dx = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D}{r_A}\right]$$

V_{AB-B} is calculated by integrating the electric field intensity due to the charge on conductor B from D to r_B

$$V_{AB-B} = \int_{D}^{r_B} E_B \, dx = \frac{-q}{2\pi\varepsilon_0} \ln\left[\frac{r_B}{D}\right]$$

The total voltage is the sum of the generated voltages V_{AB-A} and V_{AB-B}

$$V_{AB} = V_{AB-A} + V_{AB-B} = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D}{r_A}\right] - \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{r_B}{D}\right] = \frac{q}{2\pi\varepsilon_0} \ln\left[\frac{D^2}{r_A r_B}\right]$$

If the conductors have the same radius, r_A=r_B=r, then the voltage between conductors V_{AB}, and the capacitance between conductors C_{AB}, for a 1-m line length are

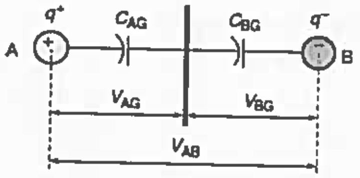
$$V_{AB} = \frac{q}{\pi\varepsilon_0} \ln\left[\frac{D}{r}\right] \ \text{(V)}$$

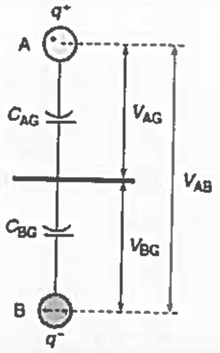
$$C_{AB} = \frac{\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \ \text{(F/m)}$$

The voltage between each conductor and ground (G) is one-half of the voltage between the two conductors. Therefore, the capacitance from either line to ground is twice the capacitance between lines (4m)

$$V_{AG} = V_{BG} = \frac{V_{AB}}{2} \ \text{(V)}$$

$$C_{AG} = \frac{q}{V_{AG}} = \frac{2\pi\varepsilon_0}{\ln\left[\frac{D}{r}\right]} \ \text{(F/m)}$$





12. Classify the types of transmission lines with model representations.

Classification of Transmission Lines - Short, Medium & Long Transmission Lines:

A. AC transmission line, and
B. DC transmission line

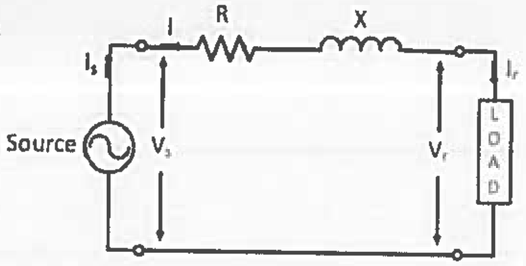
Depending upon the operating voltage and length, the overhead ac transmission lines are classified as,

1. Short transmission lines
2. Medium transmission lines
3. Long transmission lines

Short Transmission Line:

- ✓ Overhead transmission line is less than 50km
- ✓ Operating voltages of less than 20kV.

In these lines, the effect of capacitance is neglected due to smaller length and low operating voltage. Hence, the resistance and inductance effects of the line are considered while determining the performance of the short transmission line as shown below.



Equivalent circuit model of a short line

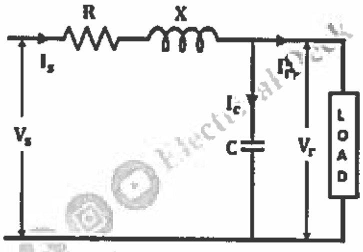
Circuit Globe

Medium Transmission Line:

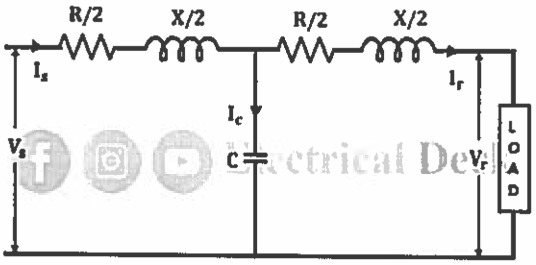
- ✓ Length of the overhead transmission line is in the range of 50-150km
- ✓ the operating voltage is greater than 20kV.

Based on the location of the capacitance at different places, the medium transmission lines have different configurations. These configurations show the different ways in which the effect of capacitance is taken into consideration. The three configurations based on the location of capacitance are,

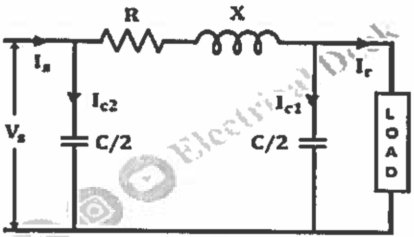
i. End condenser representation of medium transmission lines
ii. Nominal-T representation of medium transmission lines
iii. Nominal-π representation of medium transmission lines.



End Condenser Method of Medium Transmission Line



Nominal-T Method of Medium Transmission Line

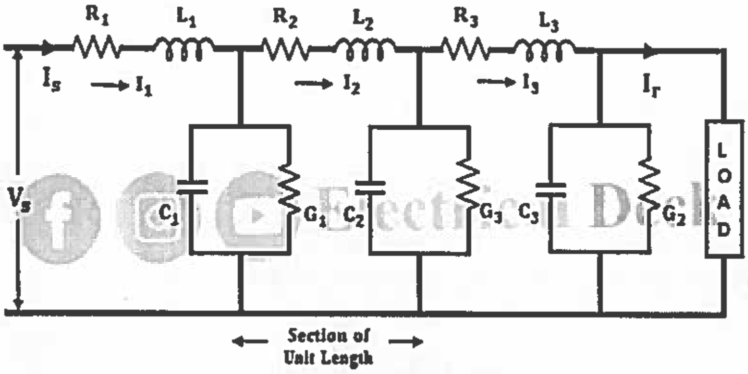


Nominal-π Method of Medium Transmission Line

Long Transmission Line:

✓ Overhead transmission lines whose length is more than 150km

✓ Operating voltage of these lines is more than 100kV



Long Transmission Line

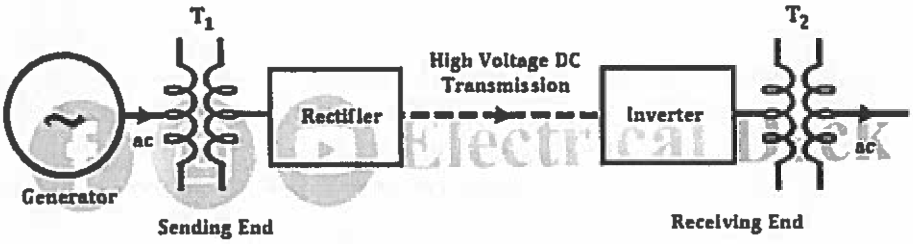
DC Transmission Line

In AC transmission lines, for transmission of power over long distances at higher voltages, the cost of transmission line and loss increases.

Also, the AC long transmission line suffers from problems like stability limits, voltage control, line compensation, interconnection of lines, ground impedance, etc due to an increase in voltage levels and distance.

The various problems associated with long-distance AC transmission have led to the development of HVDC (high voltage direct current) transmission nothing but a dc transmission line.

The use of DC power for long transmission lines has various advantages like no stability problem, absence of charging current, no skin effect, need for reactive compensation, bulk power transfer, economic power transmission, etc.



HVDC Transmission System

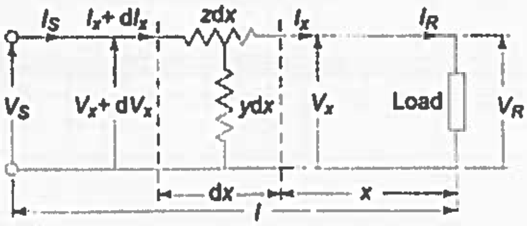
It is inefficient to use a dc transmission system for shorter and medium distances.

13. (a) Derive the expressions for the Performance of long transmission lines using rigorous method with relevant equations.

ABCD parameters in Long Transmission Line-rigorous Solution:

For rigorous solution, consider the Line Impedance and admittance uniformly distributed and not lumped.

Consider a small element of length dx situated at a distance x from receiving end. Let Z and Y denoted respectively the Series Impedance and shunt Admittance of the line per unit length.



Schematic diagram of a long line

The voltage at two ends of the element are denotes as V (towards receiving end) and V + dV (towards sending end) respectively.

Let dx be an elemental section of the line at a distance x from the receiving-end having a series impedance zdx and a shunt admittance ydx. The rise in voltage to neutral over the elemental section in the direction of increasing x is dV_x.

We can write the following differential relationships across the elemental section:

$$dVx = I_x z dx \text{ or } \frac{dV_x}{dx} = zI_x \qquad (5.14)$$

$$dIx = V_x y dx \text{ or } \frac{dI_x}{dx} = yVx \qquad (5.15)$$

It may be noticed that the kind of connection (e.g. T or π) assumed for the elemental section, does not affect these first order differential relations. Differentiating Eq. (5.14) with respect to x, we obtain

$$\frac{d^2V_x}{dx^2} = \frac{dI_x}{dx} z$$

Substituting the value of dI_x/dx from Eq. (5.15), we get

$$\frac{d^2V_x}{dx^2} = yzV_x \qquad (5.16)$$

This is a linear differential equation whose general solution can be written as follows:

$$V_x = C_1 e^{\gamma x} + C_2 e^{-\gamma x} \qquad (5.17)$$

where

$$\gamma = \sqrt{yx} \qquad (5.18)$$

and C_1 and C_2 are arbitrary constants to be evaluated.

Differentiating Eq. (5.17) with respect to x:

$$\frac{dV_x}{dx} = C_1 \gamma e^{\gamma x} - C_2 \gamma e^{-\gamma x} = zI_x$$

$$I_x = \frac{C_1}{Z_r} e^{\gamma x} - \frac{C_2}{Z_c} e^{-\gamma x} \qquad (5.19)$$

Where

$$Z_c = \left(\frac{z}{y}\right)^{1/2} \qquad (5.20)$$

The constants C_1 and C_2 may be evaluated by using the end conditions, i.e. when x = 0, $V_x = V_R$ and $I_x = I_R$. Substituting these values in Eqs. (5.17) and (5.19) gives

$$V_R = C_1 + C_2$$

$$I_R = \frac{1}{Z_c}(C_1 - C_2)$$

which upon solving yield

$$C_1 = \frac{1}{2}(V_R + Z_cI_R)$$

$$C_2 = \frac{1}{2}(V_R - Z_cI_R)$$

With C_1 and C_2 as determined above, Eqs. (5.17) and (5.19) yield the solution for V_x and I_x as

$$V_x = \left(\frac{V_R + Z_cI_R}{2}\right)e^{\gamma x} + \left(\frac{V_R - Z_cI_R}{2}\right)e^{-\gamma x}$$

$$I_x = \left(\frac{V_R/Z_c + I_R}{2}\right)e^{\gamma x} - \left(\frac{V_R/Z_c - I_R}{2}\right)e^{-\gamma x} \qquad (5.21)$$

Here Z_c is called the **characteristic impedance** of the Long Transmission Line and γ is called the **propagation constant.**

Knowing V_R, I_R and the parameters of the line, using Eq. (5.21) complex number rms values of V_x and I_x at any distance x along the line can be easily found out.

A more convenient form of expression for voltage and current is obtained by introducing hyperbolic functions. Rearranging Eq. (5.21), we get

$$V_x = V_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right) + I_RZ_c\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right)$$

$$I_x = V_R\frac{1}{Z_c}\left(\frac{e^{\gamma x} - e^{-\gamma x}}{2}\right) + I_R\left(\frac{e^{\gamma x} + e^{-\gamma x}}{2}\right)$$

These can be rewritten after introducing hyperbolic functions, as

$$V_x = V_R \cosh \gamma x + I_RZ_c \sinh \gamma x \qquad (5.22)$$

$$I_x = I_R \cosh \gamma x + V_R \frac{1}{Z_c}\sinh \gamma x$$

where x = *l*, $V_x = V_s$, $I_x = I_s$

$$\therefore \quad \begin{bmatrix} V_S \\ I_S \end{bmatrix} = \begin{bmatrix} \cosh \gamma l & Z_c \sinh \gamma l \\ \frac{1}{Z_c}\sinh \gamma l & \cosh \gamma l \end{bmatrix} \begin{bmatrix} V_R \\ I_R \end{bmatrix} \qquad (5.23)$$

Here

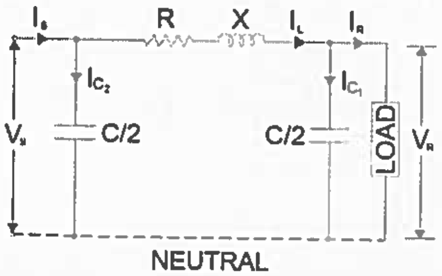
$$A = D = \cosh \gamma l$$

$$B = Z_c \sinh \gamma l \qquad (5.24)$$

$$C = \frac{1}{Z_c}\sinh \gamma l$$

13. (b) Using nominal π method, derive an expression for sending end voltage and current for a medium transmission line.

In Nominal π Method, the shunt capacitance of each line i.e. phase to neutral is divided into two equal parts. One part is lumped at the sending end while the other is lumped at receiving end as shown in figure below.



Notice that, in this method there is no effect of shunt capacitance at sending end on the line voltage drop and hence on voltage regulation but this accounts for the charging current in sending end.

Let

IR = Load Current per phase

R = Resistance per phase

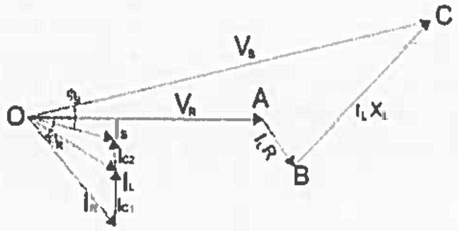
X = Reactance per phase

C = Capacitance per phase

Cos Ø R = Receiving end power factor (lagging)

Vs = Sending end voltage

Let us now draw the phasor. Assume receiving end voltage VR as reference and load current IR lagging this voltage by ØR.



Therefore,

VR= VR+ j0

and IR= IR<-ØR = IR (CosØR – jSinØR)

Charging Current at load end IC1 = j(ωC/2) VR = jπfCVR

Line Current **IL = IR + IC1** (phasor sum)

Sending end voltage Vs = VR + IL (R+jX)

Now,

Charging current at sending end IC2 = j(ωC/2)Vs = jπfCVs

Hence, Sending end current **Is = IL + IC2** (phasor sum)

Thus, sending end current and voltage is calculated as above and from these parameters the performance of line is evaluated.

14. Explain the different methods used to improve the string efficiency of insulators with necessary diagrams.

Methods of Improving String Efficiency

Method # 1. By Using Insulators with Larger Discs or by Providing Each Insulator Unit with a Metal Cap:

It is clear from the expression of string efficiency that the string efficiency increases with the decrease in value of K (i.e. the ratio of shunt capacitance to mutual capacitance). One method is to design the units such that the mutual capacitance (capacitance of each unit) is much greater than the shunt capacitance (capacitance to earth). This can be achieved by using insulators with larger discs or providing each insulator unit with a metal cap. The ratio K can be made 1/6 to 1/10 by this method.

Method # 2. By Using Longer Cross-Arms:

The ratio of shunt capacitance to mutual capacitance, K can alternatively be reduced by using longer cross-arms so that the horizontal distance from line support (pole or tower) is increased thereby decreasing the shunt capacitance. But the limitations of cost and mechanical strength of line supports do not allow the cross-arms to be too long and it has been found that in practice it is not possible to obtain the value of K less than 0.1.

Method # 3. By Capacitance Grading:

It is seen that non-uniform distribution of voltage across an insulator string is due to leakage current from the insulator pin to the supporting structure, which cannot be eliminated. However, it is possible that discs of different capacities are used such that the product of their capacitive reactance and the current flowing through the respective unit is same. This can be achieved by grading the mutual capacitance of the insulator units i.e., by having lower units of more capacitance—maximum at the line unit and minimum at the top unit, nearest to the cross-arm. By this method complete equality of voltage across the units of an insulator string can be obtained but this method needs a large number of different-sized insulator units and maintaining spares of all varieties of insulator discs. So this method is not used in practice below 200 kV.

Consider a 4-unit string. Let C be the capacitance of the top unit and let the capacitances of others units are C_2, C_3 and C_4, as shown in Fig.

Assume $C_1 = k\,C$

Applying Kirchhoff's first law to node A we get, $I_2 = I_1 + i_1$

=> $\omega C_2 v = \omega C v + \omega C_1 v$ or $C_2 = C + K C = C(1 + K)$ ... (9.11)

Applying Kirchhoff's first law to node B we get, $I_3 = I_2 + i_2$

or $\omega C_3 v = \omega C_2 v + \omega C_1 \times 2v$

(or) $C_3 = C_2 + 2KC = C(1 + K) + 2KC = C(1 + 3K)$ ... (9.12)

Applying Kirchhoff's first law to node C we get,

$I_4 = I_3 + i_3$

or $\omega C_4 v = \omega C_3 v + \omega C_1 \times 3v$

or $C_4 = C_3 + 3KC = C(1 + 3K) + 3KC = C(1 + 6K)$ ...(9.13)

Thus, it will be possible to equalize the potential across all the units, if their capacitances are in the ratio of 1: (1 + K): (1 + 3 K): (1 + 6 K) and so on.

But in practice it is impossible to obtain such units which will have their capacitances in above ratio, although nearby results can be obtained by employing standard insulators for most of the units and employing larger units adjacent to the line.

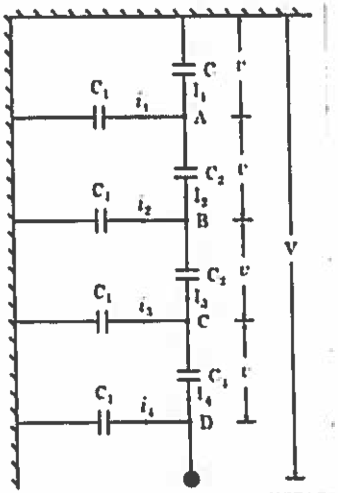
Method # 4. By Static Shielding:

In case of capacitance grading, insulator units of different capacitances are used so that the flow of different currents through the respective units produce equal voltage drop. In static shielding, pin to supporting structure charging currents are exactly cancelled so that the same current flows through the identical insulator units and produce equal voltage drops across each insulator unit.

This arrangement is given is Fig. 9.22. In this method a guard or grading ring, which usually takes the form of a large metal ring surrounding the bottom unit and electrically connected to the metal work at the bottom of this unit, and therefore to the line conductor.

The guard ring screens the lower units, reduces their earth capacitance C_1 and introduces a number of capacitances between the line conductor and the various insulator unit caps. These capacitances are greater for lower units and thus the voltages across them are reduced. With this method also it is impossible to obtain in practice an equal distribution of voltage but considerable improvements are possible.

Let the capacitances between the links and the shield be C_x, C_y and C_z respectively as shown in Fig. 9.22, and let v be the potential across each unit.



Since the capacitance of each unit is same, therefore, their charging currents I_1, I_2, I_3, and I_4 would be same, let it be I. If $C_1 = K\,C$

Applying Kirchhoff s first law to node A we get,

$I + i_x = I + i_1$ or $i_x = i_1$... (9.14)

Similarly, $i_y = i_2$... (9.15)

and $i_z = i_3$... (9.16)

Also, $i_1 = \omega\,C_1 v = \omega\,K\,C\,v$ ... (9.17)

$i_2 = 2\,\omega\,C_1\,v = 2\,\omega\,K\,C\,v$ ... (9.18)

$i_3 = 3\,\omega\,C_1 v = 3\,\omega\,K\,C\,v$ ... (9.19)

The potential causing current i_x is 3 v (voltage across three units leaving the top one).

So, $i_x = \omega\,C_x \times 3\,v = 3\,\omega\,C_x\,v$ ... (9.20)

Comparing Eqs. (9.14), (9.17) and (9.20), we have,

$3\,\omega\,C_x v = \omega\,K\,C\,v$ or $C_x = K\,C/3$ ... (9.21)

The potential causing current y is 2 v and therefore,

$I_y = 2\,\omega\,C_y\,v$ ... (9.22)

Comparing Eqs. (9.15), (9.18) and (9.22) we have,

$2\,\omega\,C_y\,v = 2\,\omega\,K\,C\,v$ or $C_y = K\,C$ ... (9.23)

The potential causing current i_z is v and therefore,

$i_z = \omega\,C_z\,v$ ... (9.24)

Comparing Eqs. (9.16), (9.19) and (9.24), we have,

$\omega\,C_z\,v = 3\,\omega\,K\,C\,v$

or $C_z = 3\,K\,C$

In general, if there are n units

$i_1 = \omega\,K\,C\,v$ and $i_x = (n - 1)\,\omega\,C_x v$

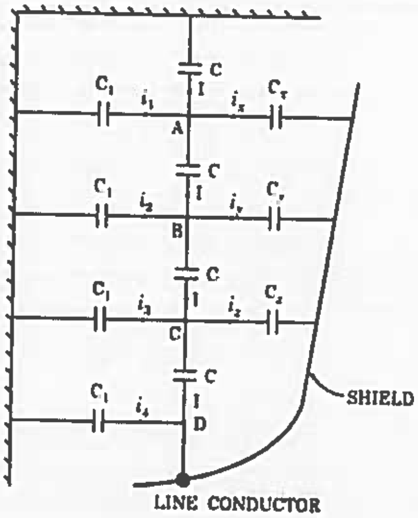
or $C_x = KC/\,(n\text{-}1)$

Similarly, $C_y = 2KC/\,(n\text{-}2)$

and $C_z = 3KC/\,(n\text{-}3)$

or The capacitance of the pth metal link to the line is given as:

$C_p = pKC/\,(n\text{-}p)$

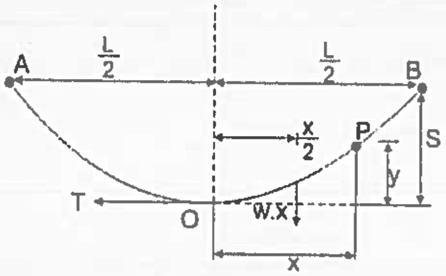


15. (a) Derive an expression for sag of a line supported between two supports of the same tower height.

Calculation of Sag in a Transmission Line:

Sag calculation for supports is at equal levels

Suppose, AOB is the conductor. A and B are points of supports. Point O is the lowest point and the midpoint.



Let, L = length of the span, i.e. AB

w = weight per unit length of the conductor

T = tension in the conductor.

We have chosen any point on the conductor, say point P.

The distance of point P from the Lowest point O is x.

y is the height from point O to point P.

Equating two moments of two forces about point O as per the figure above we get,

$$Ty = wx \times \frac{x}{2}$$

$$Now, y = \frac{wx^2}{2T},$$

The maximum dip (**sag**) is represented by the value of y at either of the supports A and B. At support A, x = l/2 and y = S.

$$Then\ S = \frac{wL^2}{8T}$$

15. (b) A 132 kV transmission line has the following data:
Wt. of conductor = 680 kg/km;
Length of span = 260 m
Ultimate strength = 3100 kg;
Safety factor = 2.
Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 m.

Sol: Weight of conductor/mt run, $w = \frac{680}{1000} = 0.68$ kg.

Working Tension, $T = \frac{\text{Ultimate Strength}}{\text{Safety factor}}$

$= \frac{3100}{2} = 1550$ kg.

Span length, $l = 260$ m.

$\therefore$ Sag $= \frac{wl^2}{8T} = \frac{0.68 \times 260^2}{8 \times 1550}$

$= 3.7$ m.

Height = GC + Sag = 10 + 3.7

= 13.7 m.

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(HOD) 7/1/23.