NSRIT

AUTONOMOUS

ANSWER KEY & SCHEME OF EVALUATION

B. Tech. (S 1)
Supplementary) ACY
2021 - 2022

ACADEMIC
REGULATION
2020

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NSRIT

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	Common to All	Academic Year	2021- 2022
Course Code	20BSX11	Test Duration	3 Hrs. Max. Marks 70	Semester	I
Course	LINEAR ALGEBRA AND DIFFERENTIAL EQUATIONS				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	If λ is an Eigen value of a matrix A, then show that λ^2 is Eigen value of A^2	20BSX11.1	L1
2	Find the rank of a matrix $\begin{bmatrix} 1 & 2 & 3 & 4 \\ -2 & 0 & 5 & 7 \end{bmatrix}$	20BSX11.2	L1
3	Solve $(2x^3 - y)dx + xdy = 0$	20BS X11.3	L2
4	Solve $\frac{d^4y}{dx^4} + 4\frac{d^2y}{dx^2} = 0$	20BS X11.4	L2
5	State Lagrange's mean value theorem	20BS X11.5	L2

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Find the eigen values and eigen vectors of the matrix: $A = \begin{bmatrix} 8 & -6 & 2 \\ -6 & 7 & -4 \\ 2 & -4 & 3 \end{bmatrix}$	6M	20BSX11.1	L2
6 (b)	Solve the system of equations $2x + 2y + 2z = 0, -2x + 5y + 2z = 1,\ 8x + y + 4z = -1$	6M	20BSX11.1	L3
	OR			
7 (a)	Solve the system of equations x + y – z + w = 0, x – y + 2z -w = 0,3x + y + w = 0	6M	20BSX11.1	L3
7 (b)	Find the rank of the matrix A by reducing it to its normal form where: A = $\begin{bmatrix} 1 & 2 & 3 & -1 \\ 2 & 1 & 3 & 1 \\ 1 & 0 & 1 & 1 \\ 0 & 1 & 1 & 1 \end{bmatrix}$	6M	20BSX11.1	L2
8	Verify Cayley -Hamilton theorem for the matrix A= $\begin{bmatrix} 2 & -1 & 1 \\ -1 & 2 & -1 \\ 1 & -1 & 2 \end{bmatrix}$ hence find A^{-1}	12M	20BSX11.2	L2
	OR			
9	Reduce the quadratic form $3x^2 + 5y^2 - 3z^2 - 2xy - 2yz + 2zx$ to canonical form by the use of an orthogonal transformation hence find the rank, index, signature and nature of the quadratic form	12M	20BSX11.2	L2
10 (a)	Solve $(1 + y^2)\ dx = (tan^{-1}y - x)dy$	6M	20BSX11.3	L2
10 (b)	Solve $(y^2 - xy)dx + x^2dy = 0$	6M	20BSX11.3	L3
	OR			
11 (a)	Find the orthogonal trajectories of the family of $curves\quad ax + y^2 = x^2$	6M	20BSX11.3	L2

11 (b)	The number N of bacteria grew at a rate proportional to N. The value of N was initially 100 and increased to 332 in one hour. What is the value of N after $1\frac{1}{2}$ hours?	6M	20BSX11.3	L3
12 (a)	Solve $(D^4 + 2D^3 - 3D^2)y = 0\,,\ where\ D = \frac{d}{dx}$	6M	20BSX11.4	L2
12 (b)	Solve $(D^2 + 6D + 9\,)y = 5^x,\ \ where\ D = \frac{d}{dx}$	6M	20BSX11.4	L3
	OR			
13 (a)	Solve$(D^2 + 3D + 2)y = sin2x,\ \ where\ D = \frac{d}{dx}$	6M	20BSX11.4	L2
13 (b)	Use method of variation of parameters to solve the differential equation: $\frac{d^2y}{dx^2} + y = sec\,x$	6M	20BSX11.4	L3
14 (a)	Verify Lagrange's mean value theorem for f(x)= log x in the interval [1, e]	6M	20BSX11.5	L2
14 (b)	Determine whether the functions u = $\frac{x-y}{x+y}$ $, v = \frac{xy}{(x+y)^2}$are functionally dependent or not.	6M	20BSX11.5	L3
	OR			
15 (a)	Expand e^x log (1 +y) in powers of x and y upto terms of third degree	6M	20BSX11.5	L2
15 (b)	Show that rectangular solid of maximum volume that can be inscribed in a given sphere is a cube	6M	20BSX11.5	L3

SCHEME OF VALUATION

1. λ is an eigen value of A

$\Rightarrow \exists\, X (\neq 0) \ni AX = \lambda X$ —— 1M

by pre-multiplying with A, we get

$A(AX) = A(\lambda X)$

$\Rightarrow A^2X = \lambda (AX) = \lambda(\lambda X) = \lambda^2 X$

$\Rightarrow A^2X = \lambda^2 X \Rightarrow \lambda^2$ is an eigen value of A^2 — 1M

2. $\begin{bmatrix} 1 & 2 & 3 & 4 \\ -2 & 0 & 5 & 7 \end{bmatrix}$

$R_2 \to R_2 + 2R_1 \sim \begin{bmatrix} 1 & 2 & 3 & 4 \\ 0 & 4 & 11 & 15 \end{bmatrix}$ —— 1M

$R_2 \to R_2/4 \sim \begin{bmatrix} 1 & 2 & 3 & 4 \\ 0 & 1 & \frac{11}{4} & \frac{15}{4} \end{bmatrix}$, which is in echelon form with 2 non-zero rows

$\therefore \rho(A) = 2.$ —— 1M

3. $(2x^3 - y)dx + x\,dy = 0. \Rightarrow 2x^3 dx + x\,dy - y\,dx = 0$

$\Rightarrow 2x\,dx + \left(\frac{x\,dy - y\,dx}{x^2}\right) = 0$

$\Rightarrow 2x\,dx + d\left(\frac{y}{x}\right) = 0$

by int-, $x^2 + \frac{y}{x} = c.$ —— 2M

4. $\frac{d^4y}{dx^4} + 4\frac{d^2y}{dx^2} = 0 \Rightarrow (D^4 + 4D^2)y = 0$

AE is $m^4 + 4m^2 = 0 \Rightarrow m^2(m^2+4) = 0$

$\Rightarrow m^2 = 0,\ m^2 = -4 = 4i^2 = (2i)^2$

$\Rightarrow m = 0, 0, 2i, -2i$ — 1M

$\therefore y = c_1 + c_2 x + c_3 \cos 2x + c_4 \sin 2x.$ — 1M

5. If $f(x)$ is defined in an interval $[a,b]$ such that

i) $f(x)$ is continuous in $[a,b]$ — 1M

& ii) $f(x)$ is derivable in (a,b), then $\exists$ atleast one real number $c \in (a,b)$ such that

$$f'(c) = \frac{f(b)-f(a)}{b-a}.$$ — 1M

6. (a) $A = \begin{bmatrix} 8 & -6 & 2 \\ -6 & 7 & -4 \\ 2 & -4 & 3 \end{bmatrix}$

The ch. eq. is $|A-\lambda I| = 0 \Rightarrow \begin{vmatrix} 8-\lambda & -6 & 2 \\ -6 & 7-\lambda & -4 \\ 2 & -4 & 3-\lambda \end{vmatrix} = 0$

$\Rightarrow (8-\lambda)\left[\lambda^2 - 10\lambda + 21 - 16\right] + 6\left[-18 + 6\lambda + 8\right]$
$+ 2\left[24 - 14 + 2\lambda\right] = 0$

$\Rightarrow (8-\lambda)(\lambda^2 - 10\lambda + 5) + 6(6\lambda - 10) + 2(2\lambda + 10) = 0$

$\Rightarrow 8\lambda^2 - 80\lambda + 40 - \lambda^3 + 10\lambda^2 - 5\lambda + 36\lambda - 60 + 4\lambda + 20 = 0$

$\Rightarrow -\lambda^3 + 18\lambda^2 - 45\lambda = 0$

$\Rightarrow -\lambda(\lambda^2 - 18\lambda + 45) = 0 \Rightarrow -\lambda(\lambda-3)(\lambda-15) = 0$

$\Rightarrow \lambda = 0, 3, 15.$

$\therefore$ Eigen values are $0, 3, 15$. — 3M

Eigen vectors:- ~~By solving~~ $(A-(0)I)X = 0$, we get $X = \begin{pmatrix} 1 \\ 2 \\ 2 \end{pmatrix}$

By solving $(A-3I)X = 0$, we get $X = \begin{pmatrix} 2 \\ 1 \\ -2 \end{pmatrix}$

By solving $(A-15I)X = 0$, we get $X = \begin{pmatrix} 2 \\ -2 \\ 1 \end{pmatrix}$

3M

6.(b) Given system is $\begin{pmatrix} 2 & 2 & 2 \\ -2 & 5 & 2 \\ 8 & 1 & 4 \end{pmatrix}\begin{pmatrix} x \\ y \\ z \end{pmatrix} = \begin{pmatrix} 0 \\ 1 \\ -1 \end{pmatrix}$

The augmented matrix is $[A|B] = \begin{bmatrix} 2 & 2 & 2 & 0 \\ -2 & 5 & 2 & 1 \\ 8 & 1 & 4 & -1 \end{bmatrix}$

$R_1 \to R_1/2 \sim \begin{bmatrix} 1 & 1 & 1 & 0 \\ -2 & 5 & 2 & 1 \\ 8 & 1 & 4 & -1 \end{bmatrix}$

$R_2 \to R_2 + 2R_1, \quad R_3 \to R_3 - 8R_1$

$\sim \begin{bmatrix} 1 & 1 & 1 & 0 \\ 0 & 7 & 4 & 1 \\ 0 & -7 & -4 & -1 \end{bmatrix} \quad R_2 \to \frac{R_2}{7} \sim \begin{bmatrix} 1 & 1 & 1 & 0 \\ 0 & 1 & 4/7 & 1/7 \\ 0 & -7 & -4 & -1 \end{bmatrix}$

$R_3 \to R_3 + 7R_2 \sim \begin{bmatrix} 1 & 1 & 1 & 0 \\ 0 & 1 & \frac{4}{7} & 0 \\ 0 & 0 & 0 & 0 \end{bmatrix}$, which is in echelon form. 2M

$\Rightarrow \rho(A) = 2, \ \rho(A|B) = 2$ 1M

$\Rightarrow \rho(A) = \rho(A|B) < 3 \Rightarrow$ ~~The system is inconsistent~~ 2M

& ~~hence~~ $\exists$ ~~no solution~~ infinite no. of solutions.

7.(a) $A = \begin{bmatrix} 1 & 1 & -1 & 1 \\ 1 & -1 & 2 & -1 \\ 3 & 1 & 0 & 1 \end{bmatrix}$ $\begin{matrix} R_2 \to R_2 - R_1 \\ R_3 \to R_3 - 3R_1 \\ \sim \end{matrix}$ $\begin{bmatrix} 1 & 1 & -1 & 1 \\ 0 & -2 & 3 & -2 \\ 0 & -2 & 3 & -2 \end{bmatrix}$

$R_2 \to \frac{R_2}{-2} \sim \begin{bmatrix} 1 & 1 & -1 & 1 \\ 0 & 1 & -3/2 & 1 \\ 0 & -2 & 3 & -2 \end{bmatrix}$

$R_3 \to R_3 + 2R_2 \sim \begin{bmatrix} 1 & 1 & -1 & 1 \\ 0 & 1 & -3/2 & 1 \\ 0 & 0 & 0 & 0 \end{bmatrix}$

The new eq'ns are $x+y-z+w=0$.

$$y-\frac{3z}{2}+w=0. \quad \text{— 4M}$$

Put $z=k_1$, $w=k_2$.

Then $y=\frac{3z}{2}-w=\frac{3}{2}k_1-k_2$

$x=-y+z-w = -\frac{3k_1}{2}+k_2+k_1-k_2$

$=-\frac{k_1}{2}$

∴ The infinite number of non-trivial solutions are given by

$x=-\frac{1}{2}k_1$, $y=\frac{3}{2}k_1-k_2$, $z=k_1$ & $w=k_2$

where k_1 & k_2 are arbitrary. — 2M

7.(b). $A=\begin{bmatrix}1&2&3&-1\\2&1&3&1\\1&0&1&1\\0&1&1&1\end{bmatrix}$

$R_2\to R_2-2R_1$, $R_3\to R_3-R_1$

$$\sim\begin{bmatrix}1&2&3&-1\\0&-3&-3&3\\0&-2&-2&2\\0&1&1&1\end{bmatrix}$$

$C_2\to C_2-2C_1$, $C_3\to C_3-3C_1$, $C_4\to C_4+C_1$

$$\sim\begin{bmatrix}1&0&0&0\\0&-3&-3&3\\0&-2&-2&2\\0&1&1&1\end{bmatrix} \quad \text{— 2M}$$

$R_2\to \frac{R_2}{-3}$ $\sim\begin{bmatrix}1&0&0&0\\0&1&1&-1\\0&-2&-2&2\\0&1&1&1\end{bmatrix}$

$\left.\begin{matrix}R_3\to R_3+2R_2\\R_4\to R_4-R_2\end{matrix}\right\}\sim\begin{bmatrix}1&0&0&0\\0&1&1&-1\\0&0&0&0\\0&0&0&2\end{bmatrix}$

$C_3\to C_3-C_2$, $C_4\to C_4+C_2$

$$\sim\begin{bmatrix}1&0&0&0\\0&1&0&0\\0&0&0&0\\0&0&0&2\end{bmatrix} \quad \text{— 2M}$$

$R_3\leftrightarrow R_4$

$$\sim\begin{bmatrix}1&0&0&0\\0&1&0&0\\0&0&0&2\\0&0&0&0\end{bmatrix}$$

$C_3\leftrightarrow C_4$

$$\sim\begin{bmatrix}1&0&0&0\\0&1&0&0\\0&0&2&0\\0&0&0&0\end{bmatrix}$$

$R_3\to \frac{R_3}{2}$

$$\sim\begin{bmatrix}1&0&0&0\\0&1&0&0\\0&0&1&0\\0&0&0&0\end{bmatrix}$$

$$=\begin{bmatrix}I_3&0\\0&0\end{bmatrix}$$

∴ $\rho(A)=3$

— 2M

8. $A = \begin{bmatrix} 2 & -1 & 1 \\ -1 & 2 & -1 \\ 1 & -1 & 2 \end{bmatrix}$

The ch. eq. of A is $|A - \lambda I| = 0 \Rightarrow \begin{vmatrix} 2-\lambda & -1 & 1 \\ -1 & 2-\lambda & -1 \\ 1 & -1 & 2-\lambda \end{vmatrix} = 0$

$\Rightarrow (2-\lambda)\left[\lambda^2 - 4\lambda + 4 - 1\right] + (1)\left[-2+\lambda+1\right] + (1)\left[1-2+\lambda\right] = 0$

$\Rightarrow (-\lambda+2)(\lambda^2 - 4\lambda + 3) + \lambda - 1 + \lambda - 1 = 0$

$\Rightarrow -\lambda^3 + 4\lambda^2 - 3\lambda + 2\lambda^2 - 8\lambda + 6 + 2\lambda - 2 = 0$

$\Rightarrow -\lambda^3 + 6\lambda^2 - 9\lambda + 4 = 0$

Claim: C–H states that $A^3 - 6A^2 + 9A - 4I = 0$ — 4M

Verif: $A^2 = A \cdot A = \begin{bmatrix} 2 & -1 & 1 \\ -1 & 2 & -1 \\ 1 & -1 & 2 \end{bmatrix} \begin{bmatrix} 2 & -1 & 1 \\ -1 & 2 & -1 \\ 1 & -1 & 2 \end{bmatrix}$

$= \begin{bmatrix} 4+1+1 & -2-2-1 & 2+1+2 \\ -2-2-1 & 1+4+1 & -1-2-2 \\ 2+1+2 & -1-2-2 & 1+1+4 \end{bmatrix} = \begin{bmatrix} 6 & -5 & 5 \\ -5 & 6 & -5 \\ 5 & -5 & 6 \end{bmatrix}$ — 2M

$A^3 = A^2 \cdot A = \begin{bmatrix} 6 & -5 & 5 \\ -5 & 6 & -5 \\ 5 & -5 & 6 \end{bmatrix} \begin{bmatrix} 2 & -1 & 1 \\ -1 & 2 & -1 \\ 1 & -1 & 2 \end{bmatrix}$

$= \begin{bmatrix} 12+5+5 & -6-10-5 & 6+5+10 \\ -10-6-5 & 5+12+5 & -5-6-10 \\ 10+5+6 & -5-10-6 & 5+5+12 \end{bmatrix}$

$= \begin{bmatrix} 22 & -21 & 21 \\ -21 & 22 & -21 \\ 21 & -21 & 22 \end{bmatrix}$ — 2M

Now $A^3 + 2A^2 - 7A - 4I = \begin{bmatrix} 22 & -21 & 21 \\ -21 & 22 & -21 \\ 21 & -21 & 22 \end{bmatrix} - 6 \begin{bmatrix} 6 & -5 & 5 \\ -5 & 6 & -5 \\ 5 & -5 & 6 \end{bmatrix}$

$$+9\begin{bmatrix}2 & -1 & 1\\ -1 & 2 & -1\\ 1 & -1 & 2\end{bmatrix} - 4\begin{bmatrix}1 & 0 & 0\\ 0 & 1 & 0\\ 0 & 0 & 1\end{bmatrix}$$

$$= \begin{bmatrix}22 & -21 & 21\\ -21 & 22 & -21\\ 21 & -21 & 22\end{bmatrix} + \begin{bmatrix}12 & -10 & 10\\ -10 & 12 & -10\\ 10 & -10 & 12\end{bmatrix} + \begin{bmatrix}-14 & 7 & -7\\ 7 & -14 & 7\\ -7 & 7 & -14\end{bmatrix} + \begin{bmatrix}-4 & 0 & 0\\ 0 & -4 & 0\\ 0 & 0 & -4\end{bmatrix}$$

$$= \begin{bmatrix}22 & -21 & 21\\ -21 & 22 & -21\\ 21 & -21 & 22\end{bmatrix} + \begin{bmatrix}-36 & 30 & -30\\ 30 & -36 & 30\\ -30 & 30 & -36\end{bmatrix} + \begin{bmatrix}18 & -9 & 9\\ -9 & 18 & -9\\ 9 & -9 & 18\end{bmatrix} + \begin{bmatrix}-4 & 0 & 0\\ 0 & -4 & 0\\ 0 & 0 & -4\end{bmatrix}$$

$$= \begin{bmatrix}0 & 0 & 0\\ 0 & 0 & 0\\ 0 & 0 & 0\end{bmatrix} = O. \qquad \text{— 2M}$$

Hence C-H Theorem is verified.

To find A^{-1}:- $A^3 - 6A^2 + 9A - 4I = 0$

$A^{-1}(A^3 - 6A^2 + 9A - 4I) = 0 \Rightarrow A^2 - 6A + 9I = 4A^{-1}$

$$\therefore A^{-1} = \frac{1}{4}\left(A^2 - 6A + 9I\right)$$

$$= \frac{1}{4}\left\{\begin{pmatrix}6 & -5 & 5\\ -5 & 6 & -5\\ 5 & -5 & 6\end{pmatrix} + \begin{pmatrix}-12 & 10 & -10\\ 10 & -12 & 10\\ -10 & 10 & -12\end{pmatrix} + \begin{pmatrix}9 & 0 & 0\\ 0 & 9 & 0\\ 0 & 0 & 9\end{pmatrix}\right\}$$

$$= \frac{1}{4}\begin{bmatrix}3 & 5 & -5\\ 5 & -6 & 5\\ -5 & 5 & -6\end{bmatrix} \qquad \text{— 2M}$$

9. ch. eq. of A is $\begin{vmatrix} 3-\lambda & -1 & 1 \\ -1 & 5-\lambda & -1 \\ 1 & -1 & 3-\lambda \end{vmatrix} = 0.$

$\Rightarrow (3-\lambda)\{\lambda^2-8\lambda+15-1\}+(1)\{\lambda-3+1\}+(1)\{1-5+\lambda\}=0$

$\Rightarrow (-\lambda+3)(\lambda^2-8\lambda+14)+\lambda-2+\lambda-4=0$

$\Rightarrow -\lambda^3+8\lambda^2-14\lambda+3\lambda^2-24\lambda+42+2\lambda-6=0$

$\Rightarrow -\lambda^3+11\lambda^2-36\lambda+36=0$

$\Rightarrow -(\lambda-2)(\lambda^2-9\lambda+18)=0 \quad \Rightarrow -(\lambda-2)(\lambda-3)(\lambda-6)=0$

$\therefore$ Eigen values are 2, 3, 6. ——— 2 M

Eigen vector corresponding to $\lambda=2$ is $\begin{pmatrix} 1 \\ 0 \\ -1 \end{pmatrix}$

" " $\lambda=3$ is $\begin{pmatrix} 1 \\ 1 \\ 1 \end{pmatrix}$

" " $\lambda=6$ is $\begin{pmatrix} 1 \\ -2 \\ 1 \end{pmatrix}$ ——— 2M

$\therefore \hat{P} = \begin{pmatrix} 1/\sqrt{2} & 1/\sqrt{3} & 1/\sqrt{6} \\ 0 & 1/\sqrt{3} & -2/\sqrt{6} \\ -1/\sqrt{2} & 1/\sqrt{3} & 1/\sqrt{6} \end{pmatrix}$ ——— 2M

Take $X=\hat{P}Y$, $Y=\begin{pmatrix} y_1 \\ y_2 \\ y_3 \end{pmatrix}$

Then $X^T A X = Y^T D Y = 2y_1^2+3y_2^2+6y_3^2$ ——— 2M

is the canonical form.

Rank, $r=3$, ——— 1 M

index, $s=3$ ——— 1 M

signature $=0$. ——— 1 M

Nature is positive definite. ——— 1 M.

10. (a) $(1+y^2)\,dx + (x - \tan^{-1}y)\,dy = 0.$

$\Rightarrow \frac{dx}{dy} + \left(\frac{1}{1+y^2}\right)x = \frac{\tan^{-1}y}{1+y^2}$ ie, $\frac{dx}{dy} + Px = Q$ — 1 M

$IF = e^{\int P\,dy} = e^{\int \frac{1}{1+y^2}dy} = e^{\tan^{-1}y}$ — 2 M

C.S is $x\,(IF) = \int Q(IF)\,dy + c$ — 1M

$\Rightarrow x \cdot e^{\tan^{-1}y} = \int e^{\tan^{-1}y} \cdot \frac{\tan^{-1}y}{1+y^2}\,dy + c$

~~Put $e^{\tan^{-1}y} = t \Rightarrow \frac{e^{\tan^{-1}y}}{1+y^2}\,dy = dt$~~

Put $\tan^{-1}y = t \Rightarrow \frac{1}{1+y^2}dy = dt$

$= \int e^t \cdot t\,dt + c$

$= (t-1)e^t + c$

$\therefore x \cdot e^{\tan^{-1}y} = (\tan^{-1}y - 1)\,e^{\tan^{-1}y} + c.$ — 2M

10. (b) $(y^2 - xy)\,dx + x^2\,dy = 0.$ — (1)

$M = y^2 - xy, \quad N = x^2$

$\frac{\partial M}{\partial y} = 2y, \quad \frac{\partial N}{\partial x} = 2x \Rightarrow$ Not exact.

M & N are homogeneous function of same degree

$\therefore I.F = \frac{1}{Mx + Ny} = \frac{1}{(y^2 - xy)x + (x^2)y} = \frac{1}{xy^2}$ — 2

(1) $\times \frac{1}{xy^2} \Rightarrow \left(\frac{y^2 - xy}{xy^2}\right)dx + \left(\frac{x^2}{xy^2}\right)dy = 0$

$\Rightarrow \left(\frac{1}{x} - \frac{1}{y}\right)dx + \left(\frac{x}{y^2}\right)dy = 0$ — 1M

$\Rightarrow M_1\,dx + N_1\,dy = 0.$ — (2)

$\therefore \frac{\partial M}{\partial y} = \frac{1}{y^2}, \ \frac{\partial N}{\partial x} = \frac{1}{y^2} \Rightarrow$ ② is exact

The G.S. of ② is $\int\limits_{y} M_x\, dx + \int \left(\begin{matrix}\text{terms of } N, \text{ not}\\ \text{containing } x\end{matrix}\right) dy = c$ — 1M

$\Rightarrow \int\limits_{y} \left(\frac{1}{x} - \frac{1}{y}\right) dx + \int (0)\, dy = c \Rightarrow \log x - \frac{x}{y} = c$ — 2M

11.(a) Given cartesian family is $ax + y^2 = x^2$. — ①

By diff w.r.to 'x', we get

$a + 2y\frac{dy}{dx} = 2x$ — ②

By sub a value in ①, we get

$(2x - 2yy_1)x + y^2 = x^2 \Rightarrow x^2 + y^2 - 2xyy_1 = 0$

This is the DE of family ①. — 2M

By replacing y_1 with $-\frac{1}{y_1}$, we get

$x^2 + y^2 - 2xy\left(-\frac{1}{y_1}\right) = 0 \Rightarrow (x^2+y^2)y_1 + 2xy = 0$

This is the DE of the orthogonal family. — 1M

$\Rightarrow (x^2+y^2)\frac{dy}{dx} + 2xy = 0$ $\cancel{\Rightarrow (x^2+y^2)dy + 2xy\, dx = 0}$

$\cancel{M = 2xy,\ N = x^2+y^2}$

~~M & N are homog. of same degree~~

$\cancel{\therefore I.F = \frac{1}{Mx+Ny} = \frac{1}{2x^2y + x^2y + y^3}}$

$\Rightarrow \frac{dx}{dy} = -\left(\frac{x^2+y^2}{2xy}\right) = -\frac{x}{2y} - \frac{y}{2x}$

$\Rightarrow \frac{dx}{dy} + \left(\frac{1}{2y}\right)x = \left(-\frac{y}{2}\right)x^{-1}$ — ③

ie, $\frac{dx}{dy} + Px = Qx^n$, a Bernoulli's DE

for $P = \frac{1}{2y}$, $Q = -\frac{y}{2}$, $n = -1$

By multiplying (3) with x, we get

$$x \frac{dx}{dy} + \frac{1}{2y}(x^2) = -\frac{y}{2}$$

Put $x^2 = z \Rightarrow 2x \frac{dx}{dy} = \frac{dz}{dy}$

$$\Rightarrow \frac{1}{2}\frac{dz}{dy} + \frac{1}{2y} z = -\frac{y}{2}$$

$$\Rightarrow \frac{dz}{dy} + \left(\frac{1}{y}\right) z = -y$$, a Leibnitz's linear DE in y

$$I.F = e^{\int \frac{1}{y} dy} = e^{\log_e y} = y$$

sol. is $z(IF) = \int (-y)(I.F)\,dy + c$

$$\Rightarrow zy = \int (-y)(y)\,dy + c = \int -y^2\,dy + c$$

$$\Rightarrow yz = -\frac{y^3}{3} + c$$

$$\Rightarrow x^2 y = -\frac{y^3}{3} + c \quad (\because z = x^2)$$ —— 3M

11. (b) $\frac{dN}{dt} \propto N \Rightarrow \frac{dN}{dt} = kN \Rightarrow \frac{dN}{N} = k\,dt$

$\Rightarrow \log_e N = kt + \log_e c \Rightarrow N = c\cdot e^{kt}$ —— (1) —— 2M

$N = 100$ at $t = 0 \Rightarrow 100 = ce^0 \Rightarrow c = 100$

$\therefore N = 100 \cdot e^{kt}$ —— (2) —— 1M

$N = 332$ at $t = 1 \Rightarrow 332 = 100 \cdot e^k$

$\Rightarrow e^k = 3.32$ —— 1M

$\therefore N = 100\,(3.32)^t$ —— (3) —— 1M

when $t = 3/2$, $N = 100\,(3.32)^{3/2} = 604.5 \cong 605$.

1M

12) (a) $(D^4+2D^3-3D^2)y=0$.

A.E is $m^4+2m^3-3m^2=0 \Rightarrow m^2(m^2+2m-3)=0$

$\Rightarrow m^2(m+3)(m-1)=0 \Rightarrow m=0,0,-3,1$ — 3M

$\therefore$ C.F $= c_1+c_2x+c_3e^{-3x}+c_4e^{x}$ — 2M

The C.S is $y=$ C.F. $\Rightarrow y=c_1+c_2x+c_3e^{-3x}+c_4e^{x}$. — 1M

12. (b) $(D^2+6D+9)y=5^x$

A.E is $m^2+6m+9=0 \Rightarrow (m+3)^2=0 \Rightarrow m=-3,-3$

$\therefore$ C.F $=(c_1+c_2x)e^{-3x}$ — 2M

$$P.I = \frac{1}{D^2+6D+9}5^x = \frac{1}{D^2+6D+9}e^{x\log 5}$$

(Put $D=\log 5$) $= \dfrac{1}{(\log 5)^2+6(\log 5)+9}\cdot e^{x\log 5}$ — 3M

The C.S. is $y=$ C.F $+$ P.I.

$$\Rightarrow y=(c_1+c_2x)e^{-3x}+\frac{5^x}{(\log 5)^2+6\log 5+9}$$ — 1M

13. (a) $(D^2+3D+2)y=\sin 2x$.

A.E is $m^2+3m+2=0 \Rightarrow (m+1)(m+2)=0 \Rightarrow m=-1,-2$

$\therefore$ C.F. $=c_1e^{-x}+c_2e^{-2x}$ — 2M

$$P.I = \frac{1}{D^2+3D+2}\sin 2x = \frac{1}{-4+3D+2}\sin 2x \quad (D^2=-2^2=-4)$$

$$=\frac{1}{3D-2}\sin 2x = \frac{3D+2}{9D^2-4}\sin 2x = \frac{3D+2}{-36-4}\sin 2x$$

$$=\frac{-1}{40}\{3(2\cos 2x)+2\sin 2x\} = \frac{-1}{20}\{3\cos 2x+\sin 2x\}$$ — 3M

The C.S is $y=$ C.F. $+$ P.I.

$$\Rightarrow y=c_1e^{-x}+c_2e^{-2x}-\frac{1}{20}\{3\cos 2x+\sin 2x\}$$ — 1M

13.(b) $\frac{d^2y}{dx^2} + y = \sec x.$

$\Rightarrow (D^2+1)y = \sec x$

AE is $m^2+1=0 \Rightarrow m = \pm i$

$\therefore$ C.F $= c_1 \cos x + c_2 \sin x.$ — 2M

Let $u = \cos x$, $v = \sin x$, $W(u,v) = \begin{vmatrix} \cos x & \sin x \\ -\sin x & \cos x \end{vmatrix} = 1 \neq 0.$

Let us assume that P.I of P.I $= Au + Bv$

Then $A = -u \int \frac{v \sec x}{W} dx$

$= -\cos x \int \frac{\sin x . \sec x}{1} dx = -\cos x \int \tan x \, dx$

$= -\cos x \log(\sec x)$

$B = v \int \frac{u \sec x}{1} dx = \sin x \int \frac{\cos x . \sec x}{1} dx$

$= \sin x \int 1 \, dx = x \sin x$

$\therefore$ P.I $= -\cos x \log(\sec x) + x \sin x$ — 3M

The C.S is $y = $ C.F + P.I.

$= c_1 \cos x + c_2 \sin x - \cos x \log(\sec x) + x \sin x.$

— 1M

14. (a) $f(x) = \log_e x$

$f'(x) = \frac{1}{x}$ exists $\forall x$ through the interval $[1,e]$

So, $f(x)$ is continuous in $[1,e]$ — 1M

& derivable in $(1,e)$ — 1M

$\Rightarrow$ LMVT is applicable & $\exists$ at least one number

$c \in (1,e)$ $\ni$ $f'(c) = \frac{f(b)-f(a)}{b-a} = \frac{f(e)-f(1)}{e-1}$

— 1M

Verification:- $f'(c) = \dfrac{f(b)-f(a)}{b-a}$

$\Rightarrow \dfrac{1}{c} = \dfrac{\log_e e - \log_e 1}{e-1}$ $\quad (\because a=1, b=e)$

$= \dfrac{1}{e-1} \Rightarrow c = e-1$ — 2M

$e = 2.718182 \Rightarrow e-1 = 1.718182 \in (1, 2.718182)$

$\Rightarrow c \in (1, e)$ — 1M

Hence, LMVT is verified.

14.1b) $u = \dfrac{x-y}{x+y}, V = \dfrac{xy}{(x+y)^2}$

$\dfrac{\partial u}{\partial x} = \dfrac{\partial}{\partial x}\left(\dfrac{x-y}{x+y}\right) = \dfrac{(x+y)(1)-(x-y)(1)}{(x+y)^2} = \dfrac{2y}{(x+y)^2}$

$\dfrac{\partial u}{\partial y} = \dfrac{\partial}{\partial y}\left(\dfrac{x-y}{x+y}\right) = \dfrac{(x+y)(-1)-(x-y)(1)}{(x+y)^2} = \dfrac{-2x}{(x+y)^2}$

$\dfrac{\partial v}{\partial x} = \dfrac{\partial}{\partial x}\left(\dfrac{xy}{(x+y)^2}\right) = \dfrac{(x+y)^2(y) - xy \cdot 2(x+y)(1)}{(x+y)^4}$

$= \dfrac{(x+y)y - 2xy}{(x+y)^3} = \dfrac{y^2-xy}{(x+y)^3}$

$\dfrac{\partial v}{\partial y} = \dfrac{\partial}{\partial y}\left[\dfrac{xy}{(x+y)^2}\right] = \dfrac{(x+y)^2 \cdot (x) - xy \cdot 2(x+y) 1}{(x+y)^4}$

$= \dfrac{(x+y)x - 2xy}{(x+y)^3} = \dfrac{x^2-xy}{(x+y)^3}$

The Jacobian of u & v $= \begin{vmatrix} \dfrac{\partial u}{\partial x} & \dfrac{\partial u}{\partial y} \\ \dfrac{\partial v}{\partial x} & \dfrac{\partial v}{\partial y} \end{vmatrix}$ — 1M

$= \begin{vmatrix} \dfrac{2y}{(x+y)^2} & \dfrac{-2x}{(x+y)^2} \\ \dfrac{y^2-xy}{(x+y)^3} & \dfrac{x^2-xy}{(x+y)^3} \end{vmatrix}$ — 2M

$$= \frac{2y(x^2-xy)}{(x+y)^5} + \frac{2x(y^2-xy)}{(x+y)^5} = \frac{2x^2y-2xy^2+2xy^2-2x^2y}{(x+y)^5}$$

$$= 0$$

$\Rightarrow$ u & v are functionally dependent — 2M

Now $u^2 = \left(\frac{x-y}{x+y}\right)^2 = \frac{x^2+y^2-2xy}{(x+y)^2} = \frac{(x+y)^2-4xy}{(x+y)^2}$

$$= 1 - 4\frac{xy}{(x+y)^2} = 1-4v$$

$\therefore \boxed{u^2 = 1-4v.}$ — 1 M

15. $f(x,y) = e^x \log(1+y)$

$f_x = e^x \log(1+y)$, $f_{xx} = e^x \log(1+y)$, $f_{xxx} = e^x \log(1+y)$

$f_y = \frac{e^x}{1+y}$, $f_{yy} = \frac{-e^x}{(1+y)^2}$, $f_{yyy} = \frac{+2e^x}{(1+y)^3}$

$f_{xy} = \frac{e^x}{1+y}$, $f_{xy^2} = \frac{-e^x}{(1+y)^2}$, $f_{xy^3} = \frac{2e^x}{(1+y)^3}$ — 2M

$f_{x^2y} = \frac{e^x}{1+y}$

At $(x,y) = (0,0)$, $f_x = 0$, $f_y = \frac{1}{1+0} = 1$

$f_{xx} = 0$, $f_{xy} = \frac{1}{1+0} = 1$, $f_{yy} = -\frac{1}{(1+0)^2} = -1$ — 2M

$f_{xxx} = 0$, $f_{x^2y} = \frac{1}{1+0} = 1$, $f_{xy^2} = -\frac{1}{(1+0)^2} = -1$, $f_{yyy} = 2$

$\therefore$ By Maclaurin's series expansion,

$$f(x,y) = f(0,0) + \frac{1}{1!}\{x f_x(0,0) + y f_y(0,0)\} + \frac{1}{2!}\{x^2 f_{xx}(0,0) + 2xy f_{xy}(0,0) + y^2 f_{yy}(0,0)\}$$

$$+ \frac{1}{3!}\{x^3 f_{xxx}(0,0) + 3x^2y f_{x^2y}(0,0) + 3xy^2 f_{xy^2}(0,0) + y^3 f_{yyy}(0,0)\}$$ — 1M

$+ \cdots$

$\Rightarrow e^x \log(1+y) = 0 + \dfrac{x(0)+y(1)}{1!} + \dfrac{x^2(0)+2xy(1)+y^2(-1)}{2!}$

$+ \dfrac{1}{3!}\left\{x^3(0) + 3x^2y(1) + 3xy^2(-1) + y^3(2)\right\}$

$+ \cdots$

$= 0 + \dfrac{0+y}{1!} + \dfrac{0+2xy-y^2}{2!} + \dfrac{0+3x^2y-3xy^2+2y^3}{3!} + \cdots$

$\Rightarrow e^x\log(1+y) = y + xy - \dfrac{y^2}{2} + \dfrac{x^2y}{2} - \dfrac{xy^2}{2} + \dfrac{y^3}{3} + \cdots$ — 1M

15.(b) Let $2x, 2y, 2z$ be the lenght, breadth, height of the rectangular solid so that its volume is $V = 8xyz$. — 1M

Let R be the radius of the sphere so that $x^2+y^2+z^2=R^2$ — 1M

Then $F(x,y,z) = 8xyz + \lambda(x^2+y^2+z^2-R^2)$ — 1M

& $\dfrac{\partial F}{\partial x} = 0$, $\dfrac{\partial F}{\partial y} = 0$ and $\dfrac{\partial F}{\partial z} = 0$ give

$8yz + 2x\lambda = 0$, $8zx + 2y\lambda = 0$, $8xy + 2z\lambda = 0$ — 1M

or $2x^2\lambda = 2y^2\lambda = 2z^2\lambda = -8xyz$

Thus for a maximum volume, $x = y = z$. — 1M

ie, the rectangular solid is a cube. — 1M

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NSRIT

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	Common to All	Academic Year	2021 - 2022
Course Code	20ESX02	Test Duration	3 Hrs. Max. Marks 70	Semester	I
Course	PROGRAMMING FOR PROBLEM SOLVING USING 'C'				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	List any four data types in C.	20ESX02.1	L1
2	Differentiate branching and looping statements.	20ESX02.2	L2
3	List any four storage classes.	20ESX02.3	L1
4	What is call-by-value?	20ESX02.4	L1
5	List any four preprocessor directives in C.	20ESX02.5	L1

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

No.	Questions (6 through 10)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain C tokens in detail	6M	20ESX02.1	L2
6 (b)	Explain the rules to create a variable.	6M	20ESX02.1	L2
	OR			
7	Write an algorithm to check whether a given number is palindrome or not and draw the flow chart for the same.	12M	20ESX02.1	L2
8 (a)	Discuss the syntax of while loop.	4M	20ESX02.2	L2
8 (b)	Write a program to display all prime numbers between a given range using while loop.	8M	20ESX02.2	L3
	OR			
9	Explain if, if-else, if-else-if ladder and nested if with suitable examples.	12M	20ESX02.2	L2
10	Write a C program to multiply two m x n and p x q size matrices.	12M	20ESX02.3	L3
	OR			
11	Explain categories of functions with examples.	12M	20ESX02.3	L2
12	Explain dynamic memory allocation functions in detail.	12M	20ESX02.4	L2
	OR			
13 (a)	Explain array of structures.	6M	20ESX02.4	L2
13 (b)	Discuss nested structures.	6M	20ESX02.4	L2
14	Explain various file handling functions with examples.	12M	20ESX02.5	L2
	OR			
15	Explain various functions to access files randomly.	12M	20ESX02.5	L2

NSRIT

N S RAJU INSTITUTE OF TECHNOLOGY

(AUTONOMOUS)

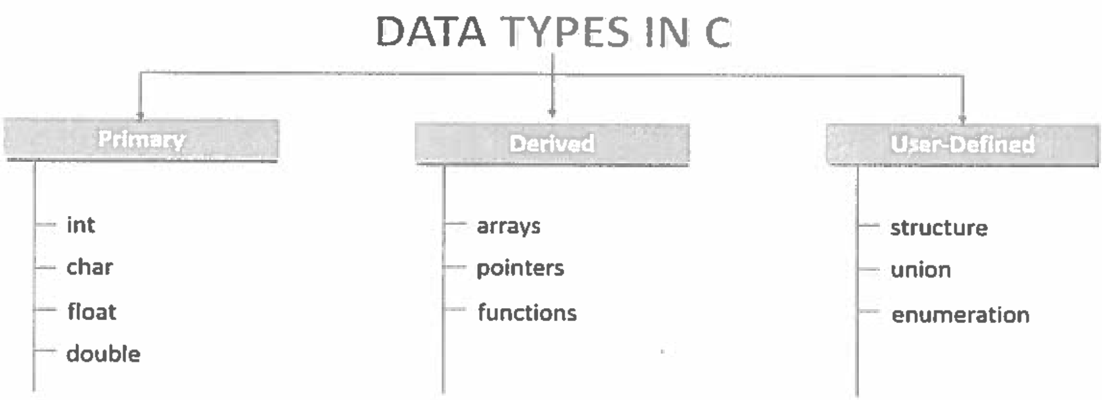
SONTYAM , ANANDAPURAM, VISAKHAPATNAM – 531 173

ANSWER KEY AND SCHEME OF EVALUATION

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

1)List any four data types in C. **(2m)**

Ans: int ,char ,float ,double



2) Differentiate branching and looping statements. 2m

Ans: Branching is deciding what actions to take and looping is deciding how many times to take a certain action.

3) List any four storage classes.

Ans : auto ,register, static ,extern **2m**

4) What is call-by-value?

Ans: The call by value method of passing arguments to a function copies the actual value of an argument into the formal parameter of the function .

5)**List any four preprocessor directives in c?**

Ans: #define ,#if,#include,#error,#else

Part B(Long Answer Questions 5x12=60 Marks)

6.a Explain C tokens in detail ?(6m)

Ans: The compiler breaks a program into the smallest possible units and proceeds to the various stages of the compilation, which is called token.

C Supports Six Types of Tokens:

1. Identifiers
2. Keywords
3. Constants
4. Strings
5. Operators
6. Special Symbols

1. IDENTIFIERS:

Identifiers are names given to different entities such as constants, variables, structures, functions, etc.

Example:

```
int amount;

double totalbalance:
```

2. **KEYWORS:**

The C Keywords must be in your information because you can not use them as a variable name.
Example:

```
#include<stdio.h>

int main()
{
  float a, b;
  printf("Showing how keywords are used.");
  return 0;
}
```

3.CONSTANTS:

Constants are like a variable, except that their value never changes during execution once defined. C Constants is the most fundamental and essential part of the C programming language.

- Constants are also called literals.
- Constants can be any of the data types.
- It is considered best practice to define constants using only upper-case names

Example:

```
#include<stdio.h>
void main()
{
```

```
const int SIDE = 10;
int area;
area = SIDE*SIDE;
printf("The area of the square with side: %d is: %d sq. units", SIDE, area);
}
```

4.OPERATORS:

C operators are symbols that are used to perform mathematical or logical manipulations. The C programming language is rich with built-in operators. Operators take part in a program for manipulating data and variables and form a part of the mathematical or logical expressions.

C programming language offers various types of operators having different functioning capabilities.

1. Arithmetic Operators
2. Relational Operators
3. Logical Operators
4. Assignment Operators
5. Increment and Decrement Operators
6. Conditional Operator
7. Bitwise Operators
8. Special Operators

5 . Special operators:

1. The Comma Operator
2. Type cast Operator
3. Reference operator or Address Operater ("&")
4. Dereference operator ("*") or Pointer Operater
5. Double Pointer operator ("**")
6. sizeof operator

6)Strings: string is a collection of characters. String syntax char name_of_the_string [size]

String is denoted with " " example:"welcome"

6.b Explain the rules to create a variable? (6m)

Variable

- It is the name for memory location that may be used to store a data value.
- A variable may take different values at different times during execution.
- A variable name may be chosen by the programmer in a meaningful way, so as to reflect its function (or) nature in the program.

For example, sum, avg, total etc.

Rules for naming variable

The rules for naming a variable are explained below −

- They must begin with a letter.
- Maximum length of variable is 31 characters in ANSI standard. But, first eight characters are significant by many compilers.
- Upper and lowercase characters are different. For example: total, TOTAL, Total are 3 different variables.
- The variable is not to be a keyword.
- White space is not allowed.

Variable declaration

The syntax and example with regards to variable declaration are explained below −

Syntax

Given below is the syntax for variable declaration −

```
Datatype v1,v2,… vn;
```

Where, v1, v2,...vn are names of variables.

For example,

```
int sum;
float a,b;
```

Variable can be declared in two ways −

- **Local declaration** − 'Local declaration' is declaring a variable inside the main block and its value is available within that block.
- **Global declaration** − 'Global declaration' is declaring a variable outside the main block and its value is available throughout the program.

Example

Following is the C program for local and global declaration of variables in C language −

```
int a, b; /* global declaration*/
main ( ){
  int c; /* local declaration*/
  - - -
```

}

Example

Given below is a C program to find the selling price (SP) and cost price (CP) of an article −

Live Demo

```
#include<stdio.h>
int main(){
  float CostPrice, SellingPrice, Amount; //variable declaration
  //costprice & sellingprice are variables and
  //float is a datatype
  printf("\n product cost price: ");
  scanf("%f", &CostPrice);
  printf("\n product selling price : ");
  scanf("%f", &SellingPrice);
  if (SellingPrice > CostPrice){
    Amount = SellingPrice - CostPrice;
    printf("\n Profit Amount = %.4f", Amount);
  }
  else if(CostPrice > SellingPrice){
    Amount = CostPrice - SellingPrice;
    printf("\n Loss Amount = %.4f", Amount);
  }
  else
    printf("\n No Profit No Loss!");
  return 0;
}
```

Output

The output is as follows −

```
product cost price : 240
product selling price : 280
Profit Amount = 40.0000
```

OR

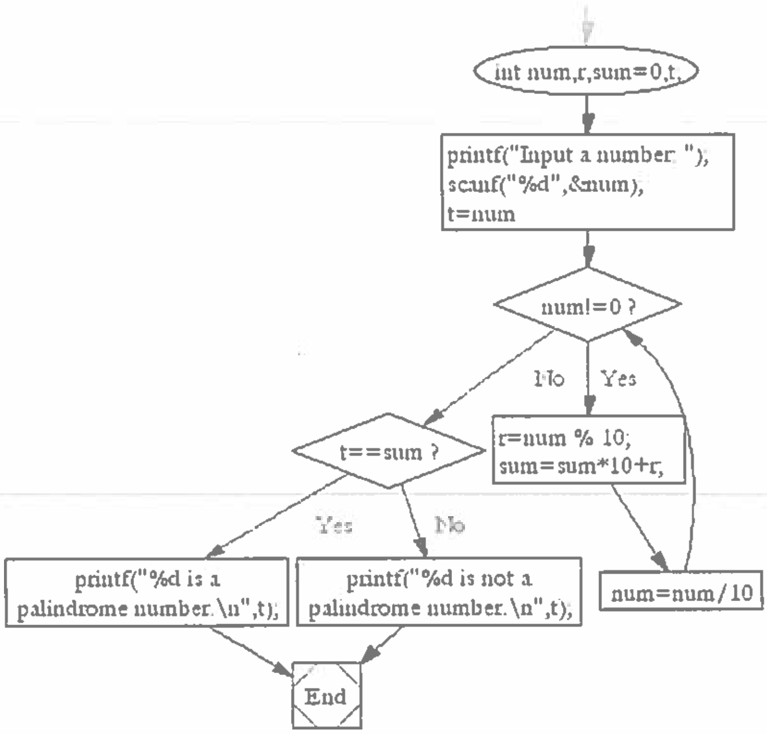
7. Write an algorithm to check whether a given number is palindrome or not and draw the flow chart for the same. (12m)

ANS: Algorithm to check whether a given number is palindrome

Palindrome number in c: A **palindrome number** is *a number that is same after reverse*. For example 121, 34543, 343, 131, 48984 are the palindrome numbers.

- STEP1: Get the number from user
- STEP2:Hold the number in temporary variable
- STEP3:Reverse the number
- STEP4:Compare the temporary number with reversed number
- STEP5:If both numbers are same, print palindrome number
- STEP6:Else print not palindrome number

Flowchart for Number Palindrome:



8.a Discuss the syntax of while loop.? (4m)

Ans: while loop executes at least once i.e. the first iteration runs without checking the condition. The condition is checked only after the first iteration has been executed

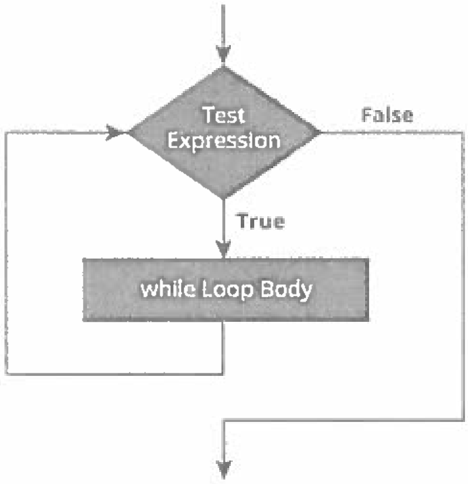
Syntax:

```
do
{
   statements1;
   statements2;
  statements3;
}while(condition);
```

Example:

```
do {
printf("Enter a number: ");
scanf("%lf", &number);
sum += number;
}
while(number != 0.0);
```

FLOWCHART:



8.b. Write a program to display all prime numbers between a given range using while loop.:?(8m)

Ans: # print all prime numbers b/w a given range using while loop:

```
#include <stdio.h>

int main()

{

 int i, a = 1, count;

 while(a <= 100)

 {

  count = 0;

  i = 2;

  while(i <= a/2)

  {

   if(a%i == 0)

   {

    count++;
```

```
        break;
    }
    i++;
  }
  if(count == 0 && a != 1 )
  {
        printf(" %d ", a);
  }
  a++;
 }
 return 0;
}
```

OR

9. Explain if, if-else, if-else-if ladder and nested if with suitable examples. (12M)

Ans:

*****IF Statement**

The if statement checks the given condition. If the condition evaluates to be true then the block of code/statements will execute otherwise not.

Syntax:

```
if(condition)
    {
        //code to be executed
    }
```

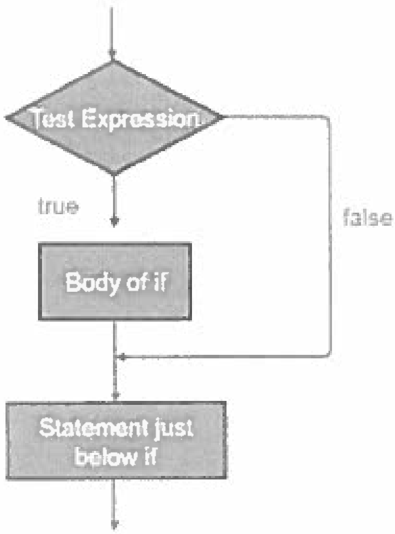
Note: If the curly brackets { } are not used with if statements than the statement just next to it is only considered associated with the if statement.

Example:

```
if (condition)
   statement 1;
statement 2;
```

In this example, only statement 1 is considered to be associated with the if statement.

Flowchart:



Example:

```
#include <stdio.h>
int main()
{
   int x = 20;
   int y = 22;
   if (x<y)
   {
      printf("Variable x is less than y");
   }
   return 0;
}
```

Output:

```
Variable x is less than y
```

IF – else Statement

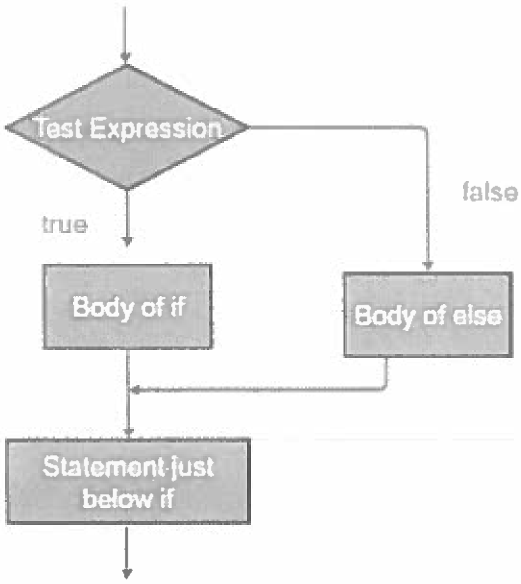
The if statement evaluates the code if the condition is true but what if the condition is not true, here comes the else statement. It tells the code what to do when the if condition is false.

Syntax:

```
if(condition)
{
  // code if condition is true
}
else
{
```

```
    // code if condition is false
  }
```

Flowchart:



Example:

```
#include <stdio.h>
int main()
{
  int age;
  printf("Enter your age:");
  scanf("%d",&age);
  if(age >=18)
  {
        /* This statement will only execute if the
         * above condition (age>=18) returns true
         */
        printf("You are eligible for voting");
  }
  else
  {
        /* This statement will only execute if the
         * condition specified in the "if" returns false.
         */
        printf("You are not eligible for voting");
  }
  return 0;
}
```

Output:

Enter your age:14

You are not eligible for voting

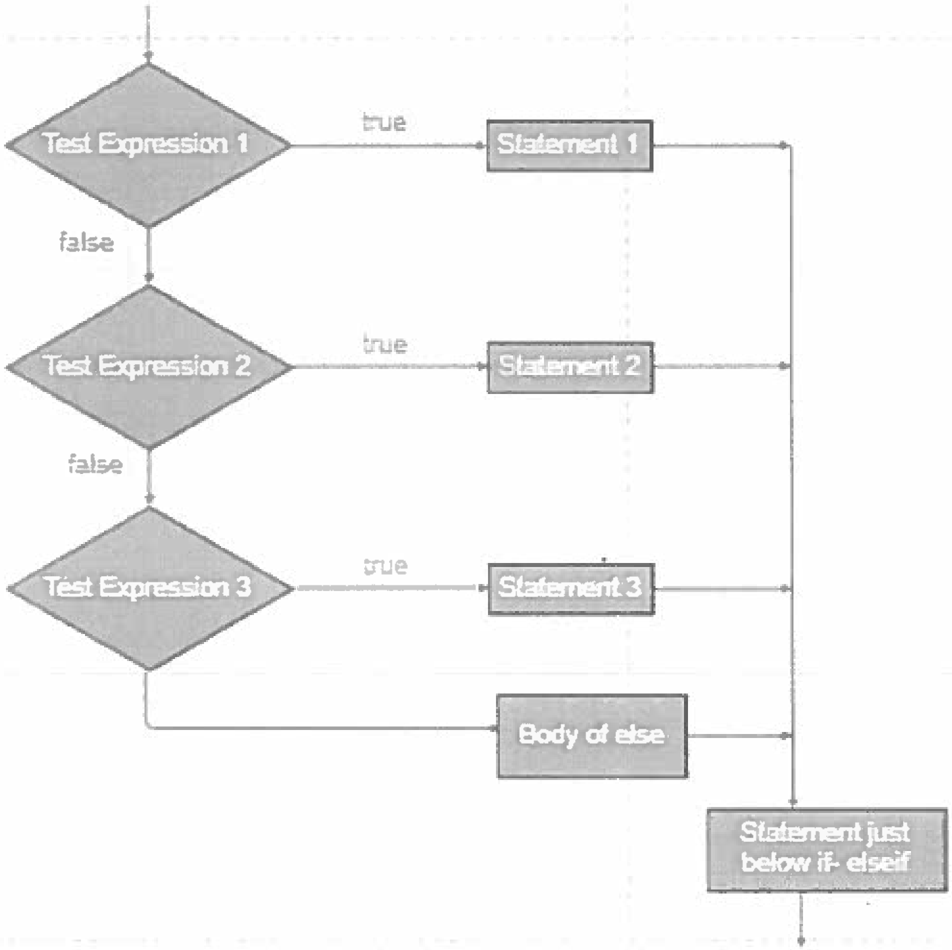
'. If – else – if ladder Statement

The if-else-if ladder statement executes one condition from multiple statements. The execution starts from top and checked for each if condition. The statement of if block will be executed which evaluates to be true. If none of the if condition evaluates to be true then the last else block is evaluated.

Syntax:

```
if(condition1)
{
    // code to be executed if condition1 is true
}
else if(condition2)
{
    // code to be executed if condition2 is true
}
else if(condition3)
{
    // code to be executed if condition3 is true
}
...
else
{
    // code to be executed if all the conditions are false
}
```

Flowchart:



Example:

```
/ C program to illustrate nested-if statement

#include <stdio.h>

int main()
{
   int i = 20;

   // Check if i is 10
   if (i == 10)
      printf("i is 10");

   // Since i is not 10
   // Check if i is 15
   else if (i == 15)
      printf("i is 15");

   // Since i is not 15
   // Check if i is 20
   else if (i == 20)
      printf("i is 20");
```

```
    // If none of the above conditions is true
    // Then execute the else statement
    else
        printf("i is not present");

    return 0;
}
```

Output:
i is 20

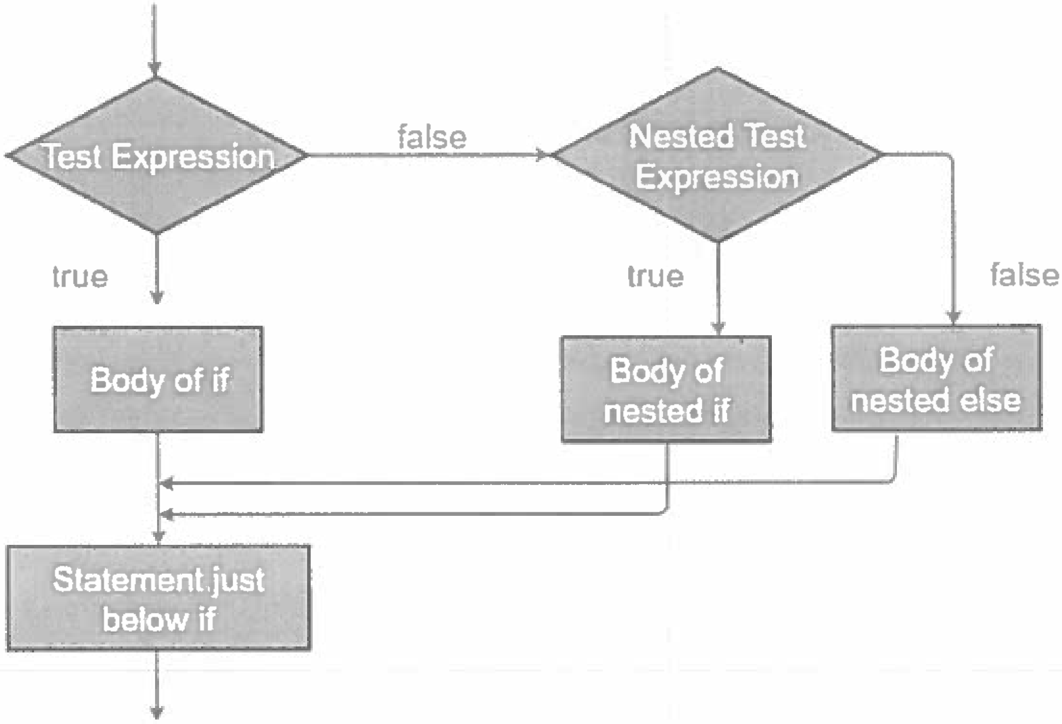
Nested – If Statement

if statement inside an if statement is known as nested if. if statement in this case is the target of another if or else statement. When more than one condition needs to be true and one of the condition is the sub-condition of parent condition, nested if can be used.

Syntax:

```
if (condition1)
{
    // code to be executed
    // if condition2 is true
    if (condition2)
    {
        // code to be executed
        // if condition2 is true
    }
}
```

Flowchart:



Example:

```
#include <stdio.h>

int main()

{

int dig1, dig2, dig3;

printf("Enter three numbers: ");

scanf("%d%d%d", &dig1, &dig2, &dig3);

if(dig1 > dig2)

{

if(dig1 > dig3)

{

printf("dig1 is the maximum");

}

else

{

printf("dig3 is the maximum");

}
```

```
}
else
{
if(dig2 > dig3)
{
printf("dig2 is the maximum");
}
else
{
printf("dig3 is the maximum");
}
}
return 0;
}
```

Output:

Greatest Digit Example 3

10. **Write a C program to multiply two m x n and p x q size matrices. (12m)**

```
 #program for multiply two matrix
#include <stdio.h>

int main()
{
        int m, n, p, q, c, d, k, sum = 0;
        int first[10][10], second[10][10], multiply[10][10];

          printf("Enter the number of rows and columns of first matrix\n");
          scanf("%d%d", &m, &n);
          printf("Enter the elements of first matrix\n");

          for (  c = 0 ; c < m ; c++ )
            for ( d = 0 ; d < n ; d++ )
              scanf("%d", &first[c][d]);

          printf("Enter the number of rows and columns of second matrix\n");
          scanf("%d%d", &p, &q);

          if ( n != p )
            printf("Matrices with entered orders can't be multiplied with each other.\n");
```

```
       else
 {
        printf("Enter the elements of second matrix\n");

        for ( c = 0 ; c < p ; c++ )
          for ( d = 0 ; d < q ; d++ )
            scanf("%d", &second[c][d]);

        for ( c = 0 ; c < m ; c++ )
        {
          for ( d = 0 ; d < q ; d++ )
          {
            for ( k = 0 ; k < p ; k++ )
            {
              sum = sum + first[c][k]*second[k][d];
            }

            multiply[c][d] = sum;
            sum = 0;
          }
        }

        printf("Product of entered matrices:-\n");

        for ( c = 0 ; c < m ; c++ )
        {
          for ( d = 0 ; d < q ; d++ )
            printf("%d\t", multiply[c][d]);

          printf("\n");
        }
      }
      return 0;
    }
```

OUTPUT:

```
Enter the number of rows and columns of first matrix 3 3
Enter the elements of first matrix
1 2 0
0 1 1
2 0 1
Enter the number of rows and columns of second matrix 3 3
Enter the elements of second matrix
1 1 2
2 1 1
1 2 1
Product of entered matrices:-
5       3       4
3       3       2
3       4       5
```

OR

11 **Explain categories of functions with examples.? (12m)**

Depending on whether arguments are present or not and whether a value is returned or not, functions are categorized into −

- Functions without arguments and without return values
- Functions without arguments and with return values
- Functions with arguments and without return values
- Functions with arguments and with return values

Functions without arguments and without return values

Calling function	Analysis	Called function

main ( )
{
--- 	
--- 	
fun ( );
}
	No arguments are passed
	
	No values are sent back
		fun ( )
		{
		---- 	
		---- 	
		}

Example

```
#include<stdio.h>
main (){
  void sum ();
  clrscr ();
  sum ();
  getch ();
}
void sum (){
  int a,b,c;
  printf("enter 2 numbers:\n");
  scanf ("%d%d", &a, &b);
  c = a+b;
  printf("sum = %d",c);
}
```

Output

Enter 2 numbers:

3

5

Sum=8

Functions without arguments and with return values

Calling function	Analysis	Called function

main ( )
{
int c;
--- 	
c= fun ( );
----- 	
----- 	
}
	No arguments are passed
	
	values are sent back
		fun ( )
		{
		---- 	
		---- 	
		return c;
		
		}

Example

```
#include<stdio.h>
main (){
  int sum ();
  int c;
  c= sum ();
  printf("sum = %d",c);
  getch ();
}
int sum (){
  int a,b,c;
  printf("enter 2 numbers");
  scanf ("%d%d", &a, &b);
  c = a+b;
  return c;
}
```

Output

Enter two numbers 10 20

30

Functions with arguments and without return values

Calling function	Analysis	Called function
main ( )	Arguments are passed	fun (int a, int b)
{		{
---		----
---	No values are sent back	----
fun (a,b);		}

}		

Example

```
#include<stdio.h>
main ( ) {
  void sum (int, int );
  int a,b;
  printf("enter 2 numbers");
  scanf("%d%d", &a,&b);
  sum (a,b);
  getch ();
}
void sum ( int a, int b) {
  int c;
  c= a+b;
  printf ("sum=%d", c);
}
```

Output

Enter two numbers 10 20

Sum=30

Functions with arguments and with return values

Calling function	Analysis	Called function
main ( )	Arguments are passed	fun (int a, int b)
{		{
int c;		----
---		----
---	value are sent back	
c= fun (a,b);		return c;
---		}

}		

Example

```
#include<stdio.h>
main (){
  int sum ( int,int);
  int a,b,c;
  printf("enter 2 numbers");
  scanf("%d%d", &a,&b);
  c= sum (a,b);
  printf ("sum=%d", c);
  getch ();
}
int sum ( int a, int b ){
  int c;
  c= a+b;
  return c;
}
```

Output

Enter two numbers 10 20

Sum=30

12. Explain dynamic memory allocation functions in detail. (12m)

C Dynamic Memory Allocation

In this tutorial, you'll learn to dynamically allocate memory in your C program using standard library functions: malloc(), calloc(), free() and realloc().

As you know, an array is a collection of a fixed number of values. Once the size of an array is declared, you cannot change it.

Sometimes the size of the array you declared may be insufficient. To solve this issue, you can allocate memory manually during run-time. This is known as dynamic memory allocation in C programming.

To allocate memory dynamically, library functions are malloc(), calloc(), realloc() and free() are used. These functions are defined in the <stdlib.h> header file.

C malloc()

The name "malloc" stands for memory allocation.

The malloc() function reserves a block of memory of the specified number of bytes. And, it returns a pointer of void which can be casted into pointers of any form.

Syntax of malloc()

ptr = (castType*) malloc(size);Example

ptr = (float*) malloc(100 * sizeof(float));

The above statement allocates 400 bytes of memory. It's because the size of `float` is 4 bytes. And, the pointer `ptr` holds the address of the first byte in the allocated memory.

The expression results in a `NULL` pointer if the memory cannot be allocated.

C calloc()

The name "calloc" stands for contiguous allocation.

The `malloc()` function allocates memory and leaves the memory uninitialized, whereas the `calloc()` function allocates memory and initializes all bits to zero.

Syntax of calloc()
ptr = (castType*)calloc(n, size);

Example:
ptr = (float*) calloc(25, sizeof(float));
The above statement allocates contiguous space in memory for 25 elements of type `float`.

C free()

Dynamically allocated memory created with either `calloc()` or `malloc()` doesn't get freed on their own. You must explicitly use `free()` to release the space.

Syntax of free()

free(ptr);

This statement frees the space allocated in the memory pointed by `ptr`.

Example 1: malloc() and free()

Here, we have dynamically allocated the memory for `n` number of `int`.

```
// Program to calculate the sum of n numbers entered by the user


#include <stdio.h>

#include <stdlib.h>

int main() {

  int n, i, *ptr, sum = 0;

 printf("Enter number of elements: ");

  scanf("%d", &n);

ptr = (int*) malloc(n * sizeof(int));

 // if memory cannot be allocated

  if(ptr == NULL) {

    printf("Error! memory not allocated.");

    exit(0);

  }

printf("Enter elements: ");

  for(i = 0; i < n; ++i) {

    scanf("%d", ptr + i);

    sum += *(ptr + i);

  }
```

```
printf("Sum = %d", sum);

 // deallocating the memory

 free(ptr);

return 0;

}
```

Run Code

Output

```
Enter number of elements: 3

Enter elements: 100

20

36

Sum = 156
```

Example 2: calloc() and free()

```
// Program to calculate the sum of n numbers entered by the user

#include <stdio.h>

#include <stdlib.h>

int main() {

 int n, i, *ptr, sum = 0;

 printf("Enter number of elements: ");

 scanf("%d", &n);


 ptr = (int*) calloc(n, sizeof(int));

 if(ptr == NULL) {

  printf("Error! memory not allocated.");
```

```
    exit(0);
  }

  printf("Enter elements: ");
  for(i = 0; i < n; ++i) {
    scanf("%d", ptr + i);
    sum += *(ptr + i);
  }
 printf("Sum = %d", sum);
  free(ptr);
  return 0;
}
```

Run Code

Output

```
Enter number of elements: 3
Enter elements: 100
20
36
Sum = 156
```

C realloc()

If the dynamically allocated memory is insufficient or more than required, you can change the size of previously allocated memory using the `realloc()` function.

Syntax of realloc()

ptr = realloc(ptr, x);

Here, `ptr` is reallocated with a new size `x`.

Example 3: realloc()

```
#include <stdio.h>
#include <stdlib.h>
int main() {
  int *ptr, i , n1, n2;
  printf("Enter size: ");
  scanf("%d", &n1);
ptr = (int*) malloc(n1 * sizeof(int));
printf("Addresses of previously allocated memory:\n");
  for(i = 0; i < n1; ++i)
    printf("%pc\n",ptr + i);
 printf("\nEnter the new size: ");
  scanf("%d", &n2);

  // rellocating the memory
  ptr = realloc(ptr, n2 * sizeof(int));
printf("Addresses of newly allocated memory:\n");
  for(i = 0; i < n2; ++i)
    printf("%pc\n", ptr + i);
  free(ptr);
return 0;
}
```

Run Code

Output

Enter size: 2

Addresses of previously allocated memory:

26855472

26855476

Enter the new size: 4

Addresses of newly allocated memory:

26855472

26855476

26855480

26855484

OR

13.a Explain array of structures.(6m)

Ans:

An array of structure in C programming is a collection of different datatype variables, grouped together under a single name.

General form of structure declaration

The structural declaration is as follows −

```
struct tagname{
  datatype member1;
  datatype member2;
  datatype member n;
};
```

Here, **struct** is the keyword

tagname specifies name of structure

member1, member2 specifies the data items that make up structure.

Example

The following example shows the usage of array of structures in C programming −

```
struct book{
  int pages;
  char author [30];
  float price;
};
```

Array of structures

- The most common use of structure in C programming is an array of structures.
- To declare an array of structure, first the structure must be defined and then an array variable of that type should be defined.
- For Example − struct book b[10]; //10 elements in an array of structures of type 'book'

Example

The following program shows the usage of array of structures.

```
#include <stdio.h>
#include <string.h>
struct student{
  int id;
  char name[30];
  float percentage;
};
int main(){
  int i;
  struct student record[2];
  // 1st student's record
  record[0].id=1;
  strcpy(record[0].name, "Bhanu");
  record[0].percentage = 86.5;
  // 2nd student's record
  record[1].id=2;
  strcpy(record[1].name, "Priya");
  record[1].percentage = 90.5;
  // 3rd student's record
  record[2].id=3;
  strcpy(record[2].name, "Hari");
  record[2].percentage = 81.5;
  for(i=0; i<3; i++){
    printf(" Records of STUDENT : %d \n", i+1);
    printf(" Id is: %d \n", record[i].id);
    printf(" Name is: %s \n", record[i].name);
    printf(" Percentage is: %f\n\n",record[i].percentage);
  }
  return 0;
}
```

Output

When the above program is executed, it produces the following result −

Records of STUDENT : 1

Id is: 1

Name is: Bhanu

Percentage is: 86.500000

Records of STUDENT : 2

Id is: 2

Name is: Priya

Percentage is: 90.500000

Records of STUDENT : 3

Id is: 3

Name is: Hari

Percentage is: 81.500000

13.b. Discuss nested structures?.(6m)

Ans:

A **nested structure** in C is a structure within structure. One structure can be declared inside another structure in the same way structure members are declared inside a structure.

Syntax:

```
struct name_1
{
   member1;
  member2;
  .
  .
  membern;
  struct name_2
  {
     member_1;
     member_2;
     .
     .
    member_n;
  }, var1
} var2;
```

The member of a nested structure can be accessed using the following syntax:

Variable name of Outer_Structure.Variable name of Nested_Structure.data member to access

Example:

```
#include<stdio.h>
struct address
{
   char city[20];
   int pin;
   char phone[14];
};
```

```
struct employee
{
   char name[20];
   struct address add;
};
void main ()
{
   struct employee emp;
   printf("Enter employee information?\n");
   scanf("%s %s %d %s",emp.name,emp.add.city, &emp.add.pin, emp.add.phone);
   printf("Printing the employee information....\n");
   printf("name: %s\nCity: %s\nPincode: %d\nPhone: %s",emp.name,emp.add.city,emp.add.pin,emp.add.phone
);
}
```

Output:

```
Enter employee information?

Arun

Delhi

110001

1234567890

Printing the employee information....

name: Arun

City: Delhi

Pincode: 110001

Phone: 1234567890
```

14. Explain various file handling functions with examples.(12m)

Ans:

File Handling in C

In programming, we may require some specific input data to be generated several numbers of times. Sometimes, it is not enough to only display the data on the console. The data to be displayed may be very large, and only a limited amount of data can be displayed on the console, and since the memory is volatile, it is impossible to recover the programmatically generated data again and again. However, if we need to do so, we may store it onto the local file system which is volatile and can be accessed every time. Here, comes the need of file handling in C.

File handling in C enables us to create, update, read, and delete the files stored on the local file system through our C program. The following operations can be performed on a file.

- Creation of the new file
- Opening an existing file
- Reading from the file
- Writing to the file
- Deleting the file

Functions for file handling

There are many functions in the C library to open, read, write, search and close the file. A list of file functions are given below:

No.	Function	Description
1	fopen()	opens new or existing file
2	fprintf()	write data into the file
3	fscanf()	reads data from the file
4	fputc()	writes a character into the file
5	fgetc()	reads a character from file
6	fclose()	closes the file
7	fseek()	sets the file pointer to given position
8	fputw()	writes an integer to file

9	fgetw()	reads an integer from file
10	ftell()	returns current position
11	rewind()	sets the file pointer to the beginning of the file

Opening File: fopen()

We must open a file before it can be read, write, or update. The fopen() function is used to open a file. The syntax of the fopen() is given below.

1. **FILE** *fopen(**const char** * filename, **const char** * mode);

The fopen() function accepts two parameters:

- The file name (string). If the file is stored at some specific location, then we must mention the path at which the file is stored. For example, a file name can be like **"c://some_folder/some_file.ext"**.
- The mode in which the file is to be opened. It is a string.

We can use one of the following modes in the fopen() function.

Mode	**Description**
r	opens a text file in read mode
w	opens a text file in write mode
a	opens a text file in append mode
r+	opens a text file in read and write mode
w+	opens a text file in read and write mode
a+	opens a text file in read and write mode
rb	opens a binary file in read mode
wb	opens a binary file in write mode

ab	opens a binary file in append mode
rb+	opens a binary file in read and write mode
wb+	opens a binary file in read and write mode
ab+	opens a binary file in read and write mode

The fopen function works in the following way.

- Firstly, It searches the file to be opened.
- Then, it loads the file from the disk and place it into the buffer. The buffer is used to provide efficiency for the read operations.
- It sets up a character pointer which points to the first character of the file.

Consider the following example which opens a file in write mode.

```
#include<stdio.h>
void main( )
{
FILE *fp ;
char ch ;
fp = fopen("file_handle.c","r") ;
while ( 1 )
{
ch = fgetc ( fp ) ;
if ( ch == EOF )
break ;
printf("%c",ch) ;
}
fclose (fp ) ;
}
```

Output

The content of the file will be printed.

```
#include<stdio.h>

void main( )
```

```
{
FILE *fp; // file pointer
char ch;
fp = fopen("file_handle.c","r");
while ( 1 )
{
ch = fgetc ( fp ); //Each character of the file is read and stored in the character file.
if ( ch == EOF )
break;
printf("%c",ch);
}
fclose (fp );
}
```

Closing File: fclose()

The fclose() function is used to close a file. The file must be closed after performing all the operations on it. The syntax of fclose() function is given below:

```
int fclose( FILE *fp );
```

OR

15. Explain various functions to access files randomly. 12m

Random accessing of files in C language can be done with the help of the following functions −

- ftell ( )
- rewind ( )
- fseek ( )

ftell ( )

It returns the current position of the file ptr.

The syntax is as follows −

```
int n = ftell (file pointer)
```

For example,

```
FILE *fp;
int n;
_____
_____
_____
n = ftell (fp);
```

Note − ftell ( ) is used for counting the number of characters which are entered into a file.

rewind ( )

It makes file ptr move to beginning of the file.

The syntax is as follows −

```
rewind (file pointer);
```

For example,

```
FILE *fp;
  -----
  -----
  rewind (fp);
  n = ftell (fp);
  printf ("%d", n);
```

Output

The output is as follows −

```
0 (always).
```

```
#include<stdio.h>
#include<conio.h>
void main(){
FILE *fp;
char c;
clrscr();
fp=fopen("file.txt","r");

while((c=fgetc(fp))!=EOF){
printf("%c",c);
}

rewind(fp);//moves the file pointer at beginning of the file
1.
while((c=fgetc(fp))!=EOF){
printf("%c",c);
}

fclose(fp);
getch();
}
```

Output:

this is a simple textthis is a simple text

fseek ( )

It is to make the file pntr point to a particular location in a file.

The syntax is as follows −

fseek(file pointer, offset, position);

Offset

- The no of positions to be moved while reading or writing.
- If can be either negative (or) positive.
  - Positive - forward direction.
  - Negative – backward direction.

Position

It can have three values, which are as follows −

- 0 – Beginning of the file.
- 1 – Current position.
- 2 – End of the file.

Example

- fseek (fp,0,2) - fp moved 0 bytes forward from the end of the file.
- fseek (fp, 0, 0) – fp moved 0 bytes forward from beginning of the file
- fseek (fp, m, 0) – fp moved m bytes forward from the beginning of the file.
- fseek (fp, -m, 2) – fp moved m bytes backward from the end of the file.

Errors

The errors related to fseek () function are as follows −

- fseek (fp, -m, 0);
- fseek(fp, +m, 2);

```
#include <stdio.h>
void main(){
  FILE *fp;

  fp = fopen("myfile.txt","w+");
  fputs("This is javatpoint", fp);
```

```
    fseek( fp, 7, SEEK_SET );
    fputs("sonoo jaiswal", fp);
    fclose(fp);
}
```

myfile.txt

This is sonoo jaiswal

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	Common to CE/ME			**Academic Year**	2021 - 2022
Course Code	20ESX01	**Test Duration**	3 Hrs.	**Max. Marks**	70	**Semester**	I
Course	**ENGINEERING DRAWING**						

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

No.	Questions (1 through 2)	Learning Outcome (s)	DoK
1	Draw the projections of the following points on a common reference line "XY". Maintaining suitable distance between them and also mention the quadrant they are in.	20ESX01.2	L1
2	A point P is 15mm above HP and 20mm in front of VP another point Q is 25mm behind VP and 40mm below hp draw p and q projections so that distance between end the projector is 90mm.Assume the position of points in suitable quadrants and draw the straight line joining a) the top views b) the front views.	20ESX01.4	L3

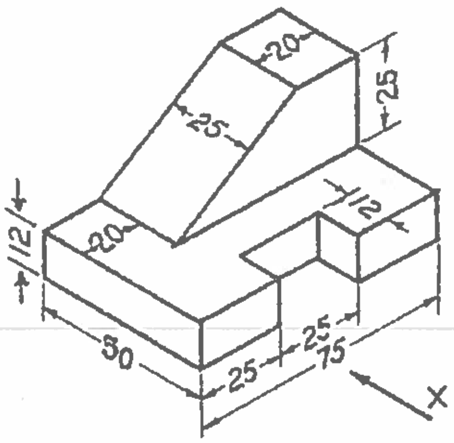
Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 10)	Learning Outcome (s)	DoK
3(a)	Construct a Pentagon of side 30mm when one side is vertical	20ESX01.1	L2
3(b)	Construct a Parabola, when the distance of focus from its directrix is 50mm and also draw a normal and tangent to the curve at a point 60mm from the Focus	20ESX01.1	L3
	OR		
4 (a)	Construct a diagonal scale of RF= 1: 32,00,000 to show kilometers and long enough to measure upto 400 kilometers. Show distance of 257 km and 333 km on your scale	20ESX01.1	L3
4(b)	A room of 1728m^3 volume is shown by a cube of 4cm side. Find R.F. and construct the scale to measure up to 50meters and indicate a distance of 37.6 meters on it.	20ESX01.1	L2
5 (a)	Assuming the position of points in suitable quadrants draw the Projections of the Following points on the same line keeping the Projectors 35mm apart. Point A – 25mm above H.P. and 45mm in front of the V.P. Point B - 40mm above the H.P. and on the V.P. Point C – on the V.P. and 35mm above the H.P.	20ESX01.2	L3
5 (b)	A line AB, 65mm long, has its end A 20 mm above the HP and 25 mm in front of the VP. The end B is 40mm above the HP and 65mm in front of the VP. Draw the projections of AB and show its inclinations with the HP and the VP?	20ESX01.2	L3
	OR		
6	The line AB of 40mm length is inclined at 30^0 to the H.P. and 60^0 to V.P. The end point A of the line is 15mm above H.P. and 20mm in front of V.P. Draw the Projections of lines?	20ESX01.2	L2
7	PQRS is a rhombus having diagonal PR=60mm and QS=40mm and they are perpendicular to each other. The plane of the rhombus is inclined at H.P such that its top view appears to be square. Draw its projections and determine inclination of the plane with the H.P.	20ESX01.3	L2
	OR		
8	A circular plate of 50mm diameter is resting on V.P. on a point on the circumference with its surface inclined at 45^0 to V.P and	20ESX01.3	L2

perpendicular to H.P. Draw its projections.

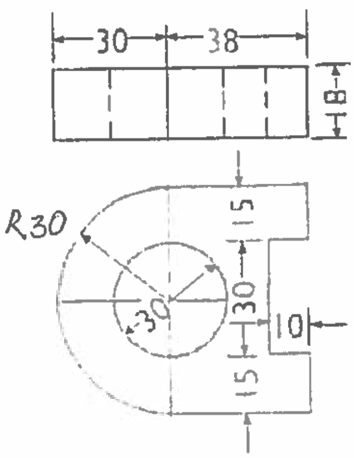
9 (a)	A hexagonal prism with side of base 25mm and 50mm long is resting on a corner of its base on HP. Draw the projections of the prism when its axis is making 30° with HP and parallel to V.P	20ESX01.4	L2
9 (b)	Draw the projections of a cylinder 75 mm diameter and 100mm long. Lying on the ground with its axis inclined at 30^0 to the VP and 45^0nclned to HP	20ESX01.4	L3
	OR		
10 (a)	Draw the projections of a cone, base 75mm diameter and axis 100mm long, lying on the ground on one of its generators with the axis parallel to the VP	20ESX01.4	L2
10 (b)	A square prism of side of base 30 mm and axis 55 mm long lies on one of its generator in the HP and its faces equally inclined to the HP. Draw its projections when its axis is inclined at an angle of 60^0 to the VP?	20ESX01.4	L3

11 Draw the front view, top view and side view from the isometric view. All dimensions are in mm. 20ESX01.5 L4



OR

12 20ESX01.5 L4

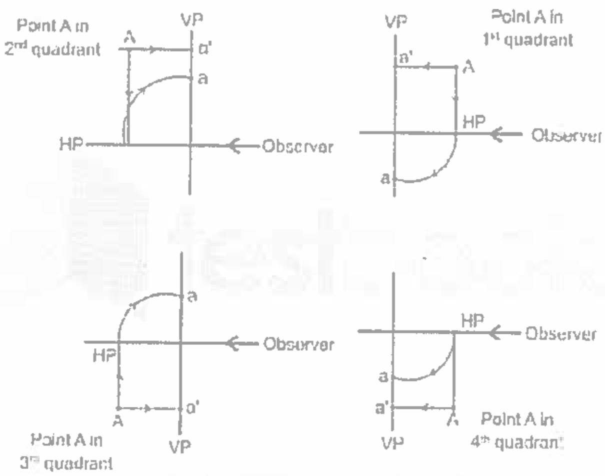


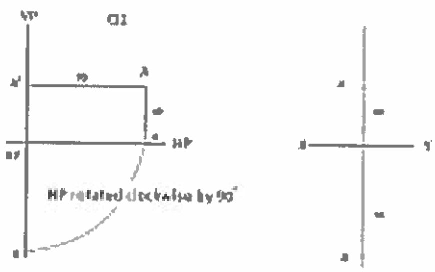
Draw the isometric view of considering the orthographic projections shown in figure.

Semester End Supplementary Examination, June, 2022

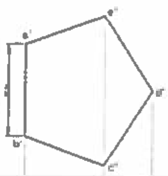
Degree	B. Tech. (U. G.)	Program	Common to CE/ME			Academic Year	2021 - 2022
Course Code	20ESX01	Test Duration	3 Hrs.	Max. Marks	70	Semester	I
Course	ENGINEERING DRAWING						

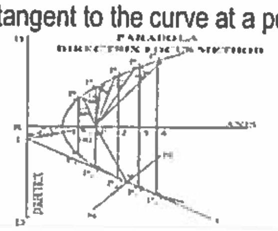
No.	Questions (1 through 2) M 5
1	Draw the projections of the following points on a common reference line "XY". Maintaining suitable distance between them and also mention the quadrant they are in.
2	A point P is 15mm above HP and 20mm in front of VP another point Q is 25mm behind VP and 40mm below hp draw p and q projections so that distance between end the projector is 90mm.Assume the position of points in suitable quadrants and draw the straight line joining a) the top views b) the front views.





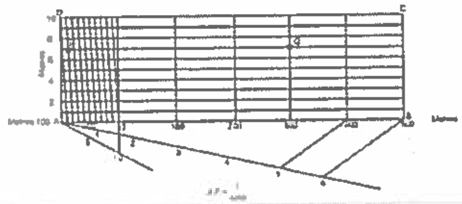
No.	Questions (6 through 10)
3(a)	Construct a Pentagon of side 30mm when one side is vertical
3(b)	Construct a Parabola, when the distance of focus from its directrix is 50mm and also draw a normal and tangent to the curve at a point 60mm from the Focus





4 (a)

Construct a diagonal scale of RF= 1: 32,00,000 to show kilometers and long enough to measure upto 400 kilometers. Show distance of 257 km and 333 km on your scale



4(b)

A room of 1728m³ volume is shown by a cube of 4cm side. Find R.F. and construct the scale to measure up to 50meters and indicate a distance of 37.6 meters on it.

$$RF = \sqrt[3]{(216)\ cm^3 / (1728 \times (100)^3)\ cm^3}$$

$$= 1/200.$$

Max Length (ML) = 42 m (6 x 7 or 7 x 6):

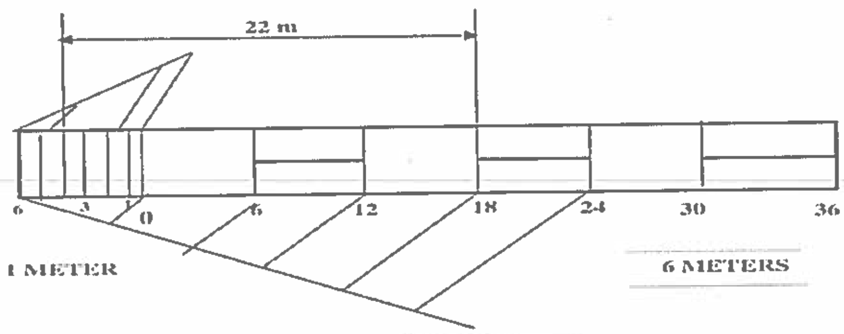
LOS is divided into 6 parts of 7 m each or 7 parts of 6 m each

Length of scale (LOS) = RF X ML (in cm)
= (1/200) x 42 x 100 cm (1 m=100cm)
= 21cm.

The length of the line that is drawn on the drawing sheet is 21 cm.

The sub scale is to be divided into 6 or 7 parts.

The length to be shown is 22 m = 18 m (on main scale) + 4 m (on sub scale)



RF=1/200

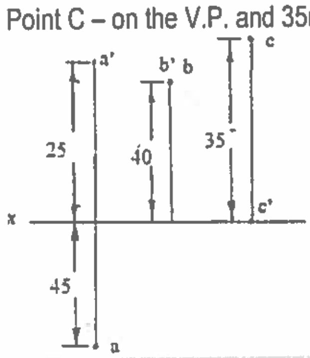
5 (a)

Assuming the position of points in suitable quadrants draw the Projections of the Following points on the same line keeping the Projectors 35mm apart.

Point A – 25mm above H.P. and 45mm in front of the V.P.

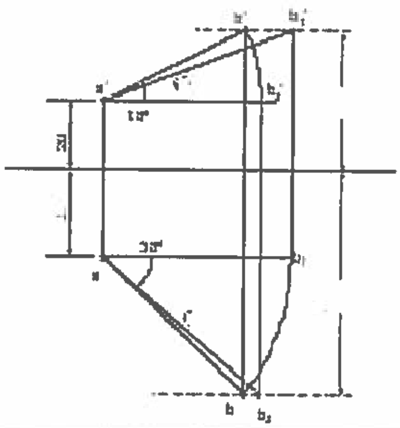
Point B - 40mm above the H.P. and on the V.P.

Point C – on the V.P. and 35mm above the H.P.

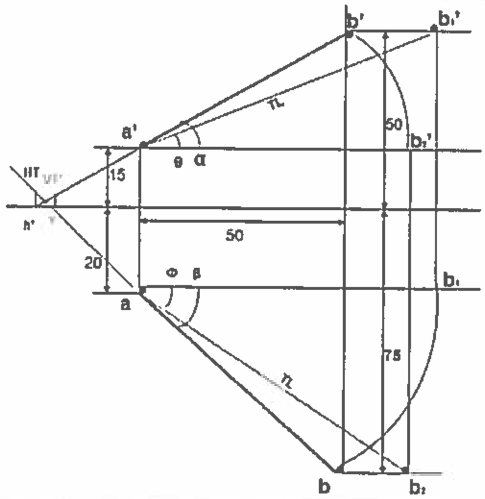


5 (b)

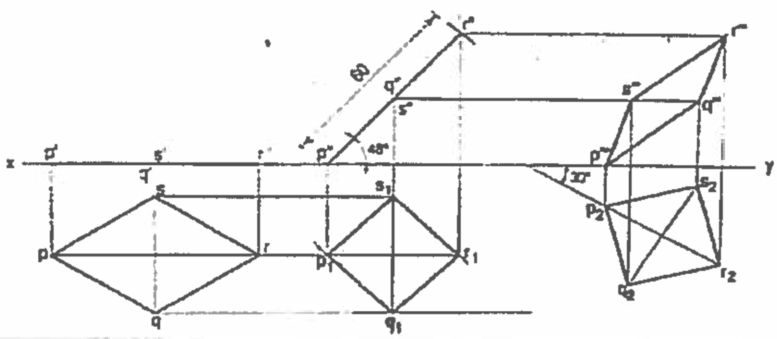
A line AB, 65mm long, has its end A 20 mm above the HP and 25 mm in front of the VP. The end B is 40mm above the HP and 65mm in front of the VP. Draw the projections of AB and show its inclinations with the HP and the VP?



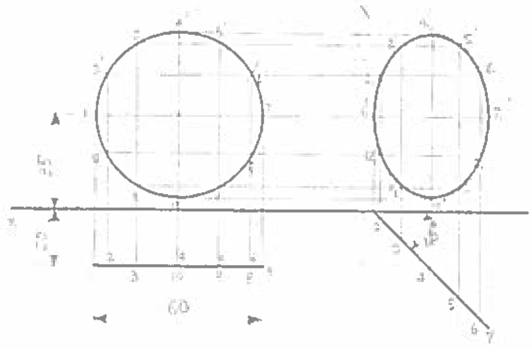
6 The line AB of 40mm length is inclined at 30⁰ to the H.P. and 60⁰ to V.P. The end point A of the line is 15mm above H.P. and 20mm in front of V.P. Draw the Projections of lines?



7 PQRS is a rhombus having diagonal PR=60mm and QS=40mm and they are perpendicular to each other. The plane of the rhombus is inclined at H.P such that its top view appears to be square. Draw its projections and determine inclination of the plane with the H.P.

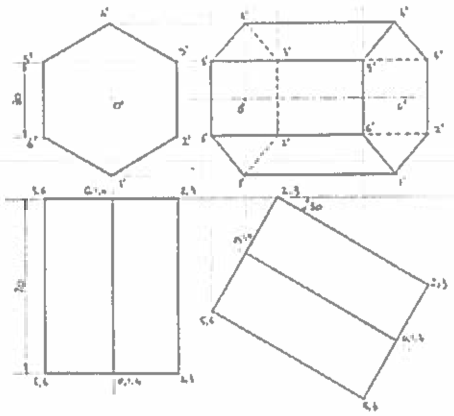


8 A circular plate of 50mm diameter is resting on V.P. on a point on the circumference with its surface inclined at 45⁰ to V.P and perpendicular to H.P. Draw its projections.



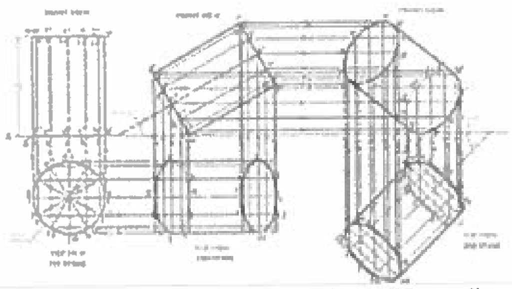
A hexagonal prism with side of base 25mm and 50mm long is resting on a corner of its base on HP. Draw the projections of the prism when its axis is making 30° with HP and parallel to V.P

9 (a)



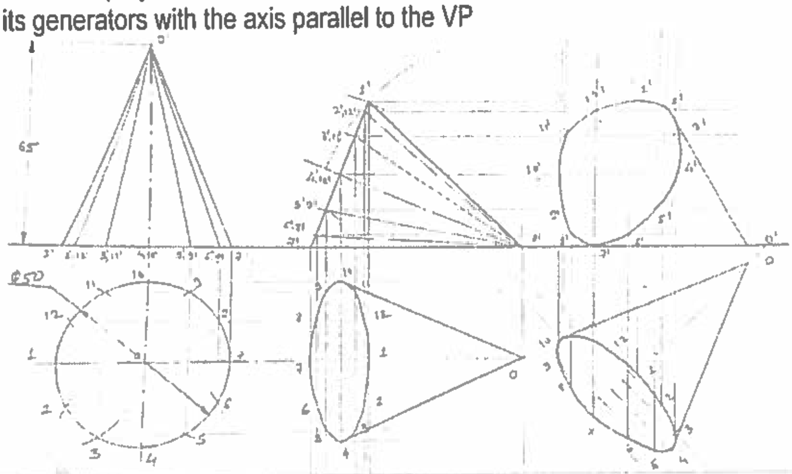
Draw the projections of a cylinder 75 mm diameter and 100mm long. Lying on the ground with its axis inclined at 30⁰ to the VP and 45⁰nclned to HP

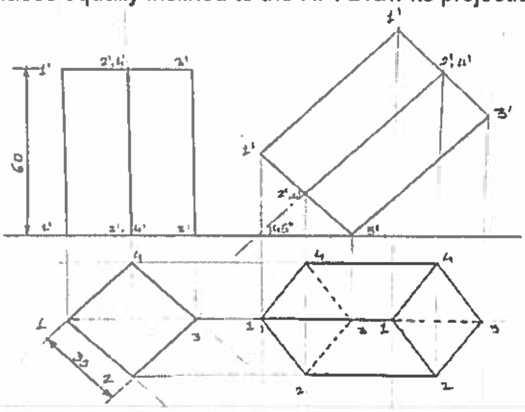
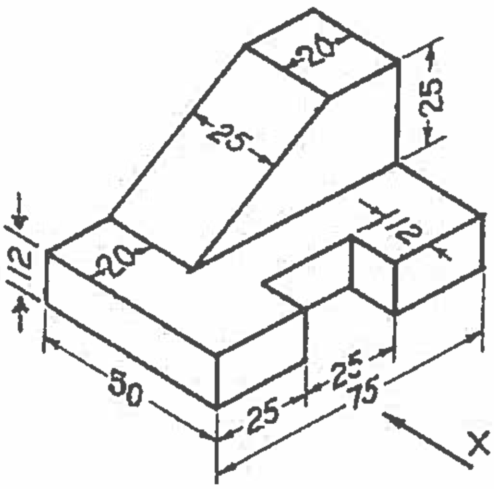
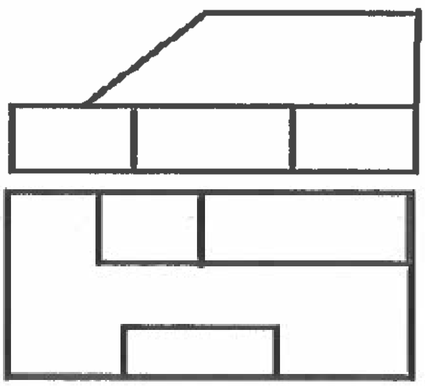
9 (b)

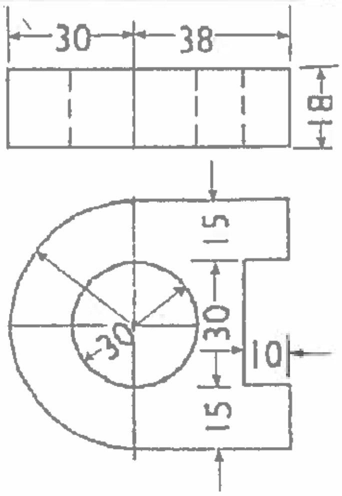


Draw the projections of a cone, base 75mm diameter and axis 100mm long, lying on the ground on one of its generators with the axis parallel to the VP

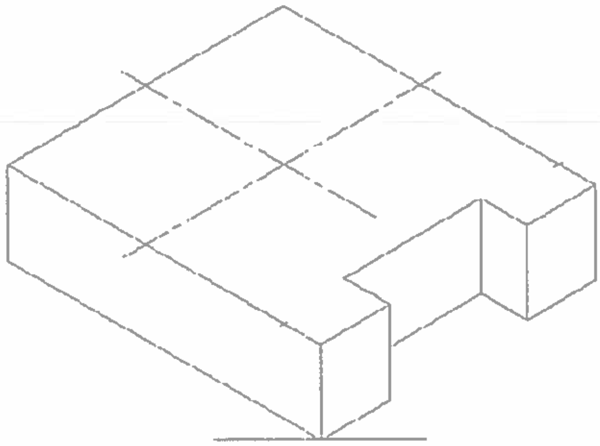
10 (a)



10 (b)	A square prism of side of base 30 mm and axis 55 mm long lies on one of its generator in the HP and its faces equally inclined to the HP. Draw its projections when its axis is inclined at an angle of 60^0 to the VP? 
11	Draw the front view, top view and side view from the isometric view. All dimensions are in mm.  



12 Draw the isometric view of considering the orthographic projections shown in figure.



NSRIT

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CSE/CSM/CSD	**Academic Year**	2021 - 2022
Course Code	20CS101	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	I
Course	FUNDAMENTALS OF COMPUTER SCIENCE				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is the difference between System and Application software?	20CS101.1	L2
2	What is meant by Installation and Assembling?	20CS101.2	L1
3	Summarize the features of LAN	20CS101.3	L5
4	List and define data types in SQL	20CS101.4	L1
5	Define Artificial Intelligence.	20CS101.5	L4

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

No.	Questions (6 through 10)	Marks	Learning Outcome (s)	DoK
6 (a)	Draw a block diagram of a computer. Explain the function of each of the blocks	8M	20CS101.1	L3
6 (b)	What is Memory? Explain types of Memory in brief	4M	20CS101.1	L3
	OR			
7 (a)	Write the history of computer? Explain the classification of computers	8M	20CS101.1	L2
7 (b)	Describe what is special purpose & General-purpose computer	4M	20CS101.1	L2
8 (a)	Explain the basic structure of a C program with an example.	6M	20CS101.2	L1
8 (b)	What is an identifier (variable)? What are the rules to construct identifier (varible)? Classify the following as valid/invalid Identifiers. i) num2 ii) $num1 iii) +add iv) a_2 v) 199_space vi) _apple vii)#12	6M	20CS101.2	L2
	OR			
9 (a)	What is type conversion? Explain two types of conversion with examples	4M	20CS101.2	L2
9 (b)	Write a calculator program in C language to do simple operations like addition, subtraction, multiplication, and division. Use switch statement in your program	8M	20CS101.2	L3
10 (a)	Explain the function of OSI reference model.	8M	20CS101.3	L1
10 (b)	List out different services of Operating Systems and explain each service	4M	20CS101.3	L1
	OR			
11 (a)	What are the objectives of computer communication networks? What are the network components? Explain	8M	20CS101.3	L2
11 (b)	Define Operating Systems and discuss its role from different perspectives	4M	20CS101.3	L2
12 (a)	List the applications of Database Systems with suitable explanation	4M	20CS101.4	L4
12 (b)	Compare the features of hierarchical, network and relational data models	8M	20CS101.4	L2
	OR			
13 (a)	What is a Database Management System? Explain various components of it	7M	20CS101.4	L1
13 (b)	Explain Three-schema architecture	5M	20CS101.4	L2
14 (a)	Explain the different domains of Artificial Intelligence	4M	20CS101.5	L1
14 (b)	Write a short note on different types of Machine Learning Algorithms	8M	20CS101.5	L2
	OR			
15 (a)	Define intelligent systems? Discuss the foundations of AI	4M	20CS101.5	L1
15 (b)	List and give brief description about basic machine learning models	8M	20CS101.5	L2

Supply key paper for fundamentals of computers

PART -A

1. What is the difference between System and Application software?

Ans:

	System Software	Application Software
1.	System Software maintains the system resources and gives the path for application software to run.	Application software is built for specific tasks.
2.	Low level languages are used to write the system software.	While high level languages are used to write the application software.
3.	Its a general purpose software.	While its a specific purpose software.
4.	Without system software, system can't run.	While without application software system always runs.

2. What is meant by Installation and Assembling?

Ans: Computer assembly is the process of building a computer from scratch by integrating multiple hardware devices into one working system.
Installation (or setup) of a computer program (including device drivers and plugins), is the act of making the program ready for execution.

3. Summarize the features of LAN?

Ans: **Network size is limited to a small geographical area, presently to a few kilometers. Data transfer rate is generally high. They range from 100 Mbps to 1000 Mbps.** In general, a LAN uses only one type of transmission medium, commonly category 5 coaxial cables.

4. List and define data types in SQL?

- Ans: Numeric data types such as int, tinyint, bigint, float, real, etc.
- Date and Time data types such as Date, Time, Datetime, etc.
- Character and String data types such as char, varchar, text, etc.
- Unicode character string data types, for example nchar, nvarchar, ntext, etc

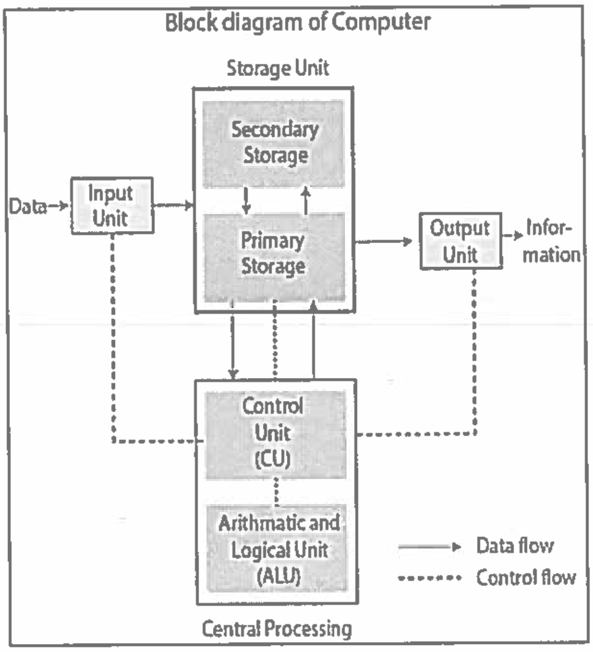
5. Define Artificial Intelligence.

Ans: Artificial intelligence (AI) is a wide-ranging branch of computer science concerned with building smart machines capable of performing tasks that typically require human intelligence.

PART-B

6. (A)Draw a block diagram of a computer. Explain the function of each of the blocks ?

ANS:



Computer Block Diagram System: Mainly computer system consists of three parts, that are central processing unit (CPU), Input Devices, and Output Devices. The Central Processing Unit (CPU) is divided into two parts again: arithmetic logic unit (ALU) and the control unit (CU). The set of instruction is in the form of raw data.

A large amount of data is stored in the computer memory with the help of primary and secondary storage devices. The CPU is like the heart/brain of the computer. The user does not get the desired output, without the necessary option taken by the CPU. The Central processing unit (CPU) is responsible for the processing of all the instructions which are given by the user to the computer system.

The data is entered through input devices such as the keyboard, mouse, etc. This set of instruction is processed by the CPU after getting the input by the user, and then the computer system produces the output. The computer can show the output with the help of output devices to the user, such as monitor, printer, etc.

- CPU (Central Processing Unit)
- Storage Unit
- ALU(Arithmetic Logic Unit)
- Control Unit

Central Processing Unit (CPU)

The computer system is nothing without the Central processing Unit so, it is also known as the brain or heat of computer. The CPU is an electronic hardware device which can perform different types of operations such as arithmetic and logical operation.

The CPU contains two parts: the arithmetic logic unit and control unit. We have discussed briefly the arithmetic unit, logical unit, and control unit which are given below:

Control Unit

The control unit (CU) controls all the activities or operations which are performed inside the computer system. It receives instructions or information directly from the main memory of the computer.
When the control unit receives an instruction set or information, it converts the instruction set to control signals then; these signals are sent to the central processor for further processing. The control unit understands which operation to execute, accurately, and in which order.

Arithmetic and Logical Unit

The arithmetic and logical unit is the combinational digital electronic circuit that can perform arithmetic operations on integer binary numbers.It presents the arithmetic and logical operation. The outputs of ALU will change asynchronously in response to the input. The basic arithmetic and bitwise logic functions are supported by ALU.

Storage Unit

The information or set of guidelines are stored in the storage unit of the computer system. The storage unit provides the space to store the data or instruction of processed data. The information or data is saved or hold in computer memory or storage device. The data storage is the core function and fundamental of the computer components.

Components of Computer System

The hardware and software exist on the computer. The information which is stored through the device is known as computer software. The hardware components of the computer system are related to electronic and mechanical parts, and the software component is related to data and computer programs. Many elements are connected to the main circuit board of the computer system called a "motherboard."

6(B) What is Memory? Explain types of Memory in brief?

ANS: Computer memory is of two basic types – **Primary memory(RAM and ROM) and Secondary memory (hard drive, CD, etc)**. Random Access Memory (RAM) is primary-volatile memory and Read-Only Memory (ROM) is primary-non-volatile memory.

There are two types of memories:

(i) primary memory
(ii) secondary memory

Primary memory: Primary storage (also known as main memory) is **the component of the computer that holds data , programs and instructions that are currently in use.** Primary storage is located on the motherboard . As a result, data can be read from and written to primary storage extremely quickly

Secondary memory:

Secondary memory is **computer memory that is non-volatile and persistent in nature and is not directly accessed by a computer/processor**. It allows a user to store data that may be instantly and easily retrieved, transported and used by applications and services. Secondary memory is also known as secondary storage.

7(a) Write the history of computer? Explain the classification of computers?
Ans: **The history of computer:**

Generations of Computers

A generation of computers refers to the specific improvements in computer technology with time. In 1946, electronic pathways called circuits were developed to perform the counting. It replaced the gears and other mechanical parts used for counting in previous computing machines.

In each new generation, the circuits became smaller and more advanced than the previous generation circuits. The miniaturization helped increase the speed, memory and power of computers. There are five generations of computers which are described below;

First Generation Computers

The first generation (1946-1959) computers were slow, huge and expensive. In these computers, vacuum tubes were used as the basic components of CPU and memory. These computers were mainly depended on batch operating system and punch cards. Magnetic tape and paper tape were used as output and input devices in this generation;

Some of the popular first generation computers are;

- **ENIAC** (Electronic Numerical Integrator and Computer)
- **EDVAC** (Electronic Discrete Variable Automatic Computer)
- **UNIVACI**(Universal Automatic Computer)
- **IBM-701**
- **IBM-650**

Second Generation Computers

The second generation (1959-1965) was the era of the transistor computers. These computers used transistors which were cheap, compact and consuming less power; it made transistor computers faster than the first generation computers.

In this generation, magnetic cores were used as the primary memory and magnetic disc and tapes were used as the secondary storage. Assembly language and programming languages like COBOL and FORTRAN, and Batch processing and multiprogramming operating systems were used in these computers.

Some of the popular second generation computers are;

- **IBM 1620**
- **IBM 7094**

- CDC 1604
- CDC 3600
- UNIVAC 1108

7(b) Describe what is special purpose & General-purpose computer?

Ans: **special purpose of a computer :**

A special purpose system is a. [s]ystem or platform that employs computing resources (i.e., hardware, firmware, and optionally software) that are physically embedded in, dedicated to, or necessary in real time for the performance of the system's mission. These computer resources are referred to as platform IT.

General purpose of a computer:

A general purpose computer is **a computer that is designed to be able to carry out many different tasks**. Desktop computers and laptops are examples of general purpose computers. Among other things, they can be used to: access the internet.

8(a). Explain the basic structure of a C program with an example.

Ans:

the basic structure of a C program. A C program is divided into different sections. There are six main sections to a basic c program.

The six sections are,

- Documentation
- Link
- Definition
- Global Declarations
- Main functions
- Subprograms

So now that the introduction is out of the way, let us jump to the main discussion. The whole code follows this outline. Each code has a similar outline. Now let us learn about each of this layer in detail.

Figure: Basic Structure Of C Program

Moving on to the next bit of this basic structure of a C program article,

Documentation Section

The documentation section is the part of the program where the programmer gives the details associated with the program. He usually gives the name of the program, the details of the author and other details like the time of coding and description. It gives anyone reading the code the overview of the code.

Example

```
/**

* File Name: Helloworld.c

* Author: Manthan Naik

* date: 09/08/2019

* description: a program to display hello world

*            no input needed

*/
```

Moving on to the next bit of this basic structure of a C program article,

Link Section

This part of the code is used to declare all the header files that will be used in the program. This leads to the compiler being told to link the header files to the system libraries.

Example

```
#include<stdio.h>
```

Moving on to the next bit of this basic structure of a C program article,

Definition Section

In this section, we define different constants. The keyword define is used in this part.

```
#define PI=3.14
```

Moving on to the next bit of this basic structure of a C program article,

Global Declaration Section

This part of the code is the part where the global variables are declared. All the global variable used are declared in this part. The user-defined functions are also declared in this part of the code.

```
float area(float r);
int a=7;
```

Moving on to the next bit of this basic structure of a C program article,

Main Function Section

Every C-programs needs to have the main function. Each main function contains 2 parts. A declaration part and an Execution part. The declaration part is the part where all the variables are declared. The execution part begins with the curly brackets and ends with the curly close bracket. Both the declaration and execution part are inside the curly braces.

```
int main(void)
{
int a=10;
printf(" %d", a);
return 0;
}
```

Moving on to the next bit of this basic structure of a C program article,

Sub Program Section

All the user-defined functions are defined in this section of the program.

```
int add(int a, int b)
{
return a+b;
}
```

Sample Program

The C program here will find the area of a circle using a user-defined function and a global variable pi holding the value of pi

```
* File Name: areaofcircle.c
* Author: Manthan Naik
* date: 09/08/2019
* description: a program to calculate area of circle
*user enters the radius
**/
#include<stdio.h>//link section
#define pi 3.14;//defination section
float area(float r);//global declaration
```

```
int main()//main function
{
float r;
printf(" Enter the radius:n");
scanf("%f",&r);
printf("the area is: %f",area(r));
return 0;
}
float area(float r)
{
return pi * r * r;//sub program
}
```

Output

Enter the radius:

7

The area is : 153.8601

This is all about basic structure of a program.

8(b) What is an identifier (variable)? What are the rules to construct identifier (varible)? Classify the following as valid/invalid Identifiers. i) num2 ii) $num1 iii) +add iv) a_2 v) 199_space vi) _apple vii)#12

Ans: "Identifiers" or "symbols" are **the names you supply for variables, types, functions, and labels in your program**. Identifier names must differ in spelling and case from any keywords. You cannot use keywords (either C or Microsoft) as identifiers; they are reserved for special use:

There are some set of rules to construct identifiers:

- A valid identifier can have letters (both uppercase and lowercase letters), digits and underscores.
- The first letter of an identifier should be either a letter or an underscore.
- You cannot use keywords like int , while etc. as identifiers.
- There is no rule on how long an identifier can be.

9(a) What is type conversion? Explain two types of conversion with examples

Ans: type conversion, type casting, type coercion, and type juggling are **different ways of changing an expression from one data type to another**. An example would be the conversion of an integer value into a floating point value or its textual representation as a string, and vice versa.

TYPES OF TYPE CONVERSIONS :

1. **Implicit Type Conversion:**

Also known as 'automatic type conversion'.

- Done by the compiler on its own, without any external trigger from the user.
- Generally takes place when in an expression more than one data type is present. In such condition type conversion (type promotion) takes place to avoid loss of data.
- All the data types of the variables are upgraded to the data type of the variable with largest data type.
-
- `bool -> char -> short int -> int ->`
- `unsigned int -> long -> unsigned ->`
- `long long -> float -> double -> long double`
- It is possible for implicit conversions to lose information, signs can be lost (when signed is implicitly converted to unsigned), and overflow can occur (when long long is implicitly converted to float).

Example of Type Implicit Conversion:

```
// An example of implicit conversion
#include<stdio.h>
int main()
{
   int x = 10;    // integer x
   char y = 'a';  // character c
```

```
    // y implicitly converted to int. ASCII
    // value of 'a' is 97
    x = x + y;

    // x is implicitly converted to float
    float z = x + 1.0;
  printf("x = %d, z = %f", x, z);
    return 0;
}
```

Output:

```
x = 107, z = 108.000000
```

2. **Explicit Type Conversion–**

This process is also called type casting and it is user defined. Here the user can type cast the result to make it of a particular data type.

The syntax in C:

```
(type) expression
```

Type indicated the data type to which the final result is converted.

```
// C program to demonstrate explicit type casting
#include<stdio.h>

int main()
{
    double x = 1.2;

    // Explicit conversion from double to int
    int sum = (int)x + 1;

    printf("sum = %d", sum);

    return 0;
}
```

Output:

```
sum = 2
```

9(b). Write a calculator program in C language to do simple operations like addition, subtraction, multiplication, and division. Use switch statement in your program?

Ans:

```
#include <stdio.h>

int main() {
```

```
char op;
double first, second;
printf("Enter an operator (+, -, *, /): ");
scanf("%c", &op);
printf("Enter two operands: ");
scanf("%lf %lf", &first, &second);

switch (op) {
  case '+':
    printf("%.1lf + %.1lf = %.1lf", first, second, first + second);
    break;
  case '-':
    printf("%.1lf - %.1lf = %.1lf", first, second, first - second);
    break;
  case '*':
    printf("%.1lf * %.1lf = %.1lf", first, second, first * second);
    break;
  case '/':
    printf("%.1lf / %.1lf = %.1lf", first, second, first / second);
    break;
  // operator doesn't match any case constant
  default:
    printf("Error! operator is not correct");
}

return 0;
}
```

```
Enter an operator (+, -, *,): *
Enter two operands: 1.5
4.5
1.5 * 4.5 = 6.8
```

10(a) Explain the function of reference mode OSI

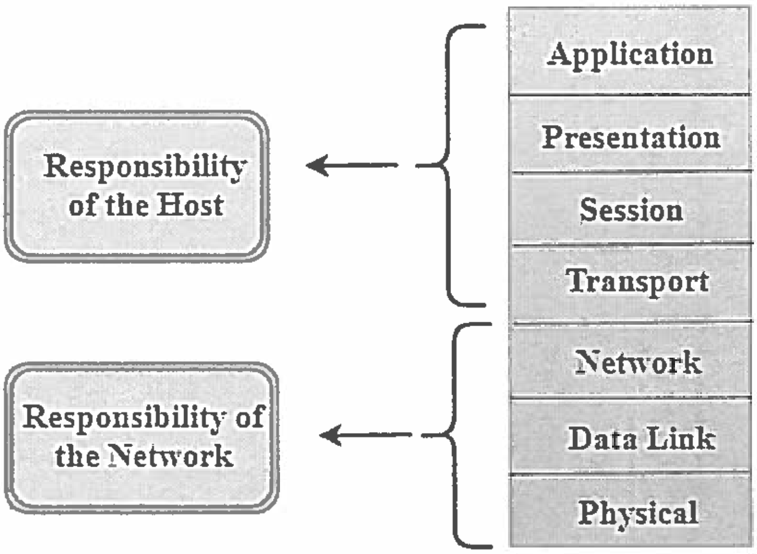
Ans:

OSI Model

- OSI stands for **Open System Interconnection** is a reference model that describes how information from a software application in one computer moves through a physical medium to the software application in another computer.

- OSI consists of seven layers, and each layer performs a particular network function.
- OSI model was developed by the International Organization for Standardization (ISO) in 1984, and it is now considered as an architectural model for the inter-computer communications.
- OSI model divides the whole task into seven smaller and manageable tasks. Each layer is assigned a particular task.
- Each layer is self-contained, so that task assigned to each layer can be performed independently.

Characteristics of OSI Model:



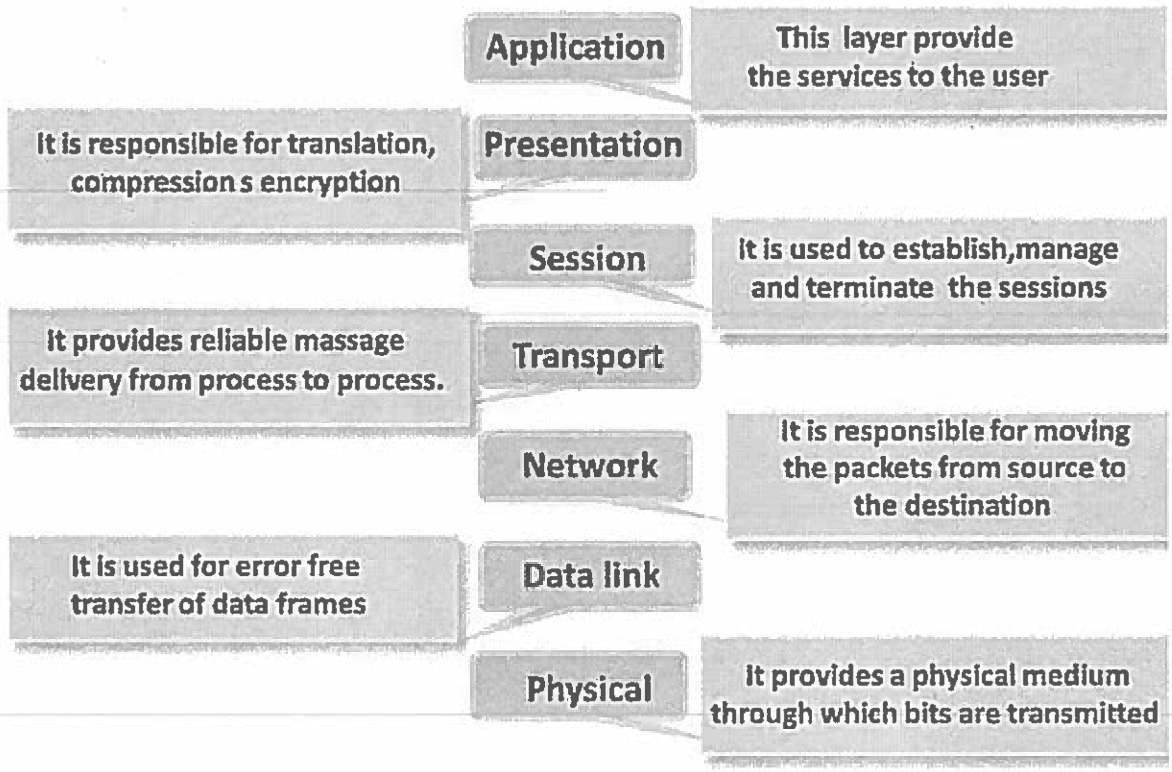
- The OSI model is divided into two layers: upper layers and lower layers.
- The upper layer of the OSI model mainly deals with the application related issues, and they are implemented only in the software. The application layer is closest to the end user. Both the end user and the application layer interact with the software applications. An upper layer refers to the layer just above another layer.
- The lower layer of the OSI model deals with the data transport issues. The data link layer and the physical layer are implemented in hardware and software. The physical layer is the lowest layer of the OSI model and is closest to the physical medium. The physical layer is mainly responsible for placing the information on the physical medium.

Functions of the OSI Layers

There are the seven OSI layers. Each layer has different functions. A list of seven layers are given below:

1. Physical Layer
2. Data-Link Layer
3. Network Layer
4. Transport Layer

5. Session Layer
6. Presentation Layer
7. Application Layer



Physical layer

- The main functionality of the physical layer is to transmit the individual bits from one node to another node.
- It is the lowest layer of the OSI model.
- It establishes, maintains and deactivates the physical connection.
- It specifies the mechanical, electrical and procedural network interface specifications.

Functions of a Physical layer:

- **Line Configuration:** It defines the way how two or more devices can be connected physically.
- **Data Transmission:** It defines the transmission mode whether it is simplex, half-duplex or full-duplex mode between the two devices on the network.
- **Topology:** It defines the way how network devices are arranged.
- **Signals:** It determines the type of the signal used for transmitting the information.

Data-Link Layer

- This layer is responsible for the error-free transfer of data frames.
- It defines the format of the data on the network.
- It provides a reliable and efficient communication between two or more devices.
- It is mainly responsible for the unique identification of each device that resides on a local network.
- It contains two sub-layers:
    - **Logical Link Control Layer**
        - It is responsible for transferring the packets to the Network layer of the receiver that is receiving.
        - It identifies the address of the network layer protocol from the header.
        - It also provides flow control.
    - **Media Access Control Layer**
        - A Media access control layer is a link between the Logical Link Control layer and the network's physical layer.
        - It is used for transferring the packets over the network.

Functions of the Data-link layer

- **Framing:** The data link layer translates the physical's raw bit stream into packets known as Frames. The Data link layer adds the header and trailer to the frame. The header which is added to the frame contains the hardware destination and source address.

- **Physical Addressing:** The Data link layer adds a header to the frame that contains a destination address. The frame is transmitted to the destination address mentioned in the header.
- **Flow Control:** Flow control is the main functionality of the Data-link layer. It is the technique through which the constant data rate is maintained on both the sides so that no data get corrupted. It ensures that the transmitting station such as a server with higher processing speed does not exceed the receiving station, with lower processing speed.
- **Error Control:** Error control is achieved by adding a calculated value CRC (Cyclic Redundancy Check) that is placed to the Data link layer's trailer which is added to the message frame before it is sent to the physical layer. If any error seems to occurr, then the receiver sends the acknowledgment for the retransmission of the corrupted frames.
- **Access Control:** When two or more devices are connected to the same communication channel, then the data link layer protocols are used to determine which device has control over the link at a given time.

Network Layer

- It is a layer 3 that manages device addressing, tracks the location of devices on the network.
- It determines the best path to move data from source to the destination based on the network conditions, the priority of service, and other factors.
- The Data link layer is responsible for routing and forwarding the packets.
- Routers are the layer 3 devices, they are specified in this layer and used to provide the routing services within an internetwork.
- The protocols used to route the network traffic are known as Network layer protocols. Examples of protocols are IP and Ipv6.

Functions of Network Layer:

- **Internetworking:** An internetworking is the main responsibility of the network layer. It provides a logical connection between different devices.
- **Addressing:** A Network layer adds the source and destination address to the header of the frame. Addressing is used to identify the device on the internet.
- **Routing:** Routing is the major component of the network layer, and it determines the best optimal path out of the multiple paths from source to the destination.
- **Packetizing:** A Network Layer receives the packets from the upper layer and converts them into packets. This process is known as Packetizing. It is achieved by internet protocol (IP).

Transport Layer

- The Transport layer is a Layer 4 ensures that messages are transmitted in the order in which they are sent and there is no duplication of data.
- The main responsibility of the transport layer is to transfer the data completely.
- It receives the data from the upper layer and converts them into smaller units known as segments.
- This layer can be termed as an end-to-end layer as it provides a point-to-point connection between source and destination to deliver the data reliably.

The two protocols used in this layer are:

- **Transmission Control Protocol**
  - It is a standard protocol that allows the systems to communicate over the internet.
  - It establishes and maintains a connection between hosts.
  - When data is sent over the TCP connection, then the TCP protocol divides the data into smaller units known as segments. Each segment travels over the internet using multiple routes, and they arrive in different orders at the destination. The transmission control protocol reorders the packets in the correct order at the receiving end.

- **User Datagram Protocol**
    - User Datagram Protocol is a transport layer protocol.
    - It is an unreliable transport protocol as in this case receiver does not send any acknowledgment when the packet is received, the sender does not wait for any acknowledgment. Therefore, this makes a protocol unreliable.

Functions of Transport Layer:

- **Service-point addressing:** Computers run several programs simultaneously due to this reason, the transmission of data from source to the destination not only from one computer to another computer but also from one process to another process. The transport layer adds the header that contains the address known as a service-point address or port address. The responsibility of the network layer is to transmit the data from one computer to another computer and the responsibility of the transport layer is to transmit the message to the correct process.
- **Segmentation and reassembly:** When the transport layer receives the message from the upper layer, it divides the message into multiple segments, and each segment is assigned with a sequence number that uniquely identifies each segment. When the message has arrived at the destination, then the transport layer reassembles the message based on their sequence numbers.
- **Connection control:** Transport layer provides two services Connection-oriented service and connectionless service. A connectionless service treats each segment as an individual packet, and they all travel in different routes to reach the destination. A connection-oriented service makes a connection with the transport layer at the destination machine before delivering the packets. In connection-oriented service, all the packets travel in the single route.
- **Flow control:** The transport layer also responsible for flow control but it is performed end-to-end rather than across a single link.
- **Error control:** The transport layer is also responsible for Error control. Error control is performed end-to-end rather than across the single link. The sender transport layer ensures that message reach at the destination without any error.

Session Layer

- It is a layer 3 in the OSI model.
- The Session layer is used to establish, maintain and synchronizes the interaction between communicating devices.

Functions of Session layer:

- **Dialog control:** Session layer acts as a dialog controller that creates a dialog between two processes or we can say that it allows the communication between two processes which can be either half-duplex or full-duplex.
- **Synchronization:** Session layer adds some checkpoints when transmitting the data in a sequence. If some error occurs in the middle of the transmission of data, then the transmission will take place again from the checkpoint. This process is known as Synchronization and recovery.

Presentation Layer

- A Presentation layer is mainly concerned with the syntax and semantics of the information exchanged between the two systems.
- It acts as a data translator for a network.
- This layer is a part of the operating system that converts the data from one presentation format to another format.
- The Presentation layer is also known as the syntax layer.

Functions of Presentation layer:

- **Translation:** The processes in two systems exchange the information in the form of character strings, numbers and so on. Different computers use different encoding methods, the presentation layer handles the interoperability between the different encoding methods. It converts the data from sender-dependent format into a common format and changes the common format into receiver-dependent format at the receiving end.
- **Encryption:** Encryption is needed to maintain privacy. Encryption is a process of converting the sender-transmitted information into another form and sends the resulting message over the network.
- **Compression:** Data compression is a process of compressing the data, i.e., it reduces the number of bits to be transmitted. Data compression is very important in multimedia such as text, audio, video.

Application Layer

- An application layer serves as a window for users and application processes to access network service.
- It handles issues such as network transparency, resource allocation, etc.
- An application layer is not an application, but it performs the application layer functions.
- This layer provides the network services to the end-users.

Functions of Application layer:

- **File transfer, access, and management (FTAM):** An application layer allows a user to access the files in a remote computer, to retrieve the files from a computer and to manage the files in a remote computer.
- **Mail services:** An application layer provides the facility for email forwarding and storage.
- Directory services: An application provides the distributed database sources and is used to provide that global information about various objects.

10(b) List out different services of Operating Systems and explain each service?
Ans:

An Operating System provides services to both the users and to the programs.

- It provides programs an environment to execute.
- It provides users the services to execute the programs in a convenient manner.

Following are a few common services provided by an operating system –

- Program execution
- I/O operations
- File System manipulation
- Communication
- Error Detection
- Resource Allocation
- Protection

I/O Operation

An I/O subsystem comprises of I/O devices and their corresponding driver software. Drivers hide the peculiarities of specific hardware devices from the users.

An Operating System manages the communication between user and device drivers.

- I/O operation means read or write operation with any file or any specific I/O device.
- Operating system provides the access to the required I/O device when required.

Resource Management

In case of multi-user or multi-tasking environment, resources such as main memory, CPU cycles and files storage are to be allocated to each user or job. Following are the major activities of an operating system with respect to resource management –

- The OS manages all kinds of resources using schedulers.
- CPU scheduling algorithms are used for better utilization of CPU.

11(a) What are the objectives of computer communication networks? What are the network components? Explain

Ans: The main goal of the computer network is **Resource Sharing**. It is to create all the programs, data and hardware accessible to anyone on the network without considering the resource's physical area and the client.

Computer Network Components

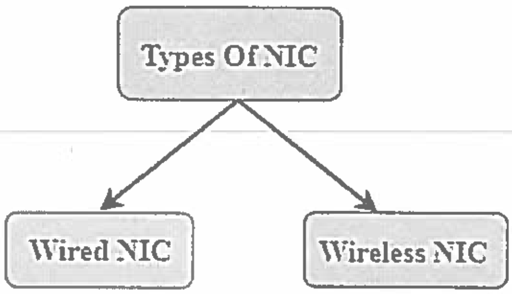
Computer network components are the *major parts* which are needed to *install the software*. Some important network components are **NIC, switch, cable, hub, router,** and **modem**. Depending on the type of network that we need to install, some network components can also be removed. For example, the wireless network does not require a cable.

Following are the major components required to install a network:

NIC

- NIC stands for network interface card.
- NIC is a hardware component used to connect a computer with another computer onto a network
- It can support a transfer rate of 10,100 to 1000 Mb/s.
- The MAC address or physical address is encoded on the network card chip which is assigned by the IEEE to identify a network card uniquely. The MAC address is stored in the PROM (Programmable read-only memory).

There are two types of NIC:



1. Wired NIC
2. Wireless NIC

Wired NIC: The Wired NIC is present inside the motherboard. Cables and connectors are used with wired NIC to transfer data.

Wireless NIC: The wireless NIC contains the antenna to obtain the connection over the wireless network. For example, laptop computer contains the wireless NIC.

Hub

A Hub is a hardware device that divides the network connection among multiple devices. When computer requests for some information from a network, it first sends the request to the Hub through cable. Hub will broadcast this request to the entire network. All the devices will check whether the request belongs to them or not. If not, the request will be dropped.

The process used by the Hub consumes more bandwidth and limits the amount of communication. Nowadays, the use of hub is obsolete, and it is replaced by more advanced computer network components such as Switches, Routers.

Switch

A switch is a hardware device that connects multiple devices on a computer network. A Switch contains more advanced features than Hub. The Switch contains the updated table that decides where the data is transmitted or not. Switch delivers the message to the correct destination based on the physical address present in the incoming message. A Switch does not broadcast the message to the entire network like the Hub. It determines the device to whom the message is to be transmitted. Therefore, we can say that switch provides a direct connection between the source and destination. It increases the speed of the network.

Router

- A router is a hardware device which is used to connect a LAN with an internet connection. It is used to receive, analyze and forward the incoming packets to another network.
- A router works in a **Layer 3 (Network layer)** of the OSI Reference model.
- A router forwards the packet based on the information available in the routing table.
- It determines the best path from the available paths for the transmission of the packet.

Advantages Of Router:

- **Security:** The information which is transmitted to the network will traverse the entire cable, but the only specified device which has been addressed can read the data.
- **Reliability:** If the server has stopped functioning, the network goes down, but no other networks are affected that are served by the router.
- **Performance:** Router enhances the overall performance of the network. Suppose there are 24 workstations in a network generates a same amount of traffic. This increases the traffic load on the network. Router splits the single network into two networks of 12 workstations each, reduces the traffic load by half.
- **Network range**

Modem

- A modem is a hardware device that allows the computer to connect to the internet over the existing telephone line.
- A modem is not integrated with the motherboard rather than it is installed on the PCI slot found on the motherboard.
- It stands for Modulator/Demodulator. It converts the digital data into an analog signal over the telephone lines.

Based on the differences in speed and transmission rate, a modem can be classified in the following categories:

- Standard PC modem or Dial-up modem
- Cellular Modem
- Cable modem

Cables and Connectors

Cable is a transmission media used for transmitting a signal.

There are three types of cables used in transmission:

- Twisted pair cable
- Coaxial cable
- Fibre-optic cable

11(b) Define Operating Systems and discuss its role from different perspectives

Ans: An **Operating System** can be defined as an **interface between user and hardware**. It is responsible for the execution of all the processes, Resource Allocation, CPU management, File Management and many other tasks.

Perspectives in which operating system is used are

1. Process Management
2. Process Synchronization
3. Memory Management
4. CPU Scheduling
5. File Management
6. Security

12(a) List the applications of Database Systems with suitable explanation?
Ans: applications of database systems:
Controls database redundancy: It can control data redundancy because it stores all the data in one single database file and that recorded data is placed in the database.
Data sharing: In DBMS, the authorized users of an organization can share the data among multiple users.
Easily Maintenance: It can be easily maintainable due to the centralized nature of the database system.
Reduce time: It reduces development time and maintenance need.
Backup: It provides backup and recovery subsystems which create automatic backup of data from hardware and software failures and restores the data if required.
multiple user interface: It provides different types of user interfaces like graphical user interfaces, application program interfaces

12(b)Compare the features of hierarchical, network and relational data models?

ans(a) Difference Between Hierarchical Network and Relational Database Model

Definition

A hierarchical model is a structure of data organized in a tree-like model using parent-child relationships while network model is a database model that allows multiple records to be linked to the same owner file. A relational model, on the other hand, is a database model to manage data as tuples grouped into relations (tables).

Basis

Hierarchical model arranges data in a tree similar structure while network model organizes data in a graph structure. In contrast, relational model arranges data in tables. Hence, this is the main difference between hierarchical network and relational database model.

Relationship

Moreover, an important difference between hierarchical network and relational database model is that while a hierarchical model represents "one to many" relationship, a network model represents "many to many" relationship. Furthermore, relational model can represent both "one to many" and "many to many" relationships.

Accessing data

Although it is difficult to access data in the hierarchical model, it is easier to access data in the network model and the relational model.

Flexibility

Also, another difference between hierarchical network and relational database model is their flexibility. The hierarchical model is less flexible, but the network model and relational model are flexible.

Conclusion

Database models help to arrange data in the databases of DBMS. The main difference between hierarchical network and relational database model is that hierarchical model organizes data in a tree-like structure while network model arranges data in a graph structure and relational database model organizes data in tables.

13(a) What is a Database Management System? Explain various components of it?

Ans: A database management system (DBMS) is **a software tool that enables users to manage a database easily**. It allows users to access and interact with the underlying data in the database. These actions can range from simply querying data to defining database schemas that fundamentally affect the database structure.

Components of database management system:

Hardware

The hardware is the actual computer system used for keeping and accessing the database. The conventional DBMS hardware consists of secondary storage devices such as hard disks. Databases run on the range of machines from micro computers to mainframes.

Software

Software is the actual DBMS between the physical database and the users of the system. All the requests from the user for accessing the database are handled by DBMS.

Data

It is an important component of the database management system. The main task of DBMS is to process the data. Databases are used to store the data, retrieved, and updated to and from the databases.

Users

There are a number of users who can access or retrieve the data on demand using the application and the interfaces provided by the DBMS.

The users of the database can be classified into different groups –

- Native Users
- Online Users
- Sophisticated Users
- Specialized Users
- Application Users
- DBA- Database Administrator

13(b) Explain Three-schema architecture?

Ans: Three schema Architecture

- The three schema architecture is also called ANSI/SPARC architecture or three-level architecture.
- This framework is used to describe the structure of a specific database system.
- The three schema architecture is also used to separate the user applications and physical database.
- The three schema architecture contains three-levels. It breaks the database down into three different categories.

the main objective of three level architecture is to enable multiple users to access the same data with a personalized view while storing the underlying data only once. Thus it separates the user's view from the physical structure of the database. This separation is desirable for the following reasons:

- Different users need different views of the same data.
- The approach in which a particular user needs to see the data may change over time.
- The users of the database should not worry about the physical implementation and internal workings of the database such as data compression and encryption techniques, hashing, optimization of the internal structures etc.
- All users should be able to access the same data according to their requirements.
- DBA should be able to change the conceptual structure of the database without affecting the user's
- Internal structure of the database should be unaffected by changes to physical aspects of the storage.

14(a) Explain the different domains of Artificial Intelligence

Ans: Artificial intelligence systems are critical for companies that wish to extract value from data by automating and optimizing processes or producing actionable insights.

There are certain domains of artificial intelligence on which we can create our expertise
Machine learning
Deep learning
Robotics

Expert systems
Fuzzy logic
Natural language processing
Computer vision

14(b) Write a short note on different types of Machine Learning Algorithms

Ans: ***Supervised learning*** is one of the most basic types of machine learning. In this type, the machine learning algorithm is trained on labeled data. Even though the data needs to be labeled accurately for this method to work, supervised learning is extremely powerful when used in the right circumstances.

In supervised learning, the ML algorithm is given a small training dataset to work with. This training dataset is a smaller part of the bigger dataset and serves to give the algorithm a basic idea of the problem, solution, and data points to be dealt with. The training dataset is also very similar to the final dataset in its characteristics and provides the algorithm with the labeled parameters required for the problem.

The algorithm then finds relationships between the parameters given, essentially establishing a cause and effect relationship between the variables in the dataset. At the end of the training, the algorithm has an idea of how the data works and the relationship between the input and the output.

This solution is then deployed for use with the final dataset, which it learns from in the same way as the training dataset. This means that supervised machine learning algorithms will continue to improve even after being deployed, discovering new patterns and relationships as it trains itself on new data.

Unsupervised machine learning holds the advantage of being able to work with unlabeled data. This means that human labor is not required to make the dataset machine-readable, allowing much larger datasets to be worked on by the program.

In supervised learning, the labels allow the algorithm to find the exact nature of the relationship between any two data points. However, unsupervised learning does not have labels to work off of, resulting in the creation of hidden structures. Relationships between data points are perceived by the algorithm in an abstract manner, with no input required from human beings.

The creation of these hidden structures is what makes unsupervised learning algorithms versatile. Instead of a defined and set problem statement, unsupervised learning algorithms can adapt to the data by dynamically changing hidden structures. This offers more post-deployment development than supervised learning algorithms.

Reinforcement learning directly takes inspiration from how human beings learn from data in their lives. It features an algorithm that improves upon itself and learns from new situations using a trial-and-error method. Favorable outputs are encouraged or 'reinforced', and non-favorable outputs are discouraged or 'punished'.

Based on the psychological concept of conditioning, reinforcement learning works by putting the algorithm in a work environment with an interpreter and a reward system. In every iteration of the algorithm, the output result is given to the interpreter, which decides whether the outcome is favorable or not.

In case of the program finding the correct solution, the interpreter reinforces the solution by providing a reward to the algorithm. If the outcome is not favorable, the algorithm is forced to reiterate until it finds a better result. In most cases, the reward system is directly tied to the effectiveness of the result.

In typical reinforcement learning use-cases, such as finding the shortest route between two points on a map, the solution is not an absolute value. Instead, it takes on a score of effectiveness, expressed in a percentage value. The higher this percentage value is, the more reward is given to the algorithm. Thus, the program is trained to give the best possible solution for the best possible reward

15(a) Define intelligent systems? Discuss the foundations of AI?

Ans: Foundations of AI commonly used techniques and theories are rule-based Fuzzy Logic neural networks station theory statistics probability theory genetic algorithm etc. since AI is interdisciplinary in nature foundation of a in various fields such as • Mathematics • Neurosciences • Control theory • Linguistics Mathematics AI system use formal logic method and Boolean logic analysis of limits to what can be computed, probability theory ,uncertainty that forms the basic for most modern approaches to Ai fuzzy logic etc Neuroscience this science of medicine helps in studying the functioning of brains. in early studies injured and abnormal people were used to understand what parts of brain work .now recent studies use accurate sensors to correlate brain activity to human thought. by monitoring individual neuron monkeys can now control a computer mouse using thought alone. Moore's law state that the computers will have as many gates as human have neurons in the year 2020 .researchers are working to low as to how to have your mechanical brain. such systems will require parallel computation remapping and interconnection to a large extent Control Theory Mission can modify their behavior In response to the environment steam engine . example water flow regulator this theory of stable feedback system helps in building systems the transition from initial state two goals it with minimum energy Linguistics speech demonstrates so much of human intelligence. analysis of human language reveal thought taking place in a in ways not understood in other settings. children can create sentences they have never heard before. languages and thoughts believed to be tightly intertwined

15(b) List and give brief description about basic machine learning models?

Ans:

Machine Learning models can be understood as a program that has been trained to find patterns within new data and make predictions. These models are represented as a mathematical function that takes requests in the form of input data, makes predictions on input data, and then provides an output in response. First, these models are trained over a set of data, and then they are provided an algorithm to reason over data, extract the pattern from feed data and learn from those data. Once these models get trained, they can be used to predict the unseen dataset.

There are various types of machine learning models available based on different business goals and data sets.

Machine learning models:

Supervised Machine Learning Models

Supervised Learning is the simplest machine learning model to understand in which input data is called training data and has a known label or result as an output. So, it works on the principle of input-output pairs. It requires creating a function that can be trained using a training data set, and then it is applied to unknown data and makes some predictive performance. Supervised learning is task-based and tested on labeled data sets.

We can implement a supervised learning model on simple real-life problems. For example, we have a dataset consisting of age and height; then, we can build a supervised learning model to predict the person's height based on their age.

Supervised Learning models are further classified into two categories:

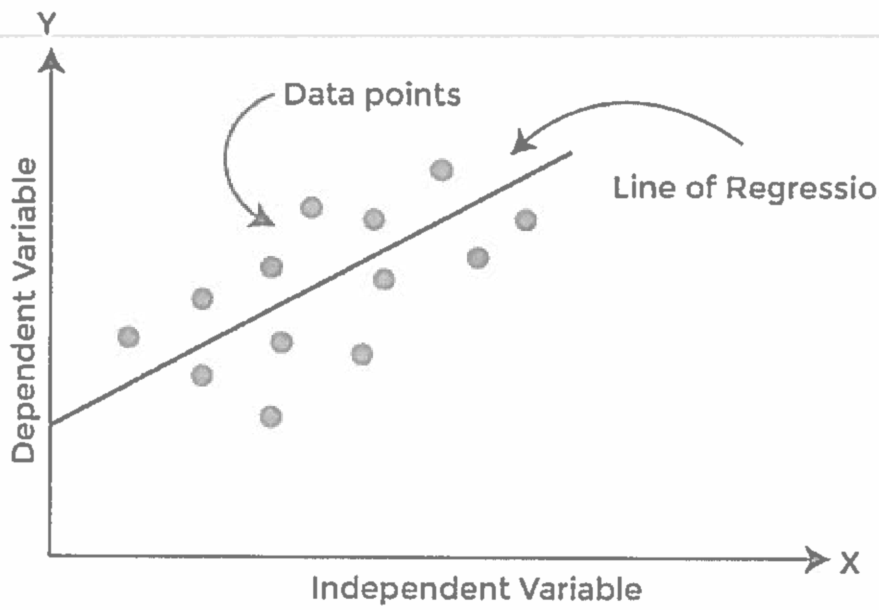
Regression

In regression problems, the output is a continuous variable. Some commonly used Regression models are as follows:

a) Linear Regression

Linear regression is the simplest machine learning model in which we try to predict one output variable using one or more input variables. The representation of linear regression is a linear equation, which combines a set of input values(x) and predicted output(y) for the set of those input values. It is represented in the form of a line:

Y = bx+ c.



The main aim of the linear regression model is to find the best fit line that best fits the data points.

Linear regression is extended to multiple linear regression (find a plane of best fit) and polynomial regression (find the best fit curve).

b) Decision Tree

Decision trees are the popular machine learning models that can be used for both regression and classification problems.

A decision tree uses a tree-like structure of decisions along with their possible consequences and outcomes. In this, each internal node is used to represent a test on an attribute; each branch is used to represent the outcome of the test. The more nodes a decision tree has, the more accurate the result will be.

The advantage of decision trees is that they are intuitive and easy to implement, but they lack accuracy.

Decision trees are widely used in **operations research, specifically in decision analysis, strategic planning**, and mainly in machine learning.

c) Random Forest

Random Forest is the ensemble learning method, which consists of a large number of decision trees. Each decision tree in a random forest predicts an outcome, and the prediction with the majority of votes is considered as the outcome.

A random forest model can be used for both regression and classification problems.

For the classification task, the outcome of the random forest is taken from the majority of votes. Whereas in the regression task, the outcome is taken from the mean or average of the predictions generated by each tree.

d) Neural Networks

Neural networks are the subset of machine learning and are also known as artificial neural networks. Neural networks are made up of artificial neurons and designed in a way that resembles the human brain structure and working. Each artificial neuron connects with many other neurons in a neural network, and such millions of connected neurons create a sophisticated cognitive structure.

Neural networks consist of a multilayer structure, containing one input layer, one or more hidden layers, and one output layer. As each neuron is connected with another neuron, it transfers data from one layer to the other neuron of the next layers. Finally, data reaches the last layer or output layer of the neural network and generates output.

Neural networks depend on training data to learn and improve their accuracy. However, a perfectly trained & accurate neural network can cluster data quickly and become a powerful machine learning and AI tool. One of the best-known neural networks is **Google's search algorithm.**

Classification

Classification models are the second type of Supervised Learning techniques, which are used to generate conclusions from observed values in the categorical form. For example, the classification model can identify if the email is spam or not; a buyer will purchase the product or not, etc. Classification algorithms are used to predict two classes and categorize the output into different groups.

In classification, a classifier model is designed that classifies the dataset into different categories, and each category is assigned a label.

There are two types of classifications in machine learning:

- **Binary classification**: If the problem has only two possible classes, called a binary classifier. For example, cat or dog, Yes or No,
- **Multi-class classification**: If the problem has more than two possible classes, it is a multi-class classifier.

Some popular classification algorithms are as below:

a) Logistic Regression

Logistic Regression is used to solve the classification problems in machine learning. They are similar to linear regression but used to predict the categorical variables. It can predict the output in either Yes or No, 0 or 1, True or False, etc. However, rather than giving the exact values, it provides the probabilistic values between 0 & 1.

b) Support Vector Machine

Support vector machine or SVM is the popular machine learning algorithm, which is widely used for classification and regression tasks. However, specifically, it is used to solve classification problems. The main aim of SVM is to find the best decision boundaries in an N-dimensional space, which can segregate data points into classes, and the best decision boundary is known as Hyperplane. SVM selects the extreme vector to find the hyperplane, and these vectors are known as support vectors.

c) Naïve Bayes

Naïve Bayes is another popular classification algorithm used in machine learning. It is called so as it is based on Bayes theorem and follows the naïve(independent) assumption between the features which is given as:

$$P(y|X) = \frac{P(X|y) * P(y)}{P(X)}$$

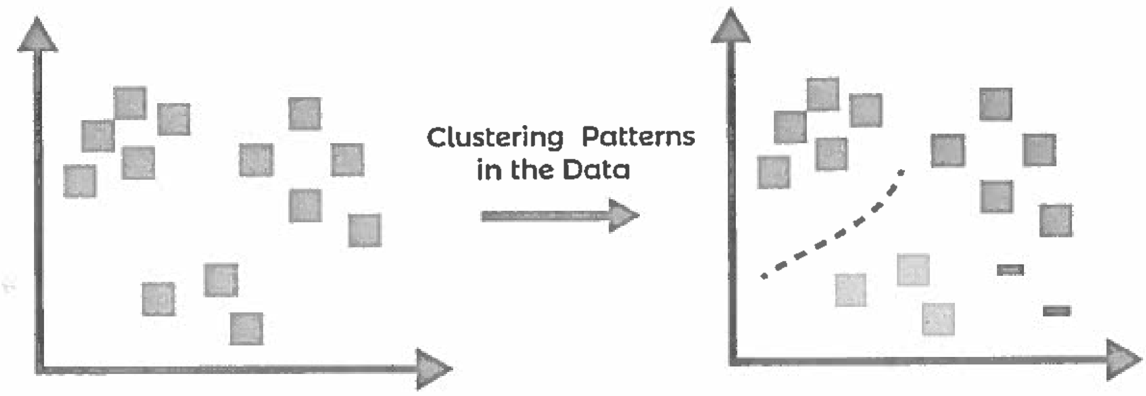
Each naïve Bayes classifier assumes that the value of a specific variable is independent of any other variable/feature. For example, if a fruit needs to be classified based on color, shape, and taste. So yellow, oval, and sweet will be recognized as mango. Here each feature is independent of other features.

2. Unsupervised Machine learning models

Unsupervised Machine learning models implement the learning process opposite to supervised learning, which means it enables the model to learn from the unlabeled training dataset. Based on the unlabeled dataset, the model predicts the output. Using unsupervised learning, the model learns hidden patterns from the dataset by itself without any supervision.

Unsupervised learning models are mainly used to perform three tasks, which are as follows:

- **Clustering**
  Clustering is an unsupervised learning technique that involves clustering or groping the data points into different clusters based on similarities and differences. The objects with the most similarities remain in the same group, and they have no or very few similarities from other groups. Clustering algorithms can be widely used in different tasks such as **Image segmentation, Statistical data analysis, Market segmentation,** etc.
  Some commonly used Clustering algorithms are *K-means Clustering, hierarchal Clustering, DBSCAN*, etc.



- **Association Rule Learning**
  Association rule learning is an unsupervised learning technique, which finds interesting relations among variables within a large dataset. The main aim of this learning algorithm is to find the dependency of one data item on another data item and map those variables accordingly so that it can generate maximum profit. This algorithm is mainly applied in **Market Basket analysis, Web usage mining, continuous production**, etc.
  Some popular algorithms of Association rule learning are ***Apriori Algorithm, Eclat, FP-growth algorithm.***
- **Dimensionality Reduction**
  The number of features/variables present in a dataset is known as the dimensionality of the dataset, and the technique used to reduce the dimensionality is known as the dimensionality reduction technique. Although more data provides more accurate results, it can also affect the performance of the model/algorithm, such as overfitting issues. In such cases, dimensionality reduction techniques are used.
  "***It is a process of converting the higher dimensions dataset into lesser dimensions dataset ensuring that it provides similar information.***"
  Different dimensionality reduction methods such ***as PCA(Principal Component Analysis), Singular Value Decomposition, etc.***

Reinforcement Learning

In reinforcement learning, the algorithm learns actions for a given set of states that lead to a goal state. It is a feedback-based learning model that takes feedback signals after each state or action by interacting with the environment. This feedback works as a reward (positive for each good action and negative for each bad action), and the agent's goal is to maximize the positive rewards to improve their performance.

The behavior of the model in reinforcement learning is similar to human learning, as humans learn things by experiences as feedback and interact with the environment.

Below are some popular algorithms that come under reinforcement learning:

- **Q-learning:** Q-learning is one of the popular model-free algorithms of reinforcement learning, which is based on the Bellman equation.

It aims to learn the policy that can help the AI agent to take the best action for maximizing the reward under a specific circumstance. It incorporates Q values for each state-action pair that indicate the reward to following a given state path, and it tries to maximize the Q-value.

- **State-Action-Reward-State-Action (SARSA):** SARSA is an On-policy algorithm based on the Markov decision process. It uses the action performed by the current policy to learn the Q-value. The SARSA algorithm stands **for State Action Reward State Action, which symbolizes the tuple (s, a, r, s', a').**
- **Deep Q Network:** DQN **or Deep Q Neural network is Q-learning** within the neural network. It is basically employed in a big state space environment where defining a Q-table would be a complex task. So, in such a case, rather than using Q-table, the neural network uses Q-values for each action based on the state.

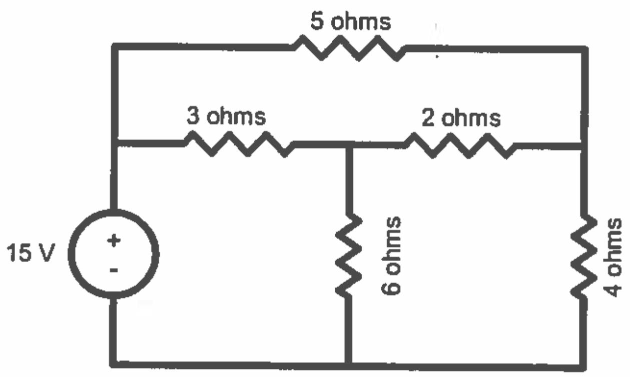
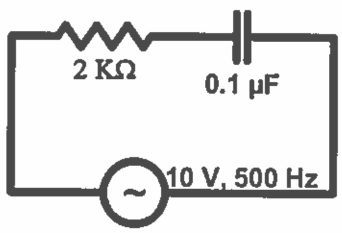
Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	ECE & EEE			Academic Year	2021 - 2022
Course Code	20ESX03	Test Duration	3 Hrs.	Max. Marks	70	Semester	I
Course	BASIC ELECTRICAL ENGINEERING						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define power factor.	20ESX03.1	L1
2	Draw OCC of a DC shunt generator.	20ESX03.2	L1
3	Define voltage regulation	20ESX03.3	L1
4	Define slip in an induction motor.	20ESX03.4	L1
5	Write any one application of a single phase induction motor.	20ESX03.5	L1

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

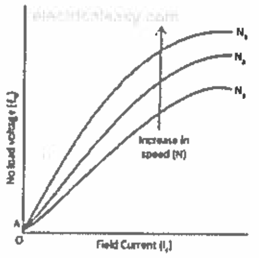
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the following with an example i) Kirchoff's current law ii) Kirchoff's voltage law iii) Voltage division rule.	6M	20ESX03.1	L2
6 (b)	Find the current flowing through 4Ω resistance 	6M	20ESX03.1	L3
	OR			
7 (a)	Define the following with respect to sinusoidal quantity: i) RMS Value ii) Average Value iii) Form factor iv) Peak factor	8M	20ESX03.1	L2
7 (b)	For the circuit shown, determine total impedance Z and current I 	4M	207ESX03.1	L3
8	Describe the constructional details of a DC generator and write the function of each part with a neat sketch	12M	20ESX03.2	L2
	OR			
9	A 6 pole DC generator with 770 wave connected armature	12M	20ESX03.2	L2

	conductors and running at 500 rpm supplies a load of 12.5Ω resistance at a terminal voltage of 250V. The armature resistance is 0.25Ω and field resistance is 250Ω. Find the armature current, induced EMF and Flux per pole.			
10	A 230 V/230 V, 3 KVA transformer gave the following results: O.C test : 230 V, 2A, 100 W S.C test: 15 V, 13 A, 120 W Determine the regulation and efficiency at full load 0.8 p.f lagging	12M	20ESX03.3	L3
	OR			
11	Why parallel operation of transformers is required? Explain the parallel operation of transformers.	12M	20ESX03.3	L2
12	Derive the EMF equation of an alternator.	12M	20ESX03.4	L3
	OR			
13	How will the rotor of a 3-Φ induction motor rotate?	12M	20ESX03.4	L3
14	Compare the 1-Phase and 3-Phase Induction Motors.	12M	20ESX03.5	L3
	OR			
15	Explain the principle of operation and construction of a single-phase induction motor	12M	20ESX03.5	L3

BASIC ELECTRICAL ENGINEERING SUPPLY EXAM KEY (JUNE 2022)

1. Define power factor ?
A. Power factor (PF) is the ratio of working power, measured in kilowatts (kW), to apparent power, measured in kilovolt amperes (kVA). Apparent power, also known as demand, is the measure of the amount of power used to run machinery and equipment during a certain period. It is found by multiplying (kVA = V x A).

2. Draw OCC of DC shunt generator?
A. The open circuit characteristics (O.C.C) or magnetization characteristics is the curve that shows the relationship between the generated EMF at no-load (E0) and the field current (If) at constant speed. It is also known as no-load saturation curve.



Open Circuit Characteristic (O.C.C.)

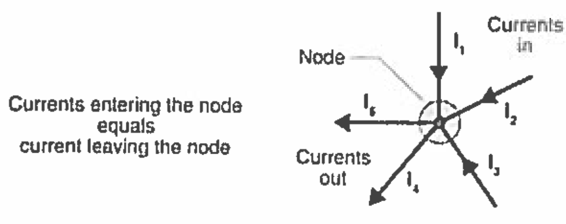
3. Define voltage regulation ?
A. The voltage regulation of the transformer is the percentage change in the output voltage from no-load to full-load. And since power factor is a determining factor in the secondary voltage, power factor influences voltage regulation. This means the voltage regulation of a transformer is a dynamic, load-dependent number

4. Define slip in an induction motor?
A. In an electrical coupling, slip is defined simply as the difference between the speeds of the two rotating members. In an induction motor, slip is a measure of the difference between the machine's synchronous speed and its shaft speed

Slip S = (Ns - N)/Ns

5. Write any one applications of single phase induction motor ?
A. These motors find use in fans, refrigerators, Air-conditioners, Vacuum cleaners, washing machines, centrifugal pumps, tools, small farming appliances, blowers etc...

6a. Kirchoff's current law :
B. Kirchhoff's Current Law goes by several names: Kirchhoff's First Law and Kirchhoff's Junction Rule. According to the Junction rule, the total of the currents in a junction is equal to the sum of currents outside the junction in a circuit.
C. The total current entering a junction or a node is equal to the charge leaving the node as no charge is lost.



$$I_1 + I_2 + I_3 + (-I_4 + -I_5) = 0$$

In the above figure, the currents I1, I2 and I3 entering the node is considered positive, likewise, the currents I4 and I5 exiting the nodes is considered negative in values. This can be expressed in the form of an equation:

$I1 + I2 + I3 - I4 - I5 = 0$

A node refers to a junction connecting two or more current-carrying routes like cables and other components. Kirchhoff's current law can also be applied to analyse parallel circuits.

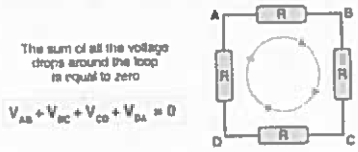
Kirchhoff's Second Law or Kirchhoff's Voltage Law:

Racnode
HOD-EEE
23/06/22

According to Kirchhoff's Voltage Law,

The voltage around a loop equals the sum of every voltage drop in the same loop for any closed network and equals zero.

Put differently, the algebraic sum of every voltage in the loop has to be equal to zero and this property of Kirchhoff's law is called conservation of energy.



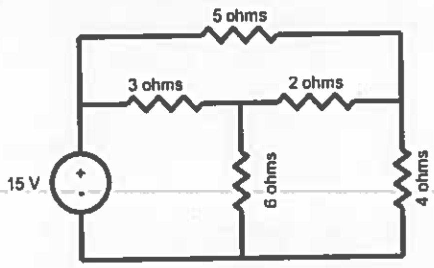
Voltage Division rule:
Voltage division rule is one of the basic rules of circuit analysis. It is applicable to all series circuits and combination circuits. Series circuit always acts as a Voltage divider. In a series circuit, the same current flows through each resistance. Hence the voltage drops across each resistor are proportional to their ohmic value. In a series circuit, current flowing will be proportional to the total resistance offered by the circuit.

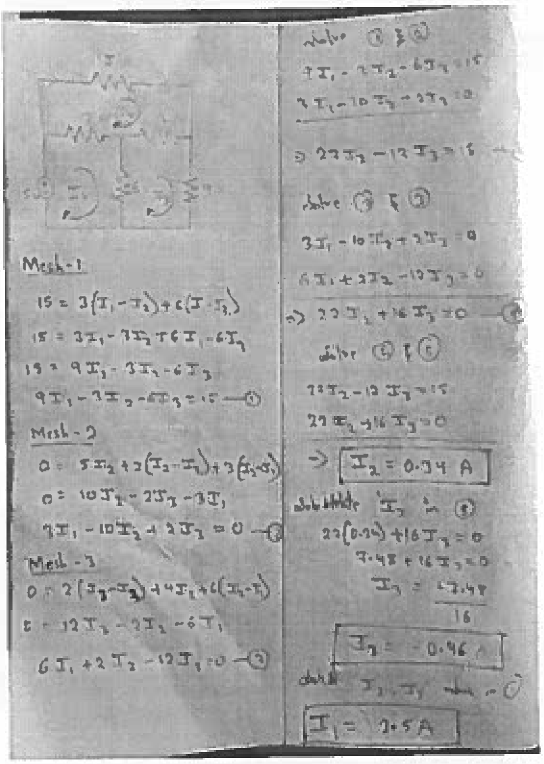
The voltage across each resistor, by ohm's law, is the current flowing through the circuit multiplied by the value of resistance. Consider a circuit with 'n' number of resistors connected in series and with a voltage V applied across it.

Voltage across R1, V1 = I.R1 = V.R1 / (R1+R2)

Voltage across R2, V2 = I.R2 = V.R2 / (R1+R2)

6b. Find current through 4 ohm resistance




Mesh-1
$15 = 3(I_1 - I_2) + 6(I_1 - I_3)$
$15 = 3I_1 - 3I_2 + 6I_1 - 6I_3$
$15 = 9I_1 - 3I_2 - 6I_3$
$9I_1 - 3I_2 - 6I_3 = 15$ — (1)
Mesh-2
$0 = 5I_2 + 2(I_2 - I_3) + 3(I_2 - I_1)$
$0 = 10I_2 - 2I_3 - 3I_1$
$3I_1 - 10I_2 + 2I_3 = 0$ — (2)
Mesh-3
$0 = 2(I_3 - I_2) + 4I_3 + 6(I_3 - I_1)$
$0 = 12I_3 - 2I_2 - 6I_1$
$6I_1 + 2I_2 - 12I_3 = 0$ — (3)

solve (1) & (2)
$9I_1 - 3I_2 - 6I_3 = 15$
$3I_1 - 10I_2 + 2I_3 = 0$
$\Rightarrow 27I_2 - 12I_3 = 15$
solve (2) & (3)
$3I_1 - 10I_2 + 2I_3 = 0$
$6I_1 + 2I_2 - 12I_3 = 0$
$\Rightarrow 22I_2 + 16I_3 = 0$ — (5)
solve (4) & (5)
$27I_2 - 12I_3 = 15$
$22I_2 + 16I_3 = 0$
$\Rightarrow \boxed{I_2 = 0.34\ A}$
substitute I_2 in (5)
$22(0.34) + 16I_3 = 0$
$7.48 + 16I_3 = 0$
$I_3 = \frac{-7.48}{16}$
$\boxed{I_3 = -0.46\ A}$
sub I_2, I_3 value in (1)
$\boxed{I_1 = 2.5\ A}$

7a.

RMS VALUE:

The term "RMS" stands for "Root-Mean-Squared". Mostly define this as the "amount of AC power that produces the same heating effect as an equivalent DC power", or something similar along these lines, but an RMS value is more than just that.
The RMS value is the square root of the mean (average) value of the squared function of the instantaneous values. The symbols used for defining an RMS value are VRMS or IRMS

The term RMS, refers to time-varying sinusoidal voltages, currents or complex waveforms were the magnitude of the waveform changes over time and is not used in DC circuit analysis or calculations were the magnitude is always constant.

When used to compare the equivalent RMS voltage value of an alternating sinusoidal waveform that supplies the same electrical power to a given load as an equivalent DC circuit, the RMS value is called the "effective value" and is generally presented as: Veff or Ieff.

In other words, the effective value is an equivalent DC value which tells you how many volts or amps of DC that a time-varying sinusoidal waveform is equal to in terms of its ability to produce the same power.

Each mid-ordinate value of a waveform (the voltage waveform in this case) is multiplied by itself (squared) and added to the next. This method gives us the "square" or Squared part of the RMS voltage expression.

Then we can define the term used to describe an rms voltage (VRMS) as being "the square root of the mean of the square of the mid-ordinates of the voltage waveform"

$$V_{rms} = \sqrt{V_0^2 + \frac{1}{2}(V_{c1}^2 + V_{c2}^2 + \cdots) + \frac{1}{2}(V_{s1}^2 + V_{s2}^2 + \cdots)}$$

Average value :

AVERAGE VALUE OF PERIODIC WAVEFORM

In general, the average value of any function v(t), with period T is given by

$$v_{av} = \frac{1}{T}\int_0^T v(t)\,dt$$

That means that the average value of a curve in the X- Y plane is the total area under the complete curve divided by the distance of the curve. The average value of a sine wave over one complete cycle is always zero. So the average value of a sine wave is defined over a half-cycle, and not a full cycle period.

The average value of the sine wave is the total area under the half-cycle curve divided by the distance of the curve.

The average value of the sine wave

$v(t) = V_p \sin \omega t$ **is given by**

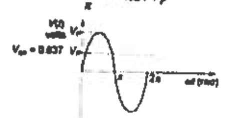
$$\text{Vav} = \frac{\text{Area under curve for half cycle}}{\text{Length of base over half cycle}}$$

$$= \frac{\int_0^{\pi} V\,d\theta}{\pi}$$

$$v_{av} = \frac{1}{\pi}\int_0^{\pi} V_p \sin \omega t\, d(\omega t)$$

$$= \frac{1}{\pi}[-V_p \cos \omega t]_0^{\pi}$$

$$= \frac{2V_p}{\pi} = 0.637\,V_p$$



Importance of Average Value:

1. The average value is used for applications like battery charging.
2. The charge transferred in capacitor circuits is measured using average values.
3. The average values of voltages and currents play an important role in analysis of the rectifier circuits.
4. The average value is indicated by d.c. ammeters and voltmeters.
5. The average value of purely sinusoidal waveform is always zero.

Form factor :

PEAK(CREST) FACTOR OF PERIODIC WAVEFORM

The peak factor of any waveform is defined as the ratio of the peak value of the wave to the rms value of the wave.

$$\text{Peak factor} = \frac{V_p}{V_{rms}}$$

$$\text{Peak factor of the sinusoidal waveform} = \frac{V_p}{V_p/\sqrt{2}} = \sqrt{2} = 1.414$$

Peak factor:

FORM FACTOR OF PERIODIC WAVEFORM

Form factor of a waveform is defined as the ratio of rms value to the average value of the wave.

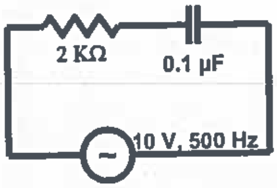
$$\text{Form factor} = \frac{rms \text{ value of the wave}}{\text{Average value of the wave}}$$

$$\text{Form factor} = \frac{V_{rms}}{V_{av}}$$

Form factor of a sinusoidal waveform can be found from the above relation.

$$\text{For the sinusoidal wave, the form factor} = \frac{V_p/\sqrt{2}}{0.637\,V_p} = 1.11$$

7b. For the circuit shown, determine total impedance Z and current I



Given $C = 0.1\,\mu F$

$R = 2k\Omega$

$V = 10V$

$X_C = \frac{1}{2\pi fC}$

$X_C = \frac{1}{2\pi \times 500 \times 0.1 \times 10^{-6}}$

$X_C = \frac{1}{31.4 \times 10^{-6}} = 318$

$Z = R + jX_C$

$Z = \sqrt{R^2 + X_C^2} = \sqrt{(2000)^2 + (318.17)^2}$

$= \sqrt{4000000 + 101124} = \sqrt{4101124} = 2025.1\,\Omega$

8a. ***Construction of a DC Machine:***

A DC generator can be used as a DC motor without any constructional changes and vice versa is also possible. Thus, a DC generator or a DC motor can be broadly termed as a DC machine. These basic constructional details are also valid for the construction of a DC motor. Hence, let's call this point as construction of a DC machine instead of just 'construction of a DC generator.

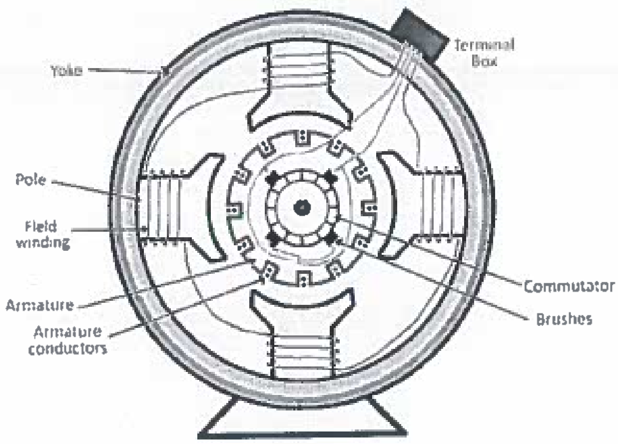


Figure : constructional details of a simple 4-pole DC machine.

The above figure shows constructional details of a simple 4-pole DC machine. A DC machine consists of two basic parts; stator and rotor. Basic constructional parts of a DC machine are described below.

1. ***Yoke:*** The outer frame of a dc machine is called as yoke. It is made up of cast iron or steel. It not only provides mechanical strength to the whole assembly but also carries the magnetic flux produced by the field winding.

2. Poles and pole shoes: Poles are joined to the yoke with the help of bolts or welding. They carry field winding and pole shoes are fastened to them. Pole shoes serve two purposes; (i) they support field coils and (ii) spread out the flux in air gap uniformly.

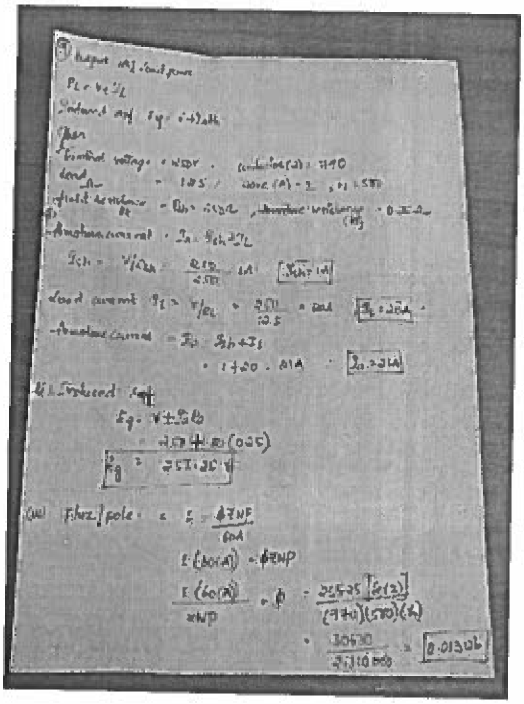
3. Field winding: They are usually made of copper. Field coils are former wound and placed on each pole and are connected in series. They are wound in such a way that, when energized, they form alternate North and South poles.

4. Armature core: Armature core is the rotor of a dc machine. It is cylindrical in shape with slots to carry armature winding. The armature is built up of thin laminated circular steel disks for reducing eddy current losses. It may be provided with air ducts for the axial air flow for cooling purposes. Armature is keyed to the shaft.

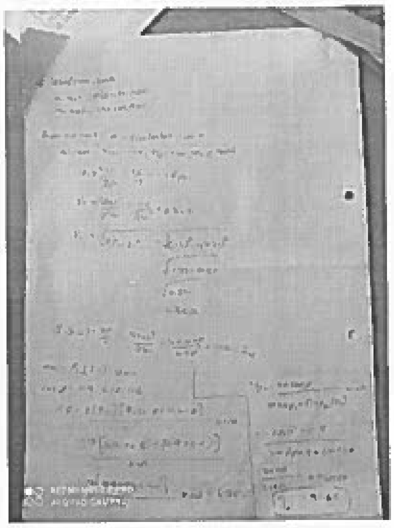
5. Armature winding: It is usually a former wound copper coil which rests in armature slots. The armature conductors are insulated from each other and also from the armature core. Armature winding can be wound by one of the two methods; lap winding or wave winding. Double layer lap or wave windings are generally used. A double layer winding means that each armature slot will carry two different coils.

6. Commutator and brushes: Physical connection to the armature winding is made through a commutator-brush arrangement. The function of a commutator, in a dc generator, is to collect the current generated in armature conductors. Whereas, in case of a dc motor, commutator helps in providing current to the armature conductors. A commutator consists of a set of copper segments which are insulated from each other. The number of segments is equal to the number of armature coils. Each segment is connected to an armature coil and the commutator is keyed to the shaft. Brushes are usually made from carbon or graphite. They rest on commutator segments and slide on the segments when the commutator rotates keeping the physical contact to collect or supply the current.

9. 6 pole DC generator with 770 wave connected armature conductors and running at 500 rpm supplies a load of 12.5Ω resistance at a terminal voltage of 250V. The armature resistance is 0.25Ω and field resistance is 250Ω. Find the armature current, induced EMF and Flux per pole.



10. 230 V/230 V, 3 KVA transformer gave the following results:
O.C test : 230 V, 2A, 100 W
S.C test: 15 V, 13 A, 120 W Determine the regulation and efficiency at full load 0.8 p.f lagging



11. Parallel operation of transformers:

The reasons for operating the transformers in parallel are as follows:

1. This is an economical method because a single large transformer is uneconomical for large load.
2. If the transformers are connected in parallel, we require extra load then we can expand the system by adding more transformers in the future.
3. Parallel operation reduces the space capacity of the substation when we connect transformers of standard size.
4. The parallel connection maximizes the electrical power system availability as we can shut down any system for maintenance without affecting other systems performance.

a_1 = turns ratio of transformer A

a_2 = turns ratio of transformer B

Z_A = equivalent impedance of transformer A referred to the secondary side.

Z_B = equivalent impedance of transformer B referred to the secondary side.

Z_L = load impedance across the secondary side.

I_A = current supplied to the load by the secondary of transformer A.

I_B = current supplied to the load by the secondary of transformer B.

V_L= load secondary voltage.

I_L= load current

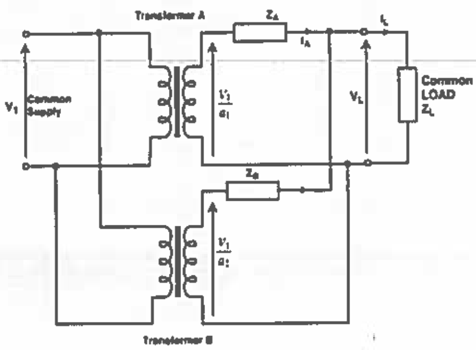


Fig: Two single-phase transformers in parallel.

By KCL,

$I_A+I_B=I_L$

BY KVL,

$$V_L = \frac{V_1}{a_1} - I_A Z_A$$

$$V_L = \frac{V_1}{a_2} - I_B Z_B = \frac{V_1}{a_2} - (I_L - I_A) Z_B$$

By solving the above two equations, we get

$$I_A = \frac{Z_B I_L}{Z_A + Z_B} + \frac{V_1(a_2 - a_1)}{a_1 a_2 (Z_A + Z_B)}$$

$$I_B = \frac{Z_A I_L}{Z_A + Z_B} - \frac{V_1(a_2 - a_1)}{a_1 a_2 (Z_A + Z_B)}$$

Each of these currents has two components; the first component represents the transformer's share of the load current and the second component is a circulating current in the secondary windings.

12. EMF equation of alternator:

EMF Equation of an Alternator :
When the rotor rotates, the stator conductors which are static in case of alternator cut by magnetic flux , they have induced EMF produced in them (according to **Faraday's law of electromagnetic induction** which states that if a conductor or coil links with any changing flux, there must be an induced EMF in it.
This induced EMF can be found by the **EMF equation of the alternator** which as follow:
Lets,
P = No. of poles
Z = No. of conductors or Coil sides in series/phase i.e. **Z** = **2T**…Where T is the number of coils or turns per phase (Note that one turn or coil has two ends or sides)
f = frequency of induced EMF in Hz
Φ = Flux per pole (Weber)
N = rotor speed (RPM)
Kd= Distribution factor =
Kc or **KP** = **Cos α/2**
If induced EMF is assumed sinusoidal then,
K*f* = Form factor = 1.11
In one revolution of the rotor i.e. in **60/N** seconds, each conductor is cut by a flux of **ΦP** Webers.
$d\Phi = \Phi P$ and also $d\Phi$ = 60/N seconds
then induced E.M.F per conductor (average)
….. *(i)* But we know that:
***f* = PN / 120** or **N= 120*f* / P**
Putting the value of N in **Equation *(i)***, we get,

Average value of EMF per conductor
If there are **Z** conductors in series per phase,
Then synchronous generator **average E.M.F per phase = 2 *f* Φ Z** Volts = **4 *f* ΦT** *Volts* ….. (Z = 2T)
Also we know that;
Form Factor = RMS Value / Average Value
RMS value = Form Factor x Average Value,
VAV = 1.11 x 4*f*ΦT = 4.44*f*ΦT Volts.
(Note that is exactly the same equation as the EMF equation of the transformer)
And the actual available voltage of generator per phase
VPH = 4.44 KC KD *f* ΦT
V = **4K*f* KC KD *f* ΦT Volts.**
Where:
- V = Actual generated Voltage per phase
- **KC** = Coil Span Factor or Pitch Factor
- **KD** = Distribution Factor
- K*f* = Form Factor
- *f* = frequency
- T = Number of coils or number of turns per phase

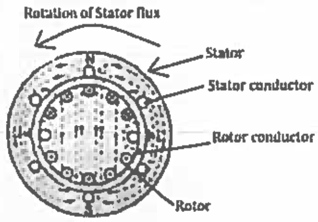
Note: If alternator or AC generator is star connected as usually the case, then the Line Voltage is √3 times the phase voltage as derived from the above equation.

13. Operation of 3 phase induction motor:

Principle of operation of 3-Φ Induction motors :

When the stator of an induction motor is connected across a balanced three phase AC supply, it draws a balanced three-phase current and these current sets up a rotating magnetic field of constant magnitude.

The magnetic flux rotating, at the synchronous speed, sweeps past the rotor conductors which are stationary at the start and induces emfs in them. The frequency of the induced emfs in the rotor conductors is same as the supply frequency.



(a)

As the rotor conductors are short circuited, the induced emfs produce three phase rotor currents which in turn produce a magnetic field that revolves at the same speed as the stator field. A starting torque is produced due to the interaction of these two fields and tends to turn the rotor in the direction of rotation of the stator field.

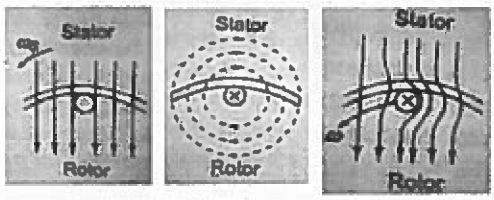
According to Lenz's law, the developed torque must oppose the cutting of flux lined by rotor conductors. Hence, the developed torque makes the rotor move in the direction of flux waves so as to reduce the relative speed between the stator flux wave and the rotor conductors and thereby reducing the cutting of flux lines by the rotor conductors.

If the starting torque is sufficient to overcome the load torque on the motor shaft, the motor will start rotating and reach its operating speed. The operating speed can never be equal to the synchronous speed N_s i.e. the rotor speed N of an induction motor is always slightly less than the synchronous speed N_s.

The distribution of flux of an induction motor is shown in the fig-(a).

If the stator winding flux rotates in anti-clockwise at an angular velocity $\omega_s = 2\pi f$, the direction of the induced emf generated in the stationary rotor conductors can be determined by Fleming's right hand rule.

The emf generated in one rotor conductor is shown in fig-(b).

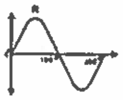
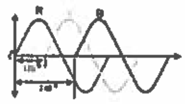
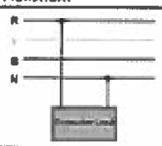
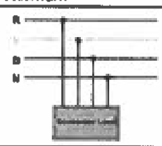


(b)

In the fig-(b), the first part shows the stator flux rotating at an angular ω_s and links the rotor conductor. The second part shows that the induced current circulating in the rotor conductor produces a flux around it in the clockwise direction which is determined by Maxwell's corkscrew rule. The third part shows the effect of the flux which strengthens the flux density on the right hand side and weakens that on the left hand side of the conductor.

Thus causes the conductor to be pushed towards the left. Thus the rotor also begins to rotate in the anti-clockwise direction i.e. in the direction of rotating magnetic field. Thus, the induction motors are self-starting.

14. Comparison of single phase and three phase induction motor:

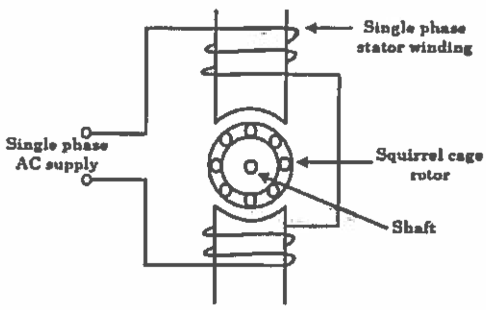
Basis for Comparison	Single Phase	Three Phase
Definition	The power supply through one conductor.	The power supply through three conductors.
Wave Shape		
Number of wire	Require two wires for completing the circuit	Requires four wires for completing the circuit
Voltage	Carry 230V	Carry 415V
Phase Name	Split phase	No other name
Network	Simple	Complicated
Loss	Maximum	Minimum
Power Supply Connection		
Efficiency	Less	High
Economical	Less	More
Uses	For home appliances.	In large industries and for running heavy loads.

15. principle of operation and construction of a single-phase induction motor:

Single phase induction motor construction:
□ Single phase induction motors consist of two main parts;
stator and rotor.

The construction of these motors is more or less similar to a three-phase squirrel-cage induction motor.

□ The stator is a stationary part and it has laminated construction, which is made up of stampings. These stampings consists of slots on its periphery to carry the stator winding. This winding is excited with a single phase AC supply.
□ The rotor is a rotating part and its construction is of squirrel cage type. The rotor consists of un insulated aluminum or copper bars which are placed in the slots.
□ These rotor bars are permanently shorted at both ends with the help of end rings as shown in figure.
□ There is no physical connection between the stator and rotor, but there is a small and uniform gap between them.
□ The rotor acts as a conductor which when placed in the stator magnetic field, an emf is induced in it, produces its own magnetic field which further interacts with stator field to produce the torque.



Principle of operation of single phase induction motor:
□ Whenever a single phase AC supply is given to the stator winding, an alternating magnetic field is produced around the stator.

- Due to the pulsating nature of the field which reverses for every half-cycle, cannot produce rotation in a stationary squirrel cage rotor.
- In case of three phase induction motor, the field produced by the supply is of rotating type and hence they are self starting motors.
- But in case of single phase motors, the field produced by the stator is not rotating (but alternating only) and hence single phase motors are not self starting.
- But, if the rotor is rotated by any other means (by hand or any tool), the induced currents in the rotor will assist with stator currents to produce revolving field. This field causes the motor to run in the direction it is started even with a single winding.
- However, it is not possible to give initial rotation every time externally if the motors are attached to loads. This problem can be avoided by converting single phase motor into a two-phase motor temporarily in order produce revolving flux. This is achieved by providing a starting winding in addition to main or running winding.
- The auxiliary or starting winding is made highly resistive whereas the main or running winding is made highly inductive.
- Due to the large phase difference between these two, the torque produced by the rotor is high enough to start it. Once the motor reaches 75 percent of its speed, the auxiliary winding may be disconnected by a centrifugal switch and the motor able to run on a single main winding.
- Single phase induction motors are used primarily for domestic and light-industrial applications where three-phase supply is generally not available.

Semester End Supplementary Examination, June, 2022

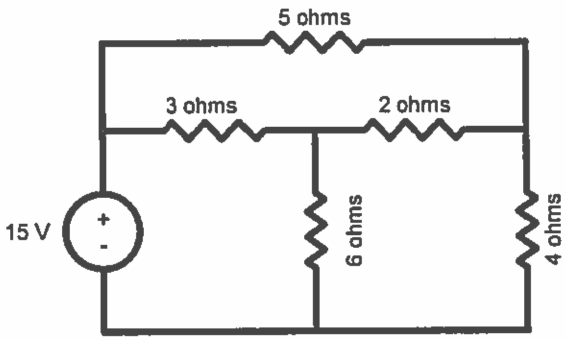
Degree	B. Tech. (U. G.)	**Program**	ECE & EEE	**Academic Year**	2021 - 2022
Course Code	20ESX03	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	I
Course	BASIC ELECTRICAL ENGINEERING				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define power factor.	20ESX03.1	L1
2	Draw OCC of a DC shunt generator.	20ESX03.2	L1
3	Define voltage regulation	20ESX03.3	L1
4	Define slip in an induction motor.	20ESX03.4	L1
5	Write any one application of a single phase induction motor.	20ESX03.5	L1

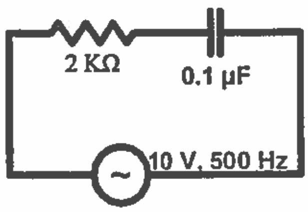
Part B (Long Answer Questions 5 x 12 = 60 Marks)

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain the following with an example i) Kirchoff's current law ii) Kirchoff's voltage law iii) Voltage division rule.	6M	20ESX03.1	L2
6 (b)	Find the current flowing through 4Ω resistance	6M	20ESX03.1	L3



OR

No.	Questions	Marks	Learning Outcome (s)	DoK
7 (a)	Define the following with respect to sinusoidal quantity: i) RMS Value ii) Average Value iii) Form factor iv) Peak factor	8M	20ESX03.1	L2
7 (b)	For the circuit shown, determine total impedance Z and current I	4M	207ESX03.1	L3



No.	Questions	Marks	Learning Outcome (s)	DoK
8	Describe the constructional details of a DC generator and write the function of each part with a neat sketch	12M	20ESX03.2	L2

OR

No.	Questions	Marks	Learning Outcome (s)	DoK
9	A 6 pole DC generator with 770 wave connected armature	12M	20ESX03.2	L2

	conductors and running at 500 rpm supplies a load of 12.5Ω resistance at a terminal voltage of 250V. The armature resistance is 0.25Ω and field resistance is 250Ω. Find the armature current, induced EMF and Flux per pole.			
10	A 230 V/230 V, 3 KVA transformer gave the following results: O.C test : 230 V, 2A, 100 W S.C test: 15 V, 13 A, 120 W Determine the regulation and efficiency at full load 0.8 p.f lagging	12M	20ESX03.3	L3
	OR			
11	Why parallel operation of transformers is required? Explain the parallel operation of transformers.	12M	20ESX03.3	L2
12	Derive the EMF equation of an alternator.	12M	20ESX03.4	L3
	OR			
13	How will the rotor of a 3-Φ induction motor rotate?	12M	20ESX03.4	L3
14	Compare the 1-Phase and 3-Phase Induction Motors.	12M	20ESX03.5	L3
	OR			
15	Explain the principle of operation and construction of a single-phase induction motor	12M	20ESX03.5	L3

NSRIT

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	ECE			Academic Year	2021 - 2022
Course Code	20BSX23	Test Duration	3 Hrs.	Max. Marks	70	Semester	I
Course	Applied Chemistry						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is conducting polymers?	20BSX23.1	L1
2	What is secondary cell?	20BSX23.2	L1
3	Write Schrodinger equation for free particle in 1D box	20BSX23.3	L1
4	How does Liquid Chromatography work in separation of mixtures of compounds?	20BSX23.4	L1
5	Define cyclodextrins in supramolecular chemistry with an example.	20BSX23.5	L1

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

No.	Questions (6 through 10)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain linear and cross-linked polymer with an example	5M	20BSX23.1	L2
6 (b)	Write any three differences of Buna-S and Buna-N polymers	7M	20BSX23.1	L1
	OR			
7 (a)	Write the preparation, properties and application of Nylon 6, 6	6M	20BSX23.1	L2
7 (b)	Write the preparation, properties and application of polyacetylene	6M	20BSX23.1	L2
8 (a)	Explain construction and working of calomel electrode	6M	20BSX23.2	L2
8 (b)	Explain potentiometric redox titration with an example	6M	20BSX23.2	L2
	OR			
9 (a)	Discuss the construction and working principle of hydrogen-oxygen fuel cell and write the disadvantages of the hydrogen-oxygen fuel cell	8M	20BSX23.2	L2
9 (b)	The Standard reduction potential of Mercury ions is 0.85V. What will be the potential of a 2M solution at temperature of 300K	4M	20BSX23.2	L2
10 (a)	What is Molecular Orbital Theory? Explain the molecular orbital energy level diagrams of O_2, CO and NO molecules.	8M	20BSX23.3	L2
10 (b)	Explain the hybridisation, geometry and magnetic properties of $[Fe(NH_3)_6]^{2+}$ and $[Co(CN)_4]^{2-}$ coordination systems. Write their magnetic nature	4M	20BSX23.3	L2
	OR			
11 (a)	What is wave function? Draw wave functions for a particle in a box at the n=1, n=2 and n=3 energy levels	6M	20BSX23.3	L2
11 (b)	Draw the energy level diagram for the probability distribution of a particle in a box at n=1 and n=2	6M	20BSX23.3	L2
12 (a)	Explain principle and instrumentation of conductivity meter with neat diagram	6M	20BSX23.4	L2
12 (b)	A solution of concentration 1 $\times 10^{-4}$ molar cyclodextrin has an absorbance at 450 nm of 0.88 in a 1.0 cm length cuvette. Calculate the molar extinction coefficient of the given solution	6M	20BSX23.4	L2
	OR			
13 (a)	Draw the ^1H and ^{13}C NMR spectra of Methyl Benzene and assign the peak positions	6M	20BSX23.4	L2
13 (b)	Explain the principle and Instrumentation of Gas Chromatography	6M	20BSX23.4	L2
14 (a)	What is supramolecular chemistry? Explain intramolecular and intermolecular hydrogen bonding interaction with an example	6M	20BSX23.5	L2
14 (b)	Discuss the importance of supramolecular systems in molecular	6M	20BSX23.5	L2

Semester End Supplementary Examination June 2022

Program : ECE (Supply) Key - Acedamic year : 2021-2022

Degree : B.Tech

Course : Applied chemistry

Semester : I

PART-A (Short Answer Questions 5X2=10 M)

1. Conducting Polymers :

- The polymers which conduct electricity are called conducting polymers.
- The conduction of the polymers may be due to unsaturation of or due to presence of externally added ingredients in them.

Eg. Poly acetylene, poly aniline, polypyrrole

Conducting polymers classified into 2 types; 1. Extrinsic C.P
2. Intrinsic C.P

2. Secondary Cell :

- A rechargable battery, storage battery or secondary cell is a type of electrical battery which can be charged, discharged into a load.
- It can be recharged many times.
- Used in electronic calculators, cordless electronic shavers etc.

3. Schrodinger equation for free particle in 1-D Box :

$$\frac{d^2\psi}{dx^2} + \frac{8\pi^2 m}{h^2}(E-V)\psi = 0$$

where ψ = Amplitude of wave
m = mass of particle
h = planck's constant
V = velocity of particle
E = Total energy.

A4. Liquid Chromatography:

- Components within a mixture are separated in a column based on each components affinity for mobile phase.
- If the components are of different polarities, mobile phase of distinct polarity is passed through columns. One component will migrate through column faster than other.

A5. Cyclo - Dextrins:

- Cyclodextrins (CDs) are macrocyclic, oligosaccharides consisting of α-1, 4- linked D- glucopyranose units [16,17].
- They have a distinctive truncated cone structure.

Ex. Cyclodextrin derivatives;
methyl - α - cyclodextrin
hydroxypropyl - α - cyclodextrin.

PART-B (5x12 = 60 M)

6(a) Explain linear and cross-linked polymer with an example.

Linear Polymer: Linear polymers are formed during addition polymerisation. They are also called branched link structures. Examples are thermoplastics. Eg:- Polyethylene.

$$n\,(CH_2=CH_2) \longrightarrow -[CH_2-CH_2]-_n$$

By using free radical initiators, low density polyethene (LDPE) is produced with 0.92 gm/cc density

$$\underset{\text{(Monomer)}}{\underset{\text{ethylene}}{n\,(CH_2=CH_2)}} \xrightarrow[80-250^\circ C]{\text{Benzyl peroxide}} \underset{\text{(Polymer)}}{\underset{\text{LDPE}}{-[CH_2-CH_2]-_n}}$$

1500 atm pressure

By using coordination chain mechanism, high density polyethene (HDPE) is produced having 0.965 gm/cc density.

$$n[CH_2=CH_2] \xrightarrow[120^\circ C]{\text{Triethyl Al}} \underset{\text{HDPE}}{-[CH_2-CH_2]-_n}$$

Cross-linked Polymer: Cross linked polymers are formed during condensation polymerisation. They are also called 3-dimentional structures. Thermosetting plastics are examples of cross-linked polymers.

Eg:- Nylon-6,6:

It is prepared by the reaction between adipic acid and hexa methylene di amine.

$$\underset{\text{adipic acid}}{n[HOOC-(CH_2)_4-COOH]} + \underset{\text{Hexamethylene diamine}}{n[H_2N-(CH_2)_6-NH_2]}$$

$$\downarrow -H_2O$$

$$-[\overset{O}{\overset{\|}{C}}-(CH_2)_4-\overset{O}{\overset{\|}{C}}-NH-(CH_2)_6-NH]-_n + 2nH_2O$$

c)
b) Write any three differences between Buna S and Buna N polymers.

Buna - N	Buna - S
* The monomers of Buna N is 1,3 butadiene and acrylonitrile.	* The monomers of Buna S is 1,3 butadiene and styrene.
* An alternative name is Nitrile-Butadiene-Rubber.	* An alternative name is Styrene-Butadiene Rubber.
* Buna-N is synthesized by the polymerization of 1,3- butadiene and acrylonitrile in the presence of sodium.	* Buna-S is synthesized by the polymerisation of 1,3 butadiene and styrene in the ratio 3:1 in the presence of sodium.
* It is used in belts, sealing of oils etc.	* It is used in automobile industries.

Buna N :

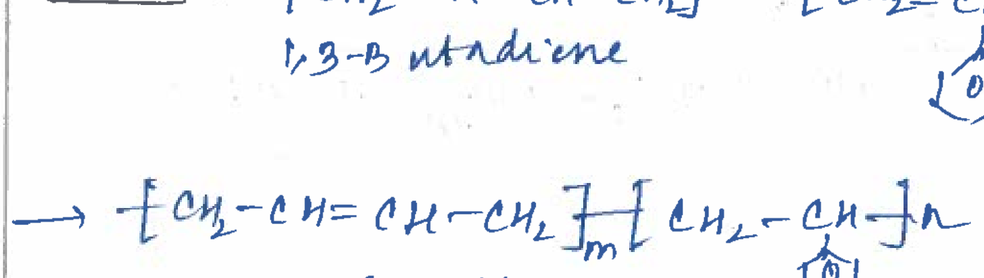
$$m[CH_2=CH-CH=CH_2] + n[CH_2=\underset{CN}{CH}] \xrightarrow[\text{catalyst}]{Na}$$

1,3- Butadiene Acrylonitrile

$$\rightarrow -[CH_2-CH=CH-CH_2]_m-[CH_2-\underset{CN}{CH}]_n$$ (Buna N Rubber)

Buna S : $m[CH_2=CH-CH-CH_2] + n[CH_2=CH]$ $\xrightarrow[]{\text{Na catalyst}}$

1,3-Butadiene   Styrene.



Buna-S Rubber

7(a) Write the preparation, properties and application of Nylon-6,6 ?

Ans It is prepared by reaction between adipic acid and hexamethylene diamine

$$n[HOOC-(CH_2)_4-COOH] + n[H_2N-(CH_2)_6-NH_2] \xrightarrow{-H_2O} [\overset{O}{\overset{\|}{C}}-(CH_2)_4-\overset{O}{\overset{\|}{C}}-NH-(CH_2)_6-NH]_n + 2nH_2O$$

Adipic acid   Hexamethylene diamine

Application for Nylon-6,6.

→ It is used for preparing fibre

→ It is used for the moulding purpose for gears bearings

→ It is used for making filaments for ropes bristles and brushes

properties for Nylon-6,6.

→ It is translucent, whitish, horny high melting point

→ It is having high temperature, stability and good scratch resistance

→ It is soluble in phenol formic acid but insoluble in organic solvents like benzene and acetone....

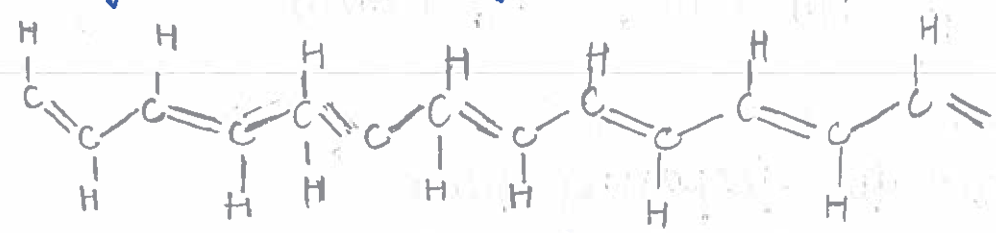
7) b) write the preparation, properties and applications of polyacetylene

Ans. The valence band is filled band and conduction band is empty. when the energy gap between these is low the electrons(e^-) from the valence band are excited to conduction band becomes mobile throughout the polymer and show conductivity.

Eg:- trans-polyacetylene, polyaniline.

$$nCH \equiv CH \xrightarrow{\text{Polymerisation}} \text{-}(CH=CH)\text{-}_n$$

acetylene $\qquad\qquad$ polyacetylene



The semiconducting polyacetylene $(CH_2)_n$ has a typical carbon backbone structure. The localized e^-s in σ bonds from the backbone of the polymer chain and dominate the medicinal properties while the e^-s in the π bonds are delocalized along the chain and responsible for the electrical & optical properties of a conjugated polymer

The σ bonds form completly filled low lying energy bands that have energy gaps than π bonds electrons. In presence of electric field the electrons constituting bonds can move along the molecular chain

polyacetylene. posses alternate signal and double bond that give rise to mobile π-e's when doped i.e becomes anistropic metallic conducters.

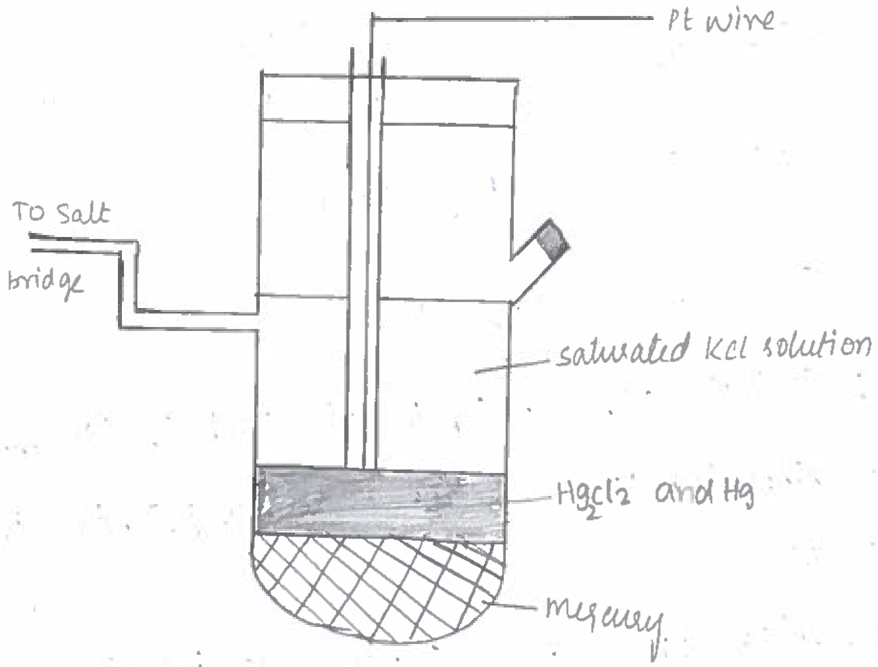
They are two types of doping :-

(i) oxidation with halogen (p-doping)

(ii) reduction with alkali (n-doping)

3(a) Explain construction and working of calomel electrode.

Saturated Calomel Electrode :



Calomel electrode

Calomel electrode is a metal-metal salt ion electrode. It consists of mercury, mercurous chloride and a solution of KCl. Mercury is placed at the bottom of a glass tube having a side tube on each side. Mercury is covered by a paste of mercurous chloride (calomel) with mercury & KCl

A solution of KCl is introduced above the paste through the side tube. A platinum wire sealed in a glass tube is dipped into mercury and used to provide the external electrical contact. The concentration whose potential is to be determined is connected to this electrode through salt bridge. The potential of electrode depends upon the concentration of KCl solution.

The net reversible electrode reaction is

$$Hg_2Cl_2 (s) + 2e^- \rightleftharpoons 2Hg(l) + 2Cl^-$$

NERST Equation:-

$$E = E^{\circ}_{Hg_2Cl_2} - \frac{2.303RT}{2F} \log [Cl^-]^2$$

$$= E^{\circ} - \frac{2.303RT}{F} \log [Cl^-]$$

$$= E^{\circ} - 0.0591 \log [Cl^-]$$

The electrode potential depends up on the conc of KCl solution

At 25°c For saturated KCl solution electrode potential is 0.2415 volts.

For 1N KCl solution standard reduction potential is 0.281 volts

For 0.1N KCl solution the reduction potential is 0.3338 volts

The electrode can be coupled with hydrogen electrode containing solution of unknown concentration.

$$Pt, H_2 (1atm) / H^+ = ? \,//\, Hg_2Cl_2(s) / Hg^+$$

The emf of cell,

$$E_{cell} = E_{right} - E_{left} = 0.2422V + 0.0592 V\,pH$$

$$pH = \frac{E_{cell} - 0.02422V}{0.0592V}$$

Applications:

1. It is used as a secondary reference electrode in the measurement of single electrode potential.
2. It is the most commonly used reference electrode in all potentiometric determinations.

8(b) Explain potentiometric redox titration with an example.

Oxidation - Reduction Titrations

The procedure adopted for oxidation titration is the same as in acid-base titration, the only difference is that the electrode reversible to hydrogen ions is replaced by a bright platinum electrode. The EMF of the electrode is determined by the activity of ratio of the substance being oxidized or reduced. For eg: Fe^{2+} titrated against $K_2Cr_2O_7$. The Fe^{2+} solution is taken in the beaker, treated with dil H_2SO_4 and pt electrode and calomel electrodes are dipped. The electrodes are connected to the potentiometer and emf of the solution after the addition of $K_2Cr_2O_7$ is recorded. On addition of $K_2Cr_2O_7$ from the burette, EMF of the cell increases first slowly, but at the equivalence point there will be sudden jump in potential. since change in ratio of Fe^{2+}/Fe^{3+} ion concentration. A graph is plotted with emf and volume of $K_2Cr_2O_7$. A sigmoid curve is obtained and the steepest portion of the curve indicates the end point of the titration.

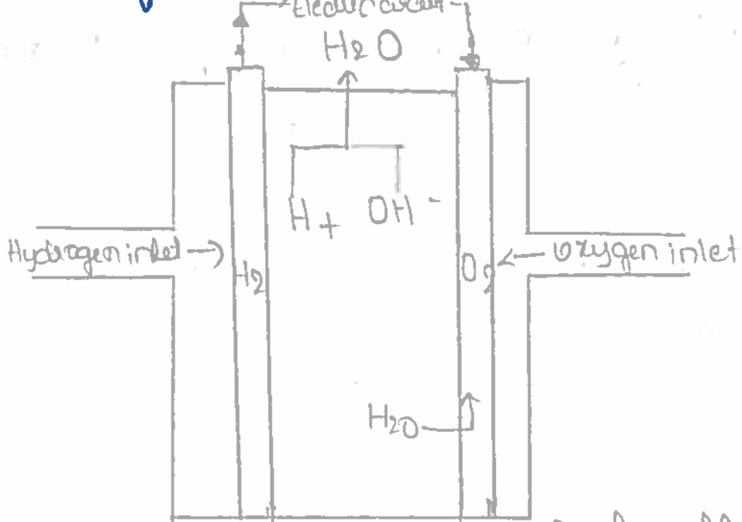
1)a) Discuss the construction and working principle of hydrogen-oxygen fuel cell and write the disadvantages of the hydrogen-oxygen fuel cell.

Definition: A fuel cell is an electrochemical which converts chemical energy contained in readily available fuel oxidant system into electrical energy.

Principle: The basic principle of fuel cell is as same as that of an electrochemical cell. The fuel cell operates like a galvanic cell. The only difference is that the fuel and the oxidant are stored outside the cell. Fuel and oxidant are supplied continously and separately to the electrodes at which they undergo redox reaction. Fuel cells are capable of supplying current as long as reactants are replenished

Fuel + Oxidant ⟶ Oxidation products + Electric Energy

Hydrogen oxygen fuel cell: This cell is a common type of fuel cell. Similar to a galvanic cell, fuel cell also have two half cells. Both half cells have porous graphite electrode with a catalyst. The electrodes are placed in the aqueous solution of NaOH or KOH which acts as an electrolyte. Hydrogen and oxygen are supplied at anode and cathode respectively at about 50 atmospheric pressure, the gases diffuse at respective electrodes. The two half-cell reactions are as follows:



Hydrogen oxygen fuel cell

At anode: $2H_2(g) + 4OH^-(aq) \rightarrow 4H_2O(l) + 4e^-$

At cathode: $O_2(g) + 2H_2O(l) + 4e^- \rightarrow 4OH^-(aq)$

The net reaction: $2H_2(g) + O_2(g) \rightarrow 2H_2O(l)$

The EMF of this cell is measured to be 1.23V. A number of such fuel cell are stacked together in series to make a battery.

Advantages:

1) The energy conversion is very high (75-82%)
2) They save fossil fuels.
3) They have quick start system
4) They have low maintenance cost
5) noise and thermal pollution are very low.

Disadvantages:

1) The major disadvantage of the fuel cell is the high cost and the problems of durability and storage of large amount of hydrogen
2) The accurate life time is also not known.

Applications:

1) The most important application of a fuel cell is its use in space flight.
2) Fuel cell batteries for automotive will be a great boom for the future.

Limitation:

1) The life time of fuel cell is not accurately known
2) Their initial cost is high.

A stream of pure hydrogen is bubbled around the platinum foil at a constant pressure of one atmosphere. The SHE may be represented as

$$H_2(1\,atm) ; Pt / H^+ (C = 1)$$

The electrode potential of SHE = zero. Depending on half-cell to which it attached hydrogen electrode can acts as a cathode or anode.

At oxidation reactions act as anode:

$$\frac{1}{2} H_2(g)(1\,atm) \rightarrow H^+ (1M) + e^-$$

At reduction reaction act as cathode:

$$H^+ (1M) + e^- \rightarrow \frac{1}{2} H_2(g)(1\,atm)$$

<u>Nernst</u> <u>Equation</u>:

The concentration effect i.e. when H^+ concentration is not 1M the EMF of the electrode alters. To calculate the potential of the standard electrode, Nernst equation is used as given below.

$$E = E^0 - 2.303\, RT/nF \log_{10}[H^+]$$

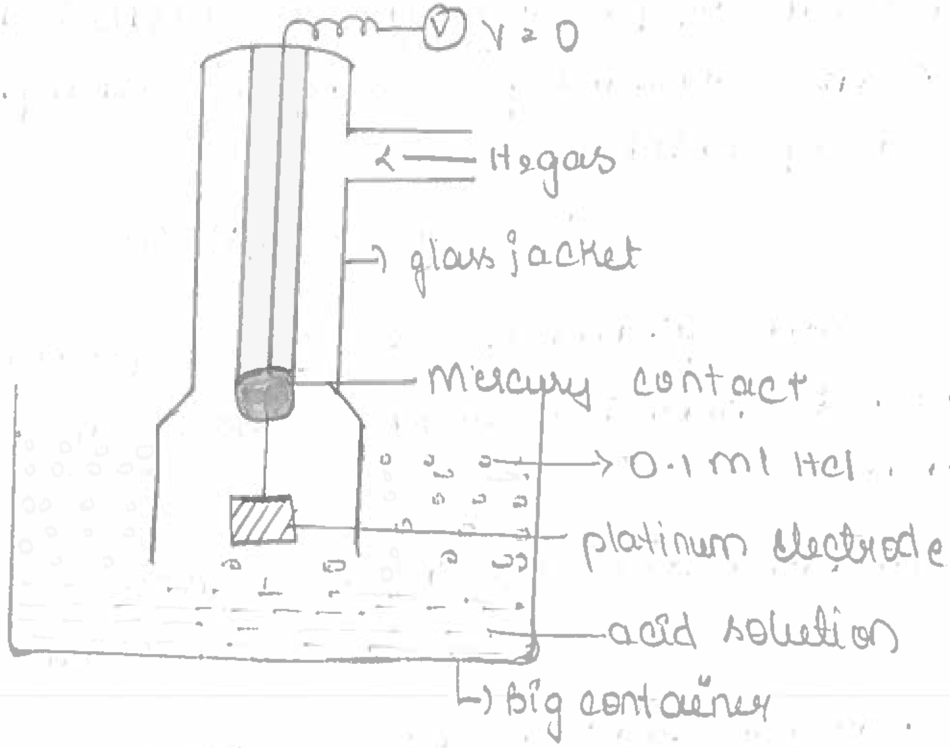
Substituting all the values we get the potential of the electrode at 25°C as

$$E = E^0 - \frac{2.303 \times 8.313 \times 298}{1 \times 96500} \log_{10}[H^+]$$

$$E = E^0 - 0.0591 \log_{10}[H^+]$$

$$E = E^0 - 0.0591\, pH$$

$$E = -E = -E^0 + \frac{0.0591}{n} \log_{10}[H^+]$$


V = 0
H₂ gas
glass jacket
Mercury contact
0.1 M HCl
platinum electrode
acid solution
Big container

10.(a) What is Molecular Orbital Theory? Explain the molecular orbital energy level diagrams of O_2, CO and NO molecules.

A. Molecular Orbital Theory:

In 1932, Hund and Mulliken had put forward the molecular orbital theory. The main features are as follows

* New orbitals termed as the molecular orbitals are formed by the overlap of atomic orbitals of the combining atoms.
* The number of molecular orbitals is equal to the number of atomic orbitals.
* The electron in an atomic orbital is influenced by just one positive nucleus of the atom i.e it is monocentric, whereas depending upon the number of atoms in the molecules of the electron of a molecular orbital is under the influence of more than one nuclei, i.e, it is polycentric.
* Atomic orbitals with comparable energies as well as proper orientations combine to form molecular orbitals.

Molecular orbital energy level diagrams of O_2, CO & NO molecules:-

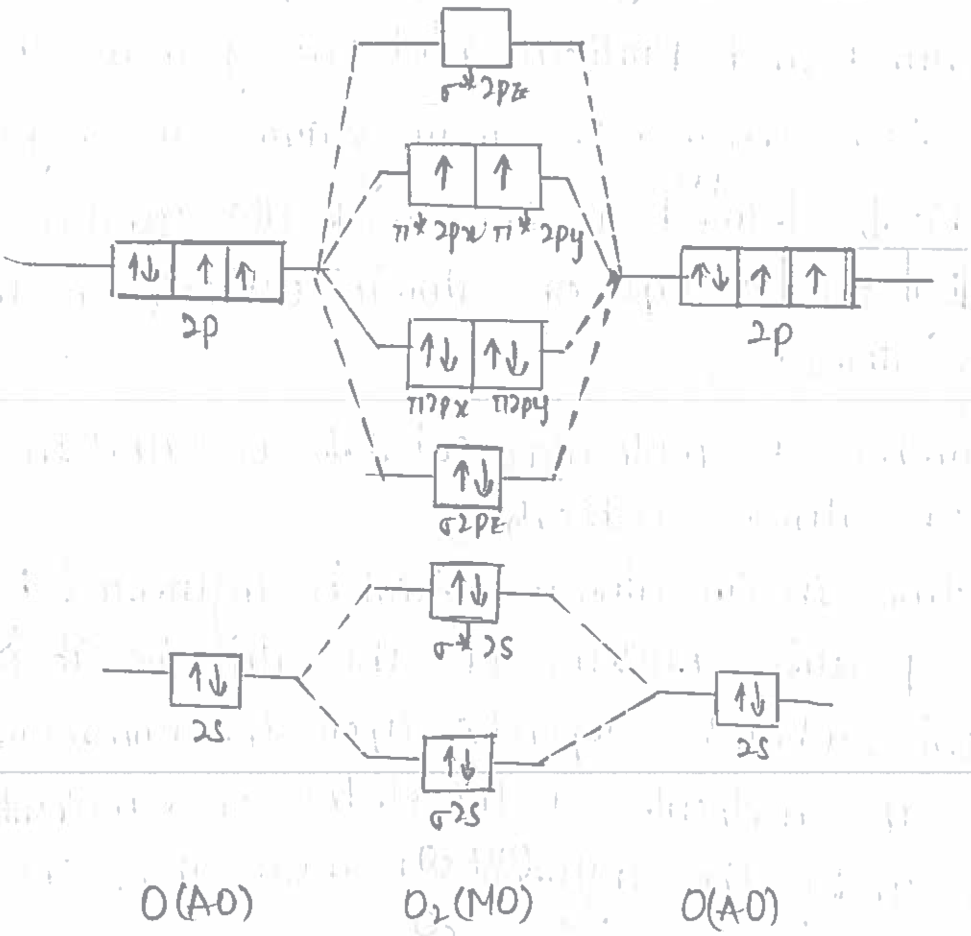
(i) Oxygen molecule (O_2):-

The electronic configuration of the molecule is

O_2: $KK(\sigma 2s)^2 (\sigma^* 2s)^2 (\sigma 2p_z)^2 (\pi 2p_x)^2 = (\pi 2p_y)^2 (\pi^* 2p_x)^1 = (\pi^* 2p_y)^1$

Bond order $= \frac{8-4}{2} = 2$

The bond dissociation energy of O_2 molecule is 495 $kJ\,mol^{-1}$ and the bond length is 121 pm.



(ii) Carbon monoxide molecule (CO):

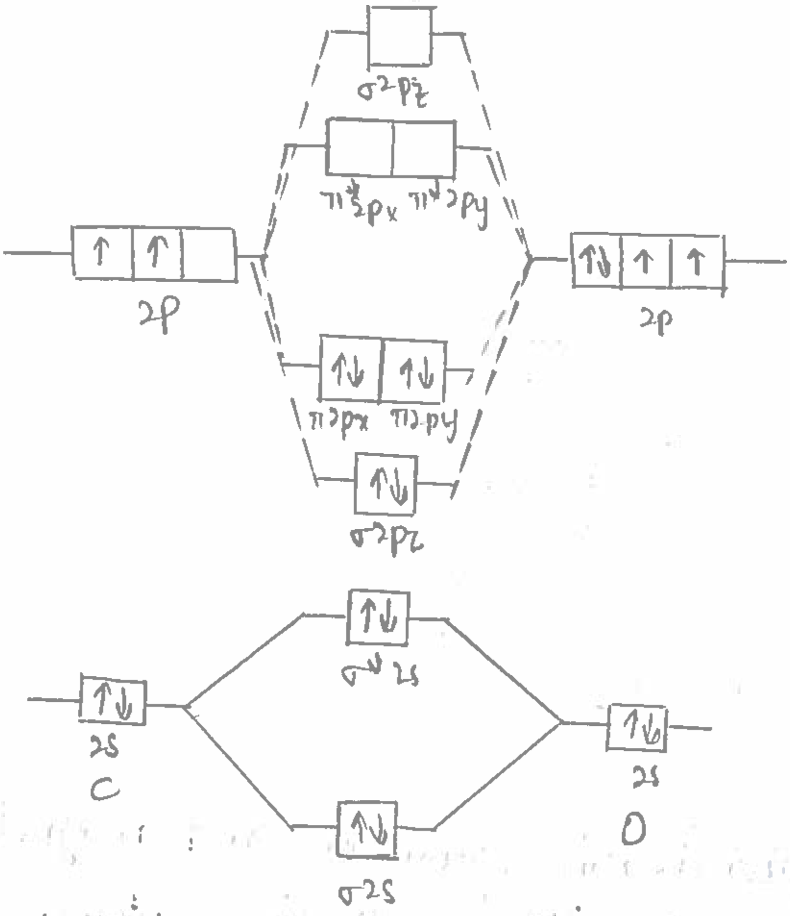
The electronic configuration of carbon and oxygen atoms are

$C = 1s^2 2s^2 2p^2$ and $O = 1s^2 2s^2 2p^4$

$CO = KK (\sigma 2s)^2 (\sigma^* 2s)^2 (\sigma 2p_z)^2 (\pi 2p_x)^2 = (\pi 2p_y)^2$

Bond order $= \frac{8-2}{2} = 3$

The molecule is diamagnetic there is no unpaired electron in it. gives the molecular orbital diagram of CO.



(iii) Nitric oxide Molecule (NO) :-

The electronic configuration of N and O atoms are

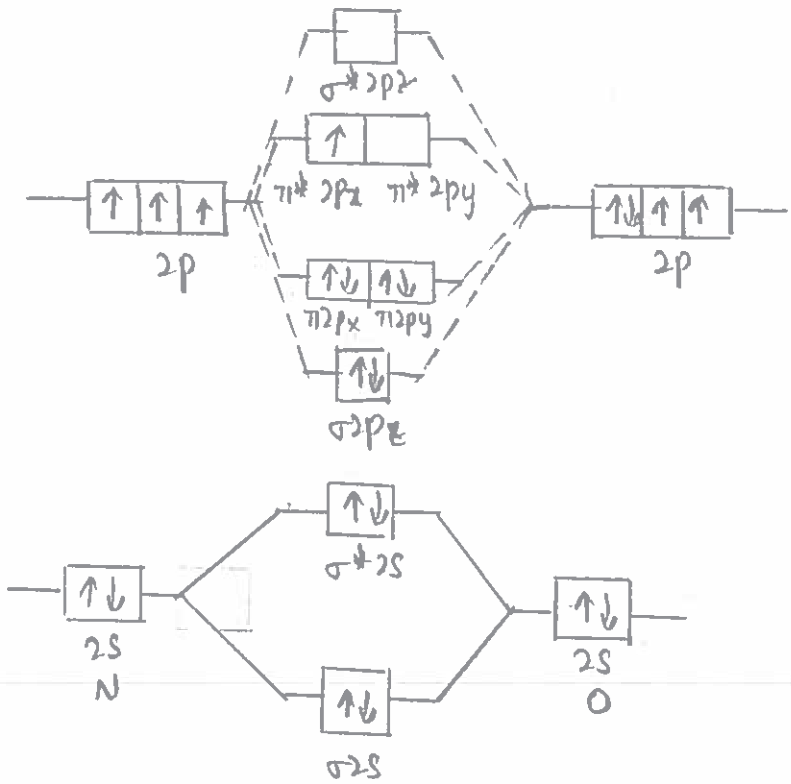
$N = 1s^2 2s^2 2p_x^1 2p_y^1 2p_z^1$

$O = 1s^2 2s^2 2p_x^2 2p_y^1 2p_z^1$

The molecular orbital electronic configuration of NO molecule is .

$NO = KK(\sigma 2s)^2 (\sigma^* 2s)^2 (\sigma 2p_z)^2 (\pi 2p_x)^2 = (\pi 2p_y)^2 (\pi^* 2p_x)^1$

Bond order $= \frac{8-3}{2} = 2\frac{1}{2}$.



10.(b) Explain the hybridization, geometry and magnetic properties of $[Fe(NH_3)_6]^{2+}$ & $[Co(CN)_4]^{2-}$ coordination systems. Write their magnetic nature.

A. Hybridization

It is defined as the concept of mixing two atomic orbitals to give rise to a new type of hybridized orbitals.

The energy corresponding to their transition corresponds to green and yellow lights which are absorbed from white light, while the blue and red portions are emitted. The solution of $[Fe(NH_3)_6]^{2+}$ is purple.

→ In octahedral complexes so vary from one metal ion to another and the nature of ligands.

→ for examples three complexes of Co^{3+} as $[Co(H_2O)_6]^{3+}$ $[Co(NH_3)_6]^{3+}$ & $[Co(CN)_6]^{3-}$.

coordination compound	wavelength of light absorbed	colour of light absorbed	colour of coordination entity
1. $[Co(CN)_6]^{3+}$	310 nm	violet	Pale yellow
2. $[Co(H_2O)_6]^{3+}$	475 nm	Blue	Red

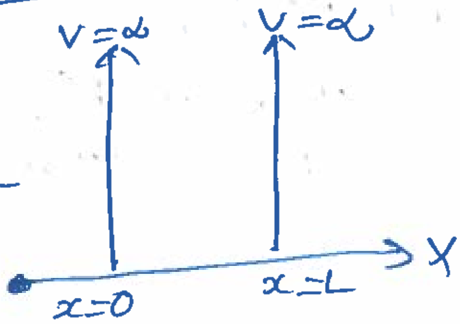
Magnetic properties of complex compounds on coordination compounds:-

→ Diamagnetic substances repell the magnetic lines of forces and decreases the flux

→ Paramagnetic compounds attract the magnetic lines of forces and increase the flux.

11)(a) ψ^2 it is also known as probability density, ψ^2 is o the probability of finding an electron at a point within the atom,

Particle in one dimensional Box



→ consider a particle that is constrained to move only in x-direction from $x=0$ to $x=L$

→ It means the particle is moving x-direction only from $x=0$ to $x=L$

inside the box within the bound axes,

→ As the particle is moving inside, this box, it can't come outside the box, so the finding probability of particle in outside the box will be 0 (zero)

→ As the particle inside the box, the potential energy inside the box potential energy (U) will be α

The schrodinger wave equation in operator form

$$\hat{H}\psi = \epsilon\psi \quad \text{—— (1)}$$

where $\hat{H}$ = Hamiltenian operator in x-direction

$$\hat{H} = \frac{-h^2}{8\pi^2 m}\frac{d^2}{dx^2} + V$$

[V_2 = potential energy, m = mass, h = plank's constant]

put the $\hat{H}$ value in eq (1)

$$= \frac{-h^2}{8\pi^2 m}\frac{d^2\psi}{dx^2} + V\psi = \epsilon\psi$$

$$= \frac{-h^2}{8\pi^2 m}\frac{d^2\psi}{dx^2} + V\psi - \epsilon\psi = 0$$

$$= -\left(\frac{-h^2}{8\pi^2 m}\frac{d^2\psi}{dx^2} + V\psi - \epsilon\psi\right) = 0$$

$$= \frac{h^2}{8\pi^2 m}\frac{d^2\psi}{dx^2} + (\epsilon - V)\psi = 0 \quad \text{—— (2)}$$

multiplying the equation (2) by $8\pi^2 m / h^2$

$$= \frac{\cancel{8\pi^2 m}}{\cancel{h^2}} \times \frac{\cancel{h^2}}{\cancel{8\pi^2 m}}\frac{d^2\psi}{dx^2} + \frac{8\pi^2 m}{h^2}(\epsilon - V)\psi = 0$$

$$= \frac{d^2\psi}{dx^2} + \frac{8\pi^2}{h^2}(\epsilon - V)\psi = 0 \quad \text{—— (3)}$$

Now appling the two conditions one is the particle is in outside the box and in inside the box

(I) outside the box:

The potential energy (V) will be α, this value put in eq (3)

$$\frac{d^2\psi}{dx^2} + \frac{8\pi^2 m}{h^2}(E-\alpha)\psi = 0$$

(II) Inside the box.

The potential energy (V) will be 0 (V=0)

This value put in eq (3)

$$\frac{d^2\psi}{dx^2} + \frac{8\pi^2 m}{h^2}(E-0)\psi = 0$$

$$\frac{d^2\psi}{dx^2} + \frac{8\pi^2 m}{h^2}E\psi = 0$$

$$\frac{d^2\psi}{dx^2} + k^2\psi = 0 \qquad \left(\because k^2 = \frac{2mE}{\hbar^2} \quad \hbar = h/2\pi\right)$$

Now, wave function can be represented as

$\psi(x) = A\sin kx + B\cos kx$ (A, B = arbitrary constant)

(ii) b) $x = L$ (at the boundary)

the finding probability at the boundary will be '0'

$x(L) = 0$

$\psi(L) = A\sin kL + B\cos kL$

$0 = A\sin kL + 0$

$A\sin kL = 0$

$kL = n\pi$

$\sin kL = 0$

$kL = n\pi$

$(\pi, 2\pi, 3\pi)$

$$K = \frac{n\pi}{L} \quad \text{but} \quad K^2 = \frac{2mE}{\hbar^2}$$

$$= \frac{n^2\pi^2}{L^2} = \frac{2mE}{\hbar^2} \quad (\hbar = h/2\pi)$$

$$= \frac{n^2\pi^2}{h^2} = \frac{2mE}{\left(\frac{h}{2\pi}\right)^2} = \frac{\frac{2mE}{h^2}}{4\pi^2} = \frac{2mE}{h^2} \times 4\pi^2$$

$$= \frac{n^2\pi^2}{L^2} = \frac{2mE}{h^2} \times 4\pi^2$$

$$= n^2 \cancel{\pi^2} h^2 = 2mE \times 4\pi^2 L^2 = 8\cancel{\pi^2} mEL^2$$

$$E = \frac{n^2 h}{8mL^2}$$

where $n=1$, $$E_1 = \frac{(1^2)\, h^2}{8mL^2}$$

$$E_1 = \frac{h^2}{8ma^2}$$

where $n=2$ $$E_2 = \frac{(n)^2 h^2}{8ma^2} = \frac{(2)^2 h^2}{8ma^2} = \frac{4h^2}{8ma^2}$$

$$E_2 = 4h^2/8ma^2$$

where $n=3$ $$E_3 = \frac{3h^2}{8ma^2}$$

where $n=4$ Energy will be indicated as

$$E_4 = \frac{16h^2}{8ma^2}$$

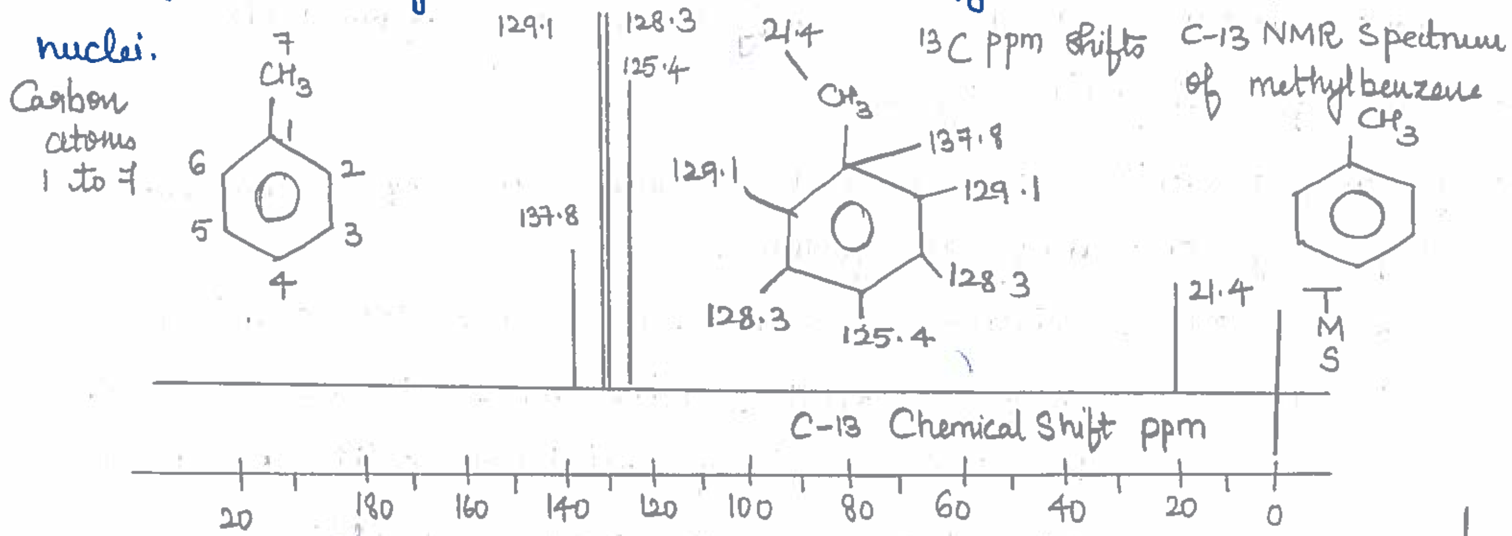
3 (a) Draw the ^1H and ^{13}C NMR Spectra of Methyl Benzene and assign peak positions.

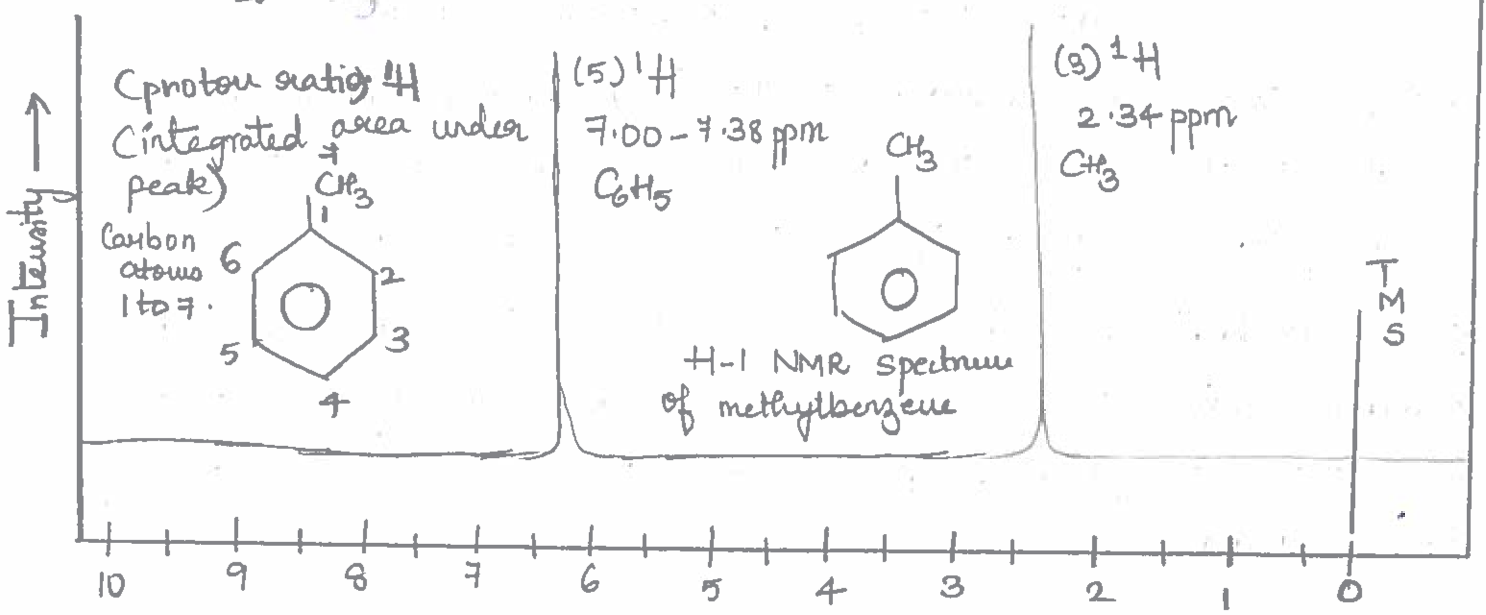
A. The NMR signals are detected with sensitive radio receivers.

- It gives the details of a molecule's individual functional group in the form of NMR signals (or) spectra.
- So NMR spectra is a plot of frequency vs intensity.
- Since the nucleus involved in the proton the spectrum is also called as proton magnetic resonance spectrum (PMRS) (^1H NMR).

→ ^{13}C NMR:

- Carbon NMR can be used to determine the number of non-equivalent carbons and to identify the types of carbon atoms.
- ^{13}C signals are spread over a much wider range than ^1H signals making it easier to identify & count individual nuclei.





13(b) Explain principle and instrumentation of gas chromatography?

A. principle:

- In GSC, the coloumn is packed with an active solid stationary phase and components of mixture are distributed between gas phase and active solid phase.

  In this due to differences in adsorption behaviour of components of mixture, separation is achieved.

- In GLC, the coloumn is packed with a porous ~~solid~~ solid (of large area) which is coated with thin layer of non-volatile liquid.

  In this separation takes place due to difference in partitioning of sample between mobile gas phase and thin layer of liquid coated on inert support

Instrumentation:

A gas chromatography consists of:

1. A supply of carrier-gas (He, Ar or N_2) from a high pressure cylinder having a pressure regulator and flow meters.
2. A sample injection system.
3. packing materials in packed coloumn or open tubular or capillary coloumns as follows.
   a. In GSC - activated carbon, silic gas or alumina
   b. In GLC - porous materials like glass beads, ground fire-bricks, finely divided celite or polymers like GIC, porapak, chromsorb etc.

   → In GLC, the coloumn containing porous material is filled with non-volatile liquid such as silicone oil, polythylene glycol, apieson oil or grease etc.

4. The detector is situated at the exist of separation coloumn which senses and measures the small amount of the separated components present in the carrier-gas leaving the coloumn.

14 (b) Discuss the importance of supramolecular systems in molecular switches and molecular elevators with example.

A. Molecular Switches:

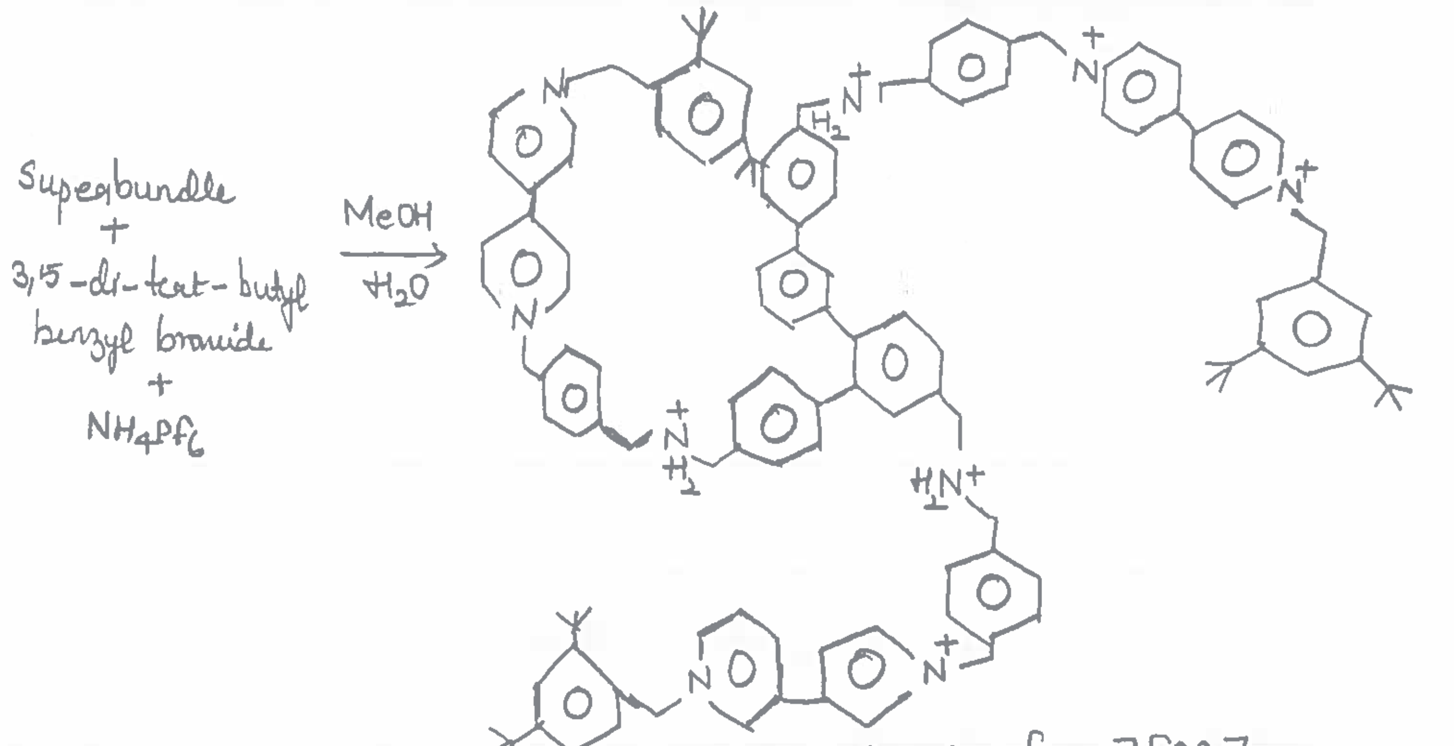
- A molecular switch is a molecule that can be reversibly shifted between two or more stable states, in response to environmental stimuli, changes such as pH, light, temperature an electric current micro-environment or in the presence of ions and ligands. The oldest forms of the synthetic molecular switches are pH indicators which exhibit the different colours due to the changes in pH of the medium.

The different types of molecular switches are listed.

1. Acidochromic molecular switches.
2. photochromic molecular switches.
3. Coordination switching.

Molecular Elevators:

- A molecular elevator is a molecular machine that behaves like a nanoscale elevator, which is made of platform like interlocked with a trifurcated rig-like component which is only 3.5 nanometers is size. Hence molecular elevator is more and better than previously reported molecular machines.



Molecular elevator $[GH_3][PF_6]_9$

5. (a) List any four applications of rotaxanes in the supramolecular systems.

A. A rotaxane is a mechanically interlocked molecular architecture consisting of a dumbell shaped molecule which is threaded through a macrocycle.

→ Applications:

1. Molecular switches (machines):

- The potential use of rotaxanes in molecular electronics on logic molecular switching elements and motor shuttles. These molecular machines are based on the movement of the macrocycle on the dumbbell molecule. The macrocycle can rotate around the shaft from one site to another; rotate like wheel, and axle to function as the molecular switch.

2. Ultrastable dyes:

- Rotaxanes potential application in long lasting dyes based on enhanced stability of the inner portion of dumbbell shaped molecule for example cyclodextrin protected azo dyes.

3. Nano recording:

- Rotaxane is deposited as a Langmuir-Blodgett film on ITO - coated glass as memory dot.

02/7/22

HOD

NSRIT

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	Program	CSE, CSM, CSD & EEE	Academic Year	2021 - 2022
Course Code	20BSX33	Test Duration	3 Hrs. Max. Marks 70	Semester	I
Course	APPLIED PHYSICS				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	In Newton's ring's experiment of interference due to reflected light, center spot is perfectly dark - Justify	20BSX33.1	L2
2	Define numerical aperture of an optical fiber.	20BSX33.2	L1
3	Define susceptibility.	20BSX33.3	L1
4	List any two merits of classical free electron theory.	20BSX33.4	L1
5	List any two differences between intrinsic semiconductors and extrinsic semiconductors.	20BSX33.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6	Derive the equations for path difference, condition for maximum intensity and minimum intensity in thin film interference due reflected light.	12M	20BSX33.1	L2
	OR			
7	Explain the construction and working of Nicol's prism with neat diagram.	12M	20BSX33.1	L2
8 (a)	Derive the equation for relation between Einstein coefficient's of spontaneous emission and stimulated emission of light.	6M	20BSX33.2	L2
8 (b)	Write any three applications of Laser.	6M	20BSX33.2	L1
	OR			
9 (a)	Derive the equation for acceptance angle of an optical fiber.	8M	20BSX33.2	L2
9 (b)	Mention any four applications/advantages of an optical fiber.	4M	20BSX33.2	L1
10 (a)	Define hysteresis curve for a ferromagnetic material with diagram.	4M	20BSX33.3	L1
10 (b)	Write the differences between soft ferromagnetic materials and hard ferromagnetic materials based on hysteresis curve.	8M	20BSX33.3	L2
	OR			
11 (a)	Derive Claussius- Mossoti equation.	6M	20BSX33.3	L2
11 (b)	List any three applications of dielectric materials.	6M	20BSX33.3	L2
12 (a)	Derive the Schrodinger's time independent wave equation.	9M	20BSX33.4	L2
12 (b)	Explain the significance of wave function.	3M	20BSX33.4	L2
	OR			
13 (a)	List out the drawbacks of classical free electron theory.	6M	20BSX33.4	L1
13 (b)	Explain the variation of Fermi function *f(E)*, with temperature.	6M	20BSX33.4	L2
14 (a)	State and explain Bloch theorem.	6M	20BSX33.5	L2
14 (b)	Explain the significance of E Vs k diagram in Kronig-Penny model with neat diagram.	6M	20BSX33.5	L2
	OR			
15 (a)	Explain how origin of formation of energy bands in solids enables to classify into conductors, semiconductors and insulators.	8M	20BSX33.5	L2
15 (b)	List any four applications of Hall Effect.	4M	20BSX33.5	L1

Part –A

1. In Newton's rings experiment at the center we will get dark spot because at center (point of contact between lens and glass plate) the thickness of air film is zero.

i.e t=0 so the path difference is equal to λ/2 which results dark region.

2. The light gathering capacity of an optical fiber is known as Numerical Aperture and it is proportional to Acceptance Angle. It is numerically equal to sine of minimum Acceptance Angle.

$$NA = \sin\theta_a$$

3. The intensity of magnetization produced in a material is directly proportional to the magnetization filed H.

$$I \propto H$$

$$\Rightarrow I = \chi H$$

$$\Rightarrow \chi = \frac{I}{H}$$

Where χ proportionality is constant, called the magnetic susceptibility of the material.

4. It is used to verify Ohm's law.

The electrical and thermal conductivities of metals can be explained.

5.

Intrinsic semiconductors	Extrinsic semiconductors
Pure in form	Impure in form
No of electrons and holes are equal	No of electrons and holes are never equal

Part –B

A. 6. The colors of thin films, soap bubbles and oil slicks can be explained as due to the phenomenon of interference by reflection. In all these cases the formation of interference pattern is by the division of amplitude.

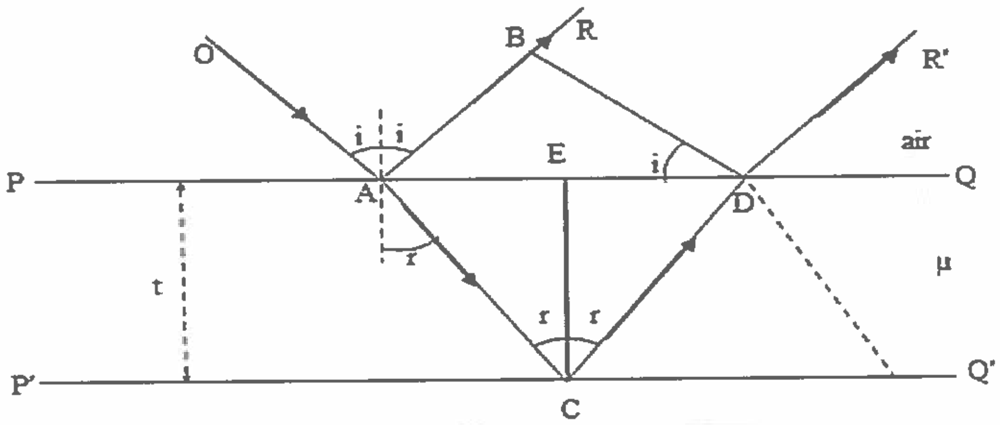
Consider two plane parallel surfaces *PQ* and *P'Q'* are separated by a distance *t*. The refractive index of the medium between the surfaces is assumed as μ.

Let a ray of light OA be incident on the surface *PQ* at A, then some part of the light is reflected into the same medium as AR ray. The remaining part is transmitted into the medium and is reflected at the lower surface *P'Q'* at C, meets the upper surface at D and emerges as *DR'* ray.

The path difference between the reflected rays AR and DR'

= path (AC+CD) in medium – path AB in air

$= \mu\,(AC+CD) - AB$ (for air $\mu = 1$) …… (1)



From Δ AEC, $\cos r = \dfrac{CE}{AC}$

$$\Rightarrow AC = \frac{CE}{\cos r}$$

$$\Rightarrow AC = \frac{t}{\cos r} \quad \ldots\ldots (2) \qquad (\because CE = t)$$

The ΔAEC and ΔCED are similar

$$\therefore \text{AC+CD} = \frac{t}{\cos r} + \frac{t}{\cos r}$$

$$\text{AC+CD} = \frac{2t}{\cos r} \quad \ldots\ldots. (3)$$

From ΔABD, $\sin i = \dfrac{AB}{AD}$

$$\Rightarrow AB = AD \sin i \quad \ldots\ldots (4)$$

According to *Snell's law*, the refractive index $\mu = \dfrac{\sin i}{\sin r}$

$$\Rightarrow \sin i = \mu \sin r \quad \ldots.. (5)$$

Substitute equation (5) in equation (4)

$$AB = AD \mu \sin r \ldots.. (6)$$

From ΔAEC, $\tan r = \dfrac{AE}{CE}$

$$\Rightarrow AE = CE \tan r$$

$$\Rightarrow AE = t \tan r \quad \ldots\ldots (7)$$

$$\therefore AD = AE + ED$$

$$\therefore AD = 2t\tan r \quad \ldots\ldots (8)$$

Substitute equation (8) in equation (6)

$$\therefore AB = 2t \times \tan r \times \mu \sin r$$

$$AB = 2t \times \mu \times \frac{\sin r}{\cos r} \times \sin r$$

$$AB = \frac{2\mu t \sin^2 r}{\cos r} \quad \ldots.. (9)$$

Therefore the path difference $\delta = \frac{2\mu t}{\cos r} - \frac{2\mu t \sin^2 r}{\cos r}$

$$\delta = \frac{2\mu t}{\cos r}\left[1 - \sin^2 r\right]$$

$$\delta = \frac{2\mu t}{\cos r}\cos^2 r$$

$$\delta = 2\mu t \cos r \quad \ldots\ldots (10)$$

Since the ray AR is reflected at the air medium (rarer- denser) interface, it undergoes a phase change of π or path increases of $\lambda/2$.

Hence the path difference between the reflected rays AR and DR' is

$$2\mu t \cos r - \frac{\lambda}{2} \ldots\ldots (11)$$

Conditions of maxima and minima in reflected light:

1. *Condition for bright band:*

The film will appear bright if the path difference

$$2\mu t \cos r - \frac{\lambda}{2} = n\lambda$$

$$2\mu t \cos r = (2n+1)\frac{\lambda}{2} \quad \text{Where n=0, 1, 2, 3 ……}$$

2. *Condition for dark band:*

The film will appear dark if the path difference

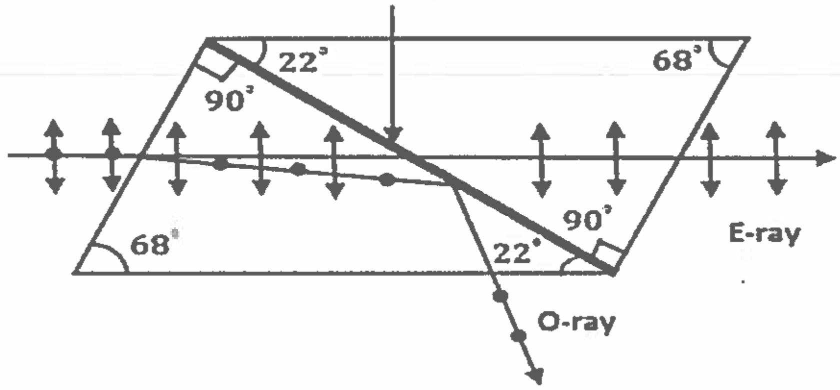
$$2\mu t \cos r - \frac{\lambda}{2} = (2n-1)\frac{\lambda}{2}$$

$$2\mu t \cos r = n\lambda$$

7. Nicol Prism:

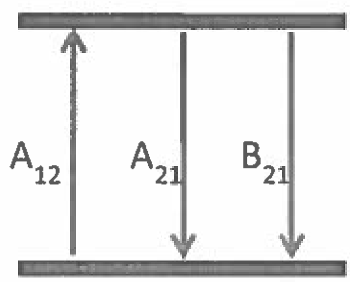
Nicol prism is an optical device made from calcite crystal used to produce plane polarized light. Calcite crystal is modified such that it eliminates one of the two refracted rays by total internal reflection.

Construction: Nicol prism is constructed as shown in figure, from calcite crystal whose length is nearly three times of its width. The end faces of the crystal are cut down such that the angles of principal section are 68° and 112° instead of 71° and 109°. The crystal is then cut diagonally and the two surfaces are polished and cemented with Canada balsam which is a transparent material. Canada balsam is optically denser for extraordinary rays and less dense for ordinary rays.



Working: When an unpolarized light incident on Nicol prism, it splits into two rays Extraordinary ray and ordinary ray. Both the rays are plane polarized. Because of construction of Nicol prism, the ordinary ray incident on the Canada Balsam interface at angle greater than critical angle and suffers total internal reflection. The extraordinary ray transmitted through Canada Balsam and emerges out of the prism. In this way plane polarized light is obtained.

8(a)



Let us consider two energy states E_1 and E_2 such that $E_2 > E_1$.

Assume N_1 and N_2 be the number of atoms in the lower and higher energy states respectively and $N_1 > N_2$.

When a photon of energy is equal to the difference between two energy states i.e., $E_2 - E_1 = h\upsilon$ be incident on any one of the atom in the lower energy state, the atom may absorb the photon and excite and moves to the higher energy state.

The Probable rate of transition from lower energy state to higher energy state by absorption process is

$$(P_{12})_{ab} = A_{12}u(\upsilon)$$

Where A_{12} is Einstein coefficient of absorption, $u(\upsilon)$ is energy density corresponding to frequency difference of energy states E_1 and E_2.

The number of atoms in the lower energy state that absorb photons and rise to higher energy state is

$$N_1(P_{12})_{ab} = N_1 A_{12} u(\upsilon)$$

The probable rate of transition from higher energy state to lower energy state by spontaneous emission process is

$$(P_{21})_{sp} = A_{21}$$

Where A_{12} is a constant called Einstein coefficient for spontaneous emission of radiation

The number of atoms that moves from the higher energy state to lower energy state is

$$N_2(P_{21})_{sp} = N_2 A_{21}$$

The probable rate of transition from higher energy state to lower energy state by stimulated emission process is

$$(P_{21})_{st} = B_{21}u(\upsilon)$$

Where B_{21} is a constant called Einstein coefficient for simulated emission of radiation, $u(\upsilon)$ is the incident energy supplied to the atoms in the lower energy state.

The number of atoms that moves from higher energy state to lower energy state is

$$N_2(P_{21})_{st} = N_2 B_{21} u(\upsilon)$$

At thermal equilibrium, the rate of absorption is equal to sum of the rate of spontaneous and stimulated emissions.

$$N_1 A_{12} u(\upsilon) = N_2 A_{21} + N_2 B_{21} u(\upsilon)$$

$$[N_1 A_{12} - N_2 B_{21}]u(\upsilon) = N_2 A_{21}$$

$$u(\upsilon) = \frac{N_2 A_{21}}{[N_1 A_{12} - N_2 B_{21}]}$$

$$u(\upsilon)=\frac{N_2A_{21}}{N_2B_{21}\left[\frac{N_1A_{12}}{N_2B_{21}}-1\right]}$$

$$u(\upsilon)=\frac{A_{21}}{B_{21}}\frac{1}{\left[\frac{N_1A_{12}}{N_2B_{21}}-1\right]}$$

We know that $\frac{N_1}{N_2}=e^{\frac{E_2-E_1}{kT}} \quad \Rightarrow \frac{N_1}{N_2}=e^{\frac{h\upsilon}{kT}}$

$$u(\upsilon)=\frac{A_{21}}{B_{21}}\frac{1}{\left[\frac{A_{12}}{B_{21}}e^{\frac{h\upsilon}{kT}}-1\right]}$$

According to Planck's radiation law, the radiation density per unit volume is

$$u(\upsilon)=\frac{8\pi h\upsilon^3}{c^3}\frac{1}{\left[e^{\frac{h\upsilon}{kT}}-1\right]}$$

From equation (5) and (6), we get

$$\frac{A_{12}}{B_{21}}=1 \Rightarrow A_{12}=B_{21}$$

The probability of stimulated emission is same as that of the probability of absorption.

$$\frac{A_{21}}{B_{21}}=\frac{8\pi h\upsilon^3}{c^3}$$

$$\frac{A_{21}}{B_{21}}\propto\upsilon^3$$

This shows that the probability of spontaneous emission and stimulated emission increases rapidly with energy difference between two states.

(b) Lasers have very wide range of applications. Lasers are used in:

Scientific studies: Isotope separation, Plasma generation and study

Defense: Laser guided missiles, RADARs

Industries: Drilling high quality holes, high quality welding. high quality cutting.

9(a) The maximum angle of incidence at the end face of an optical fiber for which the light ray can be propagated along core-cladding interface is known as maximum Acceptance angle. It is also called acceptance cone half angle.

Consider a ray of light travelling along a medium of refractive index n_0, incident at air-core interface of the optical fiber and making an angle θ_i with the axis of the fiber. It is refracted into the core of refractive index n_1 with angle of refraction θ_r. This ray makes an angle ϕ with the normal at the core-cladding interface and is totally reflected into the core as shown in figure.

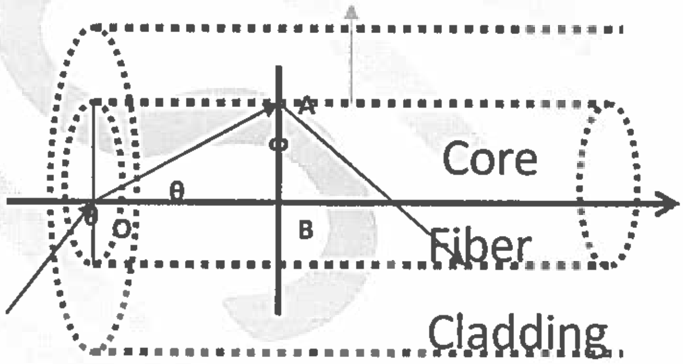
If ϕ is the greater than the critical angle θ_c, the ray undergoes total internal reflection at the interface, since $n_1>n_2$. As long as the angle ϕ is greater than $\theta_c$the light will stay within the fiber.

According to Snell's law

$$n_0 \sin\theta_i = n_1 \sin\theta_r$$

$$\sin\theta_i = \frac{n_1}{n_0}\sin\theta_r \quad .... (1)$$

If $\theta_i$is increased beyond a limit, ϕ will decrease below the critical angle θ_c and ray escapes from the side walls of the fiber.



From Δ^{le} OAB, $\quad \phi + \theta_r = 90^0$

$$\theta_r = 90^0 - \phi$$

$$\sin\theta_r = \sin(90^0 - \phi)$$

$$\sin\theta_r = \cos\phi \quad .... (2)$$

Substituting equation (2) in equation (1), we get

$$\sin\theta_i = \frac{n_1}{n_0}\cos\phi \quad .... (3)$$

When $\phi = \theta_c$

$$\sin\theta_{i(\max)} = \frac{n_1}{n_0}\cos\theta_c \qquad \text{.......... (4)}$$

But the condition for total internal reflection, $\sin\theta_c = \frac{n_2}{n_1}$

$$\cos\theta_c = \sqrt{1-\sin^2\theta_c}$$

$$\cos\theta_c = \sqrt{1-(\frac{n_2}{n_1})^2}$$

$$\cos\theta_c = \frac{\sqrt{n_1^2-n_2^2}}{n_1}$$

Substituting $\cos\theta_c$ in equation (4)

$$\sin\theta_{i(\max)} = \frac{n_1}{n_0}\frac{\sqrt{n_1^2-n_2^2}}{n_1}$$

$$\sin\theta_{i(\max)} = \frac{\sqrt{n_1^2-n_2^2}}{n_0}$$

Representing $\sin\theta_{i(\max)}$ as θ_a

$$\sin\theta_a = \frac{\sqrt{n_1^2-n_2^2}}{n_0}$$

For air medium n_0=1, then

$$\sin\theta_a = \sqrt{n_1^2-n_2^2}$$

$$\theta_a = \sin^{-1}\sqrt{n_1^2-n_2^2}$$

This is required expression for Maximum Acceptance Angle in optical fibers.

The angle $\theta_a$is called the acceptance angle of the fiber. The acceptance angle may be defined as the maximum angle of incidence that light ray makes with the axis of the fiber to get the total internal reflection. It is also called Acceptance cone half angle.

(b) 1.A much greater amount of information can be carried on an optical fiber compared to a copper cable.

2. In all cables some of the energy is lost as the signal goes along the cable. Then the signal need to be boosted using regenerators. These are required every 2 to 3 km for copper cables and every 20 to 50 km for optical fibers.

3. optical fibers are used in endoscopic applications.

4. Signal tampering is not possible.

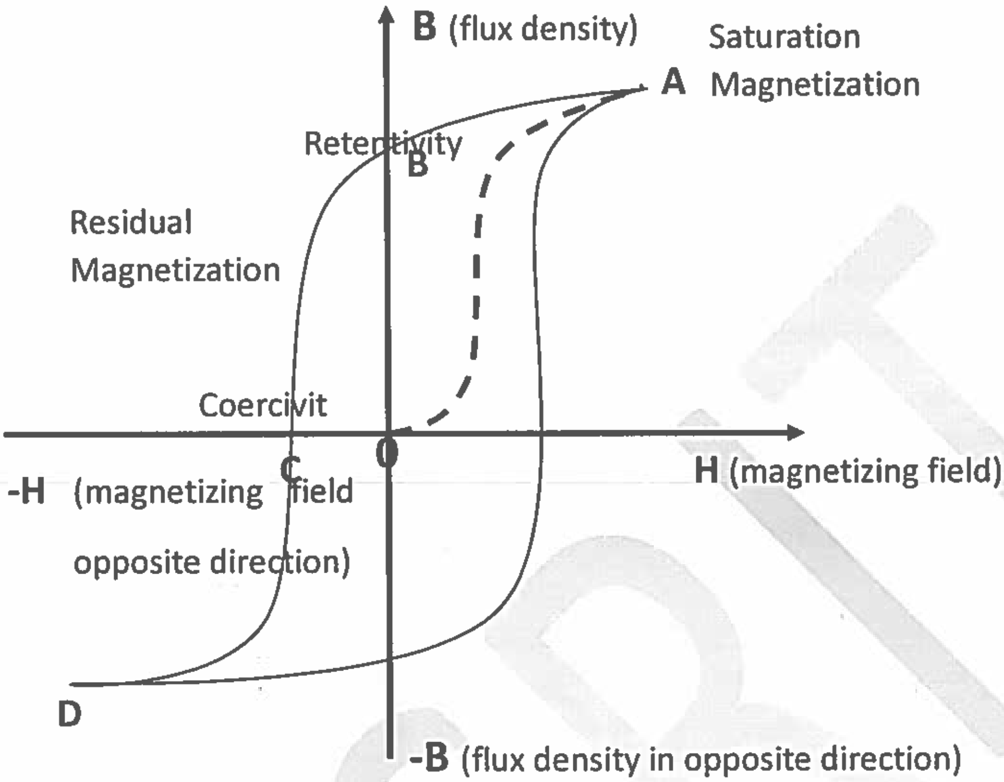
10(a)

A typical property of ferromagnetic material is hysteresis. Hysteresis may be defined as the lag in the change of magnetization behind the variation of the magnetic field. It gives the relationship between the induced magnetic flux density (B) and the magnetizing field (H), often referred as the B-H loop or I-H loop.

Consider an unmagnetized ferromagnetic material is placed in a magnetizing field. When the material is slowly magnetized and the magnetic flux density (B) increases with increase of magnetizing field (H) initially through OA and reaches saturation at A.

When H is decreased, B decreases but it does not comes to zero at H=0. The residual flux density (B) set up in the material represented by OB is called retentivity. To bring B to zero, opposite magnetizing filed is applied. This magnetizing field represented by OC is called coercivity. After reaching the saturation level D, when the magnetizing field is reversed, the curve closes to the point A, completing a cycle. The loop

OABCDEFA is called hysteresis loop.

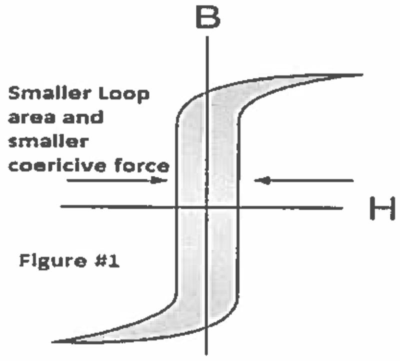


The area of the hysteresis loop gives the loss of energy due to the cycle of magnetization and demagnetization and is dissipated in the form of heat. The retentivity and coercivity of the hysteresis loop are the characteristics of different ferromagnetic materials.

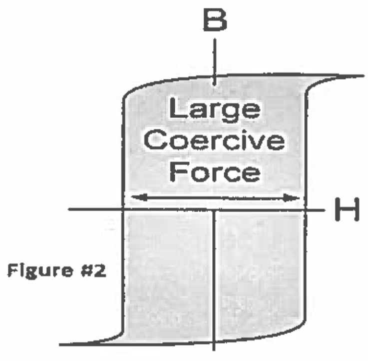
From the hysteresis loop, the following properties of a magnetic material can be determined.

(b) On the basis of magnetization process, the magnetic materials can be divided into two classes.

i. Soft magnetic materials ii. Hard magnetic materials



Soft Magnetic Material



Hard Magnetic Material

Properties of soft magnetic materials:

i. They are easily magnetized and demagnetized.

ii. They have small hysteresis loss due to small hysteresis loop area.

iii. These materials have large values of permeability and susceptibility.

iv. Coercivity and retentivity are small.

v. In the soft magnetic materials, the domain walls must be able to move easily and reversibly so that magnetization changes by large amounts for small changes in the magnetic field.

vi. They are used to make electromagnets.

Eg: Iron-Silicon alloy, Iron-Nickel alloy, Aluminum-Silicon-Iron alloy, ferrites and garnets.

Properties of Hard magnetic materials:

i. They cannot be easily magnetized and demagnetized.

ii. These materials have small values of permeability and susceptibility.

iii. These materials have large hysteresis loss due to large hysteresis loop area.

iv. The coercivity and retentivity are large.

v. The eddy current loss is low due to its high resistivity.

vi. In these materials, the domain wall moment is difficult owing to crystal imperfections and irreversible in nature.

Eg: Aluminum-Nickel-Cobalt-Copper-Iron alloy (Alnico), copper-nickel-iron alloy (Cunife), copper-nickel-cobalt alloy (Cunico), Sm-Co alloy.

11(a)

It gives the relation between dielectric constant and polarizability at atoms in a dielectric material.

The dipole moment of an atom is proportional to the internal field.

$$\mu = \alpha_e E_i$$

If there are N atoms per unit volume, then the polarization is given by

$P = N\alpha_e E_i$ and $E_i = E + \dfrac{P}{3\varepsilon_0}$

$$P = N\alpha_e (E + \frac{P}{3\varepsilon_0})$$

$$P = N\alpha_e \left(\frac{3\varepsilon_0 E + P}{3\varepsilon_0} \right)$$

$$\frac{N\alpha_e}{3\varepsilon_0} = \left(\frac{P}{3\varepsilon_0 E + P} \right) \quad \ldots\ldots (1)$$

We know that $D = \varepsilon_0 E + P \Rightarrow \dfrac{P}{E} = \dfrac{D}{E} - \varepsilon_0$

From the definition of electric displacement vector

$$D = \varepsilon E$$

$$D = \varepsilon_r \varepsilon_0 E \qquad (\because \varepsilon_r = \frac{\varepsilon}{\varepsilon_0})$$

$$\therefore \frac{P}{E} = \frac{\varepsilon_r \varepsilon_0 E}{E} - \varepsilon_0$$

$$P = \varepsilon_0 E(\varepsilon_r - 1) \quad \ldots\ldots (2)$$

From equations (1) and (2)

$$\frac{N\alpha_e}{3\varepsilon_0} = \frac{\varepsilon_0 E(\varepsilon_r - 1)}{3\varepsilon_0 E + \varepsilon_0 E(\varepsilon_r - 1)}$$

$$\frac{N\alpha_e}{3\varepsilon_0} = \frac{\varepsilon_0(\varepsilon_r - 1)\,\mathrm{E}}{\varepsilon_0(\varepsilon_r + 2)\,\mathrm{E}}$$

$$\frac{N\alpha_e}{3\varepsilon_0} = \frac{\varepsilon_r - 1}{\varepsilon_r + 2}$$

This is called Classius-Mossotti Equation

(b) 1. Insulating materials: Dielectric materials can be used as insulating materials.

The material should have low dielectric constant, low dielectric loss, high dielectric strength and high resistance.

2. Capacitors: Dielectric materials are used to prepare dielectric capacitors which have higher capacity value and also can be operated at higher voltages.

3. Dielectric heating: Dielectric loss in the material manifests in the form heat. This principle can be used to manufacture microwave ovens.

12(a)

Consider a system of stationary waves associated with a particle. Let x, y, z be the coordinates of the particle and ψ be the wave displacement for the *de-Broglie* at any time t. Here ψ is called wave function. It is assumed that ψ is finite, single valued and periodic function.

The classical differential equation of a wave motion is given by

$$\frac{\partial^2\psi}{\partial t^2} = v^2\left[\frac{\partial^2\psi}{\partial x^2}+\frac{\partial^2\psi}{\partial y^2}+\frac{\partial^2\psi}{\partial z^2}\right]$$

$$\frac{\partial^2\psi}{\partial t^2} = v^2\nabla^2\psi \qquad \text{..... (1)}$$

where $\nabla^2 = \frac{\partial^2}{\partial x^2}+\frac{\partial^2}{\partial y^2}+\frac{\partial^2}{\partial z^2}$ being a Laplacian operator and v is the wave velocity.

The solution of equation (1) is given by

$$\psi = \psi_0 e^{-i\omega t} \qquad \text{..... (2)}$$

where ω is the angular velocity of the particle

Differentiating equation (2) with respect to t

$$\frac{\partial\psi}{\partial t} = \psi_0(-i\omega)e^{-i\omega t}$$

$$\frac{\partial^2\psi}{\partial t^2} = \psi_0(i\omega)^2 e^{-i\omega t}$$

$$\frac{\partial^2\psi}{\partial t^2} = -\psi_0\omega^2 e^{-i\omega t}$$

$$\frac{\partial^2\psi}{\partial t^2} = -\omega^2\psi \qquad \text{....... (3)}$$

Substituting the equation (3) in (1)

$$-\omega^2\psi = v^2\nabla^2\psi$$

$$\Rightarrow \nabla^2\psi = \frac{-\omega^2}{v^2}\psi \quad \ldots\ldots (4)$$

We know that $\omega = 2\pi\upsilon$

$$\omega = \frac{2\pi v}{\lambda} \; (\because \upsilon = \frac{v}{\lambda})$$

$$\frac{\omega}{v} = \frac{2\pi}{\lambda}$$

$$\frac{\omega^2}{v^2} = \frac{4\pi^2}{\lambda^2} \quad \ldots\ldots (5)$$

Substituting equation (5) into equation (4), we get

$$\nabla^2\psi = -\frac{4\pi^2}{\lambda^2}\psi$$

$$\nabla^2\psi + \frac{4\pi^2}{\lambda^2}\psi = 0 \quad \ldots\ldots. (6)$$

But *de-Broglie's* wavelength $\lambda = \frac{h}{mv}$

$$\Rightarrow \frac{1}{\lambda} = \frac{mv}{h}$$

$$\Rightarrow \frac{1}{\lambda^2} = \frac{m^2v^2}{h^2} \quad \ldots\ldots (7)$$

Substituting equation (7) in equation (6), we get

$$\nabla^2\psi + \frac{4\pi^2 m^2 v^2}{h^2}\psi = 0 \quad \ldots\ldots (8)$$

If E and V are the total and potential energies of the particle respectively, then its kinetic energy is

$$\frac{1}{2}mv^2 = E - V$$

$$\Rightarrow mv^2 = 2(E-V)$$

$$\Rightarrow m^2v^2 = 2m(E-V) \quad \ldots\ldots. (9)$$

Substituting equation (9) equation (8), we get

$$\nabla^2\psi + \frac{8\pi^2 m}{h^2}(E-V)\psi = 0 \quad \ldots\ldots (10)$$

This equation is known as Schrodinger time-independent wave equation.

(b) Physical significance of wave function: According to Max Born, ψ2 gives the probability of finding the particle at a point at any instant. ψ2 may be real or imaginary .But probability must be real and hence we represent it as |ψ|2 or ψ*ψ.

Ψ must have the following properties,

(i) Ψ must be finite for all values of x,y,z.

(ii) Ψ must be single valued. i.e. for each set of x,y,z values it must have only one value.

(iii)Ψ must be continuous in all regions except where the potential energy is infinite.

(iv)Ψ is analytical i.e. it possess continuous first order derivative.

(v) Ψ must vanish at boundaries.

13(a)

1. It is a macroscopic theory.
2. It cannot explain the electrical conductivity of semiconductors and insulators properly.
3. Dual nature is not explained.
4. It cannot explain the Compton effect, Photo-electric effect.
5. The theoretical and experimental values of specific heat are not matched.
6. Atomic fine spectra could not be accounted.
7. Different types of magnetisms could not be explained satisfactorily by this theory.

(b)

Fermi-Dirac distribution function *F(E)* is used to calculate the probability of an electron occupying a certain energy level.

The distribution of electrons among different energy levels as a function of temperature is known as Fermi-Dirac distribution function.

$$F(E) = \frac{1}{1 + \exp\left(\frac{E - E_F}{k_B T}\right)}$$

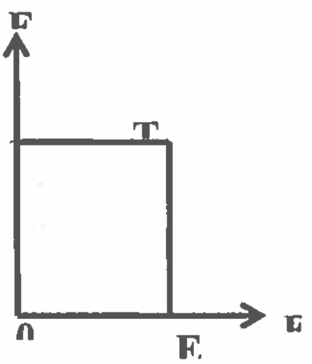
Where E = Energy of allowed state

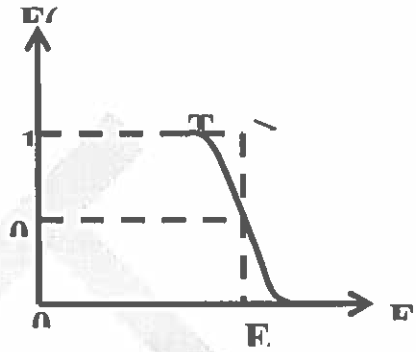
E_F= Fermi energy

k_B = Boltzmann constant

T = Temperature in K

The probability value of *F(E)* lies between 0 and 1





If $F(E) = 1$, the energy level is occupied by an electron

If $F(E) = 0$, the energy level is vacant

If $F(E) = 0.5$ or $1/2$, then there is a 50% chance for finding the electron in the energy level

Effect of temperature on Fermi-Dirac distribution function:

At *0K*, the electrons are filled up to a maximum energy level called Fermi energy level E_F. All the energy levels above the Fermi energy level are empty.

Case i: At T = 0K and $E < E_F$

$$F(E) = \frac{1}{1+e^{-\infty}} = \frac{1}{1} = 1$$

Therefore, the probability of electrons to occupy the energy level between Fermi energy level is 100%.

Case ii: At $T = 0K$ and $E > E_F$

$$F(E) = \frac{1}{1+e^{\infty}} = \frac{1}{1+\infty} = 0$$

This means that at *0K*, electrons are completely occupied below and above E_F electrons are occupied.

Case iii: At $T = 0K$ and $E = E_F$

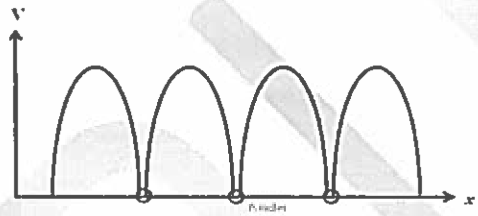
$$F(E) = \frac{1}{1+1} = \frac{1}{2} = 0.5$$

The Fermi level in a metal is the energy level for which the probability occupation is half.

14(a)

A crystalline solid consists of a lattice which is composed of a large number of positive ion cores at regular intervals and the conduction electrons can move freely throughout the lattice. If the electron moves through the crystal lattice, it experiences periodic potentials. The potential of an electron at the positive ion site is minimum and is maximum in between two ions. Since the lattice is a periodic structure of the ion arrangement in a crystal, the variation of potential V of an electron also periodic.

The behavior of an electron in crystalline solid can be described by using Schrodinger wave equation. This leads to concept of energy bands in solids.



The one dimensional Schrodinger wave equation can be written as

$$\frac{d^2\psi}{dx^2}+\frac{8\pi^2 m}{h^2}(E-V(x))\psi=0 \quad ..... (1)$$

Where the periodic potential V(x) may be defined by means of lattice constant 'a' as

$$V(x)=V(x+a)$$

Applying the periodic potential, Bloch has shown that one dimensional solution of the Schrodinger equation takes the form

$$\psi(x)=e^{ikx}u_k(x) \quad ..... (2)$$

Where $u_k(x)$ has the same periodicity of the lattice given by

$$u_k(x)=u_k(x+a)$$

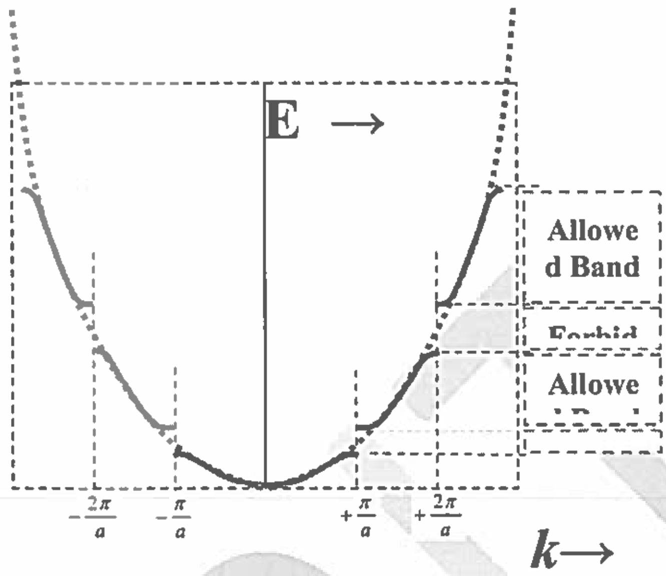
Here k represents the state of motion of the electron called propagation vector and k^{th} state corresponding to an electron having momentum $P=\frac{hk}{2\pi}$ and $k=\frac{2\pi}{\lambda}$, λ is *de-Broglie's* wavelength. The solution of equation (2) consists of a plane wave e^{ikx} modulated by the periodic function.

In three dimensions the solution of the Schrodinger wave equation is given by

$$\psi_k(r)=e^{ikr}u_k(\mathrm{r}) \quad ..... (3)$$

The equations (2) and (3) are known as Bloch functions in one and three dimensions respectively.

(b)



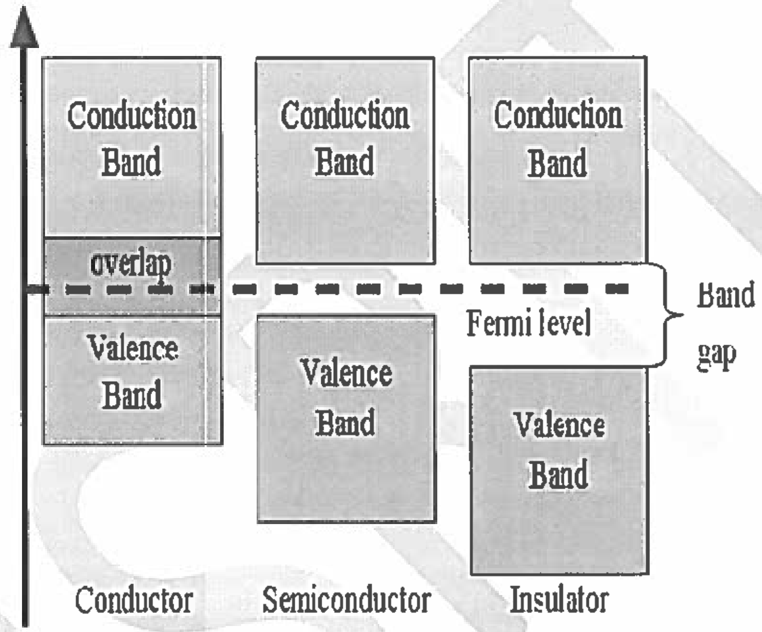
The allowed energy region between $-\frac{\pi}{a}$ to $\frac{\pi}{a}$ be known as first Brillouin zone. The allowed energy region between $-\frac{\pi}{a}$ to $\frac{-2\pi}{a}$ and $\frac{\pi}{a}$ to $\frac{2\pi}{a}$ is known as second Brillouin zone and so on. The discontinuity in energy are called forbidden region (bands). The discontinuity appears in the *E-K* graph at $k = \pm\frac{n\pi}{a}$ where $n=1, 2, 3\ldots$

15(a)

Depending on the nature of band occupation by electron and on the width of the forbidden band, the solids can be classified into conductors, semiconductors and insulators. The metals are good conductors of electricity while the insulators are bad conductors of the electricity. The electrical conductivity of semiconductors lies between that of a metal and insulators. The band theory of solids can explain the electrical conductivity of a solid. The completely filled bands and completely empty bands do not contribute to the electrical conduction. The valence band, conduction band and forbidden band are important for the electrical properties of solid.

Insulators:

In insulators, the conduction band is completely empty, the valence band is completely filled and there is a large energy gap ($E_g > 2eV$) between conduction band and valence band. When an electric field is applied, there is no new energy level available to the electron and there is no conduction of electricity. Because of the large band gap, the transition of electron from valence band to conduction band is also not possible. At room temperature, the thermal energy (k_BT) is much less than the band gap energy. The diamond is a perfect insulator having band gap of $5.5\ eV$.



Semiconductors:

In semiconductors, the conduction and valence bands are partially filled at room temperature. The energy gap between the valence band and the conduction band is small as compared to that of insulator. Due to the small energy gap, some of the valence band electrons make transitions to the conduction band by acquiring thermal energy. These electrons leave an equal number of vacant states or holes in the valence band. These holes behave like positive charge and also contribute to the conduction of electricity. The conductivity is in between that of insulators and conductors. The examples for semiconductors are silicon and germanium having band gaps $1.1eV$ and $0.7eV$ respectively. At absolute zero temperature, all the semiconductors are insulators. The conductivity increases with increase of temperature. Hence semiconductors have negative temperature coefficient of resistance.

Conductors:

In conductors, the valence band and conduction band overlap and there is no energy gap between them. At room temperature, the free electrons exist in the conduction band hence conductivity is high. The resistance increases (conductivity decreases) with increase of temperature hence conductors have a positive temperature coefficient resistance. The total current in the conductors is simply a flow of electrons. Metals are the best examples for conductors.

(b) ApplicationsofHallEffect

1. To determine thetype ofgivensemiconductor.
2. TodeterminemagneticfluxdensityB
3. Todeterminecarrierconcentrationof nandp
4. Todetermine mobilityofchargecarriers(μ)

Semester End Supplementary Examination, June, 2022

Degree	B. Tech. (U. G.)	**Program**	CE / ME			**Academic Year**	2021 - 2022
Course Code	20BSX21	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	I
Course	ENGINEERING CHEMISTRY						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define carbonate and non carbonate hardness of water.	20BSX21.1	L1
2	Write the principle of electroless plating.	20BSX21.2	L1
3	How do you rate the quality of petrol?	20BSX21.3	L1
4	Outline the synthesis of Buna-N and write any two applications.	20BSX21.4	L2
5	What is colloid? Give two examples.	20BSX21.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	With a well labeled diagram, explain break point chlorination. Give the advantages.	6M	20BSX21.1	L1
6 (b)	What is caustic embrittlement? How is it caused? Explain.	6M	20BSX21.1	L2
	OR			
7 (a)	Demonstrate the softening of water by zeolite method.	6M	20BSX21.1	L1
7 (b)	Explain the principle and procedure of the estimation of hardness of water by complexometric titration.	6M	20BSX21.1	L2
8 (a)	Define electrochemical series. How is it established? Discuss any three applications of the series.	6M	20BSX21.2	L2
8 (b)	Illustrate the construction and working of phosphoric acid fuel cell. Give its merits and limitations.	6M	20BSX21.2	L2
	OR			
9 (a)	Write the principle of cathodic protection. Explain how metals can be protected in different ways by cathodic protection.	6M	20BSX21.2	L2
9 (b)	Rate of corrosion of a metal depends on nature of metal and its environment. Substantiate.	6M	20BSX21.2	L2
10 (a)	How the percentage of carbon, hydrogen, nitrogen and sulphur in coal are determined? Explain.	6M	20BSX21.3	L2
10 (b)	Vegetable oils cannot be used directly as fuel in IC engines. Give reasons. How do you overcome this problem and convert them in to biodiesel? Explain the chemistry involved.	6M	20BSX21.3	L2
	OR			
11 (a)	Describe the refining of crude oil in to different fractions.	6M	20BSX21.3	L2
11 (b)	Define gross and net calorific value of a fuel. Why GCV is higher than NCV? Compute the HCV and LCV of a fuel having the following composition. 81% carbon, 6.2% hydrogen, 2.6% nitrogen, 1.1% sulphur and 2.4% ash.	6M	20BSX21.3	L2
12 (a)	Discuss about the mechanism of coordination polymerization.	6M	20BSX21.4	L2

12 (b)	Define and classify refractory materials. Enlist their properties and applications.	6M	20BSX21.4	L2
	OR			
13 (a)	Classify lubricants and give examples. Enumerate the functions of lubricants.	6M	20BSX21.4	L2
13 (b)	Discuss the chemistry of setting and hardening of Portland cement.	6M	20BSX21.4	L2
14 (a)	Explain the synthesis of nanometal oxides by an electrochemical method. Enumerate the medical applications of nanomaterials.	7M	20BSX21.5	L2
14 (b)	Explain, how the colloids and nanomaterials are stabilized.	5M	20BSX21.5	L2
	OR			
15 (a)	With a neatly labeled block diagram, explain the principle, instrumentation and applications of Scanning Electron Microscopy.	5M	20BSX21.5	L2
15 (b)	Discuss about nano sensors and their applications.	7M	20BSX21.5	L2

SEMESTER Question Paper

Degree	B. Tech. (U. G.)	**Program**	CE/ME	Test	I/I-Supply	**Academic Year**	2021 - 2022
Course Code	20BSX21	**Test Duration**	90 Min.	Max. Marks	70	**Semester**	I
Course	**ENGINEERING CHEMISTRY**						

Key and Scheme of Evaluation

No.	Questions (1 through 5)	Marks
1	Temporary Hardness mainly caused by the presence of dissolved bicarbonates of calcium, magnesium and other heavy metals.Permanent Hardness: It is due to the presence of dissolved chlorides and sulphates of calcium, magnesium, iron and other metals.	2
2	Electroless plating. The method of deposition of a metal from its salt solution on a catalytically active surface by a suitable reducing agent without using electrical energy is called electroless plating	2
3	The quality or rating of diesel is expressed by cetane number. Cetane is n- Hexadecane (C16H34)	2
4	It is used for manufacturing of tyres. It is used in the footwear industry for making shoe soles and footwear components	2
5	**Define cracking** decomposition of bigger hydrocarbon molecules into simpler, low boiling hydrocarbons of low molecular weight	2

No. Questions (6 through 11)

6 (a)

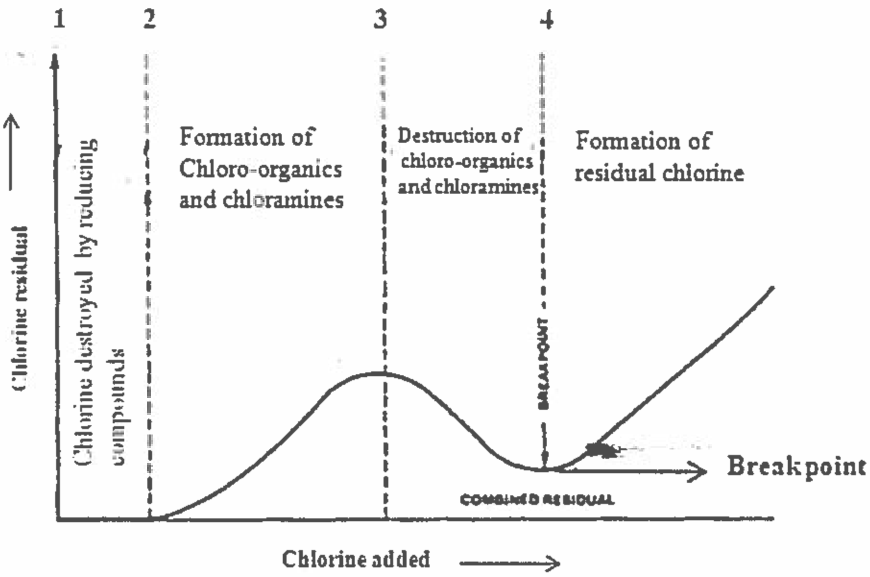
Break point chlorination:

The amount of chlorine required to kill bacteria and to remove organic matter is called break point chlorination.

A typical relationship between the amount of chlorine added to water and the experimentally determined free residual chlorine is shown in the graph below which explains the break point chlorination.

(i) First (between points 1 and 2), the water reacts with reducing compounds in the water, such as hydrogen sulfide. These compounds use up the chlorine, producing no chlorine residual.

6M



(ii) Between points 2 and 3, the chlorine reacts with organics and ammonia naturally found in the water and forms chloro-organics and chloramines.

(iii) Between points 3 and 4, the added chlorine will break down most of the chloramines in the water, leaving behind free chlorine which destroys pathogenic bacteria, actually lowering the chlorine residual.

(iv) Finally, the water reaches the breakpoint, at point 4 (minima). The breakpoint is the point at which the chlorine demand has been totally satisfied - the chlorine has reacted with all reducing agents, organics, ammonia and pathogenic bacteria in the water. After minima when more chlorine is added, it is not used in any reaction. Thus, the residual chlorine keeps on increasing in proportion to added chlorine. Hence for effective killing of microorganisms, sufficient amount of chlorine has to be added. Addition of chlorine in such dosages is known as break point chlorination.

Advantages of break –point chlorination:

1. It ensures complete destruction of organic compounds which imparts colour, bad odour and unpleasant taste to water.
2. It completely destroys all the disease producing bacteria.
3. It prevents the growth of any weeds in water.

Caustic embrittlement: Caustic embrittlement is a term used for the appearance of cracks inside the boilers particularly at those places which are under stress such as riveted joints due to the high concentration of alkali leading to failure of the boiler. .

Reasons:

It is a type of boiler corrosion caused due to the presence of alkali-metal carbonates and bicarbonates in feed water and also the presence of sodium carbonate.

$$Na_2CO_3 + H_2O \rightarrow NaOH + CO_2$$

This caustic water flows inside the boiler and causes some minute hair-cracks, by capillary action. On evaporation of water, the dissolved caustic soda increases its concentration which attacks the surrounding area, thereby dissolving iron of boiler as sodium ferroate (Na_2FeO_2). This causes embrittlement of boiler parts such as bends, joints, rivets etc, due to which the boiler gets fail. 2m

Caustic cracking can be explained by considering the following concentration cell. The iron at plane surfaces surrounded by dilute NaOH becomes cathodic and the iron at bends, joints and rivets surrounded by highly 2m

6 (b) concentrated NaOH becomes anodic which consequently gets corroded

1m

+(Anode) Iron at rivets , bends joints..etc	Conc NaOH	Dil NaOH	-(Cathode) Iron at plane surfaces

1m

Prevention:

- By maintaining the pH of water and neutralization of alkali.
- By using sodium phosphate as softening reagents instead of sodium carbonate in the external treatment of boiler feed water.

Caustic embrittlement can also be prevented by adding tannin or lignin or Sodium

OR

ZEOLITE PROCESS (Permutit Process)

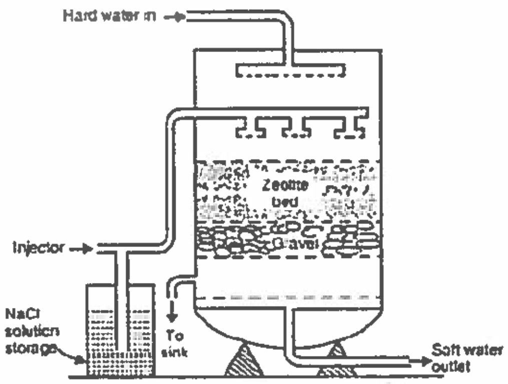
Zeolite is a 3D silicate. The chemical composition of zeolites is hydrated sodium aluminium silicate, represented as $Na_2O.Al_2O_3.xSiO_2.yH_2O$ where x = 2-10 and y = 2-6. Zeolites are capable of exchanging reversibly its sodium ions for hardness causing Ca^{2+} and Mg^{2+} in water. Hence zeolites are cation exchangers. 2m

Zeolites are of two types 2m

7 (a) 1) Natural zeolites are natural and non-porous having the composition 2m

Natural zeolite: $Na_2O\ Al_2O_3\ .4SiO_2\ .2H_2O$

Synthetic zeolites are porous and posses gel structure and are prepared from china clay, feldspar and a soda ash. Synthetic zeolites posses higher exchange capacity. Zeolites have cage like structure. It is derived from SiO_4



Zeolite Process

Process: The hard water is passed through a zeolite bed fixed in a cylinder at a specific rate.The hardness causing ions Ca^{2+}, Mg^{2+} etc are retained by the zeolite as CaZe and MgZe respectively

The reactions taking place during softening process are

$$Na_2Ze + Ca(HCO_3)_2 \rightarrow CaZe + 2NaHCO_3$$
$$Na_2Ze + Mg(HCO_3)_2 \rightarrow MgZe + 2NaHCO_3$$
$$Na_2Ze + CaCl_2 \rightarrow 2NaCl + CaZe$$
$$Na_2Ze + MgCl_2 \rightarrow 2NaCl + MgZe$$

Regeneration: After the use of this process for certain time zeolite is exhausted i.e. all Na^+ ions of the zeolite are replaced by Ca^{+2} and Mg^{+2} ions and hard water will not be further softened.

Regeneration:

Exhausted zeolite can be regenerated by heating it with brine solution (10% NaCl)

$CaZe + 2NaCl \rightarrow Na_2Ze + CaCl_2$

$MgZe + 2NaCl \rightarrow Na_2Ze + MgCl_2$

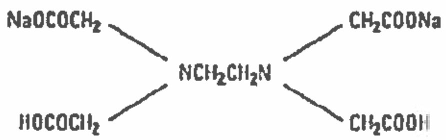
Estimation of Hardeness of Water by EDTA method (Complexometric titration)

This is a complexometric titration where ethylene diamine tetra acetic acid [EDTA] is used as a complexing agent. EDTA form complexes with different metal ions at different PH 9-10. To maintain the PH 9-10 ammonical buffer is used. An alcoholic solution of Eriochrome black –T is used as an indicator. The disodium salt of EDTA is used for complexation since solubility of EDTA is very low.

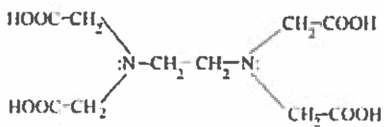
2m

Structure of EDTA:

7 (b)



Di-sodium salt of EDTA



structure of EDTA

2m

Principle: In this complexometric titration, the water sample is titrated with standard solution of Di-sodium salt of EDTA using EBT indicator. First EDTA solution is standardized with a standard solution of calcium carbonate, prepared by dissolving a known weight of cacium carbonate in dil.HCl and then making up the solution to a known volume with distilled water .The permanent hardness of water can be determined by titrating the water after boiling well to remove the temporary hardness as carbonates of calcium and magnesium against EDTA by adding EBT and buffer solution. Temporary 2m
hardness can be determined by subtracting total hardness of water from permanent hardness of water.

Reactions involved: EBT indicator when added to hard water at pH = 10, forms weak complexes with calcium and magnesium present in hard water. It results in the formation of unstable wine red Ca-EBT or Mg-EBT complexes. During titration with EDTA, calcium first reacts to form relatively stable complex followed by magnesium to give Mg^{+2}-EDTA complex releasing the free indicator(blue).The colour changes from wine-red to blue at the endpoint.

$$\left.\begin{array}{l} Ca^{2+} \\ \\ Mg^{2+} \end{array}\right\} + \text{Eriochrome black - T} \rightarrow \left[\begin{array}{l} Ca^{2+} \\ \quad \text{Eriochrome black - T} \\ Mg^{2+} \end{array}\right] \text{Complex}$$

(of water) (Unstable complex) (Wine-red)

$$\xrightarrow{\text{EDTA}} \left[\begin{array}{l} Ca^{2+} \\ \quad \text{EDTA} \\ Mg^{2+} \end{array}\right] \text{Complex} + \text{Eriochrome black - T}$$

(blue)

Various steps involved in this method are...

1. Preparation of standard hard water: Dissolve 1g of pure, dry CaCO3 in minimum quantity of dil. HCl and then evaporate the solution to dryness on a water bath. Dissolve the residue in distilled water to make 1 liter solution. Each mL of this solution thus contains 1mg of CaCO3 equivalent hardness.

1 mL hard water solution = 1mg of CaCO3 equivalent hardness.

2. Standardization of EDTA solution: Rinse and fill the burette with EDTA solution. Pipette out 50mL of standard hard water in a conical flask. Add 10-15mL of buffer solution and 4 to5 drops indicator. Titrate with EDTA solution till wine-red colour changes to clear blue.

3. Titration of unknown hard water:

Rinse and fill the burette with EDTA solution. Pipette out 50 ml of unknown sample hard water in a conical flask. Add 10-15mL of buffer solution and 4 to 5 drops of indicator. Titrate with EDTA solution till wine-red colour changes to clear blue. Let volume used be V1mL.

4. Titration of Permanent hardness: Take 250mL of the water sample in a large beaker. Boil it till the volume is reduced to about 50mL, filter it, wash the precipitate with distilled water, collect filtrate and washing in a 250mL measuring flask. Finally make up the volume to 250mL with distilled water. Then, titrate 50mL of boiled water sample just as in Step (2). Let volume used be V2mL.

2m

8 (a) ELECTROCHEMICAL SERIES: Metals arranged in the increasing order of standard electrode reduction potential or decreasing order of standard oxidation potential as compared to that of standard hydrogen electrode is called electrochemical series. These standard electrode potentials are measured at 250C

1m

1m

Electrode	Oxidation reaction	Standard potential (volts)	Nature
$Li \mid Li^+$	$Li \longrightarrow Li^+ + e$	+3.040	
$K \mid K^+$	$K \longrightarrow K^+ + e^-$	+2.924	
$Ca \mid Ca^{2+}$	$Ca \longrightarrow Ca^{2+} + 2e^-$	+2.870	
$Na \mid Na^+$	$Na \longrightarrow Na^+ + e^-$	+2.710	
$Al \mid Al^{3+}$	$Al \longrightarrow Al^{3+} + 3e^-$	+1.660	
$Zn \mid Zn^{2+}$	$Zn \longrightarrow Zn^{2+} + 2e^-$	+0.762	reducing agents
$Fe \mid Fe^{2+}$	$Fe \longrightarrow Fe^{2+} + 2e^-$	+0.441	
$Cd \mid Cd^{2+}$	$Cd \longrightarrow Cd^{2+} + 2e^-$	+0.403	
$Ni \mid Ni^{2+}$	$Ni \longrightarrow Ni^{2+} + 2e^-$	+0.236	
$Sn \mid Sn^{2+}$	$Sn \longrightarrow Sn^{2+} + 2e^-$	+0.140	
$Pb \mid Pb^{2+}$	$Pb \longrightarrow Pb^{2+} + 2e^-$	+0.126	
$Pt \mid H_2, H^+$	$H_2 \longrightarrow 2H^+ + 2e^-$	0.000	
$Cu \mid Cu^{2+}$	$Cu \longrightarrow Cu^{2+} + 2e^-$	-0.337	
$Ag \mid Ag^+$	$Ag(s) \longrightarrow Ag + e^-$	-0.799	oxidising agents
$Hg \mid Hg^+$	$Hg(l) \longrightarrow Hg^{2+} + 2e^-$	-0.920	
$Cl_2 \mid Cl^-$	$2Cl^- \longrightarrow Cl_2(g) + e^-$	1.359	

In this series, iron lies above hydrogen and copper lies below it. Hence, if an iron rod is dipped in CuSO4 solution a layer of copper metal will get deposited on the surface of the iron rod.

Fe → Fe++ + 2e-

Cu++ + 2e→ Cu

Fe + Cu++→ Fe++ + Cu.

The reverse of this reaction is not possible, i.e. a copper rod dipped in FeSO4 solution will not show redox reaction.

Applications of Electrochemical Series:

1. Comparison of oxidizing and reducing power: It gives information about the relative ease with which oxidation and reduction of metal occurs. Based on the electrochemical series the element with higher reduction potential have a greater tendency to get reduced and act as good oxidizing agents. Whereas the elements with lower reduction potential have a tendency to get oxidized and act as good reducing agents.

Ex: F2 can be reduced easily than Li+ ions. So it is a good oxidizing agent

2. Relative activities of metal: Provides information about the replacement tendencies of metals. The greater the O.P of a metal, more easily it can lose electrons and greater is its reactivity, i.e. the metal with higher O.P can displace the metal with lower O.P. in their salt solution.

Ex: Mg > Zn > Fe > Cu > Ag. Zn has lower reduction potential than Cu. Hence Zn can displace copper from CuSO4 solution

Any metal above hydrogen will displace hydrogen from dil. acid solution. For example, Na reacts with water to liberate hydrogen because

E0 (Na+/Na) = -2.714 V is less than E0(H+ / H2 = 0).

2Na + H2SO4 → Na2SO4 +H2

2K + H2SO4 → K2SO4 + H2

Ca + H2SO4 → CaSO4 +H2

4. Provides information about the relative corrosion tendencies of metals.

5. Predicts spontaneity of redox reaction. If EMF of a cell is –ve, the reaction is non spontaneous and if EMF is +ve then the reaction is spontaneous.

6. This series is also helpful in direct calculation of the EMF of cell formed between electrodes.

E0= -0.76V (reduction potential)

8 (b) 2m

2m

1m
1m

OR

Cathodic coatings are obtained by coating a more noble metal (i.e. metals having higher electrode potential like Sn, Au, Ag, Pt etc.) than the base metal. They protect the base metal as they have higher corrosion resistance than the base metal due to cathodic nature. 2m

9 (a) Cathodic coating protects the base metal only when the coating is uniform and free from pores. 2m

The formation of pores over the cathodic coating exposes the base metal (anode) to environment and a galvanic cell is set up. This causes more damage to the base metal. 2m

The rate and extent of corrosion depends upon various factors due to nature of metal and nature of corroding environment.

Nature of metal:

Purity of the metal:

If impurities are present in a metal the corrosion rate is increased due to formation of tiny electrochemical cells at the exposed parts and the anodic parts get corroded. A pure metal is more corrosion resistant than impure metal. By increasing its purity the corrosion resistance of a metal can be improved

Ex: Zn metal containing impurity undergoes corrosion of zinc, due to the formation of electrochemical cells. The rate and extent of corrosion increases with the increasing exposure and extent of the impurities.

2. Electrode potentials: Metals with higher reduction potentials do not corrode easily. They are noble metals like gold, platinum and silver. Whereas the metals with lower reduction potentials readily undergo corrosion (eg. Zn, Mg, Al etc.). 2m

3. Position of metal in galvanic series: Metals which possess low reduction potentials and occupy higher end of galvanic series undergo corrosion easily.

4. Metals which possess high reduction potentials and occupy lower end of galvanic series do not undergo corrosion and they get protected.

When two metals are in electrical contact in presence of an electrolyte, then the metal which is more active undergoes corrosion.

The rate of corrosion depends on the difference in their position in galvanic series. Greater the difference more will be the extent of corrosion at anode.

Eg: The potential difference between Fe and Cu is 0.78V which is more than that between Fe and Sn (0.30V). Therefore, Fe corrodes faster when in contact with Cu than that with Sn. On this account, the use of dissimilar metals should be avoided wherever possible (Eg. Bolt & nuts, screw & washer). 2m

9 (b)

5. Relative areas of anodic and cathodic parts: If the metal has small anodic and large cathodic area, the rate of corrosion is very high. This is because the more electrons are liberated at smaller anodic area, which are consumed at cathode. If the cathodic area is larger, the liberated electrons are rapidly consumed at cathode. This further enhances the anodic reaction leading to increase in the rate of corrosion.

When two dissimilar metals or alloys are in contact,

$$\text{corrosion at anodic area} \propto \frac{\text{Areaofcathodicpart}}{\text{Areaofanodicpart}}$$

6. Hydrogen over voltage:

The difference between the potential of the electrode at which the electrolysis actually proceeds continuously (actual decomposition potential) and the theoretical decomposition potential for the same solution is called overvoltage. 2m

When a metal which is at a high position in galvanic series (eg. Zn) is placed in 1N H2SO4, it undergoes corrosion with deposition of a film on its surface and evolution of hydrogen gas. So the initial rate of corrosion is high, but decreases after a while due to salt film and H2 film surrounding the metal which causes high over voltage and reduces the corrosion rate. However, if few drops of CuSO4 are added, the corrosion rate of Zn is accelerated because some copper gets deposited on the Zn metal, forming minute cathodes, where the hydrogen overvoltage is reduced. Hence reduction in overvoltage of the corroding metal/alloy accelerates the corrosion. So higher the over voltage, lesser is the corrosion

$$\text{Rate of corrosion } \alpha \ \frac{1}{\text{over voltage}}$$

Physical state of metal: Metals with small grain size have more tendencies to undergo corrosion. Metal with more stress/strain also undergoes corrosion easily.

Nature of surface film: If the corrosion product formed is more stable, insoluble and nonporous, it acts as protective layer and prevents further corrosion (Eg. Ti, Al and Cr). If the corrosion product is porous, volatile and soluble, it further enhances the corrosion (Fe, Zn and Mg).

Volatility of corrosion product:

If the corrosion produced volatilizes as soon as it is formed the metal surface is exposed for further attack. This creates rapid and excessive corrosion.

For example the corrosion product of molybdenum as molybdenum oxide is volatile.

Solubility of corrosion product:

If the oxide film formed as corrosion product is soluble in corroding medium the corrosion proceeds at a faster rate. The corrosion product acts as a physical barrier between the metal and environment.

For example $PbSO_4$ film formed by Pb in sulphuric acid medium.

Nature of Environment:

1. Temperature: The rate of corrosion increases with increase in temperature due to increase in diffusion rate.
2. Humidity in air: The rate of corrosion increases with the presence of moisture in atmosphere because the moisture or humidity present in atmosphere furnishes water to the electrolyte which is essential for setting up of an electrochemical cell. The oxide film formed has the tendency to absorb moisture which creates another electrochemical cell.
3. Presence of impurities: Atmosphere is contaminated with gases like CO_2, SO_2, H_2S; fumes of H_2SO_4, HCl etc. and other suspended particles in the vicinity of industrial areas. They are responsible for electrical conductivity, thereby increasing corrosion.
4. Effect of PH: pH value of the medium has the greater effect on corrosion. Generally acidic medium (i.e. $pH < 7$) is more corrosive than basic medium. Acidic pH increases the rate of corrosion. However some metals like Al, Zn, Pbetc dissolve in alkaline solutions as complex ions. Consequently, corrosion of metals, readily attacked by acid can be reduced by increasing the PH of the attacking environment.

Acidic medium: $PH < 7$ - Corrosion is more

Basic medium: $PH > 7$ - Corrosion is less

Eg: Zn which is readily corroded in acidic solutions suffers very less corrosion in alkaline medium, i.e PH =11. Al has less corrosion at pH=5.5 which corrodes rapidly at PH = 8.5.

5. Amount of oxygen in atmosphere: As the percentage of oxygen in atmosphere increases, the rate of corrosion also increases due to the formation of oxygen concentration cell. The decay of metal occurs at the anodic part and the cathodic part of the metal is protected.

10 (a)

Carbon and Hydrogen:

About 1 to 2 grams of accurately weighed coal sample is burnt in a current of oxygen in a combustion apparatus. C and H of the coal are converted into CO_2 and H_2O respectively. The gaseous products of combustion are absorbed respectively in KOH and $CaCl_2$ tubes of known weights. The increase in weights of these are then determined. 2m

$$C + O_2 \rightarrow CO_2$$

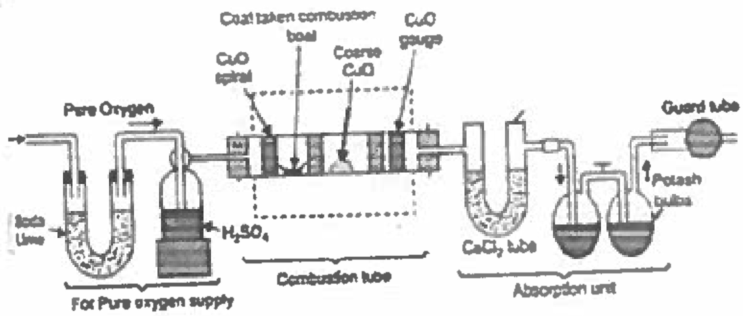
$$2KOH + CO_2 \rightarrow K_2CO_3 + H_2O$$

$$H_2 + \tfrac{1}{2} O_2 \rightarrow H_2O$$

$$CaCl_2 + 7\,H_2O \rightarrow CaCl_2.7H_2O$$

$$\% \text{ of Carbon} = \frac{\text{Increase in weight of KOH}}{\text{Weight of coal sample taken}} \times \frac{12}{44} \times 100$$

$$\% \text{ of Hydrogen} = \frac{\text{Increase in weight of CaCl2 tube}}{\text{Weight of coal sample taken}} \times \frac{2}{18} \times 100$$



Carbon and hydrogen determination

Nitrogen:

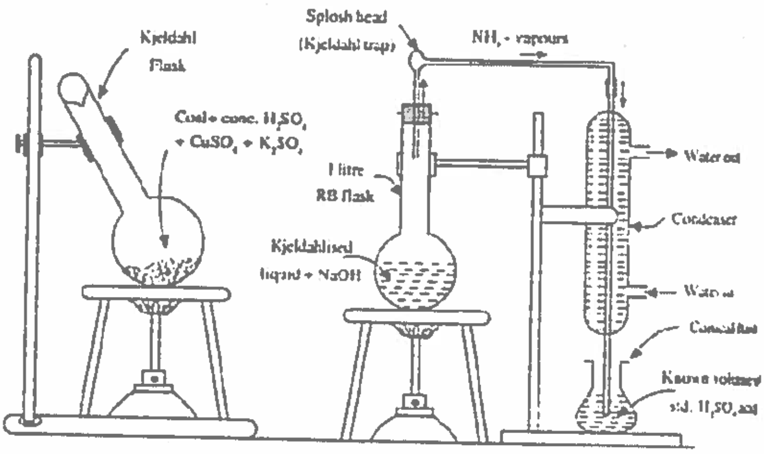
About 1 gram of accurately weighed powdered coal is heated with concentrated H2SO4 along with K2SO4 (catalyst) in a long-necked Kjeldahl's flask. After the solution becomes clear, i.e, when whole nitrogen is converted into ammonium sulphate, it is treated with excess of NaOH and the liberated ammonia is distilled over and absorbed in a known volume of standard acid solution. The unused acid is then determined by back titration with standard NaOH solution. From the volume of acid used by liberated ammonia, the percentage of N2 in coal is calculated as follows:

$$N_2 + H_2SO_4 \xrightarrow{K_2SO_4} (NH_4)_2SO_4$$

$$(NH_4)_2SO_4 + 2NaOH \longrightarrow 2NH_3\uparrow + 2Na_2SO_4 + 2H_2O$$

$$2NH_3 + H_2SO_4 \longrightarrow (NH_4)_2SO_4$$

$$\% \text{ of } N2 = \frac{\text{Volume of acid} \times \text{Normality of acid} \times 1.4}{\text{Weight of coal taken}}$$ 2m



Estimation of nitrogen by Kjeldahl's method

Sulphur:

Sulphur is determined from the washings obtained from the known mass of coal, used in bomb calorimeter for determination of calorific value. During this determination, S is

$$S \xrightarrow{O_2} SO_4^{2-} \xrightarrow{BaCl_2} BaSO_4$$

converted into sulphates. The washings are treated with barium chloride solution, when barium-sulphate is precipitated. 2m

This precipitate is filtered, washed and heated to constant weight.

$$\% \text{ of Sulphur} = \frac{\text{Weight of BaSO4 obtained} \times 32}{\text{Weight of coal sample taken} \times 233} \times 100$$

Ash: The residual coal is taken in a crucible and then heated without lid in a muffle furnace at 7500C for ½ hour. The crucible is then taken out, cooled first in air, then in a desiccator and weighed. Heating, cooling and weighing is repeated, till a constant weight is obtained. The residue is reported as ash on percentage-basis.

Thus,

$$\% \text{ of Ash} = \frac{\text{Weight of ash left}}{\text{Weight of coal taken}} \text{ X } 100$$

Oxygen:

It is determined indirectly by deducting the combined percentage of carbon, hydrogen, nitrogen, sulphur and ash from 100.

% of Oxygen = 100 – % of (C + H + S + N + Ash)

2m

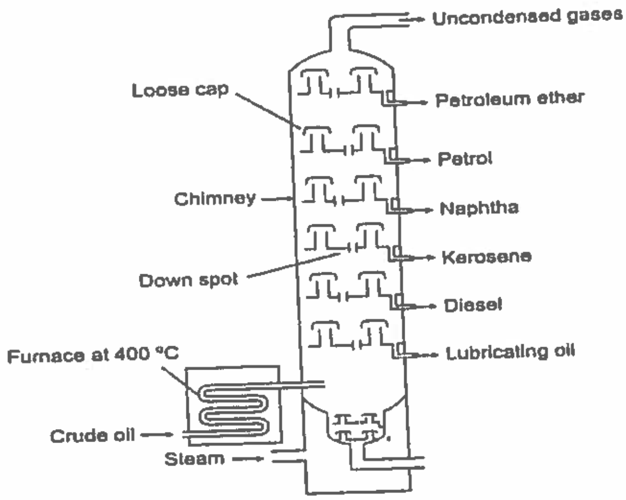
10 (b)

1. In internal combustion engine, diesel or gasoline mixed with air is used as fuel
2. The power output and efficiency of internal combustion engine depends on a factor called compression ratio
3. CR is the ratio of volume of gases at the end of suction stroke to the volume of gases at the end of compression stroke: C.R = Vs.s/Vc.s 2m
4. The efficiency of engine increases with increase in CR ratio which depends on the constituents present in petrol
5. Due to higher compression ratio fuel air mixture is heated to higher temperatures; the fuel ignites even before the regular spark occurs. This pre-ignition is called knocking
6. Some constituents of petrol, the rate of CR raises fast so that the last drops of the fuel air mixture gets instantaneously ignited producing a loud noise.
7. This rattling noise produced in the internal combustion engine is known as Knocking. 2m

OR

11 (a)

a. Fractional distillation: Heating of crude oil around 4000C in an iron retort, produces hot vapours which is allowed to pass through fractionating column. It is a tall cylindrical tower containing a number of horizontal stainless trays at short distances and is provided with small chimney covered with loose cap. As the vapours go up they get cooled gradually and fractional condensation takes place. Higher boiling fractions condenses first later the lower boiling fractions 2m



4m

Name of Fraction	Boiling range	Approx- composition in terms of hydrocarbon 'C' atoms	Uses
(1) Uncondensed gas	Below 300C	C1toC4	A domestic or industrial fuel.
(2) Petroleum ether	300-700C	C5-C7	As a solvent.
(3)Gasoline or petrol	400-1200C	C5-C9	As motor fuel, solvent, in dry cleaning
(4)Naphtha or solvent spirit	120-180oC	C9-C10	As solvent, in dry cleaning
(5) Kerosene	1800C-2500C	C10-C16	As an illuminant ,Engine fuel
(6) Diesel oil	250-320oC	C10-C18	Diesel engine fuel
(7)Heavy oil on refraction	320-400oC	C17-C30	Gasoline by cracking
(a) Lubricating oil	---	....	As lubricant
(b) petroleum jelly	—	....	As lubricant and in cosmetics and ointments

c) Paraffin wax d) Greases 8)Residue(asphalt, petroleum coke)			In candles, boot polishes As lubricant Used for making tar roads, water proof roofing

11 (b)

1) Higher Calorific value (HCV) or Gross calorific value (GCV) 2m

GCV or HCV = 1/100[8080 C + 34500 (H - O/8) + 2240 S] cal/gm

= 1/100 x 733835 = 7338.35 kcal/kg 2m

2) Lower Calorific value (LCV) or Net calorific value (NCV)

LCV = HCV – 0.09 x H x 587

= 7338.35 – 633.96 = 6704.39 kcal/kg 2m

12(a)

Co-Ordination Mechanism (or) Zeigler – Natta polymerization:

The combination of transition metal halides like TiCl3 and organometallic compounds like Al(C2H5)3 are called Ziegler-Natta catalysts. This mechanism consists of three steps. 2m

Step 1: chain initiation.

Step 2: chain propagation.

Step 3: chain termination. 2m

Step 1: Chain initiation:

In this process to prepare a catalyst by using transitional metal halides and tri alkyl aluminum.

$$TiCl_3 + Al(C_2H_5)_3 \rightarrow \underset{(Cat-R)}{Cl_3-Ti-Al-(C_2H_5)_3}$$ 2m

By using this catalyst it react with selective monomers to initiate the chain and form monomer catalyst complexes.

$$\underset{\text{(Complex catalyst)}}{Cat-R} \quad + \quad \underset{\text{(Monomer)}}{CH_2=CHCl} \rightarrow \underset{\text{(Monomer catalyst complex)}}{Cat-CH_2-CHCl-R}$$

Step 2: Chain propagation:

In this process the chain is propagated by the addition of number of monomer units to the monomer catalyst complexes.

$$Cat-CH_2-CHCl-R+n(CH_2=CHCl) \rightarrow Cat-CH_2-(CH_2-CHCl)_n-CH_2-CHCl-R$$

Step 3: Chain termination:

The chain growth is stopped at a particular point by addition of Hydrogen halide (HX).

$$Cat-CH_2-(CH_2-CHCl)_n-CH_2-CHCl-R + HX \rightarrow CH_3-(CH_2-CHCl)_n-CH_2-CH_2Cl-R + Cat-X$$

2m

12(b) **Define and classify composite materials, give their constituents. Explain why composite materials have superior properties compared to conventional materials** 2m

Description

Composite materials

Superior materal 2m

13(a) **With a neatly labeled diagram, explain the mechanism of thick film lubrication**

- Lubricants make moulding of plastic easier
- They impart flawlss, glossu finish to the products
- They prenentmoulded article from sticking to the frabrication equipment
- E.g. Waxes, olis, stearates, oleates, and soap

13(b) **Discuss about the fiber and structural reinforced composites, enlist their engineering applications**

- Combination of plastic material & solid fillers give hard plastic with mechanical strength & impact resistant is known as reinforced plastic.
- The fiber polymers with solid/fillers to impart mechanical strength & hardness without losing plasticity are known as fiber reinforced plastics (FRP). 2m 2m 2m
- Fillers like carborandum, quartz & mica – impart hardness & strength.
- Barium salt impervious to x-rays.
- Asbestos provide heat & corrosion resistant for FRP.

14(a) **What are colloids? Explain the synthesis of colloids by any one method with suitable examples**

Definition 2m

Synthesis 3m

Methods with example 2m

14(b)	How do you characterize the surface of a substance by X-ray diffraction method? Surface Diagram X-ray metod	 1m 2m 2m
15(a)	Discuss the synthesis of nanometals by chemical method. Enumerate the applications of nanomaterials in catalysis Synthesis Methods apllications	 2m 2m 1m
15(b)	What is adsorption isotherm? Give its significance Definitions significance	4m 3m

SEMESTER Question Paper

Degree	B. Tech. (U. G.)	**Program**	CE/ME	Test	I/I-Supply	**Academic Year**	2021 - 2022
Course Code	20BSX21	**Test Duration**	90 Min.	Max. Marks	70	**Semester**	I
Course	ENGINEERING CHEMISTRY						

Key and Scheme of Evaluation

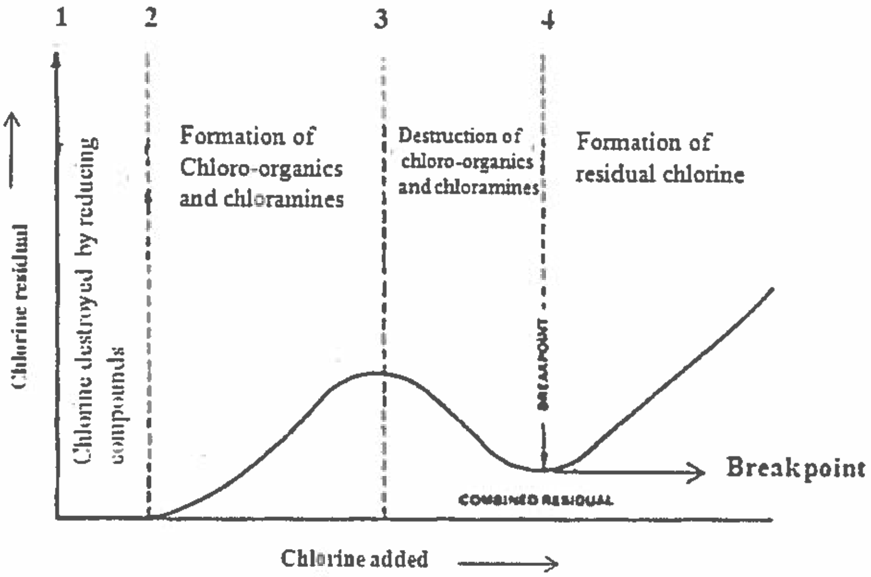
No.	Questions (1 through 5)	Marks
1	Temporary Hardness mainly caused by the presence of dissolved bicarbonates of calcium, magnesium and other heavy metals.Permanent Hardness: It is due to the presence of dissolved chlorides and sulphates of calcium, magnesium, iron and other metals.	2
2	Electroless plating. The method of deposition of a metal from its salt solution on a catalytically active surface by a suitable reducing agent without using electrical energy is called electroless plating	2
3	The quality or rating of diesel is expressed by cetane number. Cetane is n- Hexadecane (C16H34)	2
4	It is used for manufacturing of tyres. It is used in the footwear industry for making shoe soles and footwear components	2
5	**Define cracking** decomposition of bigger hydrocarbon molecules into simpler, low boiling hydrocarbons of low molecular weight	2
No.	**Questions (6 through 11)**	

6 (a)

Break point chlorination:

The amount of chlorine required to kill bacteria and to remove organic matter is called break point chlorination.
A typical relationship between the amount of chlorine added to water and the experimentally determined free residual chlorine is shown in the graph below which explains the break point chlorination.

(i) First (between points 1 and 2), the water reacts with reducing compounds in the water, such as hydrogen sulfide. These compounds use up the chlorine, producing no chlorine residual.



6M

(ii) Between points 2 and 3, the chlorine reacts with organics and ammonia naturally found in the water and forms chloro-organics and chloramines.

(iii) Between points 3 and 4, the added chlorine will break down most of the chloramines in the water, leaving behind free chlorine which destroys pathogenic bacteria, actually lowering the chlorine residual.

(iv) Finally, the water reaches the breakpoint, at point 4 (minima). The breakpoint is the point at which the chlorine demand has been totally satisfied - the chlorine has reacted with all reducing agents, organics, ammonia and pathogenic bacteria in the water. After minima when more chlorine is added, it is not used in any reaction. Thus, the residual chlorine keeps on increasing in proportion to added chlorine. Hence for effective killing of microorganisms, sufficient amount of chlorine has to be added. Addition of chlorine in such dosages is known as break point chlorination.

Advantages of break –point chlorination:

1. It ensures complete destruction of organic compounds which imparts colour, bad odour and unpleasant taste to water.
2. It completely destroys all the disease producing bacteria.
3. It prevents the growth of any weeds in water.

6 (b)

Caustic embrittlement: Caustic embrittlement is a term used for the appearance of cracks inside the boilers particularly at those places which are under stress such as riveted joints due to the high concentration of alkali leading to failure of the boiler. .

Reasons:

It is a type of boiler corrosion caused due to the presence of alkali-metal carbonates and bicarbonates in feed water and also the presence of sodium carbonate.

$$Na_2CO_3 + H_2O \rightarrow NaOH + CO_2$$

This caustic water flows inside the boiler and causes some minute hair-cracks, by capillary action. On evaporation of water, the dissolved caustic soda increases its concentration which attacks the surrounding area, thereby dissolving iron of boiler as sodium ferroate (Na_2FeO_2). This causes embrittlement of boiler parts such as bends, joints, rivets etc, due to which the boiler gets fail. 2m

Caustic cracking can be explained by considering the following concentration cell. The iron at plane surfaces surrounded by dilute NaOH becomes cathodic and the iron at bends, joints and rivets surrounded by highly concentrated NaOH becomes anodic which consequently gets corroded 2m

1m

1m

+(Anode)			-(Cathode)
Iron at rivets , bends joints..etc	Conc NaOH	Dil NaOH	Iron at plane surfaces

Prevention:

- By maintaining the pH of water and neutralization of alkali.
- By using sodium phosphate as softening reagents instead of sodium carbonate in the external treatment of boiler feed water.

Caustic embrittlement can also be prevented by adding tannin or lignin or Sodium

OR

7 (a)

ZEOLITE PROCESS (Permutit Process)

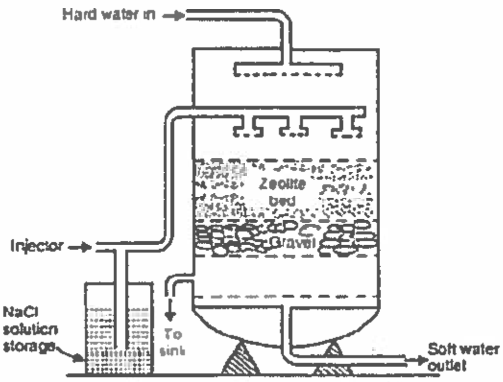
Zeolite is a 3D silicate. The chemical composition of zeolites is hydrated sodium aluminium silicate, represented as $Na_2O.Al_2O_3.xSiO_2.yH_2O$ where x = 2-10 and y = 2-6. Zeolites are capable of exchanging reversibly its sodium ions for hardness causing Ca^{2+} and Mg^{2+} in water. Hence zeolites are cation exchangers. 2m

Zeolites are of two types 2m

1) Natural zeolites are natural and non-porous having the composition 2m

Natural zeolite: $Na_2O\ Al_2O_3\ .4SiO_2\ .2H_2O$

Synthetic zeolites are porous and posses gel structure and are prepared from china clay, feldspar and a soda ash. Synthetic zeolites posses higher exchange capacity. Zeolites have cage like structure. It is derived from SiO_4



Zeolite Process

Process: The hard water is passed through a zeolite bed fixed in a cylinder at a specific rate.The hardness causing ions Ca2+, Mg2+ etc are retained by the zeolite as CaZe and MgZe respectively

The reactions taking place during softening process are

$$Na_2Ze + Ca(HCO_3)_2 \rightarrow CaZe + 2NaHCO_3$$
$$Na_2Ze + Mg(HCO_3)_2 \rightarrow MgZe + 2NaHCO_3$$
$$Na_2Ze + CaCl_2 \rightarrow 2NaCl + CaZe$$
$$Na_2Ze + MgCl_2 \rightarrow 2NaCl + MgZe$$

Regeneration: After the use of this process for certain time zeolite is exhausted i.e. all Na^+ ions of the zeolite are replaced by Ca^{+2} and Mg^{+2} ions and hard water will not be further softened.

Regeneration:

Exhausted zeolite can be regenerated by heating it with brine solution (10% NaCl)

$CaZe + 2NaCl \rightarrow Na_2Ze + CaCl_2$

$MgZe + 2NaCl \rightarrow Na_2Ze + MgCl_2$

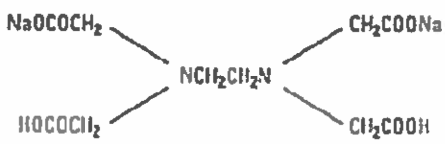
Estimation of Hardeness of Water by EDTA method (Complexometric titration)

This is a complexometric titration where ethylene diamine tetra acetic acid [EDTA] is used as a complexing agent. EDTA form complexes with different metal ions at different PH 9-10. To maintain the PH 9-10 ammonical buffer is used. An alcoholic solution of Eriochrome black –T is used as an indicator. The disodium salt of EDTA is used for complexation since solubility of EDTA is very low.

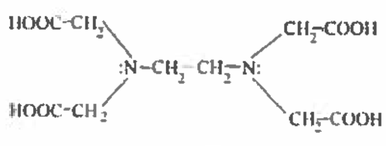
2m

Structure of EDTA:

7 (b)



Di-sodium salt of EDTA



structure of EDTA

2m

Principle: In this complexometric titration, the water sample is titrated with standard solution of Di-sodium salt of EDTA using EBT indicator. First EDTA solution is standardized with a standard solution of calcium carbonate, prepared by dissolving a known weight of cacium carbonate in dil.HCl and then making up the solution to a known volume with distilled water .The permanent hardness of water can be determined by titrating the water after boiling well to remove the temporary hardness as carbonates of calcium and magnesium against EDTA by adding EBT and buffer solution. Temporary hardness can be determined by subtracting total hardness of water from permanent hardness of water. 2m

Reactions involved: EBT indicator when added to hard water at pH = 10, forms weak complexes with calcium and magnesium present in hard water. It results in the formation of unstable wine red Ca-EBT or Mg-EBT complexes. During titration with EDTA, calcium first reacts to form relatively stable complex followed by magnesium to give Mg+2-EDTA complex releasing the free indicator(blue).The colour changes from wine-red to blue at the endpoint.

$$\left.\begin{array}{l} Ca^{2+} \\ \\ Mg^{2+} \end{array}\right\} + \text{Eriochrome black - T} \rightarrow \left[\begin{array}{l} Ca^{2+} \\ \quad \text{Eriochrome black - T} \\ Mg^{2+} \end{array}\right] \text{Complex}$$

(of water) (Unstable complex)
(Wine-red)

$$\xrightarrow{\text{EDTA}} \left[\begin{array}{l} Ca^{2+} \\ \quad \text{EDTA} \\ Mg^{2+} \end{array}\right] \text{Complex} + \text{Eriochrome black - T}$$

(blue)

Various steps involved in this method are...
1. Preparation of standard hard water: Dissolve 1g of pure, dry CaCO3 in minimum quantity of dil. HCl and then evaporate the solution to dryness on a water bath. Dissolve the residue in distilled water to make 1 liter solution. Each mL of this solution thus contains 1mg of CaCO3 equivalent hardness.
1 mL hard water solution = 1mg of CaCO3 equivalent hardness.
2. Standardization of EDTA solution: Rinse and fill the burette with EDTA solution. Pipette out 50mL of standard hard water in a conical flask. Add 10-15mL of buffer solution and 4 to5 drops indicator. Titrate with EDTA solution till wine-red colour changes to clear blue.

3. Titration of unknown hard water:
Rinse and fill the burette with EDTA solution. Pipette out 50 ml of unknown sample hard water in a conical flask. Add 10-15mL of buffer solution and 4 to 5 drops of indicator. Titrate with EDTA solution till wine-red colour changes to clear blue. Let volume used be V1mL.

4. Titration of Permanent hardness: Take 250mL of the water sample in a large beaker. Boil it till the volume is reduced to about 50mL, filter it, wash the precipitate with distilled water, collect filtrate and washing in a 250mL measuring flask. Finally make up the volume to 250mL with distilled water. Then, titrate 50mL of boiled water sample just as in Step (2). Let volume used be V2mL.

2m

8 (a) ELECTROCHEMICAL SERIES: Metals arranged in the increasing order of standard electrode reduction potential or decreasing order of standard oxidation potential as compared to that of standard hydrogen electrode is called electrochemical series. These standard electrode potentials are measured at 250C

1m
1m

Electrode	Oxidation reaction	Standard potential (volts)	Nature
$Li \mid Li^+$	$Li \longrightarrow Li^+ + e$	+3.040	
$K \mid K^+$	$K \longrightarrow K^+ + e^-$	+2.924	
$Ca \mid Ca^{2+}$	$Ca \longrightarrow Ca^{2+} + 2e^-$	+2.870	
$Na \mid Na^+$	$Na \longrightarrow Na^+ + e^-$	+2.710	
$Al \mid Al^{3+}$	$Al \longrightarrow Al^{3+} + 3e^-$	+1.660	
$Zn \mid Zn^{2+}$	$Zn \longrightarrow Zn^{2+} + 2e^-$	+0.762	reducing agents
$Fe \mid Fe^{2+}$	$Fe \longrightarrow Fe^{2+} + 2e^-$	+0.441	
$Cd \mid Cd^{2+}$	$Cd \longrightarrow Cd^{2+} + 2e^-$	+0.403	
$Ni \mid Ni^{2+}$	$Ni \longrightarrow Ni^{2+} + 2e^-$	+0.236	
$Sn \mid Sn^{2+}$	$Sn \longrightarrow Sn^{2+} + 2e^-$	+0.140	
$Pb \mid Pb^{2+}$	$Pb \longrightarrow Pb^{2+} + 2e^-$	+0.126	
$Pt \mid H_2, H^+$	$H_2 \longrightarrow 2H^+ + 2e^-$	0.000	
$Cu \mid Cu^{2+}$	$Cu \longrightarrow Cu^{2+} + 2e^-$	-0.337	
$Ag \mid Ag^+$	$Ag(s) \longrightarrow Ag + e^-$	-0.799	oxidising agents
$Hg \mid Hg^+$	$Hg(l) \longrightarrow Hg^{2+} + 2e^-$	-0.920	
$Cl_2 \mid Cl^-$	$2Cl^- \longrightarrow Cl_2(g) + e^-$	1.359	

In this series, iron lies above hydrogen and copper lies below it. Hence, if an iron rod is dipped in CuSO4 solution a layer of copper metal will get deposited on the surface of the iron rod.

$Fe \rightarrow Fe^{++} + 2e^-$

$Cu^{++} + 2e^- \rightarrow Cu$

$Fe + Cu^{++} \rightarrow Fe^{++} + Cu$.

The reverse of this reaction is not possible, i.e. a copper rod dipped in FeSO4 solution will not show redox reaction.

Applications of Electrochemical Series:

1. Comparison of oxidizing and reducing power: It gives information about the relative ease with which oxidation and reduction of metal occurs. Based on the electrochemical series the element with higher reduction potential have a greater tendency to get reduced and act as good oxidizing agents. Whereas the elements with lower reduction potential have a tendency to get oxidized and act as good reducing agents.

Ex: F2 can be reduced easily than Li+ ions. So it is a good oxidizing agent

2. Relative activities of metal: Provides information about the replacement tendencies of metals. The greater the O.P of a metal, more easily it can lose electrons and greater is its reactivity, i.e. the metal with higher O.P can displace the metal with lower O.P. in their salt solution.

Ex: Mg > Zn > Fe > Cu > Ag. Zn has lower reduction potential than Cu. Hence Zn can displace copper from CuSO4 solution

Any metal above hydrogen will displace hydrogen from dil. acid solution. For example, Na reacts with water to liberate hydrogen because

$E^0 (Na^+/Na) = -2.714$ V is less than $E^0(H^+ / H_2 = 0)$.

$2Na + H_2SO_4 \rightarrow Na_2SO_4 + H_2$

$2K + H_2SO_4 \rightarrow K_2SO_4 + H_2$

$Ca + H_2SO_4 \rightarrow CaSO_4 + H_2$

4. Provides information about the relative corrosion tendencies of metals.

5. Predicts spontaneity of redox reaction. If EMF of a cell is –ve, the reaction is non spontaneous and if EMF is +ve then the reaction is spontaneous.

6. This series is also helpful in direct calculation of the EMF of cell formed between electrodes.

$E^0 = -0.76$V (reduction potential)

8 (b) 2m

2m

		1m 1m
	OR	
9 (a)	Cathodic coatings are obtained by coating a more noble metal (i.e. metals having higher electrode potential like Sn, Au, Ag, Pt etc.) than the base metal. They protect the base metal as they have higher corrosion resistance than the base metal due to cathodic nature. Cathodic coating protects the base metal only when the coating is uniform and free from pores. The formation of pores over the cathodic coating exposes the base metal (anode) to environment and a galvanic cell is set up. This causes more damage to the base metal.	2m 2m 2m

The rate and extent of corrosion depends upon various factors due to nature of metal and nature of corroding environment.

Nature of metal:

Purity of the metal:

If impurities are present in a metal the corrosion rate is increased due to formation of tiny electrochemical cells at the exposed parts and the anodic parts get corroded. A pure metal is more corrosion resistant than impure metal. By increasing its purity the corrosion resistance of a metal can be improved

Ex: Zn metal containing impurity undergoes corrosion of zinc, due to the formation of electrochemical cells. The rate and extent of corrosion increases with the increasing exposure and extent of the impurities.

2. Electrode potentials: Metals with higher reduction potentials do not corrode easily. They are noble metals like gold, platinum and silver. Whereas the metals with lower reduction potentials readily undergo corrosion (eg. Zn, Mg, Al etc.). 2m

3. Position of metal in galvanic series: Metals which possess low reduction potentials and occupy higher end of galvanic series undergo corrosion easily.

4. Metals which possess high reduction potentials and occupy lower end of galvanic series do not undergo corrosion and they get protected.

When two metals are in electrical contact in presence of an electrolyte, then the metal which is more active undergoes corrosion.

The rate of corrosion depends on the difference in their position in galvanic series. Greater the difference more will be the extent of corrosion at anode.

Eg: The potential difference between Fe and Cu is 0.78V which is more than that between Fe and Sn (0.30V). Therefore, Fe corrodes faster when in contact with Cu than that with Sn. On this account, the use of dissimilar metals should be avoided wherever possible (Eg. Bolt & nuts, screw & washer). 2m

9 (b)

5. Relative areas of anodic and cathodic parts: If the metal has small anodic and large cathodic area, the rate of corrosion is very high. This is because the more electrons are liberated at smaller anodic area, which are consumed at cathode. If the cathodic area is larger, the liberated electrons are rapidly consumed at cathode. This further enhances the anodic reaction leading to increase in the rate of corrosion.

When two dissimilar metals or alloys are in contact,

$$\text{corrosion at anodic area} \propto \frac{\text{Areaofcathodicpart}}{\text{Areaofanodicpart}}$$

6. Hydrogen over voltage:

The difference between the potential of the electrode at which the electrolysis actually proceeds continuously (actual decomposition potential) and the theoretical decomposition potential for the same solution is called overvoltage. 2m

When a metal which is at a high position in galvanic series (eg. Zn) is placed in 1N H2SO4, it undergoes corrosion with deposition of a film on its surface and evolution of hydrogen gas. So the initial rate of corrosion is high, but decreases after a while due to salt film and H2 film surrounding the metal which causes high over voltage and reduces the corrosion rate. However, if few drops of CuSO4 are added, the corrosion rate of Zn is accelerated because some copper gets deposited on the Zn metal, forming minute cathodes, where the hydrogen overvoltage is reduced. Hence reduction in overvoltage of the corroding metal/alloy accelerates the corrosion. So higher the over voltage, lesser is the corrosion

$$\text{Rate of corrosion } \alpha \frac{1}{\text{over voltage}}$$

Physical state of metal: Metals with small grain size have more tendencies to undergo corrosion. Metal with more stress/strain also undergoes corrosion easily.

Nature of surface film: If the corrosion product formed is more stable, insoluble and nonporous, it acts as protective layer and prevents further corrosion (Eg. Ti, Al and Cr). If the corrosion product is porous, volatile and soluble, it further

enhances the corrosion (Fe, Zn and Mg).
Volatility of corrosion product:
If the corrosion produced volatilizes as soon as it is formed the metal surface is exposed for further attack. This creates rapid and excessive corrosion.
For example the corrosion product of molybdenum as molybdenum oxide is volatile.
Solubility of corrosion product:
If the oxide film formed as corrosion product is soluble in corroding medium the corrosion proceeds at a faster rate. The corrosion product acts as a physical barrier between the metal and environment.
For example $PbSO_4$ film formed by Pb in sulphuric acid medium.
Nature of Environment:
1. Temperature: The rate of corrosion increases with increase in temperature due to increase in diffusion rate.
2. Humidity in air: The rate of corrosion increases with the presence of moisture in atmosphere because the moisture or humidity present in atmosphere furnishes water to the electrolyte which is essential for setting up of an electrochemical cell. The oxide film formed has the tendency to absorb moisture which creates another electrochemical cell.
3. Presence of impurities: Atmosphere is contaminated with gases like CO_2, SO_2, H_2S; fumes of H_2SO_4, HCl etc. and other suspended particles in the vicinity of industrial areas. They are responsible for electrical conductivity, thereby increasing corrosion.
4. Effect of PH: pH value of the medium has the greater effect on corrosion. Generally acidic medium (i.e. $pH < 7$) is more corrosive than basic medium. Acidic pH increases the rate of corrosion. However some metals like Al, Zn, Pbetc dissolve in alkaline solutions as complex ions. Consequently, corrosion of metals, readily attacked by acid can be reduced by increasing the PH of the attacking environment.
Acidic medium: $PH < 7$ - Corrosion is more
Basic medium: $PH > 7$ - Corrosion is less
Eg: Zn which is readily corroded in acidic solutions suffers very less corrosion in alkaline medium, i.e PH =11. Al has less corrosion at pH=5.5 which corrodes rapidly at PH = 8.5.

5. Amount of oxygen in atmosphere: As the percentage of oxygen in atmosphere increases, the rate of corrosion also increases due to the formation of oxygen concentration cell. The decay of metal occurs at the anodic part and the cathodic part of the metal is protected.

10 (a)

Carbon and Hydrogen:

About 1 to 2 grams of accurately weighed coal sample is burnt in a current of oxygen in a combustion apparatus. C and H of the coal are converted into CO_2 and H_2O respectively. The gaseous products of combustion are absorbed respectively in KOH and $CaCl_2$ tubes of known weights. The increase in weights of these are then determined. 2m

$$C + O_2 \rightarrow CO_2$$

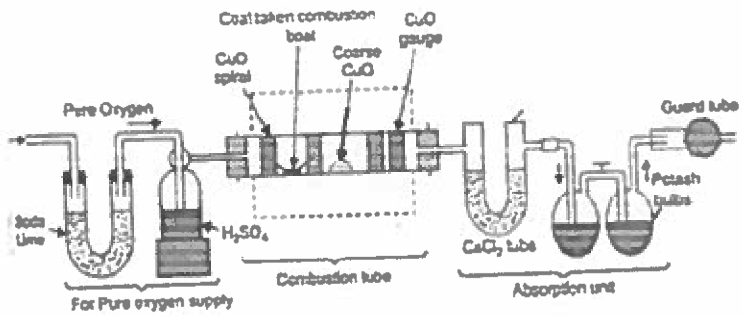
$$2KOH + CO_2 \rightarrow K_2CO_3 + H_2O$$

$$H_2 + \tfrac{1}{2} O_2 \rightarrow H_2O$$

$$CaCl_2 + 7\,H_2O \rightarrow CaCl_2.7H_2O$$

$$\%\text{ of Carbon} = \frac{\text{Increase in weight of KOH}}{\text{Weight of coal sample taken}} \times \frac{12}{44} \times 100$$

$$\%\text{ of Hydrogen} = \frac{\text{Increase in weight of CaCl2 tube}}{\text{Weight of coal sample taken}} \times \frac{2}{18} \times 100$$



Carbon and hydrogen determination

Nitrogen:

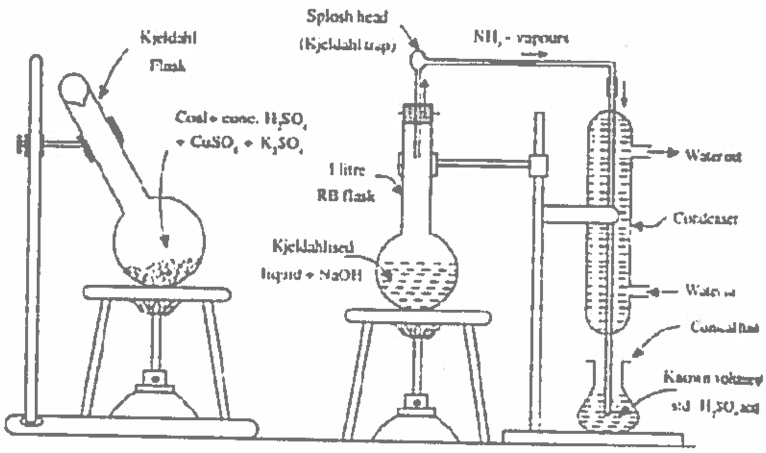
About 1 gram of accurately weighed powdered coal is heated with concentrated H2SO4 along with K2SO4 (catalyst) in a long-necked Kjeldahl's flask. After the solution becomes clear, i.e, when whole nitrogen is converted into ammonium sulphate, it is treated with excess of NaOH and the liberated ammonia is distilled over and absorbed in a known volume of standard acid solution. The unused acid is then determined by back titration with standard NaOH solution. From the volume of acid used by liberated ammonia, the percentage of N2 in coal is calculated as follows:

$$N_2 + H_2SO_4 \xrightarrow{K_2SO_4} (NH_4)_2SO_4$$

$$(NH_4)_2SO_4 + 2NaOH \longrightarrow 2NH_3\uparrow + 2Na_2SO_4 + 2H_2O$$

$$2NH_3 + H_2SO_4 \longrightarrow (NH_4)_2SO_4$$

$$\% \text{ of } N2 = \frac{\text{Volume of acid} \times \text{Normality of acid} \times 1.4}{\text{Weight of coal taken}}$$ 2m



Estimation of nitrogen by Kjeldahl's method

Sulphur:

Sulphur is determined from the washings obtained from the known mass of coal, used in bomb calorimeter for determination of a calorific value. During this determination, S is

$$S \xrightarrow{O_2} SO_4^{2-} \xrightarrow{BaCl_2} BaSO_4$$

converted into sulphates. The washings are treated with barium chloride solution, when barium-sulphate is precipitated. 2m

This precipitate is filtered, washed and heated to constant weight.

$$\% \text{ of Sulphur} = \frac{\text{Weight of BaSO4 obtained} \times 32}{\text{Weight of coal sample taken} \times 233} \times 100$$

Ash: The residual coal is taken in a crucible and then heated without lid in a muffle furnace at 7500C for ½ hour. The crucible is then taken out, cooled first in air, then in a desiccator and weighed. Heating, cooling and weighing is repeated, till a constant weight is obtained. The residue is reported as ash on percentage-basis.

Thus,

$$\% \text{ of Ash} = \frac{\text{Weight of ash left}}{\text{Weight of coal taken}} \text{ X } 100$$

Oxygen:

It is determined indirectly by deducting the combined percentage of carbon, hydrogen, nitrogen, sulphur and ash from 100.

% of Oxygen = 100 – % of (C + H + S + N + Ash)

2m

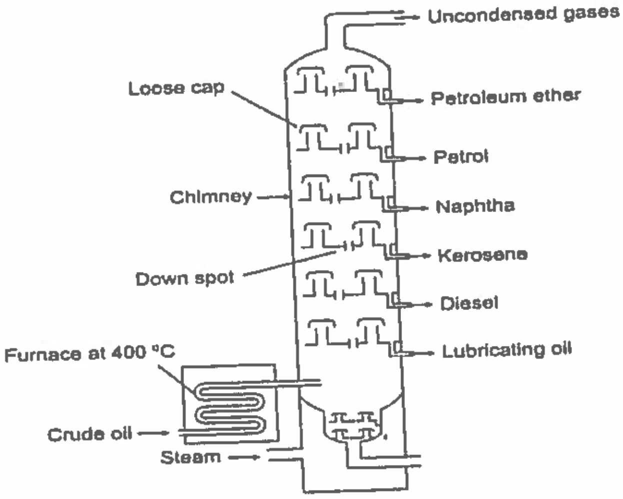
10 (b)

1. In internal combustion engine, diesel or gasoline mixed with air is used as fuel
2. The power output and efficiency of internal combustion engine depends on a factor called compression ratio
3. CR is the ratio of volume of gases at the end of suction stroke to the volume of gases at the end of compression stroke: C.R = Vs.s/Vc.s 2m
4. The efficiency of engine increases with increase in CR ratio which depends on the constituents present in petrol
5. Due to higher compression ratio fuel air mixture is heated to higher temperatures; the fuel ignites even before the regular spark occurs. This pre-ignition is called knocking
6. Some constituents of petrol, the rate of CR raises fast so that the last drops of the fuel air mixture gets instantaneously ignited producing a loud noise.
7. This rattling noise produced in the internal combustion engine is known as Knocking. 2m

OR

11 (a)

a. Fractional distillation: Heating of crude oil around 4000C in an iron retort, produces hot vapours which is allowed to pass through fractionating column. It is a tall cylindrical tower containing a number of horizontal stainless trays at short distances and is provided with small chimney covered with loose cap. As the vapours go up they get cooled gradually and fractional condensation takes place. Higher boiling fractions condenses first later the lower boiling fractions 2m



4m

Name of Fraction	Boiling range	Approx- composition in terms of hydrocarbon 'C' atoms	Uses
(1) Uncondensed gas	Below 300C	C1toC4	A domestic or industrial fuel.
(2) Petroleum ether	300-700C	C5-C7	As a solvent.
(3)Gasoline or petrol	400-1200C	C5-C9	As motor fuel, solvent, in dry cleaning
(4)Naphtha or solvent spirit	120-180oC	C9-C10	As solvent, in dry cleaning
(5) Kerosene	1800C-2500C	C10-C16	As an illuminant ,Engine fuel
(6) Diesel oil	250-320oC	C10-C18	Diesel engine fuel
(7)Heavy oil on refraction	320-400oC	C17-C30	Gasoline by cracking
(a) Lubricating oil	—	....	As lubricant
(b) petroleum jelly	——	....	As lubricant and in cosmetics and ointments

c) Paraffin wax d) Greases 8)Residue(asphalt, petroleum coke)			In candles, boot polishes As lubricant Used for making tar roads, water proof roofing

11 (b)

1) Higher Calorific value (HCV) or Gross calorific value (GCV) 2m

GCV or HCV = 1/100[8080 C + 34500 (H - O/8) + 2240 S] cal/gm

= 1/100 x 733835 = 7338.35 kcal/kg 2m

2) Lower Calorific value (LCV) or Net calorific value (NCV)

LCV = HCV – 0.09 x H x 587

= 7338.35 – 633.96 = 6704.39 kcal/kg 2m

12(a)

Co-Ordination Mechanism (or) Zeigler – Natta polymerization:

The combination of transition metal halides like TiCl3 and organometallic compounds like Al(C2H5)3 are called Ziegler-Natta catalysts. This mechanism consists of three steps. 2m

Step 1: chain initiation.

Step 2: chain propagation.

Step 3: chain termination. 2m

Step 1: Chain initiation:

In this process to prepare a catalyst by using transitional metal halides and tri alkyl aluminum.

$$TiCl_3 + Al(C_2H_5)_3 \rightarrow \underset{(Cat-R)}{Cl_3-Ti-Al-(C_2H_5)_3}$$ 2m

By using this catalyst it react with selective monomers to initiate the chain and form monomer catalyst complexes.

$$\underset{\text{(Complex catalyst)}}{Cat-R} + \underset{\text{(Monomer)}}{CH_2=CHCl} \rightarrow \underset{\text{(Monomer catalyst complex)}}{Cat-CH_2-CHCl-R}$$

Step 2: Chain propagation:

In this process the chain is propagated by the addition of number of monomer units to the monomer catalyst complexes.

$$Cat-CH_2-CHCl-R + n(CH_2=CHCl) \rightarrow Cat-CH_2-(CH_2-CHCl)_n-CH_2-CHCl-R$$

Step 3: Chain termination:

The chain growth is stopped at a particular point by addition of Hydrogen halide (HX).

$$Cat-CH_2-(CH_2-CHCl)_n-CH_2-CHCl-R + HX \rightarrow CH_3-(CH_2-CHCl)_n-CH_2-CH_2Cl-R + Cat-X$$

2m

12(b) **Define and classify composite materials, give their constituents. Explain why composite materials have superior properties compared to conventional materials** 2m

Description

Composite materials

Superior materal 2m

13(a) **With a neatly labeled diagram, explain the mechanism of thick film lubrication**

- Lubricants make moulding of plastic easier
- They impart flawlss, glossu finish to the products
- They prenentmoulded article from sticking to the frabrication equipment
- E.g. Waxes, olis, stearates, oleates, and soap

13(b) **Discuss about the fiber and structural reinforced composites, enlist their engineering applications**

- Combination of plastic material & solid fillers give hard plastic with mechanical strength & impact resistant is known as reinforced plastic.
- The fiber polymers with solid/fillers to impart mechanical strength & hardness without losing plasticity are known as fiber reinforced plastics (FRP). 2m 2m
- Fillers like carborandum, quartz & mica – impart hardness & strength. 2m
- Barium salt impervious to x-rays.
- Asbestos provide heat & corrosion resistant for FRP.

14(a) **What are colloids? Explain the synthesis of colloids by any one method with suitable examples**

Definition 2m

Synthesis 3m

Methods with example 2m

14(b)	How do you characterize the surface of a substance by X-ray diffraction method? Surface Diagram X-ray metod	 1m 2m 2m
15(a)	Discuss the synthesis of nanometals by chemical method. Enumerate the applications of nanomaterials in catalysis Synthesis Methods apllications	 2m 2m 1m
15(b)	What is adsorption isotherm? Give its significance Definitions significance	4m 3m