NSRIT

AUTONOMOUS

ANSWER KEY & SCHEME OF EVALUATION

First Year B. Tech.
(Sem. II)

ACADEMIC REGULATION 2020

Academic Year
2020 - 2021

Semester End Examination, October, 2021

Degree	B. Tech. (U. G.)	Program	Common to All			Academic Year	2020 - 2021
Course Code	20BSX12	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	PARTIAL DIFFERENTIAL EQUATIONS AND VECTOR CALCULAS						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Form the PDE by eliminating arbitrary constants a and b from $z = ax + by + a^2 + b^2$	20BSX12.1	L1
2	Solve $(D - D')(D + D' - 3)z = 0$	20HSX12.2	L2
3	Compute $\beta(\frac{3}{2}, \frac{3}{2})$	20HSX12.3	L2
4	Define Solenoidal and Irrotational vectors	20HSX12.4	L1
5	Write the Statement of Stoke's Theorem	20HSX12.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Find the differential equation of all spheres whose centres lie on the z - axis.	6M	20BSX12.1	L2
6 (b)	Solve $x^2 (y - z)p + y^2(z - x)q = z^2 (x - y)$	6M	20BSX12.1	L3
	OR			
7 (a)	Solve $(\frac{p}{2} + x)^2 + (\frac{q}{2} + y)^2 = 1$	8M	20BSX12.1	L3
7 (b)	Solve $z = px + qy - \sqrt{2pq}$	4M	20BSX12.1	L2
8 (a)	Solve$(D^2 - 2DD' + D'^2)z = e^x + 4$	6M	20BSX12.2	L3
8 (b)	Solve $(4D^2 - 4DD' + D'^2)z = 16\log(x + 2y)$	6M	20BSX12.2	L2
	OR			
9 (a)	$(D + D' - 1)(D + 2D' - 3)z = 4 + 3x + 6y$	6M	20BSX12.2	L2
9 (b)	Solve $x^2 \frac{\partial z}{\partial x} + y^2 \frac{\partial z}{\partial y} = 0$ by the method of separation variables	6M	20BSX12.2	L3
10 (a)	Prove that $\int_0^{\frac{\pi}{2}} \sqrt{cot\theta}\, d\theta = \frac{1}{2}\Gamma\left(\frac{1}{4}\right)\Gamma\left(\frac{3}{4}\right)$	6M	20BSX12.3	L3
10 (b)	Evaluate $\int_0^a \int_0^x \int_0^{x+y} e^{x+y+z}\, dzdydx$	6M	20BSX12.3	L2
	OR			
11 (a)	Prove that $\int_0^1 \frac{x}{\sqrt{1-x^5}} dx = \frac{1}{5}\beta\left(\frac{2}{5}, \frac{1}{2}\right)$	6M	20BSX12.3	L3
11 (b)	Evaluate $\int_0^1 \int_0^{\sqrt{1-x^2}} \frac{dy\, dx}{1 + x^2 + y^2}$	6M	20BSX12.3	L2
12 (a)	Find the Directional Derivative of the function $f = x^4 + y^4 + z^4$ at the point (1, -2, 1) in the direction AB where B is (2, 6, - 1). Also find the maximum directional derivative of f at $(1, -2, 1)$	6M	20BSX12.4	L3

12 (b)	Show that $(x^2 - yz)\,\bar{i} + (y^2 - zx)\,\bar{j} - (z^2 - xy)\bar{k}$ is irrotational and hence find scalar potential	6M	20BSX12.4	L3
	OR			
13 (a)	If $\bar{F} = x^2yz\bar{i} + xy^2z\,\bar{j} + x\,y\,z^2\bar{k}$, find div $\bar{F}$ and curl $\bar{F}$ at the point (1, 2, 3)	6M	20BSX12.4	L3
13 (b)	Prove that $div(grad\, r^m) = m(m+1)r^{m-2}$	6M	20BSX12.4	L2
14	Verify Green's theorem for $\int_c [xy + y^2]dx + x^2\,dy$, where C is bounded by $y = x\ and\ y = x^2$	12M	20BSX12.5	L3
	OR			
15	Verify Gauss divergence theorem for the function $\bar{F} = y\bar{i} + x\bar{j} + z^2\bar{k}$ over the cylindrical region bounded by $x^2 + y^2 = 9, z = 0, z = 2$	12M	20BSX12.5	L3

SEMESTER END EXAM

Degree	B. Tech. (U. G.)	Program	Common to All	Test	I/II	Academic Year	2020 - 2021
Course Code	20BSX12	Test Duration	180 Min.	Max. Marks	70	Semester	I
Course	Partial Differential equations & vector Calculus						

Key and Scheme of Evaluation

No.	Questions (1 through 5)	Marks
1	By differentiating w.r.to x partially, we get p=a By differentiating w.r.to y partially, we get q=b	1
	By substituting in the given relation , we get z = p x + q y + p^2 + q^2	1
2	By comparing (D-D^1) (D+D^1-3) with (D-m_1D^1-α_1) (D-m_2D^1-α_2), we have $m_1 = 1,\ \ m_2 = -1,\ \alpha_1 = 0,\ \ \alpha_2$= 3.	1
	The solution is given by z= f_1(y + x) + $e^{3x}f_2$(y + 3x)	1
3	.β(m,n)= $\frac{\Gamma(m)\Gamma(n)}{\Gamma(m+n)}$ .In this put m=n=1/2.	1
	Then we get $\beta(\frac{3}{2},\frac{3}{2}) = \frac{\Gamma(\frac{3}{2})\Gamma(\frac{3}{2})}{\Gamma(3)} = \frac{1}{4}\frac{\Gamma(\frac{1}{2})\Gamma(\frac{1}{2})}{2} = \frac{\pi}{8}$	1
4	A Vector $\overline{F}$ is said to be Solenoidal if Div $\overline{F}$ =0 and	1
	Irrotational if curl $\overline{F} = \overline{0}$	1
5	Stoke's Theorem: If S is an open surface bounded by a closed curve C and if $\overline{F}$ is a continuously differentiable vector point function possessing continuous first order partial derivatives, then	1
	$\int_C \overline{F}.\overline{dr} = \iint_S \overline{F}.\overline{n}$ dS, where $\overline{n}$ is an unit outward drawn normal vector to the surface	1

No.	Questions (6 through 11)	
6(a)	The equation of the family of spheres having centers on z – axis is given by x^2+y^2+$(z-a)^2$=b^2	1
	By differentiating w.r.to x partially, we get 2x + 2(z-a) p =0 which gives x = -(z-a)p	2
	By differentiating w.r.to y partially, we get 2y + 2(z-a) q =0 which gives y = -(z-a)q	2
	So, $\frac{x}{y} = \frac{p}{q}$ or p y – q x = 0 is the required PDE.	1
6 (b)	The subsidiary equations are $\frac{dx}{x^2(y-z)} = \frac{dy}{y^2(z-x)} = \frac{dz}{z^2(x-y)} = \frac{ldx+mdy+ndz}{lx^2(y-z)+my^2(z-x)+nz^2(x-y)}$	1
	For the multipliers l=$\frac{1}{x^2}$, m= $\frac{1}{y^2}$, n= $\frac{1}{z^2}$, Denominator o the fourth fraction =0. So, we have Nr=o means $ldx + mdy + ndz$ =0 which gives us $\frac{1}{x^2}dx + \frac{1}{y^2}dy + \frac{1}{z^2}dz$ =0 On integration , we get $\frac{1}{x} + \frac{1}{y} + \frac{1}{z}$ = a.	2

Again for the multipliers l=$\frac{1}{x}$, m= $\frac{1}{y}$, n= $\frac{1}{z}$, Denominator o the fourth fraction =0.
So, we have Nr=o means $ldx + mdy + ndz$ =0
which gives us$\frac{1}{x}dx + \frac{1}{y}dy + \frac{1}{z}dz$ =0 2
On integration, we get x y z=b

Hence the general solution of the given PDE is given by f($\frac{1}{x}$ + $\frac{1}{y}$ + $\frac{1}{z}$, x y z) = 0. 1

$(\frac{p}{2} + x)^2 + (\frac{q}{2} + y)^2$ =1 is a non-linear PDE of first order of the form f (x , p) = g (y, q) 1

$\left(\frac{p}{2} + x\right)^2 = 1 - \left(\frac{q}{2} + y\right)^2$. 2
Let both the sides be equal to k, some constant.
Then we have $(\frac{p}{2} + x)^2$ =k and 1 - $(\frac{q}{2} + y)^2 = k$.

7 (a) Which gives p = 2 ($\sqrt{k} - x$), q = 2($\sqrt{(1-k)}$ – y)
We have dz = p dx + q dy. 2
By substituting above values of p and q , we get
dz =2 ($\sqrt{k} - x$) dx + 2($\sqrt{(1-k)}$ – y)dy

On integration, we get z= 2 ($\sqrt{k}\, x$) - x^2 + 2($\sqrt{(1-k)}$ y) - y^2 + c. 1

The given PDE is of the form z = px + qy + f (p,q), which is a Clairaut's equation. 2

The general solution will be obtained by replacing p with a and q with b where a and b are arbitrary constants. 2

7 (b)

So, the general solution is given by z =a x + b y +$\sqrt{2ab}$, where a and b are arbitrary constants. 2

$$\left(D^2 - 2DD' + D'^2\right)z = e^x + 4$$

The A.E. is $(m-1)^2$ =0, which gives m = 1, 1 2
The C.F. is given by f(y+ x) + x g(y+ x)
The Particular integral is given by P.I. =$\frac{1}{D^2-2DD'+D'^2}(e^x + 4)$
=$\frac{1}{D^2-2DD'+D'^2}(e^x)$ + $\frac{1}{D^2-2DD'+D'^2}(4e^{(0)x})$ = $\mathrm{P.I_1}$+$\mathrm{P.I_2}$(say)

8 (a) $\mathrm{P.I_1}$ = $\frac{1}{D^2-2DD'+D'^2}(e^x)$
Put D=l, D'=0 3
=e^x
$\mathrm{P.I_2}$ = $\frac{1}{D^2-2DD'+D'^2}(4e^{(0)x})$ = $\frac{4}{0}$, which is a failure case.
=4 $\frac{1}{2D-2D\prime}$ $e^{(0)x}$= $\frac{4}{0}$, which is a failure case
=4$\frac{1}{2}$ $e^{(0)x}$ =2
So, P.I. = $e^x + 2$
The complete solution is given by z= C.F. + P.I. = f(y+ x) + x g(y+ x) + $e^x + 2$ 1

$$\left(4D^2 - 4DD' + D'^2\right)z = 16\log(x+2y)$$

8 (b) The A.E. is $(2m-1)^2$ =0, which gives m = $\frac{1}{2}$, $\frac{1}{2}$ 2
The C.F. is given by f(y+ $\frac{x}{2}$) + x g(y+ $\frac{x}{2}$)

The Particular integral is given by P.I. = $\frac{1}{4D^2-24D'+D'^2}$ $16\log(x+2y)$

$=4 \quad \frac{1}{D-\frac{D'}{2}}\frac{1}{D-\frac{D'}{2}}\log(x+2y)$

$=4 \quad \frac{1}{D-\frac{D'}{2}}\int \log(2c)dx$ where c= $y+\frac{x}{2}$ 3

$=4 \quad \frac{1}{D-\frac{D'}{2}} x\log(2c) = 4 \quad \frac{1}{D-\frac{D'}{2}} x\log(x+2y) = 4\int x\log(2c)dx = 2x^2\log(2c) =$

$2x^2\log(x+2y)$

The complete solution is given by z= C.F. + P.I. = f(y+$\frac{x}{2}$) + x g(y+$\frac{x}{2}$) + $2x^2\log(x+2y)$ 1

9 (a)

Given PDE is (D + D'-1) (D +2 D'-3) z = 4 + 3x + 6y

By comparing (D +$D^1 - 1$) (D+2D^1-3) with (D-$m_1 D^1$-α_1) (D-$m_2 D^1$-α_2),

We have $m_1 = 1$, $m_2 = 2$, $\alpha_1 = 1$, $\alpha_2 = 3$. 2

The C. F. is given by z= $e^x f_1$(y + x) + $e^{3x} f_2$(y + 2x)

P.I. = $\frac{1}{D+D^1-1}\frac{1}{D+2D^1-3}$ (4 + 3x + 6y)

$=[1-(D+D')]^{-1}[1-(\frac{D+2D'}{3})]^{-1}$ (4 + 3x + 6y) 1

= $[1+D+D']$ $[1+\frac{D}{3}+\frac{2}{3}D']$ (4 + 3x + 6y)

after neglecting all partial derivatives from second order

$=[1+D+D'+\frac{D}{3}+\frac{2}{3}D']$ (4 + 3x + 6y) 1

$=[1+\frac{4}{3}D+\frac{5}{3}D']$ (4 + 3x + 6y)

=4 + 3x + 6y +$\frac{4}{3}$ D(4 + 3x + 6y) + $\frac{5}{3}D'$ (4 + 3x + 6y)

= 4 + 3x + 6y +4 + 10 1

=18 + 3x + 6y.

The complete solution is given by z= C.F. + P. I.

Z= $e^x f_1$(y + x) + $e^{3x} f_2$(y + 2x) + 18 + 3x + 6y. 1

9 (b)

Solve $x^2\frac{\partial z}{\partial x} + y^2\frac{\partial z}{\partial y} = 0$ by the method of separation variables.

Let us assume the complete solution as z=X(x).Y(y) 1

Then given PDE becomes $x^2X^1Y + y^2XY^1 = 0$

$\frac{x^2X^1}{X} = -\frac{y^2Y^1}{Y}$= k (say)

$\frac{x^2\frac{dX}{dx}}{X} = -\frac{y^2\frac{dY}{dy}}{Y}$ = k 2

Which gives $\frac{dX}{X}$ = k$\frac{dx}{x^2}$&$\frac{dY}{Y}$ = -k$\frac{dy}{y^2}$

By integrating, we get log X = $-\frac{k}{x}$& log Y = $\frac{k}{y}$

Which implies X = $e^{-\frac{k}{x}}$& Y = $e^{\frac{k}{y}}$ 2

So, the complete solution is z= $e^{-\frac{k}{x}}e^{\frac{k}{y}}$ 1

$\int_0^{\frac{\pi}{2}} \sqrt{\cot\theta}\, d\theta = \int_0^{\frac{\pi}{2}} \sqrt{\frac{\cos\theta}{\sin\theta}}\, d\theta = \int_0^{\frac{\pi}{2}} \cos^{\frac{1}{2}}\theta \sin^{\frac{-1}{2}}\theta\, d\theta$ 1

We have $\beta(m,n) = 2\int_0^{\frac{\pi}{2}} \cos^{2m-1}\theta \sin^{2n-1}\theta\, d\theta$ 1

10 (a) If we put 2m-1=p and 2n-1=q, then m= $\frac{p+1}{2}$ and n== $\frac{q+1}{2}$.

Then $\int_0^{\frac{\pi}{2}} \cos^{p}\theta \sin^{q}\theta\, d\theta = \frac{1}{2}\beta(\frac{p+1}{2},\frac{q+1}{2}) = \frac{1}{2}\beta(\frac{\frac{1}{2}+1}{2},\frac{\frac{-1}{2}+1}{2}) = \frac{1}{2}\beta(\frac{3}{4},\frac{1}{4})$ 2

$= \frac{1}{2}\frac{\Gamma(\frac{3}{4})\Gamma(\frac{1}{4})}{\Gamma(1)} = \frac{1}{2}\Gamma\left(\frac{3}{4}\right)\Gamma(\frac{1}{4})$ 2

$\int_0^{a}\int_0^{x}\int_0^{x+y} e^{x+y+z}\, dz\, dy\, dx =$

$\int_0^{a}\int_0^{x} (e^{2x+2y} - e^{x+y})\, dy\, dx$ 2

10 (b)

$= \int_0^{a}(\frac{e^{4x}}{2} - 3\frac{e^{2x}}{4} + e^{x})\, dx$ 2

$= \frac{e^{4a}}{8} - 3\frac{e^{2a}}{4} + e^{a} - \frac{3}{8}$ 2

$\int_0^1 \frac{x}{\sqrt{1-x^5}}$ dx. In this integral put $x^5 = y$. Then x = $y^{\frac{1}{5}}$ and dx = $\frac{1}{5}y^{\frac{-4}{5}}$ dy 1

x=0 gives y=0, x =1 gives y=1. 1

11 (a) So, given integral changes to $\int_0^1 \frac{y^{\frac{1}{5}}}{\sqrt{1-y}}\frac{1}{5}y^{\frac{-4}{5}}\, dy = \int_0^1 y^{\frac{-3}{5}}(1-y)^{\frac{-1}{2}}\, dy$ 1

By comparing this with $\int_0^1 y^{m-1}(1-y)^{n-1}$ dy, we have m = $\frac{2}{5}$ and n = $\frac{1}{2}$. 1

But we know that $\int_0^1 y^{m-1}(1-y)^{n-1}$ dy, = β(m,n) 2

And hence , given integral = $\beta(\frac{2}{5},\frac{1}{2})$.

The given double integral is $\int_0^1\int_0^{\sqrt{(1+x^2)}} \frac{dy\,dx}{1+x^2+y^2}$. Since y is possessing variable limits, first we have to integrate w.r.to y.

11 (b) 2

Let us assume $\sqrt{(1+x^2)}$ = k, a constant . Let us assume $\sqrt{(1+x^2)}$ = k, a constant . Then $1+x^2+y^2$ becomes k^2+y^2

So, the D.I. = $\int_0^1\int_0^{k}\frac{dy\,dx}{k^2+y^2} = \int_0^1 \frac{1}{k}\tan^{-1}(1)\, dx = \frac{\pi}{4}\int_0^1 \frac{1}{\sqrt{(1+x^2)}}dx$ 2

$= \frac{\pi}{4}\sinh^{-1}(1)$

$= \frac{\pi}{4}\log(1+\sqrt{2})$. 2

The normal to the surface $\emptyset$ =0 at any point is given by

$\nabla f = \nabla(x^4+y^4+z^4) = \bar{\imath}(4x^3) + \bar{\jmath}(4y^3) + \bar{k}(4z^3)$

At the point A(1, -2, 1), $\nabla f = 4\bar{\imath} - 32\bar{\jmath} + 4\bar{k}$ 3

The vector $\overline{AB}$ where B= (2,6 -1) is given by $\bar{\imath}+8\bar{\jmath}-2\bar{k}$

12(a) The directional derivative of f at A in the direction of $\overline{AB}$ is $\frac{\nabla f.\overline{AB}.}{|\overline{AB}|}$ 1

$= \frac{(4\bar{\imath}-32\bar{\jmath}+4\bar{k})(.\bar{\imath}+8\bar{\jmath}-2\bar{k})}{\sqrt{69}} = \frac{4+256-8}{\sqrt{69}} = \frac{252}{\sqrt{69}}$ 2

Curl ($\overline{F}$)= $\begin{vmatrix} \bar{\imath} & \bar{\jmath} & \bar{\bar{k}} \\ \frac{\partial}{\partial x} & \frac{\partial}{\partial y} & \frac{\partial}{\partial z} \\ x^2 - yz & y^2 - zx & z^2 - xy \end{vmatrix}$

= $\bar{\imath}\,[\frac{\partial}{\partial y}(z^2 - xy) - \frac{\partial}{\partial z}(y^2 - zx)] + \bar{\jmath}\,[\frac{\partial}{\partial z}(x^2 - yz) - \frac{\partial}{\partial x}(z^2 - xy)] + \bar{k}\,[\frac{\partial}{\partial x}(y^2 -$ 2

$zx) - \frac{\partial}{\partial y}(x^2 - yz)\,]$

= $\bar{\imath}(-x + x) + \bar{\jmath}(-y + y) + \bar{k}(-z + -z) = \bar{\imath}(0) + \bar{\jmath}(-0) + \bar{k}(0) = \bar{0}$

Hence, $\overline{F}$ is irrotational.

Let $\emptyset$ be the scalar potential of $\overline{F}$. Then we have $\nabla\emptyset = \overline{F}$

12(b) i.e. $\bar{\imath}\frac{\partial\emptyset}{\partial x} + \bar{\jmath}\frac{\partial\emptyset}{\partial y} + \bar{k}\frac{\partial\emptyset}{\partial z} = .\ \bar{\imath}(x^2 - yz) + \bar{\jmath}(y^2 - zx) + \bar{k}(z^2 - xy)$ 1

On comparing the coefficients, we get $\frac{\partial\emptyset}{\partial x} = x^2 - yz$, $\frac{\partial\emptyset}{\partial y} = y^2 - zx$, $\frac{\partial\emptyset}{\partial z} = z^2 - xy$ 1

By partial integration, we get

$\emptyset = \frac{x^3}{3} - xyz + f(y, z)$

$\emptyset = \frac{y^3}{3} - xyz + g(y, z)$ 1

$\emptyset = \frac{z^3}{3} - xyz + h(y, z)$

From these three, we conclude that $\emptyset = \frac{x^3}{3} + \frac{y^3}{3} + \frac{z^3}{3} - xyz$ 1

Given vector is $\overline{F} = x^2yz\,\bar{\imath} + xy^2z\,\bar{\jmath} + xyz^2\,\bar{k}$

Div ($\overline{F}$) = $\frac{\partial}{\partial x}(x^2yz) + \frac{\partial}{\partial y}(xy^2z) + \frac{\partial}{\partial z}(xyz^2) = 2xyz + 2xyz + 2xyz = 6xyz.$ 3

At the point (1,2,3), Div$\overline{F}$ =6(1)(2)(3) =36

13(a) Curl ($\overline{F}$)= $\begin{vmatrix} \bar{\imath} & \bar{\jmath} & \bar{\bar{k}} \\ \frac{\partial}{\partial x} & \frac{\partial}{\partial y} & \frac{\partial}{\partial z} \\ x^2yz & xy^2z & xyz^2 \end{vmatrix}$ 1

= $\bar{\imath}\,[\frac{\partial}{\partial y}(xyz^2) - \frac{\partial}{\partial z}(xy^2z)] + \bar{\jmath}\,[\frac{\partial}{\partial z}(x^2yz) - \frac{\partial}{\partial x}(xyz^2)] + \bar{k}\,[\frac{\partial}{\partial x}(xy^2z) - \frac{\partial}{\partial y}(x^2yz)\,]$

= $\bar{\imath}(xz^2 - xy^2) + \bar{\jmath}((x^2y - yz^2) + \bar{k}(y^2z - x^2z)$ 1

At the point (1,2,3),Curl$\overline{F}$ = $5\bar{\imath} - 16\bar{\jmath} + 9\bar{k}$ 1

$\bar{r} = x\bar{\imath} + y\bar{\jmath} + z\bar{k}$ gives us r = $\sqrt{x^2 + y^2 + z^2}$ 1

$\frac{\partial r}{\partial x} = \frac{x}{r}, \frac{\partial r}{\partial y} = \frac{y}{r}, \frac{\partial r}{\partial z} = \frac{z}{r}$

Grad (r^m) = $mr^{m-1}\frac{\bar{r}}{r}$ = m $r^{m-1}\frac{x\bar{\imath} + y\bar{\jmath} + z\bar{k}}{r}$ = m $r^{m-2}(x\bar{\imath} + y\bar{\jmath} + z\bar{k})$ 2

13(b) Div(grad r^m) = $\sum\frac{\partial}{\partial x}(m\, r^{m-2}x)$ = m r^{m-2} + m (m-2) $r^{m-3}(\frac{x}{r})x$ 1

= $\sum[\mathrm{m}\, r^{m-2} + \mathrm{m}\,(\mathrm{m}-2)\, r^{m-4}x^2]$ = $3\mathrm{m}\, r^{m-2} + \mathrm{m}\,(\mathrm{m}-2)\, r^{m-2}$ 2

= m (m+1) r^{m-2}

LHS : Evaluating the line integral $\oint(xy + y^2)$ dx + $x^2\, dy$:

The line integral is the sum of the line integrals along y = x^2 and along y= x .
Along y = x^2, dy = 2x dx and x : 0 → 1. So, the line integral along y = x^2 becomes
$\int_0^1(x^3+x^4)$ dx +2 x^3 dx = $\int_0^1(3x^3+x^4)$ dx =$(\frac{3x^4}{4}+\frac{x^5}{5})$ where x : 0 → 1

14(a) Which gives us $\frac{3}{4}+\frac{1}{5}=\frac{19}{20}$. 6

Along y = x, from A to O , dy =dx and x :1→ 0 .
So, the line integral along y = x becomes $\int_0^1 3x^2$ dx
Which is same as $x^3$where x : 1→ 0 which gives us -1
So, LHS = $\frac{19}{20}$-1 = $\frac{-1}{20}$.
Evaluating RHS :
M= xy + x^2 and N = x^2. So, $\frac{\partial N}{\partial x} = 2x$, $\frac{\partial M}{\partial y}$ = x + 2y which gives
$\frac{\partial N}{\partial x} = 2x$, $\frac{\partial M}{\partial y}$ = x – 2y
BY Green's Theorem, $\int_C Mdx + N\,dy$ = $\iint_R(\frac{\partial N}{\partial x} - \frac{\partial M}{\partial y})$dy dx

6

So, $\iint_R \left(\frac{\partial N}{\partial x} - \frac{\partial M}{\partial y}\right)$dy dx = $\int_0^1 \int_{x^2}^{x} (x - 2y)dy\,dx$ = $\int_0^1 (x^4 - x^3)$ dx

Which on simplification gives $\frac{-1}{20}$.

Hence LHS = RHS, which means that Green's theorem is verified.

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Semester End Examination, Sept./Oct., 2021

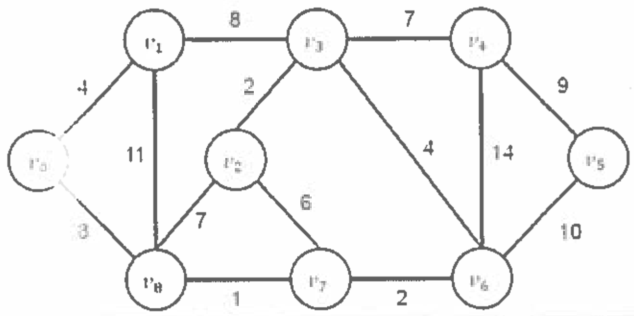
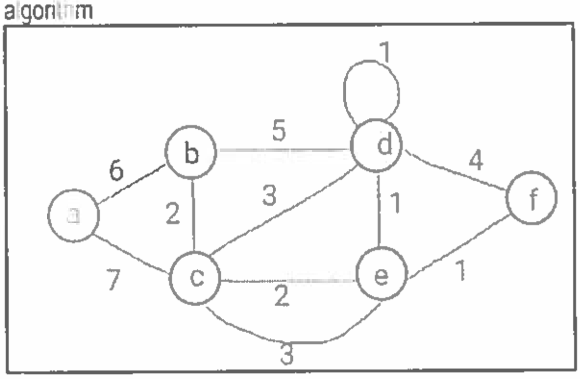
Degree	B. Tech. (U. G.)	**Program**	CSE/CSM/CSD			**Academic Year**	2020 - 2021
Course Code	20CS201	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	II
Course	**Data Structures using 'C'**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Compare Linear and Binary Search techniques	20CS201.1	L1
2	List any two advantages and disadvantages of using a Linked List	20CS201.2	L1
3	What are the conditions to be checked while using stack?	20CS201.3	L1
4	State the following terms: 1. Depth of a node 2. Height of a Tree	20CS201.4	L1
5	List any two applications of graphs	20CS201.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Outline the steps to perform binary search on a sorted array of 'N' numbers. Write the algorithm. Trace your algorithm with an example	6M	20CS201.1	L2
6 (b)	Show how the following numbers are sorted using quick sort: 8, 1, 4, 9, 0, 3, 5, 2, 7, 6	6M	20CS201.1	L3
	OR			
7 (a)	Outline the steps to perform linear search on an array of 'N' numbers. Write the algorithm. Trace your algorithm with an example	6M	20CS201.1	L2
7 (b)	Illustrate by sorting the given array using Bubble Sort. A={10,40,25,30,20}. Pictorially show the sorting iteration by iteration	6M	20CS201.1	L3
8	Explain all possible insertion operations on single linked list with corresponding algorithm using 10, 20, 30, 40, 50. And sketch stepwise procedure from start to end	12M	20CS201.2	L2
	OR			
9 (a)	Explain the algorithm to insert at front and delete at front operations on Doubly Linked List	8M	20CS201.2	L2
9 (b)	Compare Singly Linked List with Doubly Linked List	4M	20CS201.2	L2
10 (a)	Convert the given infix to postfix expression: A/B ^ C + D \ E - A \ C. Give the algorithm for the same	6M	20CS201.3	L2
10 (b)	Explain push and pop operations of stack in its array implementation	6M	20CS201.3	L2
	OR			
11	Explain the linked list implementation of Queue with necessary diagrams and algorithms	12M	20CS201.3	L2
12 (a)	Construct a binary tree and perform traversal for performing preorder, inorder and post-order on the following sequence 8->7->6->9->11->10->12	6M	20CS201.4	L1
12 (b)	Explain representation of binary tree in memory with a neat sketch	6M	20CS201.4	L2

<table>
<tr><td colspan="5">OR</td></tr>
<tr><td>13</td><td>Explain the construction of binary search tree with example. Draw the diagram showing the step by step process</td><td>12M</td><td>20CS201.4</td><td>L1</td></tr>
<tr><td>14 (a)</td><td>Find the minimum spanning tree for the given graph using Kruskal algorithm

</td><td>6M</td><td>20CS201.5</td><td>L2</td></tr>
<tr><td>14 (b)</td><td>Explain the Breadth First traversal with an example</td><td>6M</td><td>20CS201.5</td><td>L2</td></tr>
<tr><td colspan="5">OR</td></tr>
<tr><td>15 (a)</td><td>Find the minimum spanning tree for the given graph using prim's algorithm

</td><td>6M</td><td>20CS201.5</td><td>L2</td></tr>
<tr><td>15 (b)</td><td>Explain the Depth First traversal with an example</td><td>6M</td><td>20CS201.5</td><td>L2</td></tr>
</table>

Semester Question Paper

Degree: B.Tech (U.G) | Program: CSE/CS M/CSD | Test: I/II | Acd.y 2020-21

Course code: 20CS201 | Test Duration: 90 min | Max. Marks: 40 | Semester II

Course: Datastructures using 'C'

Key and Scheme of Evaluation

No. Questions (1 through 5)

1) Compare Linear and Binary Search techniques

- I/p datata needs to be Sorted in Binary Search and not in Linear Search
- Linear Search does the Sequential acces whereas Binary Search acces data randomly
- Time Complexity of linear Search - O(n), Binary Search has time complexity O(log n).
- Linear Search performs equality Comparisons and Binary Search performs Ordering Comparisons

2) List any two advantages and disadvantages of using a linked list Advantages of Linked List

- Dynamic Data Structure
- Insertion and deletion
- No Memory Wastage

Disadvantages of Linked List

- Memory Usage
- Traversal

- Reverse Traversing

3) What are the Conditions to be checked while Using stack?

Stack is a linear data structure that follows a particular order in which the operations are performed. The order ma out) or FILO (first in last OUT)

Mainly the following three basic operations are performed in the stack:

- Push: Adds an item in the stack. If the stack is full, then it is said to be an Overflow Condition.
- Pop: Removes an item from the stack. The items are popped in the reversed order in which they are pushed then it is said to be an Underflow Condition.
- Peek or Top: Returns the top element of the stack.
- is Empty: Returns true if the stack is empty, else false.

4) State the following terms: 1. Depth of a Node 2. Height of a Tree.

- The depth of a Node is the no. of edges from the node to the tree's root node.
  A root node will have a depth of 0.
- The height of a node is the no of edges on the longest path from the node to a leaf.
  A leaf node will have a height of 0.

5) List any two applications of graphs.

In Computer Science graph theory is used for the study of algorithms like

- Dijkstra's Algorithm
- Prim's Algorithm

PRINT "VALUE IS NOT PRESENT IN THE ARRAY" [END OF IF]

Step 6: EXIT

In 1st step:

BEG = 0

END = 8 ~~ron~~

MID = 4

a[mid] = a[4] = 13 < 23, therefore

In second step:

beg = mid + 1 = 5

End = 8

mid = 13/2 = 6

a[mid] = a[6] = 20 < 23, therefore;

in third Step:

beg = mid + 1 = 7

End = 8

mid = 15/2 = 7

a[mid] = a[7]

a[7] = 23 = item.

therefore, set location = mid;

The location of the item will be 7.

Item to be searched = 23

Step 1

1	5	7	8	13	19	20	23	29
0	1	2	3	4	5	6	7	8

a[mid] = 13

13 < 23

beg = mid + 1 = 5

end = 8

mid = (beg + end)/2 = 13/2 = 6

- Kruskal's Algorithm

- Graphs are Used to define the flow of Computation.
- Graphs are Used to represent networks of Communication.
- Graphs are used to represent data Organization.
- Graph transformation System work on rule-based in-memory manipulation of graphs ensure transaction - Safe, persistent storing and querying of graph Structured data.
- Graph theory is Used to find shortest path in road or a network.
- In google Maps, Various locations are represented as vertices or nodes and the roads are re and graph theory is Used to find the shortest path between two nodes.

No. Questions (6 through 12)

6(a) Outline the Steps to perform binary Search on a Sorted array of 'N' numbers. Write the algorithm. Trace your algorithm with example (6)

Binary_Search (A, lower_bound, upper_bound, VAL)

Step1: [Initialize] SET BEG = lower_bound END = upper_bound, POS = -1

Step2: Repeat Step 3 and 4 while BEG <= END

Step3: SET MID = (BEG + END) / 2

Step4: IF A[MID] = VAL SET POS = MID PRINT POS

Go to Step 6 ELSE IF A [MID] > VAL

SET END = MID - 1

ELSE SET BEG = MID + 1 [END OF IF]

[END OF LOOP] Step 5: IF POS = -1

step 2 →

1	5	7	8	13	~~19~~	~~20~~	23	~~24~~
0	1	2	3	4	5	6	7	8

a [mid] = 20

20 < 23

beg = mid + 1 = 7

end = 8

mid = (beg + end)/2 = 15/2 = 7

Step 3 →

1	5	7	8	13	19	20	~~23~~	~~24~~
0	1	2	3	4	5	6	7	8

a [mid] = 23

23 = 23

loc = mid

Return location 7

6(b) Show how the following numbers are sorted using quick sort: 8, 1, 4, 9, 0, 3, 5, 2, 7, 6 (6M)

8 1 4 9 0 3 5 2 7 6
Pi j

8 1 4 9 0 3 5 2 7 6
p ~~i~~ ~~i~~ i j

8 1 4 6 0 3 5 2 7 9
p ~~i~~ ~~i~~ ~~i~~ ~~i~~ ~~i~~ j i ~~j~~

7 1 4 6 0 3 5 2 | 8 | 9
Pi j p

7 1 4 6 0 3 5 2 8 9
p ~~j~~ ~~i~~ ~~i~~ ~~i~~ ~~i~~ ~~i~~ j i

2 1 4 6 0 3 5 | 7 | 8 9
p

2 1 4 6 0 3 5 7 8 9
Pi j

2 1 4 6 0 3 5 7 8 9
p i i ~~j~~ ~~j~~ ~~j~~

2 1 0 6 4 3 5 7 8 9
p i j

2 1 0 6 4 3 5 7 8 9

p ij ij j

0 1 | 2 | 6 4 3 5 7 8 9

p

0 1 2 | 6 4 3 5 | 7 8 9

pi i i ji

0 1 2 5 4 3 | 6 | 7 8 9

p

0 1 2 3 4 5 6 7 8 9

5 4 3

pi i j

3 4 5

p

(OR)

7 (a) Outline the steps to perform linear search on an array of 'N' numbers. Write the algorithm. Trace your algorithm

linear_search (A, N, VAL)

Step 1: [INITIALIZE] SET POS = -1

Step 2: [INITIALIZE] SET I = 1

Step 3: Repeat Step 4 while I <= N

Step 4: IF A[I] = VAL

SET POS = I PRINT POS

Go to step 6 [END OF IF]

SET I = I + 1 [END OF LOOP]

Step 5: IF POS = -1

PRINT "VALUE IS NOT PRESENT IN THE ARRAY"

[END OF IF] Step 6: EXIT

list

0	1	2	3	4	5	6	7	8
10	12	20	32	50	55	65	80	99

Search element 12

Step1 :
search element (12) is Compared with middle element (50)

list

0	1	2	3	4	5	6	7	8
10	12	20	32	~~50~~	55	65	80	99
				12				

Both are not matching. And 12 is Smaller than 50. So we Search only in the left sublist (i.e 10, 12, 20 & 32)

list

0	1	2	3
10	12	20	32

Step2 :
Search element (12) is Compared with middle Element (12)

list

0	1	2	3
10	~~12~~	20	32
	12		

Both are matching. So the result is "Element found at index 1"

Search element 80

Step1 :
Search element (80) is Compared with middle Element (50)

list

0	1	2	3	4	5	6	7	8
10	12	20	32	~~50~~	55	65	80	99
				80				

Both are not matching. And 80 is larger than 50. So we Search only in the right Sublist (i.e 55, 65, 80 & 99).

list

5	6	7	8
55	~~65~~	80	99
	80		

Both are not matching. And 80 is larger than 65. So we Search only in the right sublist (i.e 80 & 99).

list

7	8
80	99

Step3
Search Element (80) is Compared with middle Element (80)

list

7	8
~~80~~	99

Both are matching. So the result 80 is "Element found at index 7"

7(b) Illustrate by Sorting the given array using bubble sort.
A = {10, 40, 25, 30, 20} : Pictorially show the sorting iteration.

A = {10, 40, 25, 30, 20}

Iteration 1

10, 40, 25, 30, 20

10, 25, 40, 30, 20

10, 25, 30, 40, 20

10, 25, 30, 20, $\boxed{40}$

Iteration 2

10, 25, 30, 20, 40

10, 25, 30, 20, 40

10, 25, 30, 20, 40

10, 25, 20, $\boxed{30}$, $\boxed{40}$

Iteration 3

10, 25, 20, 30, 40

10, 25, 20, 30, 40

10, 20, $\boxed{25}$, $\boxed{30}$, $\boxed{40}$

Iteration 4

10, 20, 25, 30, 40

10, 20, 25, 30, 40

8) Explain all possible insertion operations on single linked list with corresponding algorithm using 10, 20, 3

Stepwise procedure from start to end (12)

Inserting a new node in a linked list

A new node is added into an already existing linked list.

Case 1: The new node is inserted at the begining.

Step 1: If Avail = Null
write Overflow
Go to step 7
[End of if]

Step2: Set New_node = Avail

Step 3: Set Avail = Avail → Next

Step4: Set New_node → Data = Val

Step 5: Set New_node → next = Start

Step 6: Set Start = New_node

Step 7: Exit

Case 2: The new node is inserted at the end.

Step 1: If Avail = Null
write overflow
Go to Step 10
[End of If]

Step 2: Set new_node = Avail

Step 3: Set Avail = Avail → next

Step 4: Set New_node → data = Val

Step 5: Set new_node → next = null

Step 6: Set PTR = Start

Step 7: Repeat Step 8 while PTR → next != Null

Step 8: Set PTR = PTR → next
[END of loop]

Step 9: Set PTR → next = new_node

Step 10: Exit

Case 3:- The ~~Node~~ new node is inserted after a given node.

Step 1: If Avail = Null
Write overflow go to Step 12
[End of If]

Step 2: Set New_node = Avail

Step 3: Set Avail = Avail → next

Step 4: Set New_node → Data = Val

Step 5: Set PTR = start

Step 6: Set preptr = PTR

Step 7: Repeat Steps 8 and 9 while preptr → data

!= Num

Step 8: Set preptr = Ptr

Step 9: Set PTR = PTR → next

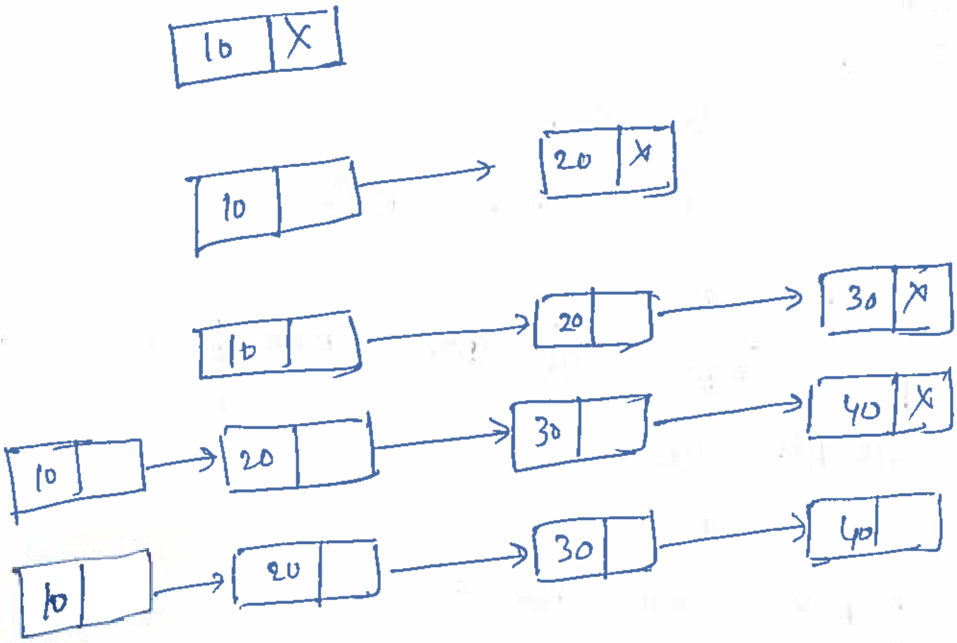
[End of Loop]

Step 10: Preptr → next = new_node

Step 11: Set New_node → next = PTR

Step 12: Exit

10 20 30 40 50



9(a). Explain the algorithm to insert at front and delete at front operations on Doubly linked list (8)

Inserting a Node at the beginning of a linked list

Step 1 : If Avail = Null
write overflow
Go to step 7
[End of If]

Step 2 : Set New_Node = Avail

Step 3 : Set Avail = Avail → Next

Step 4 : Set New_Node → Data = Val

Step 5 : Set New_Node → Next = Start

Step 6 : Set Start = New_Node

Step 7 : Exit

Inserting a Node at the End of a linked list

Step 1 : If Avail = Null
write overflow
Go to step 10
[End of If]

Step 2 : Set New_Node = Avail

Step 3 : Set Avail = Avail → Next

Step 4 : Set New_Node → Data = val

Step 5 : Set New_Node → Next = Null

Step 6 : Set PTR = START

Step 7 : Repeat Step 8 while PTR → Next != Null

Step 8 : Set PTR = PTR → Next
[End of Loop]

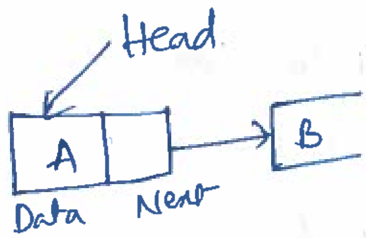
Step 9: Set PTR → Next = New_Node

Step 10: Exit

9 (b) Compare Singly Linked list with Doubly Linked list (4)

Singly linked list (SLL)

SLL nodes contain 2 field - data field and nex link field



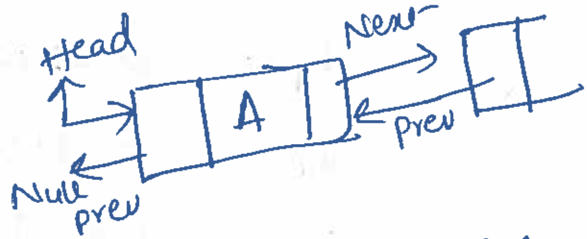
In SLL, the traversal can be done using the next node link only. Thus traversal is possible in one direction only

The SLL occupies less memory than DLL as it has only 2 fields

Complexity of insertion and deletion at a given positions is O(n)

Doubly linked list (DLL)

DLL nodes contain 3 fields data field, a previous link field and a next link field.



In DLL, the traversal can be done using the previous node link & the next node link. Thus traversal is possible in both directions (fwd & bckwd)

The DLL occupies more memory than SLL as it has 3 fields.

Complexity of insertion and deletion at a given position is O(1).

10(a)) Convert the given infix to postfix expression?

A/B^C+D/E-A/C. Give the algorithm for the same (6)

Step by step O/P for " expression.

Input String	Output Stack	Operator Stack
A/B^C+D/E-A/C	A	
A/B^C+D/E-A/C	A	/
A/B^C+D/E-A/C	AB	/
A/B^C+D/E-A/C	ABC	/^
A/B^C+D/E-A/C	ABC^/	+
A/B^C+D/E-A/C	ABC^/D	+
A/B^C+D/E-A/C	ABC^/D	+/
A/B^C+D/E-A/C	ABC^/DE	+/
A/B^C+D/E-A/C	ABC^/DE/+	-
A/B^C+D/E-A/C	ABC^/DE/+A	-
A/B^C+D/E-A/C	ABC^/DE/+A	-/
A/B^C+D/E-A/C	ABC^/DE/+AC	-/
A/B^C+D/E-A/C	ABC^/DE/+AC/-	

Step 1: Add ")" to the end of the infix expression

Step 2: push "(" on to the stack

Step 3: Repeat until each character in the infix notation is scanned

If a "(" is encountered, push it on the stack

If an operand (whether a digit or a character) is

(6)

encountered, add it to the post-fix expression.

If a ")" is encountered, then

a. Repeatedly pop from stack and add it to the postfix expression until a "(" is encountered.

b. Discard the "(". That is, remove the "(" from stack and do not add it to the postfix expression.

If an operator O is encountered, then

a. Repeatedly pop from stack and add each operator (popped from the stack) to 1 postfix expression which has the same precedence or a higher precedence than

b. Push the operator O to the stack

[End of If]

Step 4: Repeatedly pop from the stack and add it to the postfix expression until the stack i

Step 5: Exit

10.b) Explain Push and pop operations of stack in its array implementation (6)

Push Operation

Step 1 : If top = max-1
Print "Overflow"
Goto Step 4
[End of If]

Step 2 : Set Top = Top + 1

Step 3 : ~~set~~ stack [Top] = value

Step 4 : END

Pop Operation

Step1 : If Top = Null
Print "Underflow"
Goto Step 4
[End of if]

Step2 : Set Val = Stack [Top]

Step3 : Set top = top -1

Step 4 : End.

(OR)

11) Explain the linked list implementation of Queue with necessary diagrams and algorithms (12)

Operations on linked Queues

Step1 : Allocate memory for the new node and name it as PTR

Step2: Set PTR → Data = Val

Step3 : If front = Null
Set front = Rear = PTR
Set front → Next = Rear → Next = Null

Else
Set Rear → Next = PTR
Set Rear → PTR
Set Rear → Next = Null
[End of If]

Step 4 : END

Delete Operation

Step1 : If front = Null
write "Under flow"
Go to Step 5
[End of if]

Step2: Set PTR = front

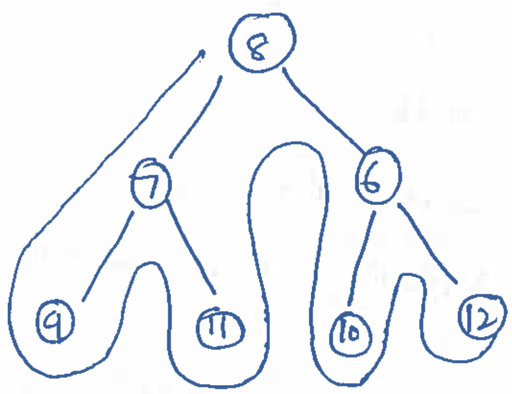
Step3: Set front = front → Next

Step4: Free PTR

Step5 : END.

12a) Construct a binary tree and perform traversal for ~~performance~~ performing preorder, inorder and post-order on the following sequence 8 → 7 → 6 → 9 → 11 → 10 → 12 (6)

8, ~~7~~, 6, 9, 11, 10, 12



Pre order:

8, 7, 9, 11, 6, 10, 12

In Order:

9, 7, 11 8 10 6 12

Post order:

9, 11, 7, 10, 12, 6, 8

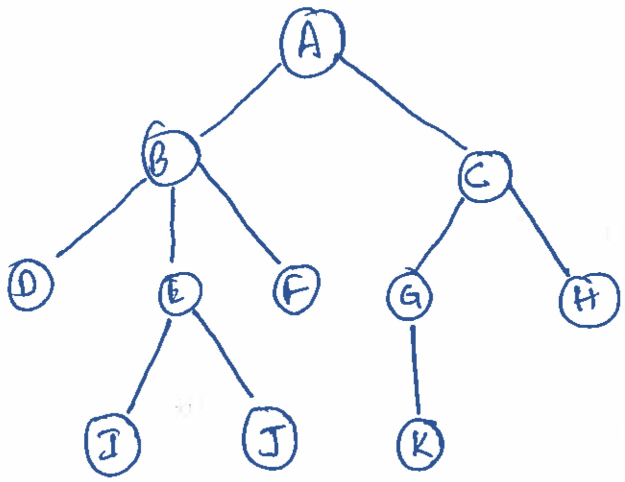
12 b) Explain representation of binary tree in memory with a neat sketch (6)

Tree Representations

A tree data structure can be represented in two methods. Those methods are as follows...

1. List Representation
2. Left child - Right Sibling Representation

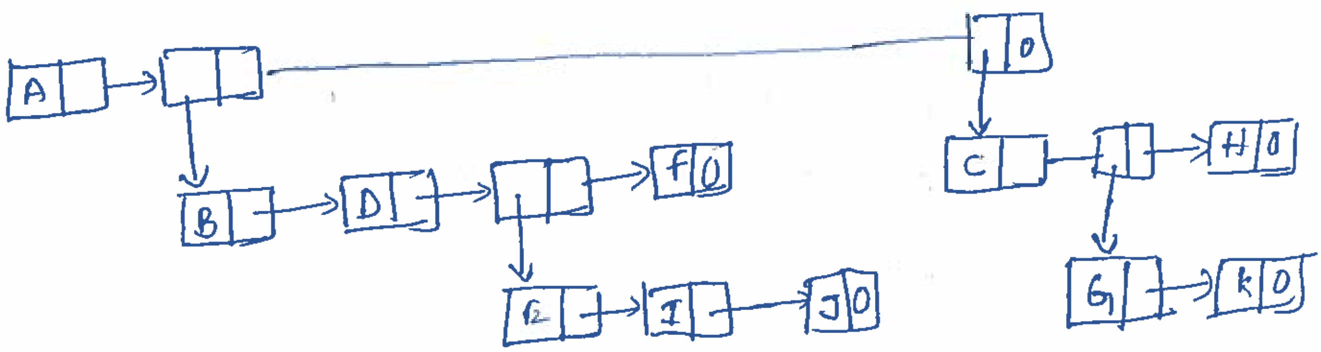
Consider the following tree...



TREE with 11 nodes and 10 Edges

- In any tree with 'N' nodes there will be max of 'N-1' edges
- In a tree every individual element is called as 'Node'.

1. List Representation.



13) Explain the construction of binary search tree with example. Draw the diagram showing the step by step process

Binary Search Tree

1. Binary Search tree can be defined as a class of binary tree, in which the nodes are arranged in a specific ordered binary tree.

2. In a binary Search tree, the Value of all the nodes on the left Sub-tree is less than the Value of the root.
3. Similarly value of all the nodes in the right Sub-tree is greater than or equal to the Value of the root.
4. This rule will be recursively applied to all the left and right Sub-trees of the root.

43, 10, 79, 90, 12, 54, 11, 9, 50

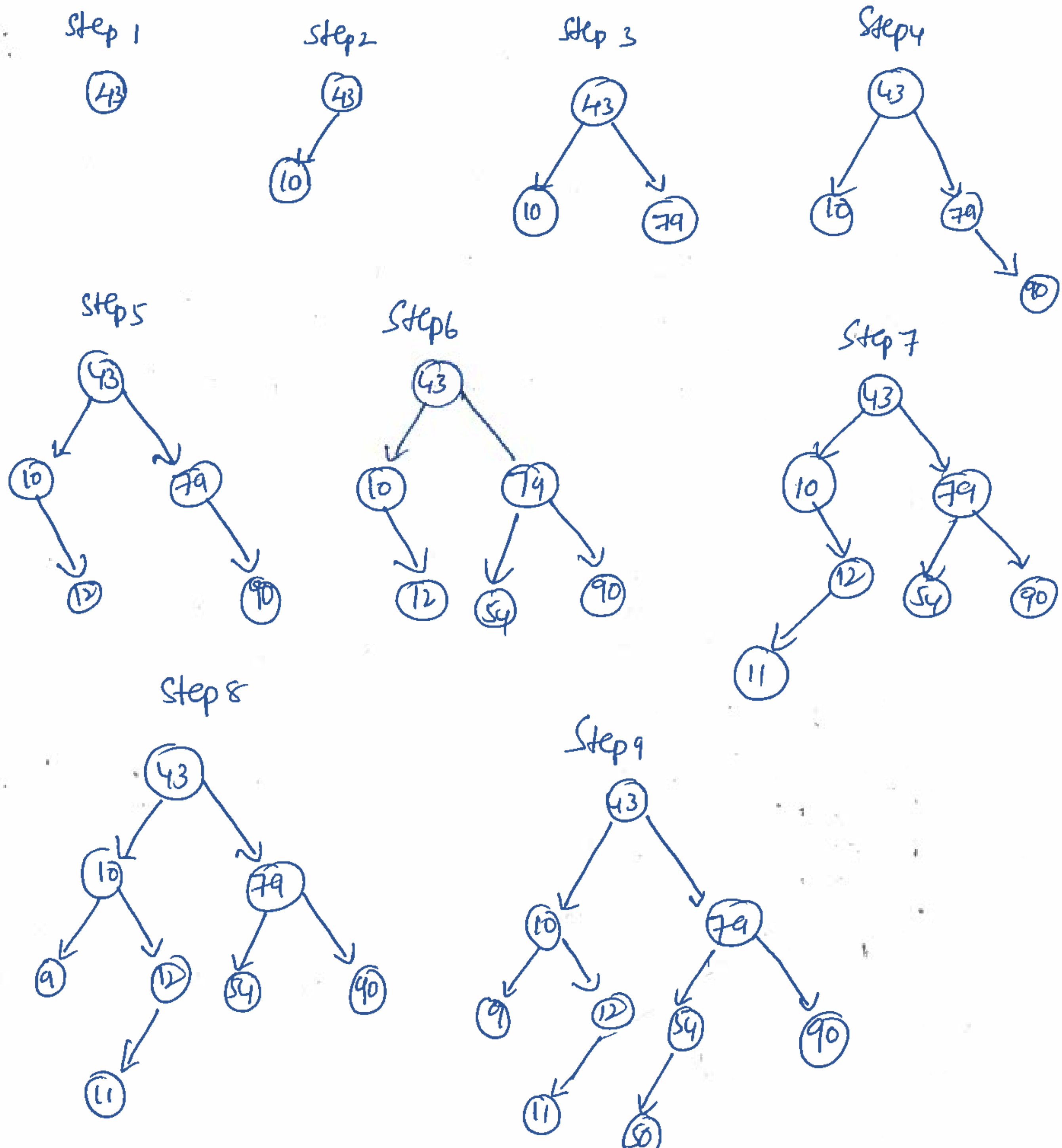
1. Insert 43 into the tree as the root of the tree.
2. Read the next element, if it is lesser than the root node element, insert it as the root of the left Sub-tree.
3. Otherwise, insert it as the root of the right of the right Sub-tree

- The process of Creating BST by using the given Elemen is shown in the image below.

43, 10, 79, 90, 12, 54, 11, 9, 50

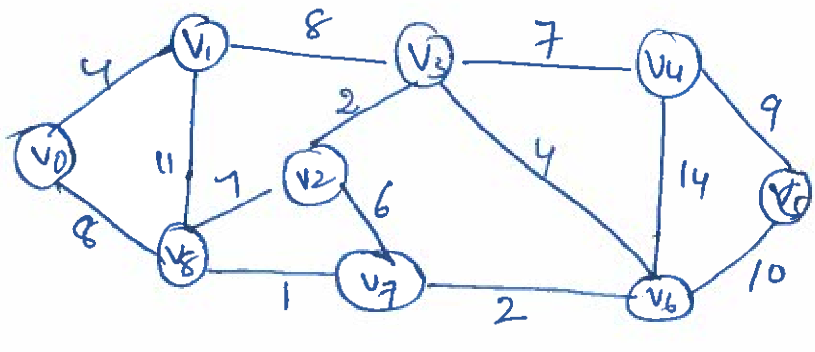
1. Insert 43 into the tree as the root of the tree
2. Read the next Element, if it is lesser than the root node Element, insert it
3. Otherwise, insert it as the root of the right of the right Sub-tree.

The process of Creating BST by Using the given elements, is shown in the image below.



Binary Search Tree Creation.

4(a) Find the minimum Spanning tree for the given graph using kruskal algorithm (6)



Step 01:

Sort all the edges from low weight to high weight.

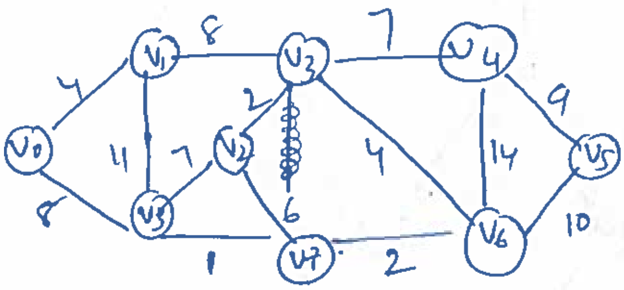
Step 2:

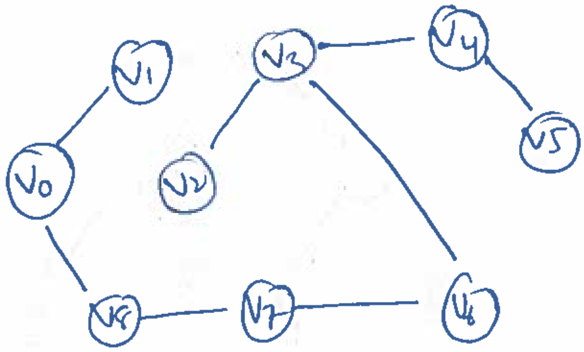
- Take the edge with the lowest weight and use it to connect the vertices of graph.
- If adding an edge create a cycle, then reject that edge and go for the next least weight

Step 3:

- Keep adding edges untill all the vertices are connected and a minimum Spanning Tree

Kruskal algorithm





V_0V_1	V_0V_8	V_1V_8	V_1V_3	V_3V_2
4	8	11	8	2
V_3V_4	V_3V_6	V_4V_5	V_4V_6	V_5V_6
7	4	9	14	10
V_6V_7	V_7V_8	V_8V_2	V_2V_8	
2	1	7	6	

V_7V_8	V_6V_7	V_3V_2	V_0V_1	V_3V_6	V_2V_3
1	2	2	4	4	6
V_8V_2	V_3V_4	V_0V_8	V_1V_3	V_4V_5	
7	7	8	8	9	
V_5V_6	V_1V_8	V_4V_6			
10	11	14			

14b) Explain the breadth first traversal with an example (6)

BFS (Breadth first Search)

BFS traversal of a graph produces a spanning tree as final result. Spanning Tree is a graph without loops. Structure with maximum size of total number of vertices in the graph to implement we use the following steps to implement BFS traversal...

Step1:- Define a queue of size total number of vertices in the graph.

Step2:- Select any vertex as starting point for traversal. Visit that vertex and insert it into the queue.

Step3:- Visit all the non-visited adjacent vertices of the vertex which is at front of the queue and insert them into the queue.

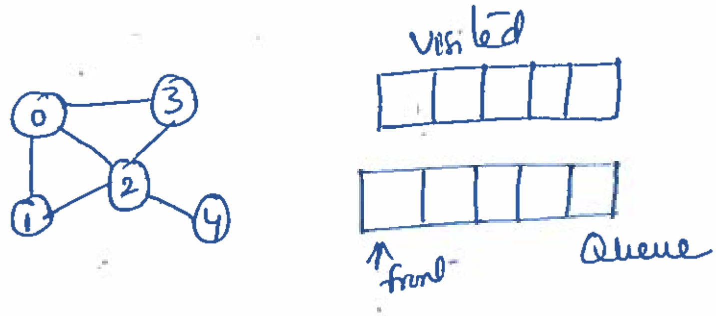
Step4:- When there is no new vertex to be visited from the vertex which is at front of the Queue then delete that vertex.

Step 5:- Repeat step 3 & 4 until queue becomes empty.

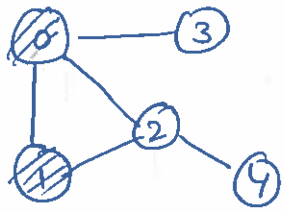
Step6:- When queue becomes empty, then produce final spanning tree by removing unused edges from the graph.

BFS

Step 1



Step2:



0				

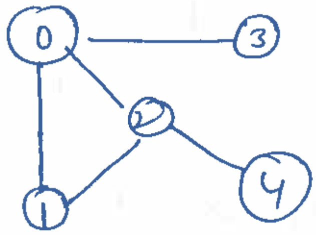
visited

1	2	3		

↑
f

Queue

Step 3:-



0	1	

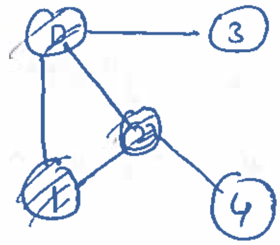
visited

2	3	

↑
front

Queue

step 4:-

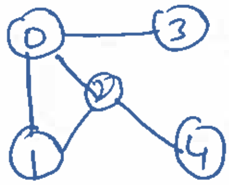


0	1	2	

3	4	

↑
front

step 5:-



0	1	2	3	

visited

4				

Queue

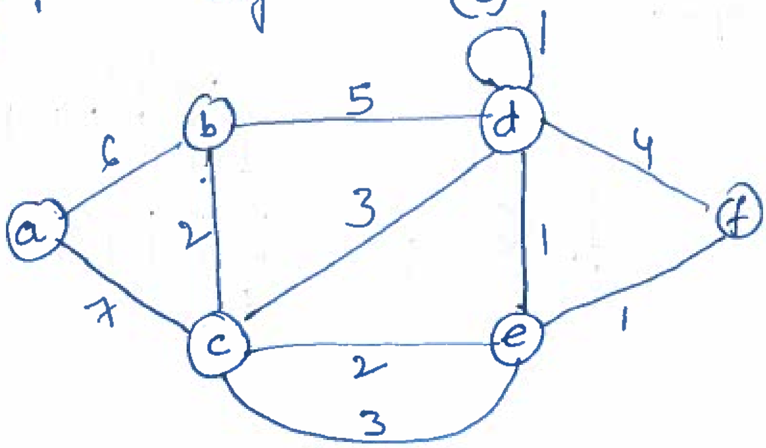
Step 6:-

visited

0	1	2	3	4

Queue

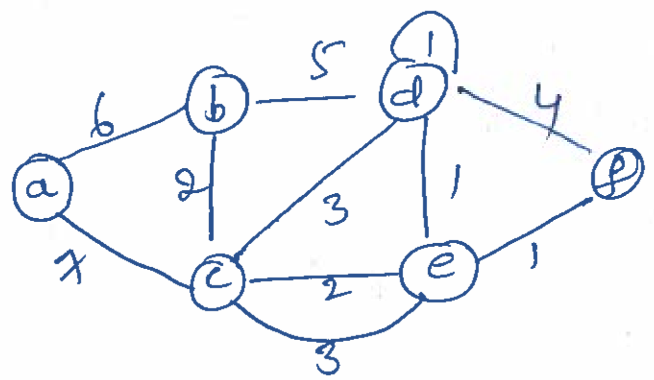
15 a) find the minimum Spanning tree for the given graph using prim's algorithm (6)

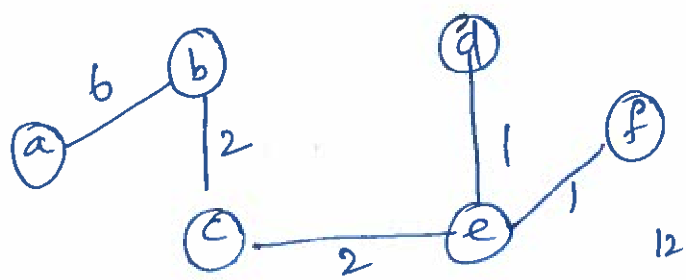


Step 01: • Randomly choose any vertex
• The vertex connecting to the edge having least weight is usually selected.

Step 02: • Find all the edges that connected the tree to new vertices
• Find the least weight edge among those edges and include it in the existing tree
• If including that edge creates a cycle, then reject that edge and look for the next least weight edge.

Step 03: Keep repeating Step-02 until all the vertices are included and Minimum Spanning Tree (MST) is obtained.





12 units

15 b) Explain the Depth first traversal with an example (6).

DFS (Depth first Search)

DFS traversal of a graph produces a spanning tree as final result. Spanning Tree is a graph without data structure with maximum size of total no of vertices in the graph to implement

We use the following steps to implement DFS traversal..

Step 1 :- Define a stack of size total no of vertices in the graph

Step 2 :- Select any vertex as starting point for traversal. Visit that vertex and push it on to the

Step 3 :- Visit any one of the non-visited adjacent vertices of a vertex which is at the top of sta the stack.

Step 4 :- Repeat step 3 until there is no new vertex to to be visited from the vertex which is at the

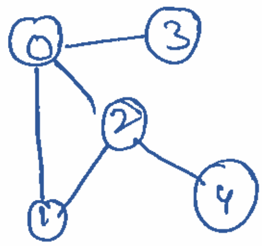
Step 5 :- When there is no new vertex to visit then use back tracking and pop one vertex from

Step 6 :- Repeat step 3, 4 and 5 until stack becomes empty.

Step 7 :- When stack becomes Empty, then produce final spanning tree by removing unused edge.

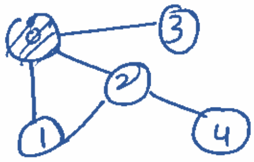
DFS

Step 1



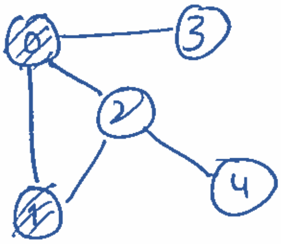
visited					
stack					

Step 2



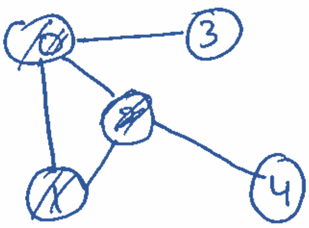
visited	0		
stack	1	2	3

Step 3



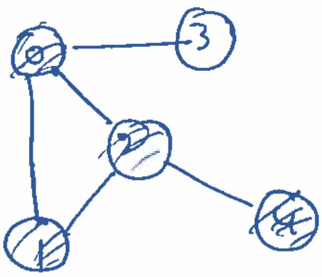
visited	0	1
stack	2	3

Step 4



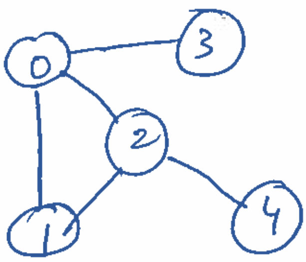
visited	0	1	2
stack	4	3	

Step 5



visited	0	1	2	4
stack	3			

Step 6



visited	0	1	2	4	3
stack					

N. Viswanatha Reddy

NSRIT

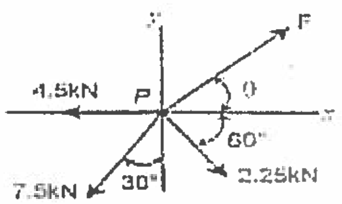
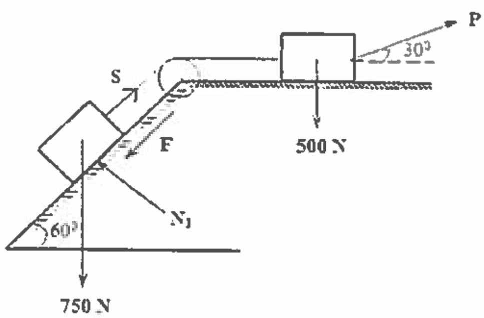
Semester End Examination, October, 2021

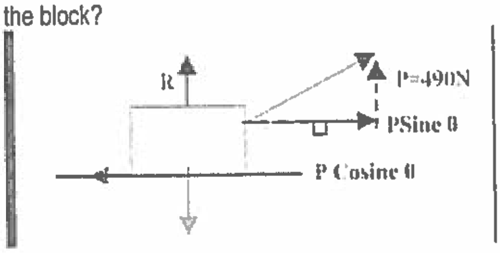
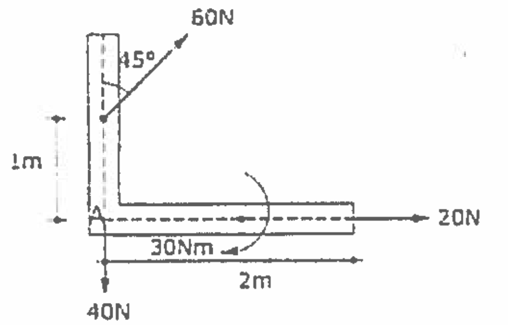
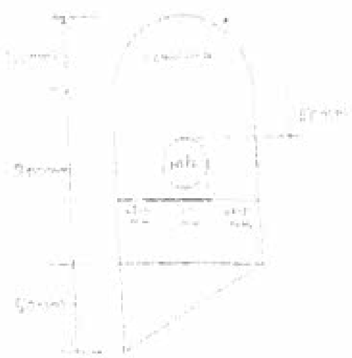
Degree	B. Tech. (U. G.)	**Program**	CE, EEE & ME	**Academic Year**	2020 - 2021
Course Code	20ESX04	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	II
Course	**ENGINEERING MECHANICS**				

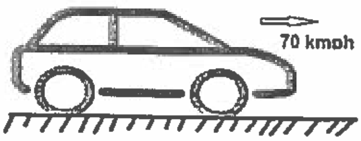
Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Lami's Theorm,	20ESX04.1	L1
2	Write any four advantages and limitations of friction	20ESX04.2	L1
3	Differentiate between moment of inertia and polar moment of inertia	20ESX04.3	L2
4	Define and mention units for velocity of projection	20ESX04.4	L1
5	Write Impulse Momentum Method.	20ESX04.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive and explain about Parallelogram Law	6M	20ESX04.1	L2
6 (b)	State and prove Triangular law of forces	6M	20ESX04.1	L3
	OR			
7 (a)	State and Explain the concept of Equilibrium	4M	20ESX04.1	L2
7 (b)	Determine the magnitude and angle and F so that particle shown in figure, is in Equilibrium 	8M	20ESX04.1	L2
8 (a)	What is the value of P in the system shown in the figure to cause the motion to impend? Assume the pulley is smooth and coefficient of friction between the other two contact surfaces is 0.20 	8M	20ESX04.2	L3
8 (b)	Define the following (i) Law of transmissibility (ii) Converse of the Law of Polygon of Forces	4M	20ESX04.2	L2
	OR			

9 (a)	A pull of 490 N inclined at 30° to the horizontal is necessary to move a block of wood on a horizontal table. If the coefficient of friction between two bodies in contact is 0.2. What is the mass of the block? 	7M	20ESX04.2	L2
9 (b)	Differentiate between the angle of repose and angle of friction	5 M	20ESX04.2	L3
10 (a)	Locate the centroid of L – section shown in figure 	7M	20ESX04.3	L3
10(b)	Explain briefly about Centre of Gravity using Varigmon's theorem	5M	20ESX04.3	L2
	OR			
11 (a)	Determine the centroid of a triangle having base width b and height h	6M	20ESX04.3	L3
11(b)	Locate the centroid of the following figure 	6M	20ESX04.3	L2
12 (a)	A man weight 100 Newton entered a lift, which moves with an acceleration of 5 m/sec². Find the force exerted by the man on the floor of lift when a) Lift is moving downward b) Lift is moving upward	5M	20ESX04.4	L3
12(b)	A motorist travelling at a speed of 80 kmph, suddenly applies brakes and halts after 70 m. Determine a) The time required to stop the car b) The coefficient of friction between the tyres and the road	7M	20ESX04.4	L3



	OR			
13(a)	A Particle is projected vertically upwards from the ground with an initial velocity of 10 m/sec. find a) The time taken to reach the maximum height b) The maximum height reached c) Time required for descending d) Velocity when it strikes the ground. Consider the upward motion of the particle	6M	20ESX04.4	L3
13(b)	A small Steel ball is shot vertically upwards from the top of a building 45m above the ground with an initial velocity of 28 m/sec a) In what time, it will reach the maximum height. b) How high above the building will the ball rise	6M	20ESX04.4	L3
14	Find the Power of a locomotive, drawing a train whose weight including that of engine is 500 kN up an incline 1 in 135 at a steady speed of 60 kmph, the frictional resistance being 6 N/kN. While the train is ascending the incline, the steam is shut off. Find how far it will move before coming to rest, assuming that the resistance to motion remains the same	12M	20ESX04.5	L3
	OR			
15	Derive the Work Energy equation for translation about Fixed Axis	12M	20ESX04.5	L3

Engineering Mechanics 20ESX04

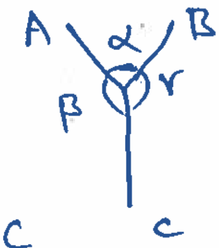
Semester End Examination

October, 2021

Part-A Short Answers $5 \times 2 = 10$ M

1. Define Lami's theorem (2m)

Sol: Three forces concentration at the point with three inclinations.



$$\frac{A}{\sin \gamma} = \frac{B}{\sin \beta} = \frac{C}{\sin \alpha}$$

2. Write any four advantages and limitations of friction. (2m)

Sol: 1) We can walk, stop the car, possible to transfer energy.
2) disadvantage heat, wear, thermal effect. etc.

3. (2m)

$$I = \int r^2 \, dA. \quad kg\text{-}m^2$$

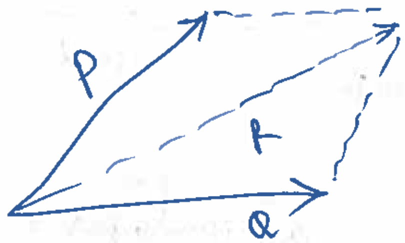
$$I = \int r^2 \, dA \quad m^4.$$

4. units = m/sec (2 m)

5. J = ΔP – momentum (2 m)

Part – B (Long Answer Questions)

6 (a)



6 M

$$R = \sqrt{P^2 + Q^2 + 2PQ\cos\theta}$$

explanation. $\alpha = \tan^{-1}\left[\frac{Q\sin\theta}{P + Q\cos\theta}\right]$

(b)

Explanation that Triangle force

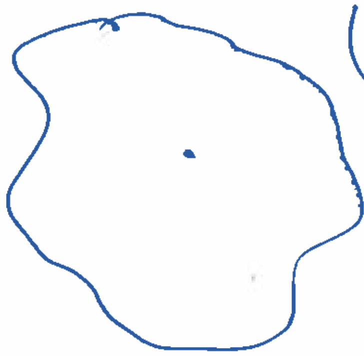
$$R = \sqrt{(\Sigma H)^2 + (\Sigma V)^2}$$ (6 m)

$$\theta = \tan^{-1}\left(\Sigma V / \Sigma H\right)$$

Q7.(a) $\Sigma H = 0$

$\Sigma V = 0$

$\Sigma m = 0.$



(4m)

(b) $\Sigma H = \Sigma H_1 + \Sigma H_2 + \Sigma H_3 + \Sigma H_4$

$\Sigma V = \Sigma V_1 + \Sigma V_2 + \Sigma V_3 + \Sigma V_4.$ (8m)

$$R = \sqrt{(\Sigma H)^2 + (\Sigma V)^2}$$

$$\theta = \tan^{-1}\left(\Sigma V / \Sigma H\right)$$

8(a) $\Sigma H = 0$ $\quad T - W\cos 30 - F_1 = 0.$

$T = 649.51 + 0.2 N_1$ (8m)

$\Sigma Y = 0.$

$N_1 - W\sin 30^\circ = 0.$ $\quad \boxed{N_1 = 375}$

$$\boxed{T = 727.51\,N}$$

$\Sigma H = 0.$

$P = 724.5 + (0.2)\, N_2$

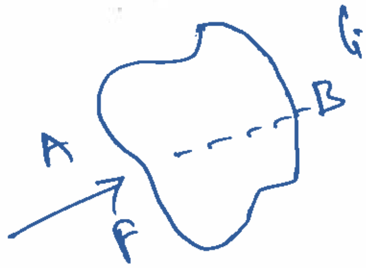
$\Sigma V = 0.$

$N_2 = -P/2 + 500.$

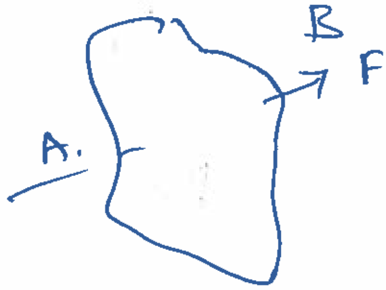
$\boxed{P = 853.57\ N}$

8(b) (4m)

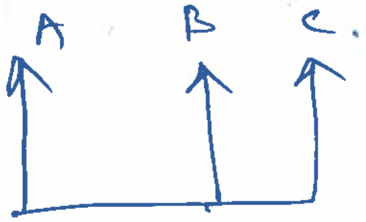
Law of Transmissibility.



line of action of force.



Law of polygon



$\Sigma V = A + B + C$

$\Sigma H = 0.$

9 (a) $\Sigma H = 0$ (7M)

$\Sigma V = 0.$

$P = ?$

(b) Angle of friction. (5M)

$\tan \phi = F/N = \mu$

Angle of Repose $\alpha = \phi.$

10 (a). <u>Centroid</u> (7M)

$$C = \left(\frac{A_1 x_1 + A_2 x_2}{A_1 + A_2} , \frac{A_1 y_1 + A_2 y_2}{A_1 + A_2} \right)$$

(b) <u>Centre of Gravity</u>

$$W = W_1 + W_2 + \cdots$$
$$= \Sigma W_i$$

$W\bar{x} = \Sigma W_i x_i$ (5m)

$W\bar{y} = \Sigma W_i y_i$

$W\bar{z} = \Sigma W_i z_i$

Mass $M\bar{x} = \Sigma m_i x_i$

$M\bar{y} = \Sigma m_i y_i$

$M\bar{z} = \Sigma m_i z_i$

Same of volume also

$\boxed{W = V_i \gamma}$

11 (a) $\left(h/3 \, , \, b/3 \right)$ (6m)

Centroid

(b)

(6m)

Centroid $= \left(\dfrac{A_1 x_1 + A_2 x_2}{A_1 + A_2} , \dfrac{A_1 y_1 + A_2 y_2}{A_1 + A_2} \right)$

NSRIT

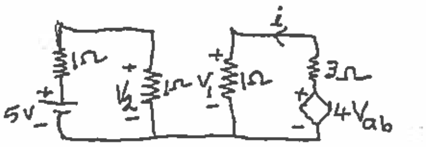
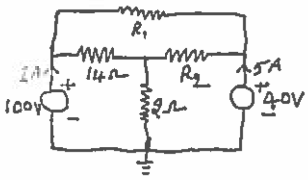
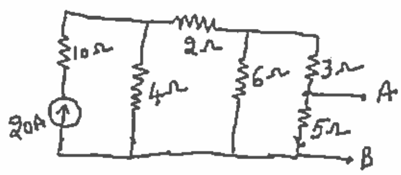
Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	**Program**	ECE	**Academic Year**	2020 - 2021
Course Code	20EE201	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	II
Course	Network Analysis and Synthesis				

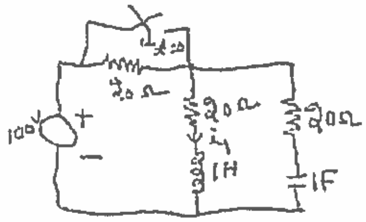
Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define series and parallel combinations	20EE201.1	L1
2	State Reciprocity theorem	20EE201.2	L1
3	Define steady state response	20EE201.3	L1
4	Define Resonance circuit	20EE201.4	L1
5	Write the reciprocity conditions for Z and Y parameters	20EE201.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Find the value of ' i ' current in the given circuit  if V_2 = 2.5V	6M	20ESX05.1	L3
6 (b)	What is the necessity of the Super Node and Super mesh analysis? Explain with one example	6M	20ESX05.1	L2
	OR			
7 (a)	Write the properties of Cut-set matrix and Tie-set matrix of network graph	6M	20ESX05.1	L2
7 (b)	Find the values of resistors R_1 and R_2 from the given circuit 	6M	20ESX05.1	L3
8 (a)	Write the statement of Nortan's theorem and procedure to find Nortan's resistance	6M	20ESX05.2	L2
8(b)	Find the current through the 5 ohms resistance using Nortan's theorem 	6M	20ESX05.2	L3
	OR			
9 (a)	Write the statement of super position theorem and write the limitations of super position theorem	6M	20ESX05.2	L3
9 (b)	State Millman's theorem and explain with one example	6M	20ESX05.2	L2

10 Find $\frac{d_i}{d_t}$ at $t = 0^{+}$

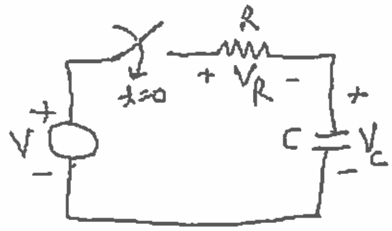


12M 20ESX05.3 L3

OR

11 (a) Evaluate the initial conditions procedure for RL and RC circuit — 8M 20ESX05.3 L2

11(b) Write the expressions for Voltage across R (V_R), Voltage across capacitor (V_c) with graphical representations



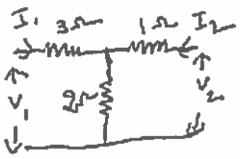
4M 20ESX05.3 L3

12 Derive the expression for frequency at which the voltage across capacitor is maximum in RLC series circuit — 12M 20ESX05.4 L2

OR

13 Derive the expression for Bandwidth and Quality factor of RLC series circuit and write the relation between bandwidth and Q — 12M 20ESX05.4 L3

14 Find the Y- parameters of the given network

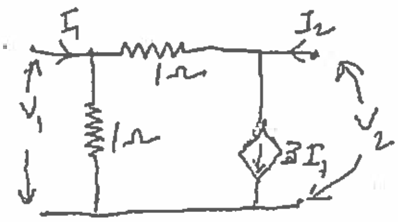


12M 20ESX05.5 L3

OR

15(a) Derive the relation between Z-parameters and ABCD-parameters of a two port networks — 6M 20ESX05.5 L2

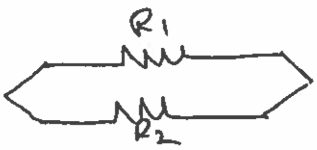
15(b) Find the Y parameters of the given network



6M 20ESX05.5 L2

1). series :- When 2nd terminal of 1st element connected to 1st terminal of 2nd element then both elements are said to be in series.

Parallel:- When all 1st terminals connected to single commo point and all 2nd terminals connected to another common point, then the elements are said to be in parallel.



2) Reciprocity theorem:- In a linear bilateral network, the ratio of excitation to response is equal in the case even though the positions of excitation and response are inter changed.

3) A steady state response is the behaviour of a circuit after a long time when steady conditions have been reached after an external excitation.

4) In a circuit, the state in which the current is maximum is called resonance. In other words, when net total current in an electrical circuit is in phase with applied voltage, then circuit is said to be in resonance

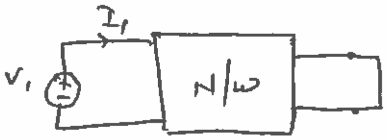
HOD-EEE
5/10/21.

5) Condition of reciprocity for z-parameter

z-parameter equation

$V_1 = z_{11} I_1 + z_{12} I_2$

$V_2 = z_{21} I_1 + z_{22} I_2$



$V_1 = V_s \quad V_2 = 0$

$I_2' = -I_2$

$$\boxed{z_{12} = z_{21}}$$

Condition of reciprocity for y-parameter:-

y-parmeter equation.

$I_1 = y_{11} V_1 + y_{12} V_2$

$I_2 = y_{21} V_1 + y_{22} V_2$

$$\boxed{y_{12} = y_{21}}$$

6b) Necessity of super Node and super mesh:-

Super node:- Super node analysis is used when a voltage source is connected between two non-reference nodes and any elements connected in parallel with it.

Super mesh:- Super mesh is used to analysis a complex electric circuit where two meshes have a common current source.

7(a) Properties Of Cut-set matrix:-

i) A cut has to intersect only one twig.
ii) The remaining can be links.
iii) The current direction of cut is same as the current direction in twig.
iv) The number of cuts is equal to number of twigs.

Properties of tie-set matrix:-

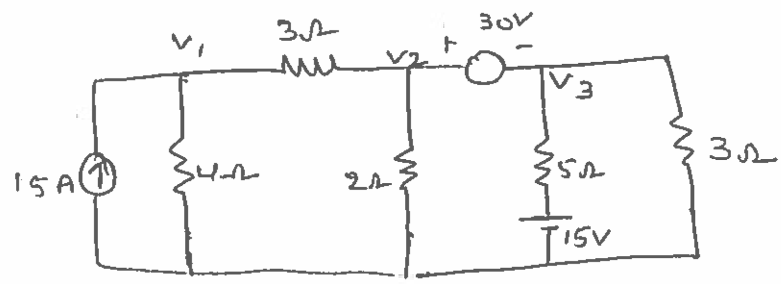
i) The number of loops will be equal to number of links.
ii) Each loop must contain only one link and remaining can be twigs.
iii) The loop direction will be same as the current direction in link.

8a) Norton's Theorem:- Any two terminal linear network with current sources, voltage sources and resistances can be replaced by an equivalent circuit consisting of a current source in parallel with a resistance.

Procedure to find Norton's resistance

i) Temporarily remove load resistance and replace with short circuit across them.
ii) Calculate short-circuit current Isc or IN
iii) Calculate the resistance of whole network as seen at two terminals, after all voltage sources replaced by short circuit and current sources replaced by open circuit.
The resistance obtained is Norton's resistance RN.

example of supernode:-



sol: Applying KCL at node-1:-

$$15 = \frac{V_1}{4} + \frac{V_1 - V_2}{3}$$

$$0.583V_1 - 0.33V_2 = 15$$

nodes (2) and (3) super node equation.

$$\frac{V_2 - V_1}{3} + \frac{V_2}{2} + \frac{V_3 - 15}{5} + \frac{V_3}{3} = 0$$

$$-\frac{V_1}{3} + V_2\left[\frac{1}{3} + \frac{1}{3}\right] + V_3\left[\frac{1}{5} + \frac{1}{3}\right] = 3$$

$$-0.33V_1 + 0.62V_2 + 0.5V_3 = 3$$

voltage between nodes (2) and (3) is given by

$$V_2 - V_3 = 30$$

current through 5Ω resistor $I_5 = \frac{V_3 - 15}{5}$

$$\cancel{V_3} = \frac{-\cancel{2.697} - 15}{5}.$$

$$V_3 = -2.697V$$

$$I_5 = \frac{-2.697 - 15}{5} = -3.539A$$

Current flows through node 3

1) a) Super position theorem:- In any linear network with several independent and dependent sources, the overall response in any part of network is equal to the sum of individual responses due to each independent source with all other independent sources reduced to zero.

Limitations:-

1) It is not applicable to network consisting of non-linear elements like transistors, diodes, etc,

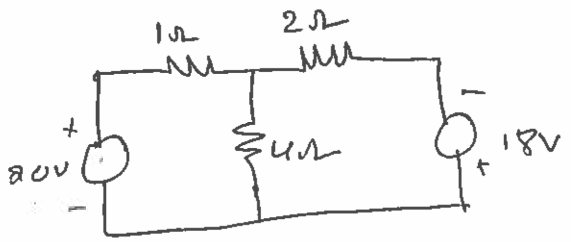
2) It is not applicable to network consisting of dependent sources

3) It is not applicable for calculation of power

4) It is not applicable for network consisting less than two independent sources.

1 b) Millman's Theorem:- Any parallel circuit which consists of one voltage source ts in series with internal resistanc in each branch can be converted into an equivalent circuit which consists of one voltage source in series with its equivalent resistance.

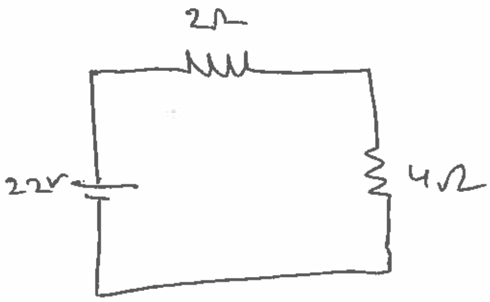
example. Current through 4Ω resistor



⇒ according to millman's theorem

$$V = \frac{\Sigma EG}{\Sigma G}$$

$$\Sigma GE = 20(1) + (-18)\left(\tfrac{1}{2}\right)$$
$$= 20 - 9$$
$$= 11$$



$$\Sigma G = G_1 \times G_2$$
$$= 1 \times \tfrac{1}{2} = 0.5V$$

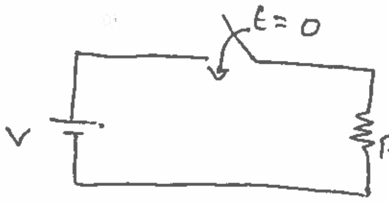
$$R_{eq} = \frac{1}{\Sigma G} = \frac{1}{0.5} = 2\Omega$$

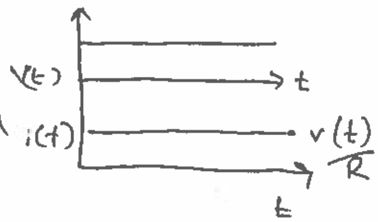
$$\text{So, } V = \frac{\Sigma EG}{\Sigma G} = \frac{11}{0.5} = 22V$$

$$\boxed{V = 22V}$$

11) a) <u>Initial Conditions:</u>

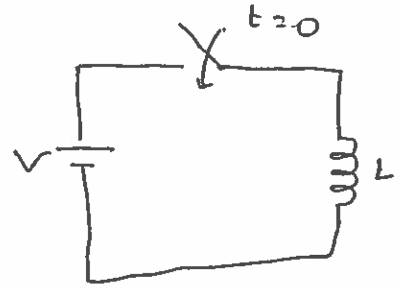
<u>Resistor:</u>

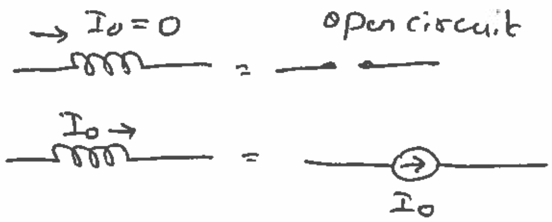




If a step voltage applied to resistor network, the current will have the same waveform as the input but will be altered in magnitude. Thus voltage and current across the resistors changes instantaneously. There is no transient period.

Inductor





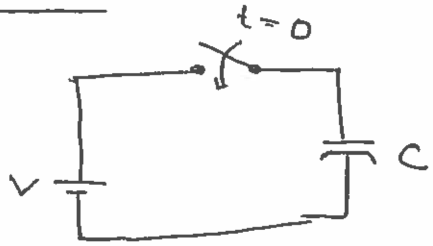
current through inductor cannot change instantaneously

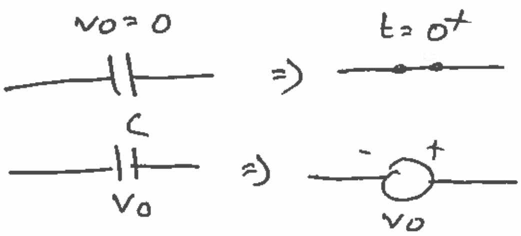
An energy source being suddenly connected to an inductor will not cause current to flow initially. and inductor acts as open circuit.

Volt-ampere relation

$$V = L \frac{di}{dt}$$

Capacitors:-



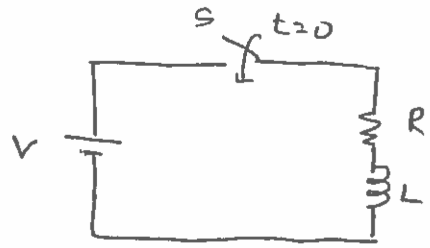


$$i = \int c \frac{dv}{dt}$$

→ voltage across capacitor cannot change instantaneously.

→ If an uncharged capacitor is switched on to DC source the current will flow instantaneously. and capacitor acts as short-circuited.

R-L circuit:



$$Ri + L\frac{di}{dt} = V$$

$$\left(\frac{R}{L} + \frac{d}{dt}\right) i = V/L$$

$$(D + R/L) i = V/L \qquad \boxed{D = d/dt}$$

Comparing with non-homogeneous differential equation

$$\left(\frac{d}{dt} + P\right) x = K \Rightarrow P = R/L, \; x = i, \; K = V/L$$

$$x = e^{-Pt} \int K e^{Pt} \, dt + C e^{-Pt}$$

for current $i = C e^{-(R/L)t} + e^{-(R/L)t} \int \frac{V}{L} e^{(R/L)t} dt$

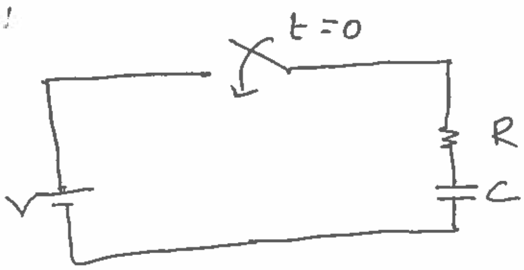
$$i = C e^{-(R/L)t} + V/R$$

$$C + V/R = 0 \Rightarrow C = -V/R$$

$$i = V/R - V/R \, e^{-(R/L)t}$$

$$\boxed{i = V/R \left(1 - e^{-(R/L)t}\right)}$$

R-C circuit:



by applying KVL

$$\frac{1}{C} \int i \, dt + Ri = V$$

$$R \cdot \frac{di}{dt} + i/C = 0$$

$$\frac{di}{dt} + 1/RC \; i = 0$$

$$(D + 1/RC) i = 0$$

at S.C, $t=0^+$ $i = V/R$

$c = V/R$

$i = \frac{V}{R} e^{-t/RC}$

when s is closed, then time constant function $\frac{V}{R} e^{-t/RC}$

$\tau = RC$ sec

$V_R = iR = R \times \frac{V}{R} e^{-t/RC}$

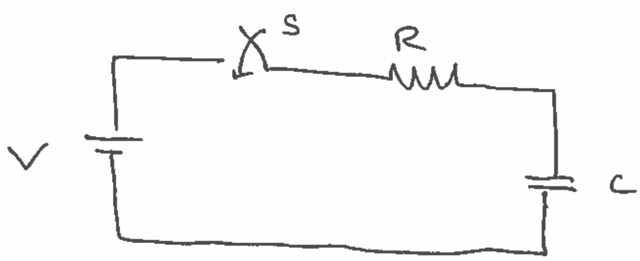
$V_R = Ve^{-t/RC}$

$V_c = \frac{1}{C} \int i dt = \frac{1}{C} \int \frac{V}{R} e^{-t/RC}$

$= -\left(\frac{V}{RC} \times RC e^{-t/RC}\right) + C$

$\boxed{V_c = -Ve^{-t/RC} + C}$

11(b) Voltage across Resistor and capacitor.



$V = iR + \frac{1}{C} \int i dt$

$0 = R \frac{di}{dt} + i/C$

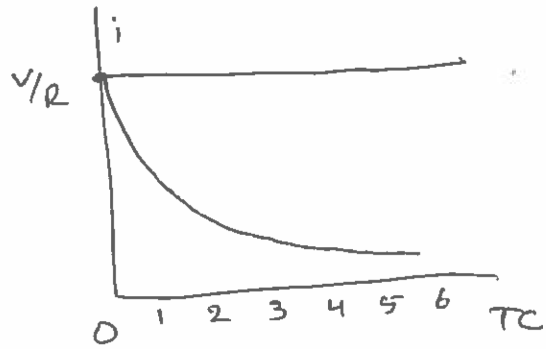
$\frac{di}{dt} + \frac{1}{RC} i = 0$

$i = ce^{-t/RC}.$

$t=0$, current $i = V/R$

$V/R = C$

$i = \frac{V}{R} e^{-t/RC}$



$V_R = iR = R \times \frac{V}{R} e^{-(1/RC)t}$

$V_R = Ve^{-t/RC}$

Voltage across capacitor:-

$V_c = \frac{1}{c} \int i\,dt$

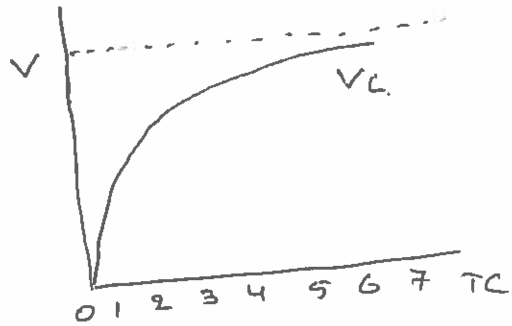
$= \frac{1}{c} \int \frac{V}{R} e^{-t/RC} dt$

$= \left(\frac{V}{RC} \times RC\, e^{-t/RC}\right) + c = -Ve^{-t/RC} + c$

At $t=0$, $V_c = 0$,

$c = V$

$V_c = V\left(1 - e^{-t/RC}\right)$



12) equation for frequency:

max. voltage drop across inductance occurs at

$$f = f_L$$

$$V_L = I X_L$$
$$= I(\omega L)$$

Current $I = V/Z$, $Z = \sqrt{R^2 + \left(\omega L - \frac{1}{\omega C}\right)^2}$

$$V_L = \frac{V\omega L}{R^2 + \left(\omega L - \frac{1}{\omega C}\right)^2}$$

V_L is maximum when $\frac{dV_L}{d\omega} = 0$

$$\frac{d}{d\omega}\left[\frac{V\omega L}{R^2 + \left(\omega L - \frac{1}{\omega C}\right)^2}\right] = 0$$

$$\omega_L = \frac{1}{\sqrt{LC}}\left[\frac{-2}{2 - \frac{R^2 C}{L}}\right]$$

$\omega_L = 2\pi f_0 L$ $\quad f_L = \frac{1}{2\pi\sqrt{LC}}\left(\sqrt{\frac{1}{1 - \frac{R^2 C}{2L}}}\right.$

$$V_C = I X_C$$
$$= I\left(\frac{1}{\omega C}\right]$$

$$I = V/Z = \frac{V}{\sqrt{R^2 + \left(\omega C - \frac{1}{\omega C}\right)^2}}$$

$$V_C = \frac{V}{\sqrt{R^2 + \left(\omega L - \frac{1}{\omega C}\right)^2}} \times \frac{1}{\omega C}$$

V_C is max. $\frac{dV_C}{d\omega} = 0$, differentiating V_C

$$\omega_C = \sqrt{\frac{1}{LC} - \frac{R^2}{2L}} = \sqrt{\frac{1}{LC} - \frac{R^2}{2L}}$$

$$f_C = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{2L}}$$

13) Expression for Bandwidth and Quality factor:

Quality factor:- Quality factor of RLC is defined as ratio of resonant frequency to bandwidth.

$$Q = \frac{f_0}{f_2 - f_1}$$

we know, $Q = C/V \Rightarrow Q = \frac{\omega L}{R}$

$$Q = \frac{I X_L \text{ or } I X_C}{V}$$

$$Q = \frac{\omega_0 L}{R} \text{ (or) } \frac{1}{\omega_0 C R}$$

$$Q = 2\pi\left(\frac{\text{Energy stored}}{\text{Energy dissipated}}\right)$$

$$\text{max. energy} = \frac{1}{2} L I_0^2 \text{ (or) } \frac{1}{2} C V^2$$

Energy dissipated/cycle = Avg power/(cycle × Period)

$$= \left(\frac{I_0}{\sqrt{2}}\right)^2 \times R \times \frac{1}{f_0}$$

$$= \frac{I_0^2 R}{2 f_0}$$

$$Q = 2\pi\left(\frac{\frac{1}{2} L I_0^2}{I_0^2 R / 2 f_0}\right)$$

$$Q = \frac{\omega_0 L}{R}$$

9) a) Super position theorem:- In any linear network with several independent and dependent sources, the overall response in any part of network is equal to the sum of individual responses due to each independent source with all other independent sources reduced to zero.

Limitations:-

1) It is not applicable to network consisting of non-linear elements like transistors, diodes, etc.,
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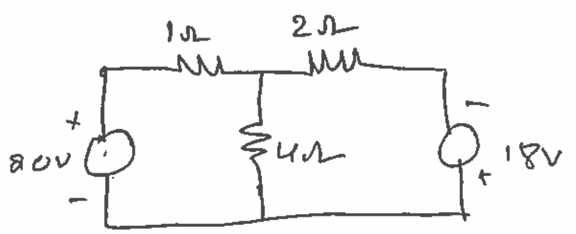
9 b) Millman's Theorem:- Any parallel circuit which consists of one voltage source is in series with internal resistances in each branch can be converted into an equivalent circuit which consists of one voltage source in series with its equivalent resistance.

example. Current through 4Ω resistor



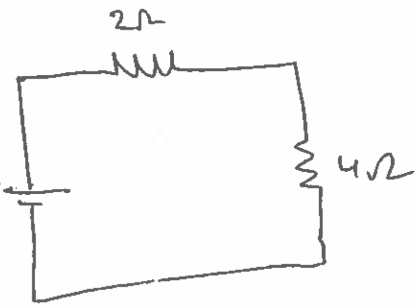
Sol: According to millman's theorem

$$V = \frac{\Sigma EG}{\Sigma G}$$

$$\Sigma GE = 20(1) + (-18)(\tfrac{1}{2})$$

$$= 20 - 9$$

$$= 11$$



$$\Sigma G = G_1 + G_2$$

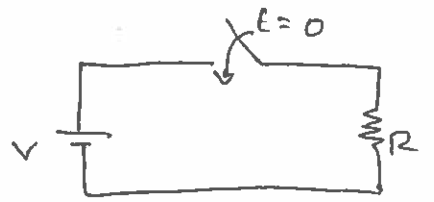
$$= 1 \times \tfrac{1}{2} = 0.5V$$

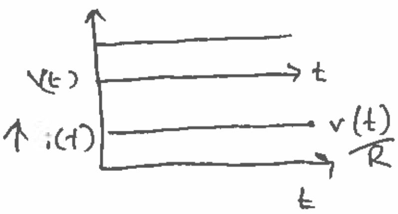
$$R_{eq} = \frac{1}{\Sigma G} = \frac{1}{0.5} = 2\Omega$$

So, $V = \frac{\Sigma EG}{\Sigma G} = \frac{11}{0.5} = 22V$

11)a) Initial Conditions:

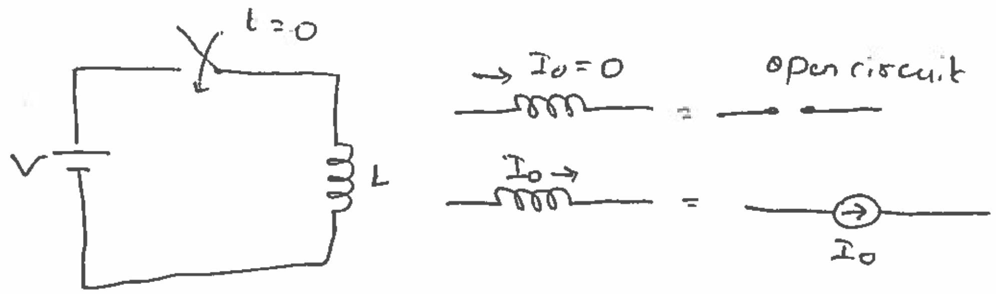
Resistor:





If a step voltage applied to resistor network, the current will have the same waveform as the input but will be altered in magnitude. Thus voltage and current across the resistors changes instantaneously. There is no transient period.

Inductor



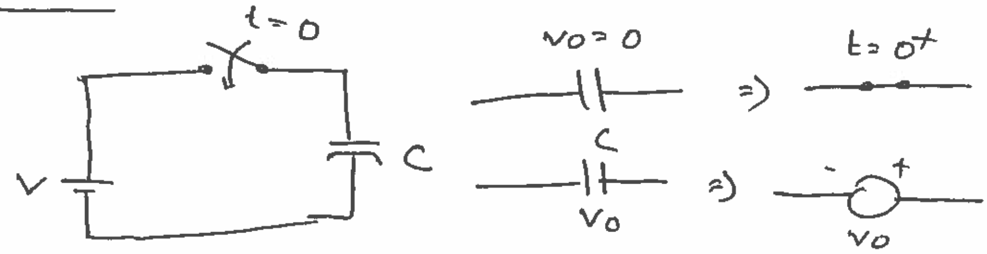
current through inductor cannot change instantaneously

An energy source being suddenly connected to an inductor will not cause current to flow initially. and inductor acts as open circuit.

Volt-ampere relation

$$V = L \frac{di}{dt}$$

Capacitors:-

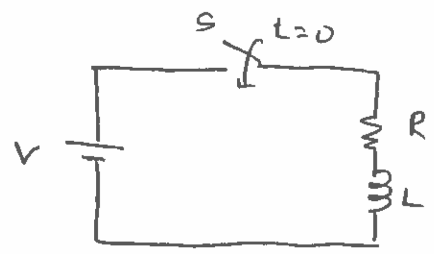


$$i = \int c \frac{dv}{dt}$$

→ voltage across capacitor cannot change instantaneously.

→ If an uncharged capacitor is switched on to DC source the current will flow instantaneously. and capacitor acts as short-circuited.

R-L circuit:



$$Ri + L\frac{di}{dt} = V$$

$$\left(\frac{R}{L} + \frac{d}{dt}\right) i = V/L$$

$$\left(D + R/L\right) i = V/L \qquad \boxed{D = d/dt}$$

Comparing with non-homogeneous differential equation

$$\left(\frac{d}{dt} + P\right)x = K \Rightarrow P = R/L, \; x = i, \; K = V/L$$

$$x = e^{-Pt}\int K e^{Pt} \, dt + C e^{-Pt}$$

for current $i = C e^{-(R/L)t} + e^{-(R/L)t}\int \frac{V}{L} e^{(R/L)t} dt$

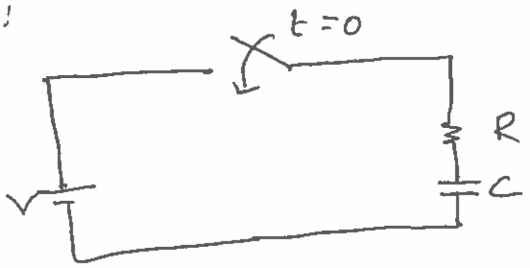
$$i = C e^{-(R/L)t} + V/R$$

$$C + V/R = 0 \Rightarrow C = -V/R$$

$$i = V/R - V/R \, e^{-(R/L)t}$$

$$i = V/R\left(1 - e^{-(R/L)t}\right)$$

R-C circuit:



by applying KVL

$$\frac{1}{C}\int i \, dt + Ri = V$$

$$R\frac{di}{dt} + i/C = 0$$

$$\frac{di}{dt} + \frac{1}{RC} i = 0$$

$$\left(D + 1/RC\right) i = 0$$

at s.c, $t=0^+$ $\quad i = V/R$

$$c = V/R$$

$$i = \frac{V}{R} e^{-t/RC}$$

when s is closed, then time constant function $\frac{V}{R} e^{-t/RC}$

$$\tau = RC \text{ sec}$$

$$V_R = iR = R \times \frac{V}{R} e^{-t/RC}$$

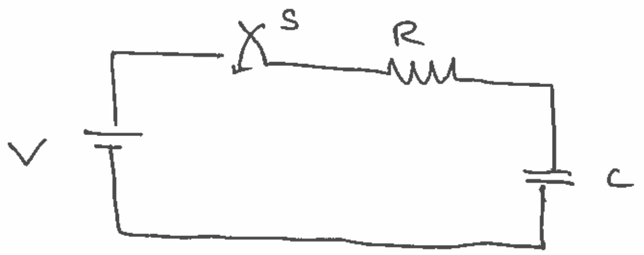
$$V_R = Ve^{-t/RC}$$

$$V_c = \frac{1}{C} \int i\,dt = \frac{1}{C} \int \frac{V}{R} e^{-t/RC}$$

$$= -\left(\frac{V}{RC} \times RC\, e^{-t/RC}\right) + c$$

$$\boxed{V_c = -Ve^{-t/RC} + C}$$

11(b) Voltage across Resistor and capacitor.



$$V = iR + \frac{1}{C} \int i\,dt$$

$$0 = R\frac{di}{dt} + \frac{i}{C}$$

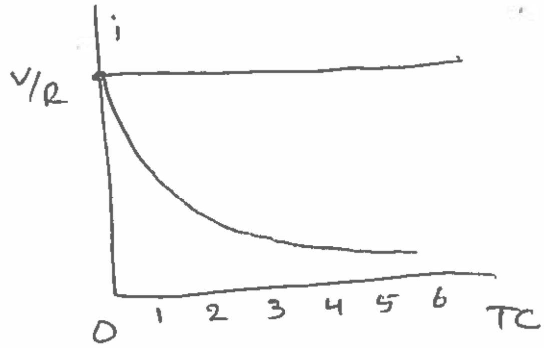
$$\frac{di}{dt} + \frac{1}{RC} i = 0$$

$$i = ce^{-t/RC}.$$

$t=0$, current $i = V/R$

$$V/R = C$$

$$i = \frac{V}{R} e^{-t/RC}$$



$$V_R = iR = R \times \frac{V}{R} e^{-(1/RC)t}$$

$$V_R = Ve^{-t/RC}$$

<u>Voltage across capacitor:-</u>

$$V_c = \frac{1}{C}\int i\,dt$$

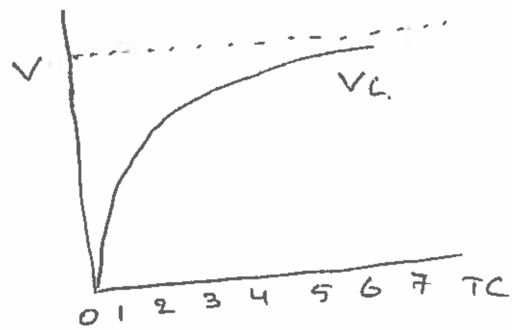
$$= \frac{1}{C}\int \frac{V}{R} e^{-t/RC}\,dt$$

$$= \left(\frac{V}{RC} \times RC\, e^{-t/RC}\right) + C = -Ve^{-t/RC} + C$$

At $t=0$, $V_c = 0$,

$$C = V$$

$$V_c = V\left(1 - e^{-t/RC}\right)$$



12) equation for frequency:

max. voltage drop across inductance occurs at

$$f = f_L$$

$$V_L = I X_L$$
$$= I(\omega L)$$

Current $I = V/Z$, $Z = \sqrt{R^2 + (\omega L - \frac{1}{\omega C})^2}$

$$V_L = \frac{V\omega L}{R^2 + (\omega L - \frac{1}{\omega C})^2}$$

V_L is maximum when $\frac{dV_L}{d\omega} = 0$

$$\frac{d}{d\omega}\left[\frac{V\omega L}{R^2 + (\omega L - \frac{1}{\omega C})^2}\right] = 0$$

$$\omega L = \frac{1}{\sqrt{LC}}\left[\frac{-2}{2 - \frac{R^2 C}{L}}\right]$$

$\omega L = 2\pi f_0 L$ $\quad f_L = \frac{1}{2\pi\sqrt{LC}}\left(\sqrt{\frac{1}{1 - \frac{R^2 C}{2L}}}\right.$

$$V_C = I X_C$$
$$= I\left(\frac{1}{\omega C}\right)$$

$$I = V/Z = \frac{V}{\sqrt{R^2 + (\omega L - \frac{1}{\omega C})^2}}$$

$$V_C = \frac{V}{\sqrt{R^2 + (\omega L - \frac{1}{\omega C})^2}} \times \frac{1}{\omega C}$$

V_C is max. $\frac{dV_C}{d\omega} = 0$, differentiating V_C

$$\omega C = \sqrt{\frac{1}{LC} - \frac{R^2}{2L}} = \sqrt{\frac{1}{LC} - \frac{R^2}{2L}}$$

$$f_C = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{\ldots}}$$

13) Expression for Bandwidth and Quality factor:

Quality factor:

Quality factor of RLC is defined as ratio of resonant frequency to bandwidth.

$$Q = \frac{f_0}{f_2 - f_1}$$

we know, $Q = \frac{C}{V} \Rightarrow Q = \frac{\omega L}{R}$

$$Q = \frac{IX_L \text{ or } IX_C}{V}$$

$$Q = \frac{\omega_0 L}{R} \text{ (or) } \frac{1}{\omega_0 CR}$$

$$Q = 2\pi\left(\frac{\text{Energy stored}}{\text{Energy dissipate}}\right)$$

max. energy $= \frac{1}{2} L I_0^2$ (or) $\frac{1}{2} C V^2$

Energy dissipated/cycle = Avg power/(cycle x period)

$$= \left(\frac{I_0}{\sqrt{2}}\right)^2 \times R \times \frac{1}{f_0}$$

$$= \frac{I_0^2 R}{2 f_0}$$

$$Q = 2\pi\left(\frac{\frac{1}{2} L I_0^2}{I_0^2 R / 2 f_0}\right)$$

$$Q = \frac{\omega_0 L}{R}$$

Expression for Bandwidth:

Consider frequency response of series RLC circuit. The capacitive reactance is greater than inductive reactance.

$$I_0 = V/R$$

$$\frac{I_0}{\sqrt{2}} = \frac{V}{\sqrt{2}R}$$

$$Z = \sqrt{R^2+\left(\omega_x L - \frac{1}{\omega_x C}\right)^2}$$

$$\frac{I_0}{\sqrt{2}} = \frac{V}{\sqrt{R^2+\left(\omega_x L - \frac{1}{\omega_x C}\right)^2}}$$

$$\sqrt{2}R = R^2+\left(\omega_x L - \frac{1}{\omega_x C}\right)^2$$

$$R = \omega_x L - \frac{1}{\omega_x C}$$

$$\omega_x \left[LC\omega_x - CR\right] = \pm 1$$

$$\Rightarrow \omega_x = \pm \frac{R}{2L} \pm \sqrt{\frac{R^2}{4L^2} + \frac{1}{LC}}$$

neglect $\rightarrow \frac{R^2}{4L^2}$

$$\omega_x = \pm \frac{R}{2L} \pm \sqrt{1/LC}$$

$$\omega_2 = \omega_0 + R/2L$$

$$\omega_1 = \omega_0 - R/2L$$

$$\text{Bandwidth} = \omega_2 - \omega_1 = \frac{R}{2L} - \left[-R/2L\right] = R/L$$

$\omega_2 = 2\pi f_2$ and $\omega_1 = 2\pi f_1$

$$2\pi(f_2 - f_1) = R/L$$

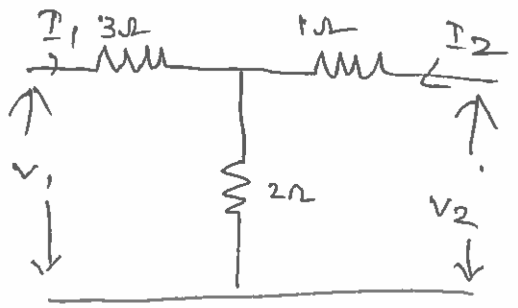
$$f_2 - f_1 = R/2\pi L$$

$\Delta f = R/2\pi f$. Bandwidth $= \frac{R}{2\pi L}$

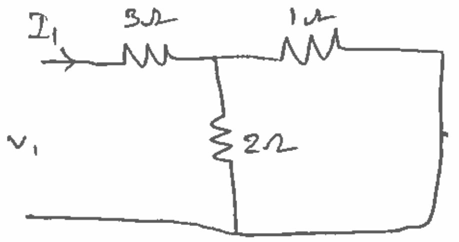
Quality factor $Q = \frac{f_0}{f_2 - f_1} = \frac{f_0}{R/2\pi L} = f_0\left(\frac{2\pi L}{R}\right)$

$$Q = \frac{2\pi f_0 L}{R} \Rightarrow \boxed{Q = \frac{\omega_0 L}{R}}$$

4.)



step - 1 By short circuiting the output port. $V_2 = 0$



from above 1Ω and 2Ω are parallel.

$$R_{12} = \frac{1 \times 2}{1+2} = 2/3 = 0.66\Omega$$

0.66Ω and 3Ω are in series.

$$R_{eq} = 0.66 + 3 = 3.66\Omega$$

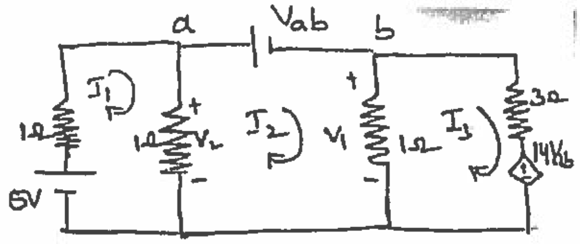
$$V_1 = I_1(3.66)$$

$$\boxed{\frac{I_1}{V_1} = \frac{1}{3.66} = 0.273\,\mho = Y_{11}}$$

$$I_2 = -I_1 \times \frac{2}{3}$$

$$I_2 = -0.66\, I_1$$

6(a) Sol: Given, $V_2 = 2.5\,V$



Consider Mesh analysis, we get

@ loop - 1:

$-5 + I_1(1) + (I_1 - I_2)(1) = 0 \Rightarrow I_1 + (I_1 - I_2) = 5 \longrightarrow ①$

@ loop - 2:

$(I_2 - I_1)(1) + V_{ab} + (I_2 - I_3)(1) = 0 \Rightarrow 2I_2 - I_1 - I_3 = -V_{ab}$

$I_1 - 2I_2 + I_3 = V_{ab} \longrightarrow ②$

@ loop - 3:

$(I_3 - I_2)(1) + 3I_3 + 14\,V_{ab} = 0 \Rightarrow I_2 - 4I_3 = 14\,V_{ab}$

$\Rightarrow 0.072 I_2 - 0.28 I_3 = V_{ab} \longrightarrow ③$

$\Rightarrow$ We know, $V_2 = 2.5V$

$\Rightarrow I_1 - I_2 = \frac{V_2}{R} = \frac{2.5}{1} = 2.5\,A.$

$\Rightarrow$ Eq - ①, $I_1 + (2.5) = 5$

$\Rightarrow \boxed{I_1 = 2.5\,A}$

$I_1 - I_2 = 0$

$I_2 = I_1 - 2.5 = 2.5 - 2.5$

$= 0$

Eq ② ~ Eq ③

$I_1 - 2I_2 + I_3 = \cancel{V_{ab}}$

$- \quad 0.072 I_2 \underset{+}{-} 0.28 I_3 = \underset{-}{\cancel{V_{ab}}}$

$I_1 - 2.072 I_2 + 1.28 I_3 = 0. \longrightarrow (a)$

Substitute I_1, I_2 values in eq-(a), we get

$2.5 - 2.072(0) + 1.28 I_3 = 0$

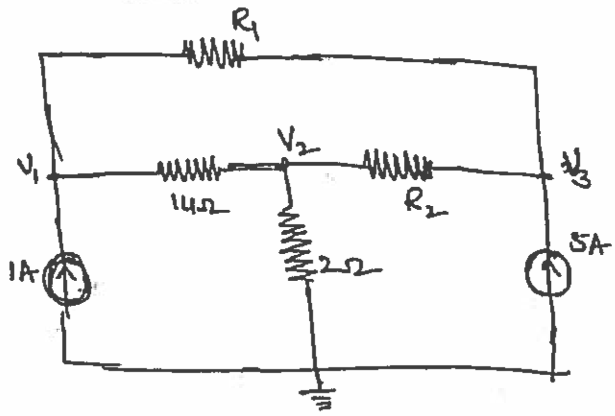
$1.28 I_3 = -2.5$

$I_3 = \frac{-2.5}{1.28} = -1.95\,A. \Rightarrow \boxed{I_3 = -1.95\,A}$

$i = -I_3 \Rightarrow -(-1.95\,A)$

$\boxed{I_3 = 1.95A}$

7(b) Considering nodal analysis,



@ node:1

$$1 = \frac{V_1 - V_3}{R_1} + \frac{V_1 - V_2}{14} \longrightarrow ①$$

@ node - 2

$$\frac{V_2 - V_1}{14} + \frac{V_2 - V_3}{R_2} = \frac{V_2 - 0}{2}$$

$$\frac{V_2 - V_1}{14} + \frac{V_2 - V_3}{R_2} = \frac{V_2}{2} \longrightarrow ②$$

@ node - 3 :

$$5 = \frac{V_3 - V_1}{R_1} + \frac{V_3 - V_2}{R_2} \longrightarrow ③,$$

Now, we know that, $V_1 = 100\,V$ & $V_3 = 40\,V$, then eq 1,2,3

Eq ①

$$1 = \frac{100 - 40}{R_1} + \frac{100 - V_2}{14} = \frac{60}{R_1} + \frac{50}{7} - \frac{V_2}{14}$$

$$\Rightarrow \frac{60}{R_1} + \frac{50}{7} - \frac{V_2}{14} = 1 \longrightarrow (a)$$

Eq ②

$$\frac{V_2 - 100}{14} + \frac{V_2 - 40}{R_2} = \frac{V_2}{2}$$

$$\Rightarrow \frac{V_2}{14} - \frac{50}{7} + \frac{V_2}{R_2} - \frac{40}{R_2} = \frac{V_2}{2} \longrightarrow (b)$$

Eq ③

$$\frac{40 - 100}{R_1} + \frac{40 - V_2}{R_2} = 5$$

$$\frac{-60}{R_1} + \frac{40}{R_2} - \frac{V_2}{R_2} = 5 \longrightarrow (c)$$

↳ From eq -(c), we get

$$\frac{V_2}{R_2} = -\frac{60}{R_1} + \frac{40}{R_2} - 5 \longrightarrow (d)$$

→ Substitute eq-(d) in eq-(b), we get,

$$\frac{V_2}{14} - \frac{50}{7} + \left[-\frac{60}{R_1} + \cancel{\frac{40}{R_1}} - 5\right] - \cancel{\frac{40}{R_2}} = \frac{V_2}{2}$$

$$\frac{V_2}{14} - \frac{V_2}{2} - \frac{50}{7} - \frac{60}{R_1} - 5 = 0$$

$$\frac{-3V_2}{7} - \frac{50}{7} - \frac{60}{R_1} - 5 = 0$$

$$\Rightarrow \frac{3V_2}{7} + \frac{50}{7} + \frac{60}{R_1} + 5 = 0 \longrightarrow (e)$$

↳ From eq-(a), we get

$$\frac{60}{R_1} + \frac{50}{7} = 1 + \frac{V_2}{14}$$

→ From eq-(e), we get

$$\frac{3V_2}{7} + \left[1 + \frac{V_2}{14}\right] + 5 = 0$$

$$\Rightarrow \left[\frac{3}{7} + \frac{1}{14}\right] V_2 + 6 = 0$$

$$V_2 = -6 \times \frac{1}{\left[\frac{42+7}{14 \times 7}\right]}$$

$$= -6 \times \left(\frac{\cancel{14} \times \cancel{7}}{\cancel{49}}\right)$$

$$\boxed{V_2 = -12\,V}$$

Substitute V_2 in eq-(e), we get

$$\frac{3(-12)}{7} + \frac{50}{7} + \frac{60}{R_1} + 5 = 0$$

$$\frac{-36+50}{7} + \frac{60}{R_1} + 5 = 0$$

$$\boxed{R_1 = -8.57\,\Omega}$$

→ From eq -(c), we get

$$\frac{-60}{R_1} + \frac{40 - V_2}{R_2} = 5$$

$$\Rightarrow \frac{\cancel{-}60}{\cancel{-}8.57} + \frac{40-(-12)}{R_2} = 5$$

$$7 + \frac{52}{R_2} = 5$$

$$\Rightarrow \frac{52}{R_2} = -2 \qquad \Rightarrow R_2 = \frac{-52}{2} = -26\,\Omega$$

$$\boxed{R_1 = -8.57\,\Omega}$$

$$\boxed{R_2 = -26\,\Omega}$$

15) a) Relation between Z and ABCD parameters:-

Defining equation of ABCD

$V_1 = AV_2 - BI_2$

$I_1 = CV_2 - DI_2$

Defining equation of Z

$V_1 = Z_{11} I_1 + Z_{12} I_2$

$V_2 = Z_{21} I_1 + Z_{22} I_2$

when $I_2 = 0$, we get

$A = \frac{V_1}{V_2}$

similarly $I_2 = 0$

$V_1 = Z_{11} I_1$ and $V_2 = Z_{21} I_1$

$$A = \frac{Z_{11} I_1}{Z_{21} I_1} = \frac{Z_{11}}{Z_{21}}$$

$$\boxed{A = \frac{Z_{11}}{Z_{21}}}$$

when $I_2 = 0$, $C = \frac{I_1}{V_2}$ $\Rightarrow V_2 = Z_{21} I_1$

$\Rightarrow \frac{I_1}{V_2} = \frac{1}{Z_{21}}$

$$\boxed{C = \frac{1}{Z_{21}}}$$

when $V_2 = 0$ $\Rightarrow B = \frac{-V_1}{I_2}$ $\Rightarrow$ $\boxed{B = \frac{\Delta Z}{Z_{21}}}$

$$D = \frac{-I_1}{I_2} = \frac{-Z_{22}/\cancel{\Delta Z}}{-Z_{21}/\cancel{\Delta Z}} = \frac{Z_{22}}{Z_{21}}$$

$$\boxed{D = \frac{Z_{22}}{Z_{21}}}$$

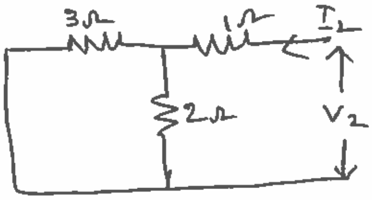
$$I_2 = -0.66\left(\frac{V_1}{3.66}\right)$$

$$\frac{I_2}{V_1} = \frac{-0.66}{3.66}$$

$$\boxed{Y_{21} = \frac{I_2}{V_1} = -0.18\,\mho}$$

Step-2:- Short circuit the input port.

$V_1 = 0$



3Ω and 2Ω are parallel

$$\Rightarrow \frac{3\times 2}{3+2} = \frac{6}{5} = 1.2\,\Omega$$

1.2Ω and 1Ω are in series.

$$\Rightarrow R_{eq} = 1.2+1 = 2.2\,\Omega$$

$$V_2 = I_2(2.2)$$

$$\boxed{\frac{I_2}{V_2} = \frac{1}{2.2} = 0.45\,\mho = Y_{22}}$$

$$\Rightarrow I_1 = -I_2 \times \frac{2}{5}$$

$$I_1 = -\left(\frac{2}{5}\right)\left(\frac{V_2}{2.2}\right)$$

$$\frac{I_1}{V_2} = -\left(\frac{2}{5}\right)\left(\frac{1}{2.2}\right)$$

$$\boxed{Y_{12} = -0.18}$$

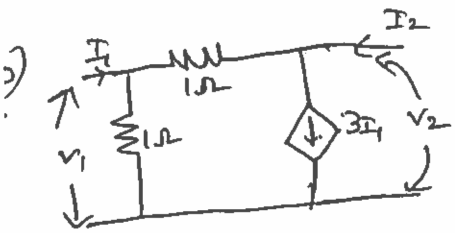
$Y_{11} = 0.273\,\mho$

$Y_{12} = -0.18\,\mho$

$Y_{21} = -0.18\,\mho$

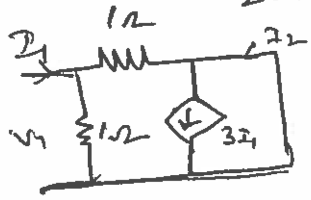
$Y_{22} = 0.45\,\mho$

)



Step-1 : short circuit o/p port.

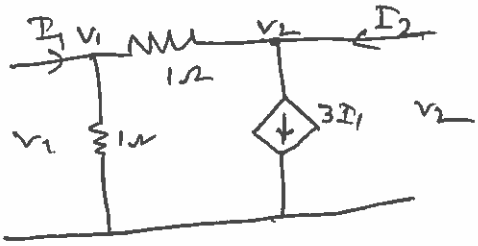
$V_2 = 0$



$Y_{11} = I_1/V_1$

$Y_{21} = I_2/V_1$

By applying nodal analysis



$\Rightarrow \frac{V_1}{1} + \frac{V_1 - V_2}{1} = I_1$

$\Rightarrow \boxed{2V_1 - V_2 = I_1}$

$\frac{V_2 - V_1}{1} + 3I_1 = I_2$

$V_2 - V_1 + 3(2V_1 - V_2) = I_2$

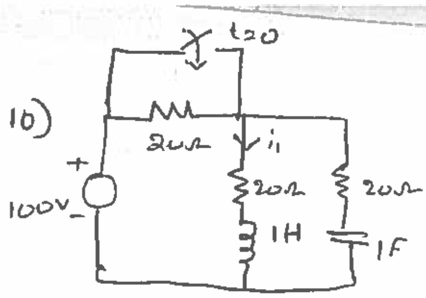
$V_2 - V_1 + 6V_1 - 3V_2 = I_2$

$\boxed{5V_1 - 2V_2 = I_2}$

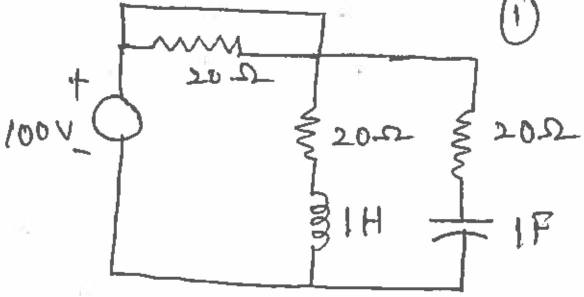
$2V_1 - V_2 = I_1 \Rightarrow \boxed{\frac{I_1}{V_1} = 2 = Y_{11}}$ $\boxed{\frac{I_1}{V_2} = -1 = Y_{12}}$

$5V_1 - 2V_2 = I_2 \Rightarrow \boxed{\frac{I_2}{V_1} = 5 = Y_{21}}$ $\boxed{\frac{I_2}{V_2} = -2 = Y_{22}}$

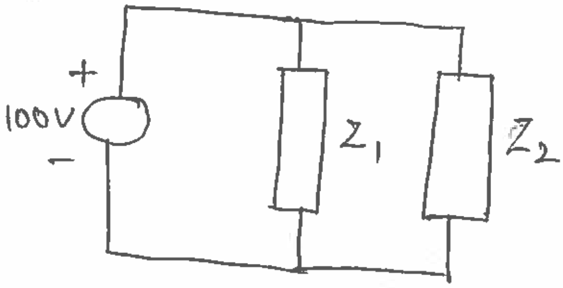
10)



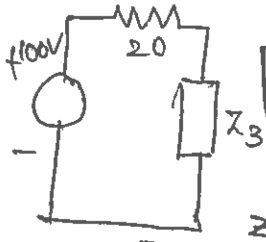
$t = 0^+$ ①



②



$t = 0^-$



$Z_4 = 20 + Z_3$

$\boxed{Z_4 = 20 + 19.82 + j1.}$

$= 39.82 + j1.3$

$Z_3 = 39.84 \angle 3.9$

$i(t^{0^-}) = \frac{100}{39.84} \angle -3.9 = 2.5 A$

$Z_1 = 20 + j2 \cdot \pi \cdot 50 \cdot 1$

$= 20 + j314$

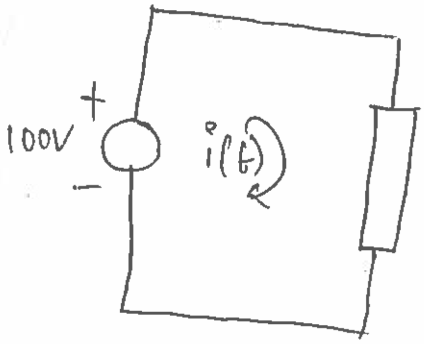
$Z_1 = \sqrt{20^2 + (314)^2} \angle \tan^{-1}\left(\frac{314}{20}\right)$

$\boxed{Z_1 = 314.64 \angle 86.35}$

$Z_2 = 20 + j\frac{1}{2\pi f C}$

$= 20 + j\,0.0032$

$\boxed{Z_2 = 20 \angle 0.344}$



$Z_1 \| Z_2 = Z_3.$

$Z_3 = \frac{Z_1 Z_2}{Z_1 + Z_2}$

$\Rightarrow Z_3 = \frac{314.64 \times 20 \angle 86.64}{40 + j314.0032}$

$\Rightarrow Z_3 = \frac{6292.8 \angle 86.64}{\cancel{100178.00}\ 316.54 \angle 82.74}$

$Z_3 = 19.88 \angle 3.9$

$\boxed{Z_3 = 19.82 + j1.35}$

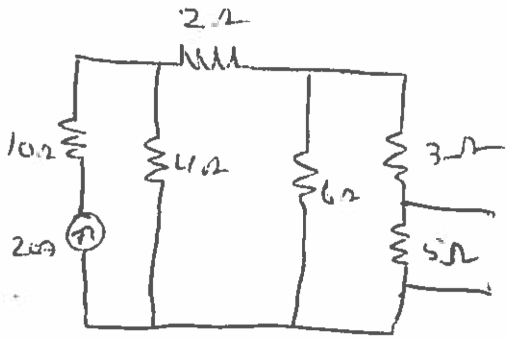
$i(t) = \frac{100}{Z_3} = \frac{100}{19.88} \angle -3.9$

$= 5.03 \angle -3.9 A$

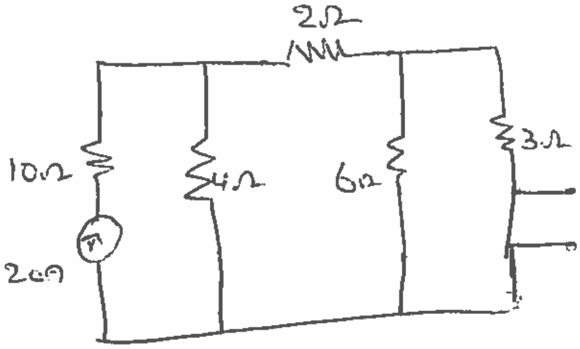
$i(t) = 5.03 A$

$\frac{di}{dt}$ @ $t = 0^+ \Rightarrow$ 2.5A

8(b)



Step-1 Remove load resistance and replace S.C terminal.



from diagram

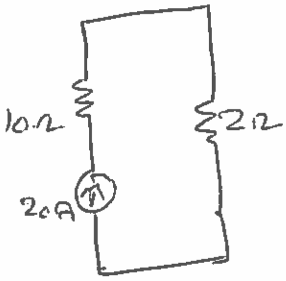
$6\Omega // 3\Omega \Rightarrow \frac{6\times 3}{6+3}$

$= \frac{18}{9} = 2\Omega$

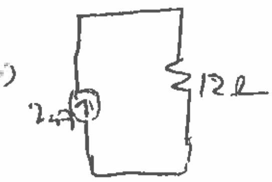
2Ω is series with 2Ω

$2+2 = 4\Omega$

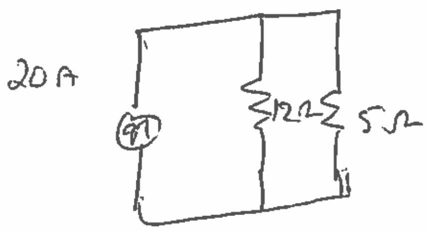
$4\Omega // 4\Omega \Rightarrow \frac{4\times 4}{4+4} = \frac{16}{8} = 2\Omega$



$\Rightarrow$



Now add load resistance.



$\Rightarrow I_{5\Omega} = 20 \times \frac{12}{12+5}$

$= 20 \times \frac{12}{17}$

~~$I_{5\Omega} = 14.$~~

$\boxed{I_{5\Omega} = 14.11A}$

12 (a) $\Sigma V = 0.$ (5m)

$R - W - F = 0.$ Lift moving Downwards

$$\boxed{R = W/g\,(a) + W}$$

$\Sigma V = 0.$

$R - W + F = 0.$ Lift moving upwards.

$$\boxed{R = -W/g\,(a) + W}$$

(b) $v^2 - u^2 = 2as.$ (7m)

$a =$

$v = u + at.$

$\Sigma H = 0.$ $\boxed{N = W}$

$$\boxed{\mu = a/g}$$

13 (a) Time Taken (i) $v = u + at.$ (6m)

$v^2 - u^2 = 2as.$

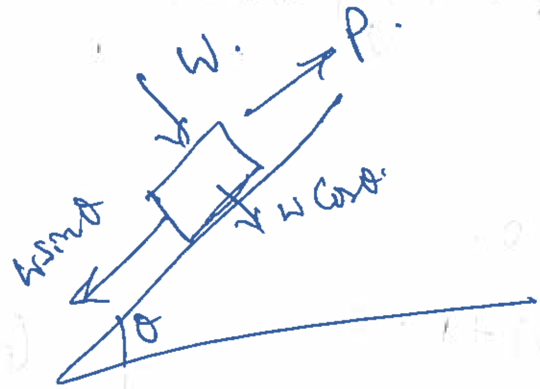
$\boxed{t = u/g}$ $\boxed{h = u^2/2g}$

13 (b) $v^2 - u^2 = 2as$.

(6m)

$V = u + at$

Q. 14



(12m)

$F = \mu N$

$\Sigma H = 0$.

$P - W \sin\theta - F = 0$.

$\boxed{P =}$

Energy = KE

$P \times S = \frac{1}{2} m v^2$.

$\boxed{S =}$

Q. 15. Work Energy Equation

(12m)

$R = \Sigma F_x$.

$R = W/g\,(a)$

$$\int_0^S R\,ds = \int_u^v W/g\,(v)\,dv.$$

$$\frac{W}{2g}(v^2 - u^2)$$

$$\boxed{R\,S = W/2g\,(v^2 - u^2)}$$

NSRIT

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	**Program**	CSE,CSM & CSD		**Academic Year**	2020 - 2021
Course Code	**20EC203**	**Test Duration**	3 Hrs.	Max. Marks 70	**Semester**	II
Course	**Digital logic Design**					

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Convert $(0.625)_{10}$ decimal number to binary number $(?)_2$ using successive multiplication method	20EC203.1	L3
2	State the absorption law of Boolean algebra	20EC203.2	L1
3	Give the general procedure for converting a Boolean expression in to multilevel NAND diagram?	20EC203.3	L1
4	What are the three types of fundamental PLDS?	20EC203.4	L1
5	What is race around condition?	20EC203.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Convert the given binary 1010110 to gray code	6M	20EC203.1	L2
6 (b)	Convert the following: (i) 57_{10} to binary (ii) 743_8 to binary (iii) $A9B_{16}$ to binary	6M	20EC203.1	L2
	OR			
7 (a)	Convert $(567.1875)_{10}$ into hexadecimal	4M	20EC203.1	L2
7 (b)	Design an 8421 to gray code converter	8M	20EC203.1	L2
8 (a)	Develop the given function $Y(M, N, O, P) = \Sigma_m(0, 2, 4, 6, 9, 13)$ Draw the K-map and Implement the simplified expression using basic gates	6M	20EC203.2	L2
8 (b)	Analyze the basic rules (laws) that are used in Boolean expressions with few examples	6M	20EC203.2	L2
	OR			
9 (a)	Simplify the following Boolean expression in i) SOP using Karnaugh map AC'+B'D+A'CD+ABCD	6M	20EC203.2	L2
9 (b)	Simplify the following Boolean expression in ii) POS using Karnaugh map AC'+B'D+A'CD+ABCD	6M	20EC203.2	L2
10 (a)	Explain how a full adder can be built using two half adders	6M	20EC203.3	L6
10 (b)	Design a 4-bit carry full adder circuit	6M	20EC203.3	L6
	OR			
11 (a)	Using 8 to 1 multiplexer, realize the Boolean function $T = f(w, x, y, z) = \Sigma(0,1,2,4,5,7,8,9,12,13)$	8M	20EC203.3	L6
11 (b)	Distinguish between a combinational logic circuit and a sequential logic circuit	4M	20EC203.3	L6
12 (a)	Show and implement the following function using a PROM $F(w,x,y,z)= \Sigma m(1,8,9,15)$ $G(w,x,y,z)= \Sigma m(0,1,2,3,4,5,7,8,10,11,12,13,14,15)$	6M	20EC203.4	L2
12 (b)	Explain the functions of JK flip flop	6M	20EC203.4	L2

	OR			
13 (a)	Implement the following Boolean function using 3×4×2 PLA, $F1(x, y, z) = \Sigma(0, 1, 3, 5)$ and $F2(x, y, z) = \Sigma(3, 5, 7)$.	6M	20EC203.4	L3
13 (b)	Realize a JK flip flop using SR flip flop	6M	20EC203.4	L3
14 (a)	Explain in detail SR latch using NAND	6M	20EC203.5	L2
14 (b)	Explain in detail SR latch using NOR	6M	20EC203.5	L3
	OR			
15 (a)	Convert the SR Flip Flop to T Flip Flop	6M	20EC203.5	L3
15 (b)	Convert the JK Flip Flop to D Flip Flop	6M	20EC203.5	L3

End Semester Examinations

I B.Tech – II Sem (CSE, CSME, CSD) AY :- 2020-21

Course Code 20EC203 Duration :- 180 Mins Max Mark :- 70M

Subject :- Digital Logic Design

Key and Scheme of Evaluation

Questions (1 through 5) Part-A (Short Answer Questions 5 X 2 = 10 M)

1. Convert $(0.625)_{10}$ decimal number to binary number $(?)_2$ using successive multiplication method — (2M)

Sol:- Given decimal number is 0.625

$0.625 = 0.625 \times 2 = 1.25 - 1 = 0.25$

$0.25 = 0.25 \times 2 = 0.5 - 0 = 0.5$

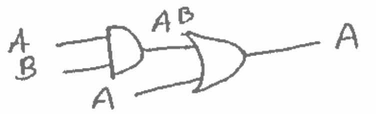
$0.5 = 0.5 \times 2 = 1 - 1 = 0$

$\therefore 0.625_{10} = (0.101)_2$

2. State the absorption law of Boolean Algebra — (2M)

Sol:- There are two laws

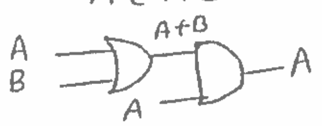
Law 1:- $A + A.B = A$



A	B	AB	A+AB
0	0	0	0
0	1	0	0
1	0	0	1
1	1	1	1

Proof $A + AB = A(1+B)$
$= A(1)$
$= A$

Law 2:- $A(A+B) = A$



A	B	A+B	(A+B)A
0	0	0	0
0	1	1	0
1	0	1	1
1	1	1	1

Proof:
$A(A+B) = A.A + A.B$
$= A + AB$
$= A(1+B)$
$= A$

3. Give the general procedure for converting a Boolean expression in to multilevel NAND diagrams. — (2M)

Sol:- 1) First, draw the AND-OR schematic for Boolean Algebra

2) AND gate will be converted into AND-INVERT and OR will be converted into INVERT-OR

3) Double bubbles along a single line cancel each other and a single bubble along a line

should be compensated by inserting an inverter in that line

4) Then redraw the whole schematic using all NAND gates by replacing AND-INVERT and INVERT-OR with NAND.

4. What are the three types of fundamental PLDS? (2M)

Sol:- 1) PROM – Programmable Read Only Memory
2) PLA – Programmable Logic Arrays
3) PAL – Programmable AND Array logic

5. What is race around condition? —(2M)

Sol:- For J-K flip flop, if J=K=1 and also the clock=1 Q_{n+1} will toggle for a long period, then output as long as CLK remains high which makes the output unstable or uncertain.

This is called a race around condition in J-K flip flop.

Part-B (Long Answer Questions 5X12 = 60 Marks)

6(a). Convert the given binary 1010110 to gray code (6M)

Sol:- The given binary is 1010110

	B_1	B_2	B_3	B_4	B_5	B_6	B_7
Binary Code	1	0	1	0	1	1	0

$G_1 = B_1 = 1$

$G_2 = B_1 \oplus B_2 = 1 \oplus 0 = 1$

$G_3 = B_2 \oplus B_3 = 0 \oplus 1 = 1$

$G_4 = B_3 \oplus B_4 = 1 \oplus 0 = 1$

$G_5 = B_4 \oplus B_5 = 0 \oplus 1 = 1$

$G_6 = B_5 \oplus B_6 = 1 \oplus 1 = 0$

$G_7 = B_6 \oplus B_7 = 1 \oplus 0 = 1$

$\therefore$ Gray Code = 1 1 1 1 1 0 1

6(b) Convert the following (i) 57_{10} to Binary (ii) 743_8 to Binary (iii) $A9B_{16}$ to Binary

Sol:- (i) 57_{10} to Binary — (2 M)

$$\begin{array}{r|l}
2 & 57 \\
\hline
2 & 28 - 1 \\
\hline
2 & 14 - 0 \\
\hline
2 & 7 - 0 \\
\hline
2 & 3 - 1 \\
\hline
 & 1 - 1
\end{array}$$

$\therefore 57_{10} = 111001_2$

(ii) 743_8 to Binary — (2M)

Sol:- For Each octal digit write equivalent 3-bit binary number.

$$\begin{array}{ccc}
7 & 4 & 3 \\
\downarrow & \downarrow & \downarrow \\
111 & 100 & 011
\end{array}$$

$\therefore 743_8 = 111100011_2$

(iii) $A9B_{16}$ to Binary — (2M)

For Each Hexadecimal digit write equivalent 4-bit binary number.

$$\begin{array}{ccc}
A & 9 & B \\
\downarrow & \downarrow & \downarrow \\
1010 & 1001 & 1011
\end{array}$$

$\therefore A9B_{16} = 101010011011_2$

7(a) Convert $(567.1875)_{10}$ into hexadecimal. — 6M

Sol The given decimal is 567.1875_{10}

Successive division by 16 to decimal integer part

$$\begin{array}{r|l}
16 & 567 \\
\hline
16 & 35 - 7 \\
\hline
 & 2 - 3
\end{array}$$

$567_{10} = 237_{16}$

Successive multiplication by 16 to decimal Fraction part

$0.1875 = 0.1875 \times 16 = 2.8125 - 2 = 0.8125$

$0.8125 = 0.8125 \times 16 = 13$

$0.1875 \times 16 = 3$

$\therefore 567.1875_{10} = 237.3_{16}$

7(b) Design an 8421 to gray code converter. — (6M)

Binary

Sol:- BCD to Gray Code Converter

BCD Code				Gray Code			
B_3	B_2	B_1	B_0	G_3	G_2	G_1	G_0
0	0	0	0	0	0	0	0
0	0	0	1	0	0	0	1
0	0	1	0	0	0	1	1
0	0	1	1	0	0	1	0
0	1	0	0	0	1	1	0
0	1	0	1	0	1	1	1
0	1	1	0	0	1	0	1
0	1	1	1	0	1	0	0
1	0	0	0	1	1	0	0
1	0	0	1	1	1	0	1

K-Map for G_3

$B_3B_2 \backslash B_1B_0$	00	01	11	10
00				
01				
11	X	X	X	X
10	1	1	X	X

$G_3 = B_3$

K-Map for G_2

$B_3B_2 \backslash B_1B_0$	00	01	11	10
00				
01	1	1	1	1
11	X	X	X	X
10	1	1	X	X

$G_2 = B_3 + B_2$

K-Map for G_1

$B_3B_2 \backslash B_1B_0$	00	01	11	10
00			1	1
01	1	1		
11	X	X	X	X
10			X	X

$G_1 = B_2\bar{B}_1 + \bar{B}_2 B_1$

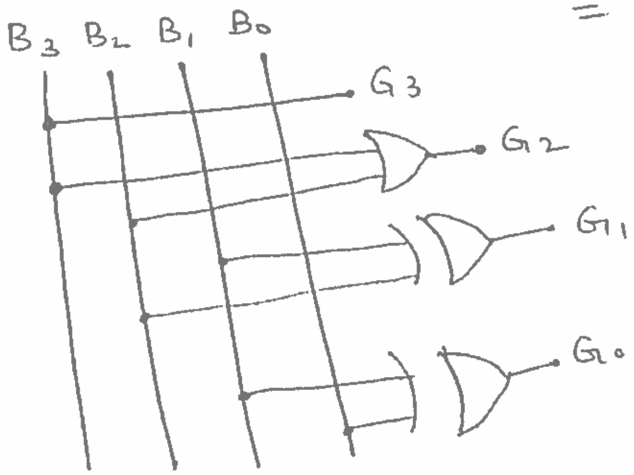
$= B_1 \oplus B_2$

K-Map for G_0

$B_3B_2 \backslash B_1B_0$	00	01	11	10
00		1		1
01		1		1
11	X	X	X	X
10		1	X	X

$G_0 = \bar{B}_1 B_0 + B_1 \bar{B}_0$

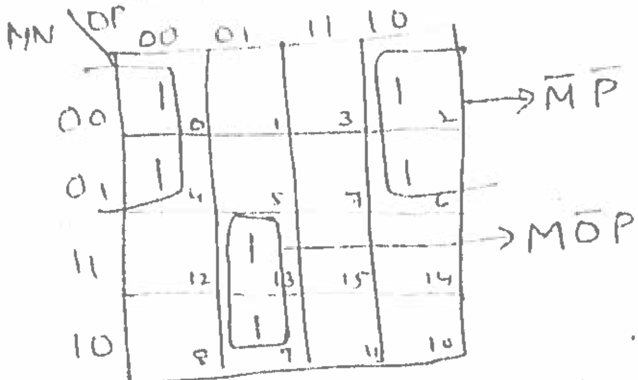
$= B_1 \oplus B_0$



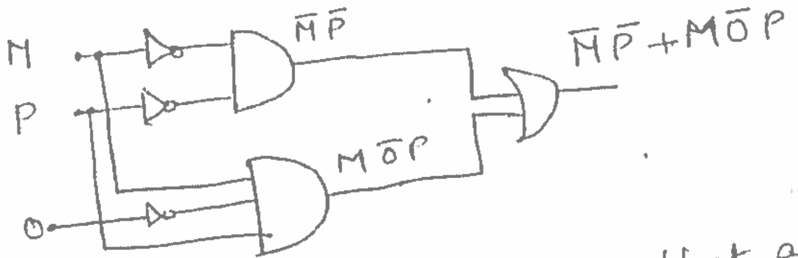
8(a) Develop the given function $Y(M,N,O,P) = \Sigma m(0,2,4,6,9,13)$ Draw the K-map and Implement the simplified expression using basic gates. —— (6M)

Sol:- Given function is

$$Y(M,N,OP) = \Sigma m(0,2,4,6,9,13)$$



$$Y(M,N,O,P) = \bar{M}\bar{P} + M\bar{O}P$$



8(b) Analyze the Basic rules (laws) that are used in Boolean expressions with few examples —(6M)

Sol!- Any three laws

(i) Complementation laws

$\bar{0} = 1$

$\bar{1} = 0$

If $A = 0$ then $\bar{A} = 1$

If $A = 1$ then $\bar{A} = 0$

$\bar{\bar{A}} = A$

(ii) AND Law

$A \cdot 0 = 0$

$A \cdot 1 = A$

$A \cdot A = A$

$A \cdot \bar{A} = 0$

(iii) OR Law:-

$A + 0 = A$

$A + 1 = 1$

$A + A = A$

$A + \bar{A} = 1$

(iv) Commutative law

$A + B = B + A$

A	B	A+B
0	0	0
0	1	1
1	0	1
1	1	1

$\cong$

B	A	B+A
0	0	0
0	1	1
1	0	1
1	1	1

(v) Associative law:-

$(A+B)+C = A+(B+C)$

$(A \cdot B) \cdot C = A \cdot (B \cdot C)$

(vi) Distributive law

$A(B+C) = AB + AC$

$ABC(D+E) = ABCD + ABCE$

$A + BC = (A+B)(A+C)$

(vii) $A + \bar{A}B = A + B$

$A(\bar{A}+B) = AB$

9(a) Simplify the following Boolean expression in SOP using Karnaugh map $AC' + B'D + A'CD + ABCD$ —(6M)

Sol:- Given Boolean expression in SOP is

$$Y = A\bar{C} + \bar{B}D + \bar{A}CD + ABCD$$

AB \ CD	00	01	11	10
00		1	1	
01			1	
11	1	1	1	
10	1	1	1	

$\rightarrow \bar{B}D$

$\rightarrow \therefore CD$

$\rightarrow A\bar{C}$

$$\therefore Y = A\bar{C} + \bar{B}D + CD$$

$$= A\bar{C} + D(\bar{B}+C)$$

9(b) Simplify the following Boolean expression in POS using Karnaugh map $A\bar{C} + \bar{B}D + \bar{A}CD + ABCD$

Sol:- Given Boolean expression in POS is

$$A\bar{C} + \bar{B}D + \bar{A}CD + ABCD$$

AB \ CD	00	01	11	10
00	0			0
01	0	0		0
11				0
10				0

$\rightarrow$ ~~A+B~~ $A + D$

$A + \bar{B} + C$

$\bar{C} + D$

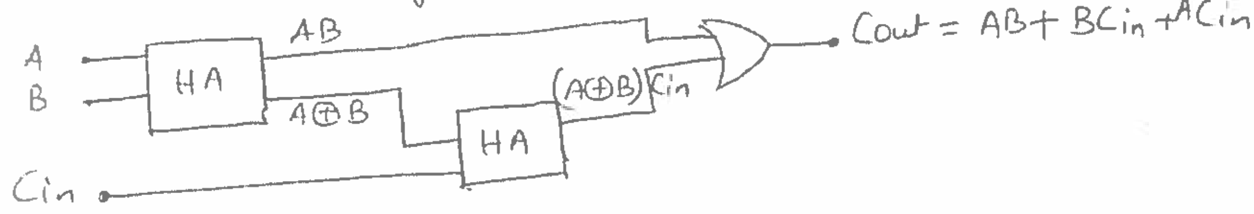
~~$\therefore Y = A + D \cdot \bar{C}$~~

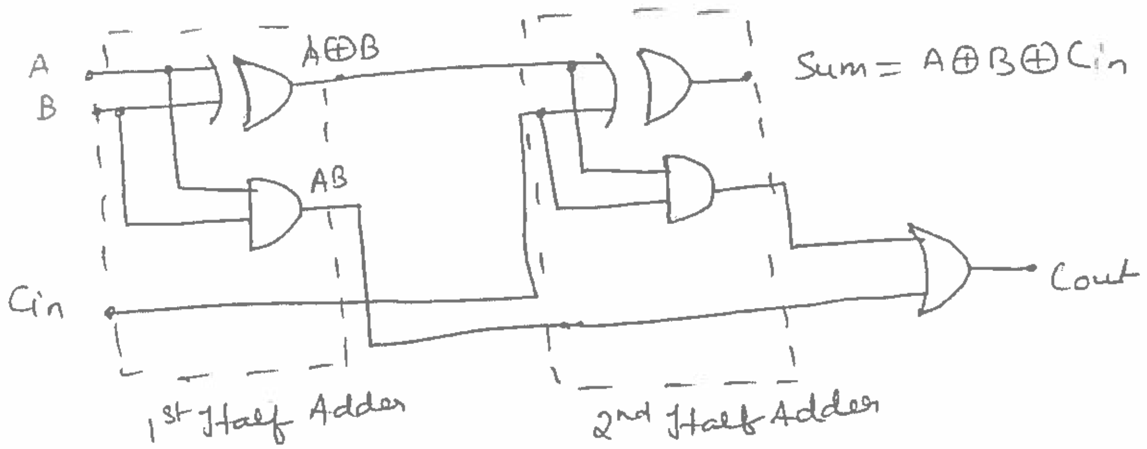
$$\therefore Y = (A+D)(\bar{C}+D)(A+\bar{B}+C)$$

10 (a) Explain how a full Adder can be built using two half Adders —— (6M)

Sol:- A Full Adder can be built using two Half Adders is as follows.

A Full adder can be implemented with two half adders and an OR-gate.





$Cout = AB + (A \oplus B)Cin$

$= AB + Cin(A\overline{B} + \overline{A}B)$

$= AB + A\overline{B}Cin + \overline{A}BCin$

$= AB(Cin+1) + A\overline{B}Cin + \overline{A}BCin$ $[\because Cin+1 = 1]$

$= ABCin + AB + A\overline{B}Cin + \overline{A}BCin$

$= AB + ACin(B+\overline{B}) + \overline{A}BCin$

$= AB + ACin + \overline{A}BCin$ $[\because Cin+1 = 1]$

$= AB(Cin+1) + ACin + \overline{A}BCin$

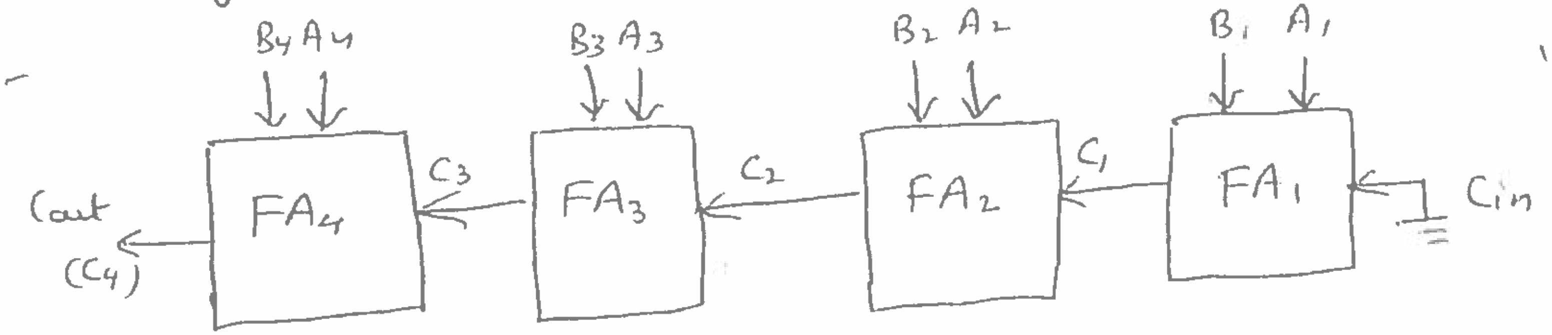
$= ABCin + AB + ACin + \overline{A}BCin$

$= BCin(A+\overline{A}) + AB + ACin$ $[\because A+\overline{A} = 1]$

$= AB + BCin + ACin$

10(b) Design a 4-bit Carry full Adder circuit.

Sol:-

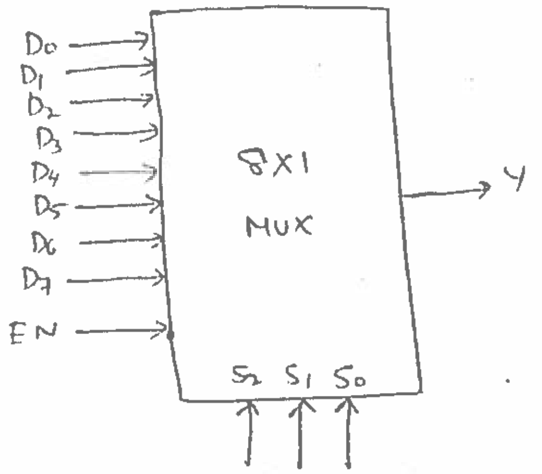


Explanation part of the design and also can be consider Carry look-a head Adder also.

11(a) Using 8 to 1 multiplexer, realize the Boolean function

$$T = f(w,x,y,z) = \Sigma(0,1,2,4,5,7,8,9,12,13)$$

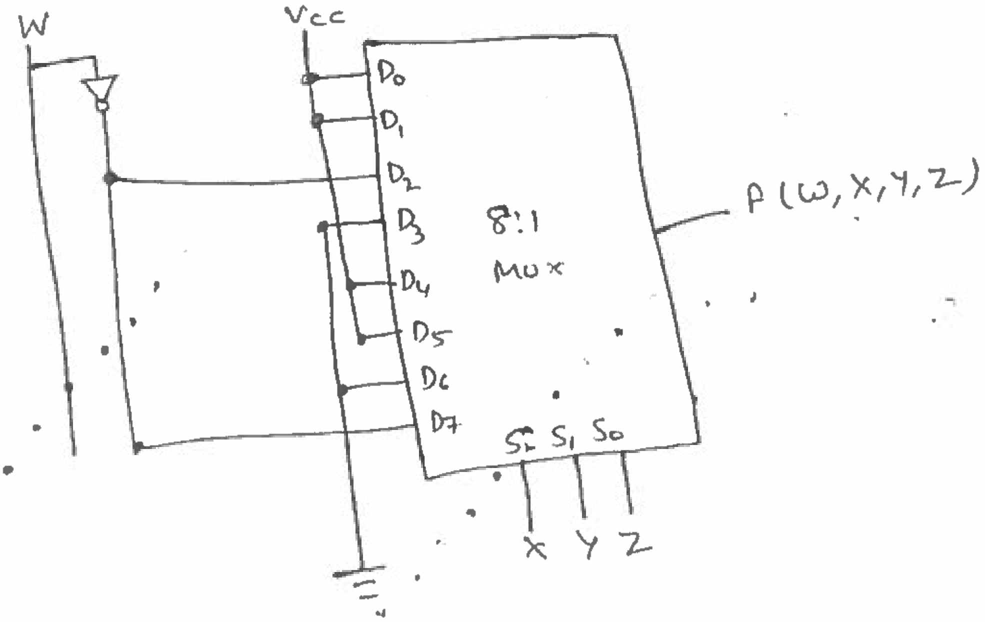
Sol:-



Given Boolean function is

$$T = f(w,x,y,z) = \Sigma(0,1,2,4,5,7,8,9,12,13)$$

	D_0	D_1	D_2	D_3	D_4	D_5	D_6	D_7
$\bar{w}$	0	1	2	3	4	5	6	7
w	8	9	10	11	12	13	14	15
	1	1	$\bar{w}$	0	1	1	0	$\bar{w}$



11(b) Distinguish between a Combinational logic circuit and a sequential logic circuit. —(6 M)

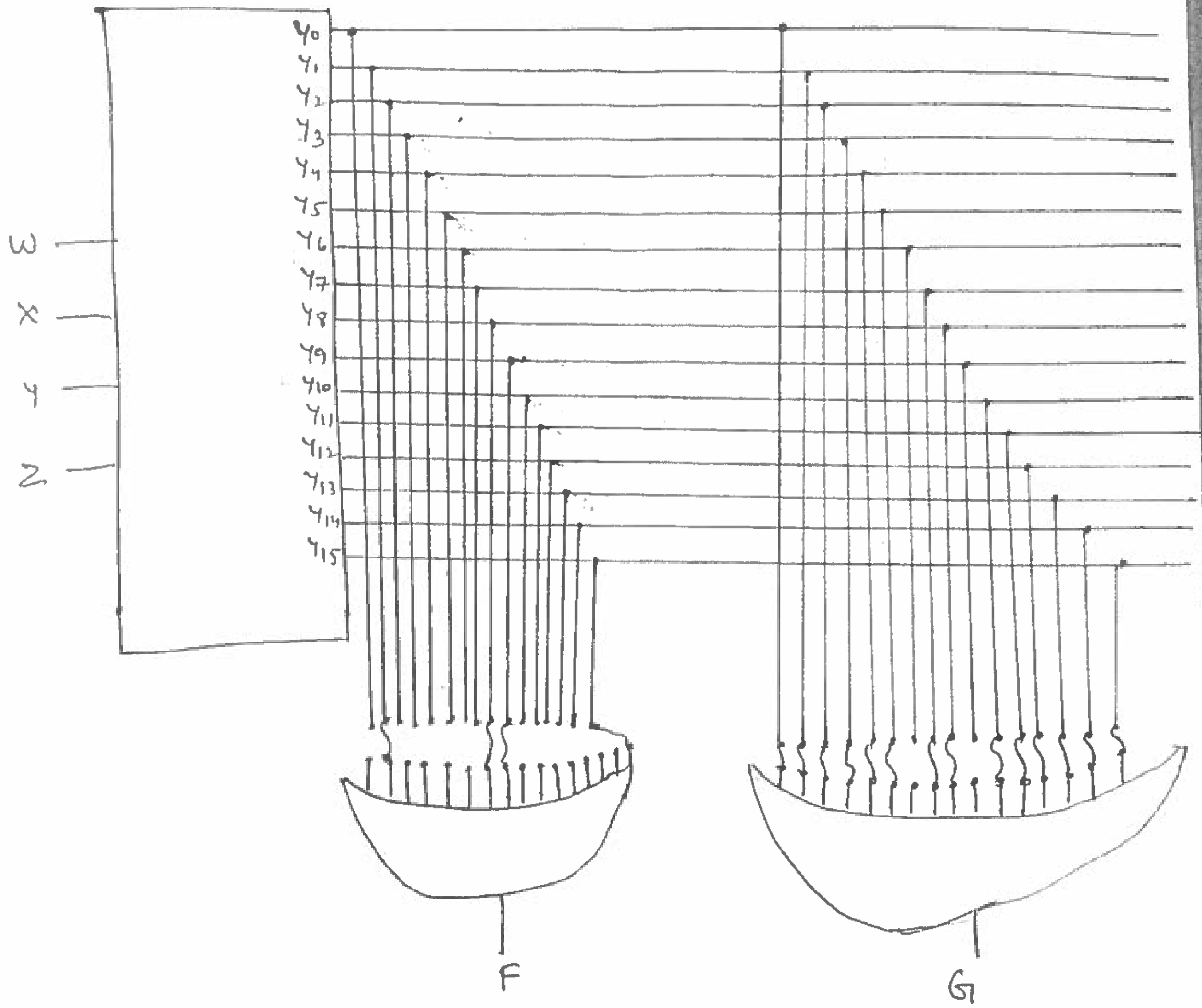
Sol:

Combinational logic Circuit	Sequential logic Circuit
1) The o/p variables at any instant of time are dependent only on the present input variables	1) The o/p variables at any instant of time are dependent not only on the present input variables but also on the present state i.e past history of the system.
2) Memory unit is not required	2) Memory unit is required to store the past history of the input variables in sequential circuits
3) Combinational circuits are faster because the delay between the i/p and the o/p is due to propogation delay of gates only	3) Sequential circuits are slower than the Combination circuits.
4) Easy to Design	4) Harder to design

12(a) Show and Implement the following function using a PROM $F(W,X,Y,Z) = \Sigma m(1,8,9,15)$

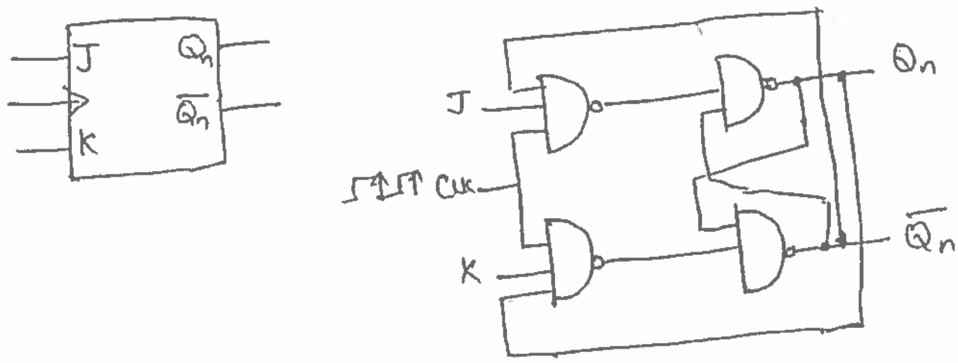
$G(W,X,Y,Z) = \Sigma m(0,1,2,3,4,5,7,8,10,11,12,13,14,15)$

12(a) Sol:-



12(b) Explain the Function of JK Flip Flop — (6M)

Sol:- <u>JK Flip Flop</u>



<u>JK Flip Flop</u> :- The uncertainty in the State of SR flip flow when S=1 & R=1 can be eliminated by converting it into a JK Flip Flop.

Working:-

case 1: when $J=0, K=0, Clk=1$, then the output

of $G_1 = \overline{0\cdot 1\cdot \bar{Q}_n} = \bar{0} = 1$

$G_2 = \overline{0\cdot 1\cdot Q_n} = \bar{0} = 1$

$G_3 = \overline{1\cdot \bar{Q}_n} = \bar{\bar{Q}}_n = Q_n = Q_{n+1}$

$G_4 = \overline{1\cdot Q_n} = \bar{Q}_n = \bar{Q}_{n+1}$

case 2: when $J=0, K=1$ then the output are

$G_1 = \overline{0.1.\bar{Q}_n} = \overline{0.\bar{Q}_n} = \bar{0} = 1$

$G_2 = \overline{1.1.Q_n} = \bar{Q}_n$

$G_3 = \overline{1.\bar{Q}_n} = \bar{\bar{Q}}_n = Q_n = Q_{n+1}$

$G_4 = \overline{\bar{Q}_n \cdot Q_n} = \bar{0} = 1 = \bar{Q}_{n+1}$

$\Rightarrow Q_{n+1} = 0$

case 3: when $J=1, K=0$, then the output are.

$G_1 = \overline{1.1.\bar{Q}_n} = \bar{\bar{Q}}_n = Q_n$

$G_2 = \overline{0.1.Q_n} = \bar{0} = 1$

$G_3 = Q_{n+1} = \overline{Q_n \cdot \bar{Q}_n} = \bar{0} = 1$

$G_4 = \bar{Q}_{n+1} = \overline{1.Q_n} = \bar{1} = 0$

Case 4: when $J=1, K=1$, then the outputs are

$G_1 = \overline{1.1.\bar{Q}_n} = \bar{\bar{Q}}_n = Q_n$

$G_2 = \overline{1.1.Q_n} = \bar{Q}_n$

$G_3 = Q_{n+1} = \overline{Q_n \cdot \bar{Q}_n} = \bar{0} = 1$ (for explaination only)

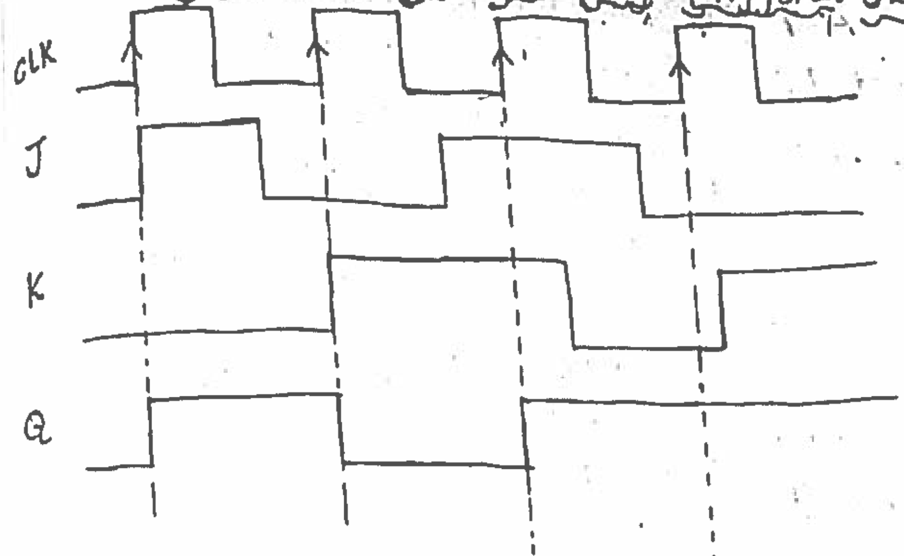
$G_4 = \bar{Q}_{n+1} = \overline{Q_n \cdot \bar{Q}_n} = \bar{0} = 1$

$G_3 = Q_{n+1} = \overline{Q_n \cdot 1} = \bar{Q}_n$

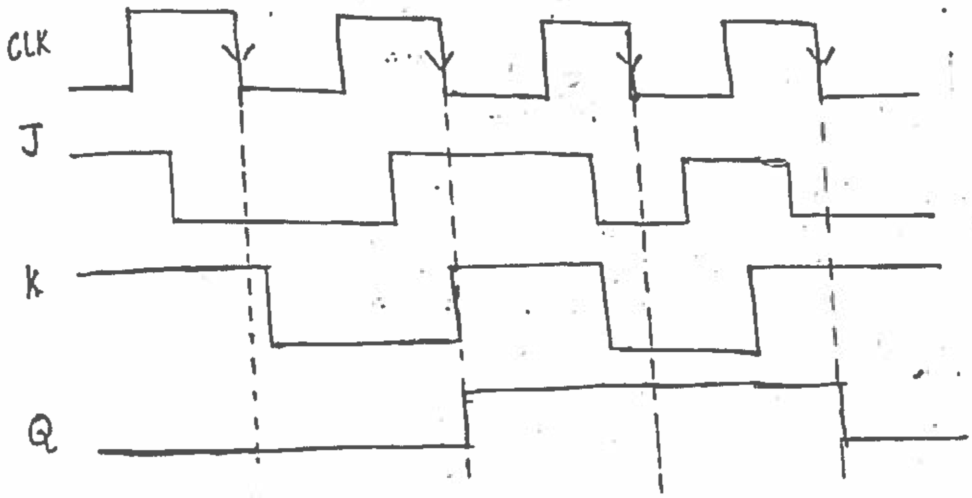
Truth Table:

CLK	J	K	Q_n	Q_{n+1}	State
↑	0	0	0	0	No change
↑	0	0	1	1	
↑	0	1	0	0	Reset
↑	0	1	1	0	
↑	1	0	0	1	Set
↑	1	0	1	1	
↑	1	1	0	1	Toggle
↑	1	1	1	0	
0	x	x	0	0	No change
0	x	x	1	1	

waveforms for +ve edge triggered JK FF



waveforms for -ve edge triggered JK FF



characteristic equation of JK FF ..

J \ KQ_n	00	01	11	10
0		1		
1	1	1		1

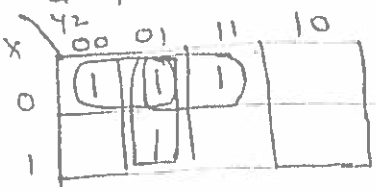
$$\boxed{Q_{n+1} = J\bar{Q}_n + \bar{K}Q_n}$$

Excitation Table: . :

PS Q_n	NS Q_{n+1}	Req. Inputs J	K
0	0	0	X
0	1	1	X
1	0	X	1
1	1	X	0

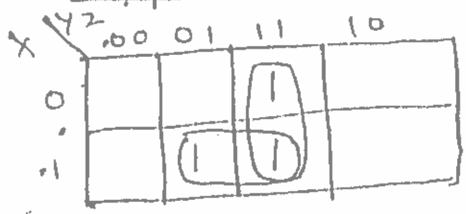
13(a) Implement the following Boolean function using 3×4×2 PLA, $F_1(x,y,z) = \Sigma(0,1,3,5)$ and $F_2(x,y,z) = \Sigma(3,5,7)$ — (6M)

Sol:- K-Map for F_1



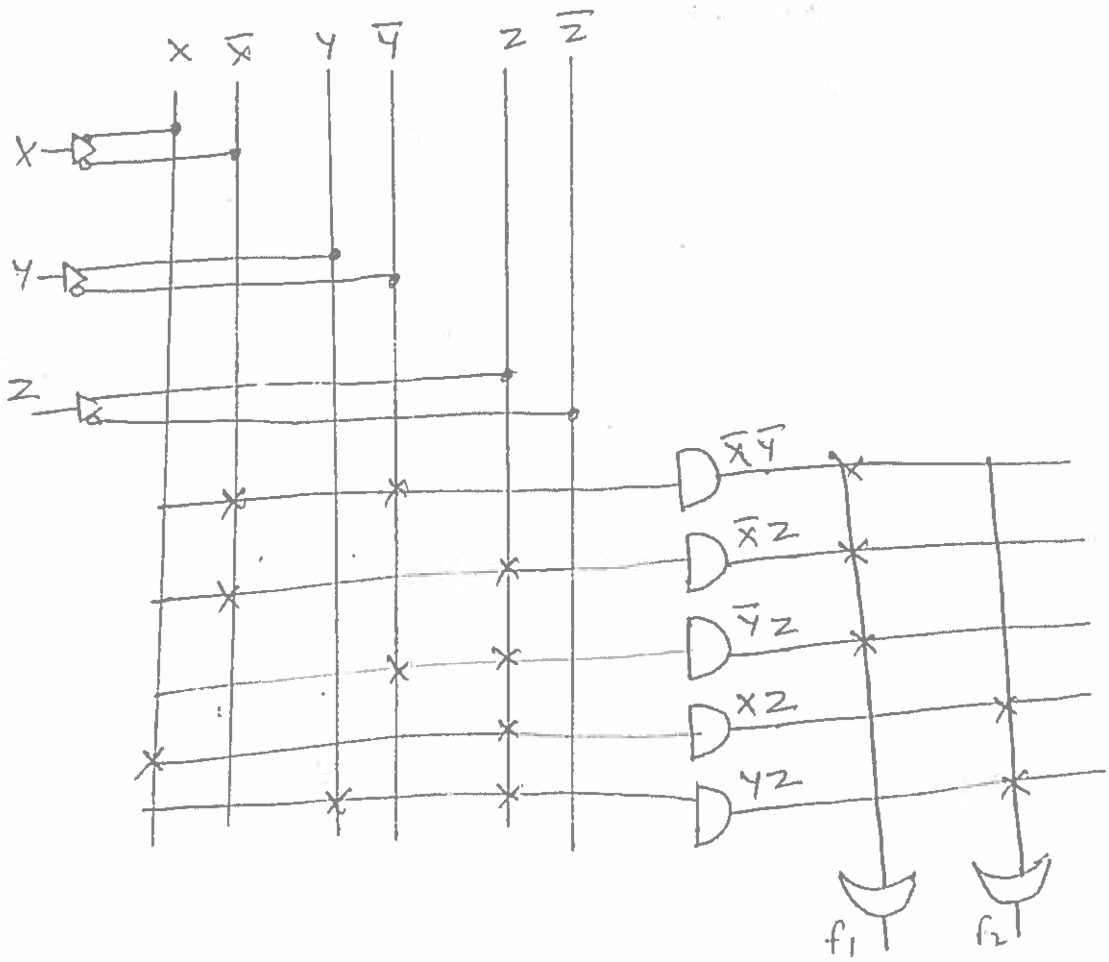
$F_1 = \bar{X}\bar{Y} + \bar{X}Z + \bar{Y}Z$

K-Map For F_2



$F_2 = XZ + YZ$

Product term	Input x,y,z	Output f_1 f_2
$\bar{X}\bar{Y}$	0 0 –	1 –
$\bar{X}Z$	0 – 1	1 –
$\bar{Y}Z$	– 0 1	1 –
XZ	1 – 1	– 1
YZ	– 1 1	– 1



13(b) Realize a JK flip flop using SR flip flop

JK Flip-flop to SR Flip-flop

Input S	Input R	PS Q_n	NS Q_{n+1}	FF inputs J	FF inputs K
0	0	0	0	0	X
0	0	1	1	X	0
0	1	0	0	0	X
0	1	1	0	X	1
1	0	0	1	1	X
1	0	1	1	X	0
1	1	0	X	X	X
1	1	1	X	X	X

K-Map for J

S \ RQ_n	00	01	11	10
0		X	X	
1	1	X	X	X

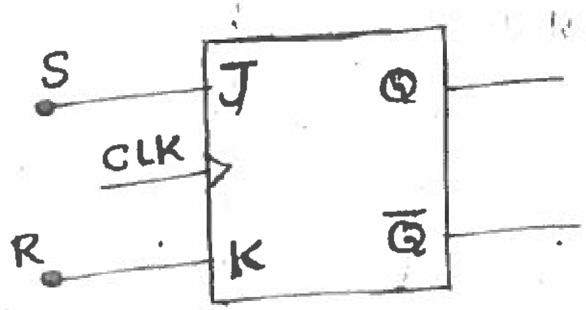
$J = S$

K-Map for K

S \ RQ_n	00	01	11	10
0	X		1	X
1	X		X	X

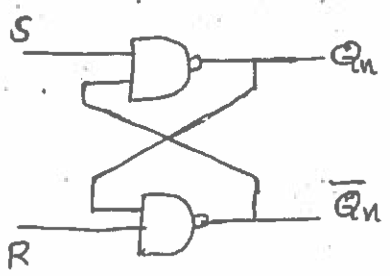
$K = R$

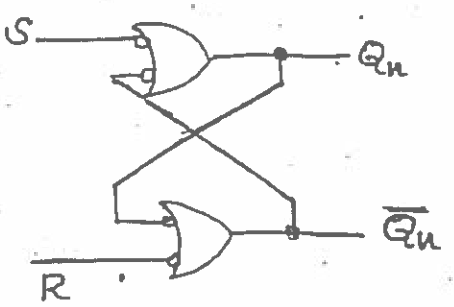
Logic symbol



NAND gate S-R Latch (Active low SR Latch):

Logic diagram and truth table of active low SR latch is shown below. Since the NAND gate is equivalent to an active - LOW OR gate, an active LOW SR latch using OR gates may also be represented ..





Working: Case 1: when $S=0$ $R=0$

$$Q_{n+1} = \overline{S\overline{Q_n}} = \overline{0\overline{Q_n}} = \overline{0} = 1$$
$$\overline{Q_{n+1}} = \overline{RQ_n} = \overline{0Q_n} = \overline{0} = 1$$

Case 2:- when $S=0$ $R=1$

$$Q_{n+1} = \overline{S\overline{Q_n}} = \overline{0\overline{Q_n}} = \overline{0} = 1$$
$$\overline{Q_{n+1}} = \overline{RQ_n} = \overline{1Q_n} = \overline{Q_n} = 0$$

Case 3: when $S=1$ $R=0$.

$$Q_{n+1} = \overline{S\overline{Q_n}} = \overline{1\cdot\overline{Q_n}} = \overline{\overline{Q_n}} = Q_n = 0$$
$$\overline{Q_{n+1}} = \overline{RQ_n} = \overline{0Q_n} = \overline{0} = 1$$

Case 4:- when $S=1$ $R=1$

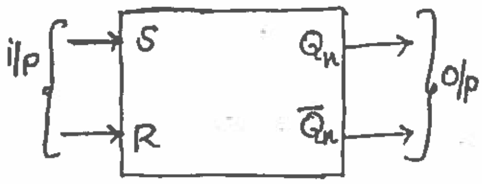
$$Q_{n+1} = \overline{S\overline{Q_n}} = \overline{1\overline{Q_n}} = \overline{\overline{Q_n}} = Q_n$$
$$\overline{Q_{n+1}} = \overline{RQ_n} = \overline{1Q_n} = \overline{Q_n}$$

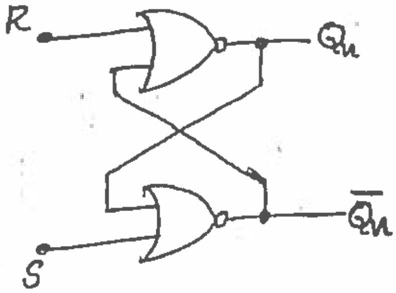
Truth Table:

S	R	Q_n	Q_{n+1}	Status
0	0	0	X	Invalid
0	0	1	X	
0	1	0	1	Set
0	1	1	1	
1	0	0	0	Reset
1	0	1	0	
1	1	0	0	No change
1	1	1	1	

S	R	Q_{n+1}
0	0	Invalid
0	1	Set
1	0	Reset
1	1	No change

NOR gate S-R Latch (Active-High SR Latch):-





Q_n represents the state before applying the inputs and Q_{n+1} represents the state after the application of the inputs.

Working: Case 1:- when $S=0, R=0$

$$Q_{n+1} = \overline{R+\overline{Q_n}} = \overline{0+\overline{Q_n}} = \overline{\overline{Q_n}} = Q_n$$

$$\overline{Q_{n+1}} = \overline{S+Q_n} = \overline{0+Q_n} = \overline{Q_n}$$

Case 2: when $R=0, S=1$

$$Q_{n+1} = \overline{R+\overline{Q_n}} = \overline{0+\overline{Q_n}} = \overline{\overline{Q_n}} = Q_n$$

$$\overline{Q_{n+1}} = \overline{S+Q_n} = \overline{1+Q_n} = \overline{1} = 0$$

$$\Rightarrow Q_{n+1} = 1.$$

Case 3: when $R=1$ $S=0$

$$Q_{n+1} = \overline{R+\overline{Q_n}} = \overline{1+\overline{Q_n}} = \overline{1} = 0.$$

$$\overline{Q_{n+1}} = \overline{S+Q_n} = \overline{0+Q_n} = \overline{Q_n} = 1$$

Case 4: when $R=1$ $S=1$

$$Q_{n+1} = \overline{R+\overline{Q_n}} = \overline{1+\overline{Q_n}} = \overline{1} = 0$$

$$\overline{Q_{n+1}} = \overline{S+Q_n} = \overline{1+Q_n} = \overline{1} = 0$$

Truth Table

R	S	Q_n	Q_{n+1}	Status
0	0	0	0	No change
0	0	1	1	
0	1	0	1	Set
0	1	1	1	
1	0	0	0	Reset
1	0	1	0	
1	1	0	x	Invalid
1	1	1	x	

R	S	Q_{n+1}
0	0	Q_n
0	1	1
1	0	0
1	1	Invalid

SR Flip-flop to T Flip-flop:-

The excitation table for this conversion is as follows:

Input	PS	NS	Flip-flop i/p	
T	Q_n	Q_{n+1}	S	R
0	0	0	0	X
0	1	1	X	0
1	0	1	1	0
1	1	0	0	1

K-Map for S

$T \backslash Q_n$	0	1
0		X
1	1	

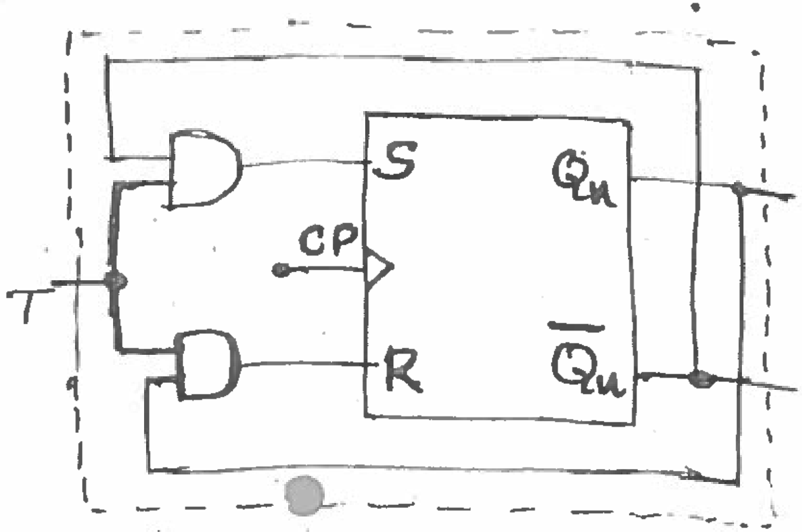
$S = T\overline{Q_n}$

K-Map for R

$T \backslash Q_n$	0	1
0	X	
1		1

$R = TQ_n$

Logic symbol



JK FF to D FF:

The excitation table for this conversion is as follows

Input	PS	NS	Flip-flop inputs	
D	Q_n	Q_{n+1}	J	K
0	0	0	0	X
0	1	0	X	1
1	0	1	1	X
1	1	1	X	0

(31)

K-Map for J

$D \backslash Q_n$	0	1
0		X
1	1	X

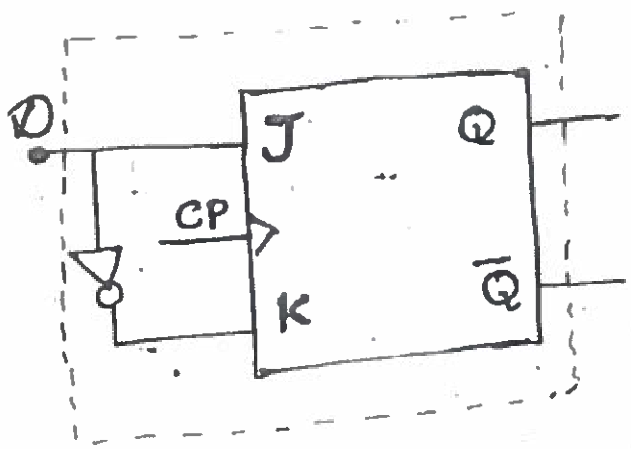
$J = D$

K-Map for K

$D \backslash Q_n$	0	1
0	X	1
1	X	

$K = \bar{D}$

Logic symbol



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(HOD-ECE)
Dr. B. SIVAPRASAD

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	Program	ECE			Academic Year	2020 - 2021
Course Code	20EC201	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	Principles of Electronics & Communication Systems						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Fermi level	20EC201.1	L1
2	What is CMRR?	20EC201.2	L1
3	What is the need for modulation?	20EC201.3	L1
4	Define PAM and PPM	20EC201.4	L1
5	Define TIR	20EC201.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain Insulator, Semiconductor & conductor with help of energy band structure	6M	20EC201.1	L2
6 (b)	Differentiate between intrinsic and extrinsic semiconductor	6M	20EC201.1	L2
	OR			
7 (a)	Explain n-type semiconductor	6M	20EC403.1	L2
7 (b)	Derive the expression for current generated due to drifting of charge carriers in semiconductors in the presence of electric field	6M	20EC403.1	L2
8 (a)	Explain application of op-amp as integrator and differentiator	6M	20EC201.2	L2
8 (b)	Explain ac characteristics of op-amp	6M	20EC201.2	L2
	OR			
9 (a)	Draw and explain the pin diagram IC741 op-amp	6M	20EC201.2	L2
9 (b)	Derive the gain for non-inverting op-amp	6M	20EC201.2	L2
10 (a)	State and explain properties of continuous signals	8M	20EC201.3	L2
10 (b)	List any four applications of FM system	4M	20EC201.3	L2
	OR			
11 (a)	Define amplitude modulation. Derive an expression for the AM wave	8M	20EC201.3	L2
11 (b)	Write about am voltage distribution	4M	20EC201.3	L2
12 (a)	State and prove sampling theorem	6M	20EC201.4	L2
12 (b)	Describe the basic principles of PCM system and PCM transmitter	6M	20EC201.4	L2
	OR			
13 (a)	Explain the basic Elements of Digital Communication System	6M	20EC201.4	L2
13(b)	With a neat diagram explain the Generation of PCM & DPCM	6M	20EC201.4	L2
14(a)	Draw and explain the working principle of an Optical transmitter	6M	20EC201.5	L2
14 (b)	Explain about LED and its type	6M	20EC201.5	L2
	OR			
15(a)	Explain the working principle of GSM	6M	20EC201.6	L2
15(b)	Explain Cellular Telephone Systems	6M	20EC201.6	L2

Degree : B.Tech   Program : ECE   Semester : II

Course Code : 20EC201   Course : Principles of Electronics & Communication Systems

Key and Scheme of Evaluation

PART - A

1. Define Fermi level ? (2M)

Ans. The Fermi level E_F indicates the probability of occupancy of an energy level by an electron.

2. What is CMRR ? (2M)

Ans. It is defined as the ratio of the differential Voltage gain A_d to the Common mode Voltage gain A_{cm}.

$$CMRR = A_d / A_{cm}$$

This parameter indicates the Capability of the op-amp to reject noise.

3. What is the need for Modulation ? (2M)

Ans : 1) To reduce the antenna height.
2) for Multiplexing of Signals.
3) To increase the range of Communication.
4) To reduce Noise and interference.

4. Define PAM and PPM. (PAM - 1M PPM - 1M)

Ans. PAM : Pulse Amplitude Modulation is a process of Changing the amplitude of high frequency periodic rectangular pulse in accordance with the amplitude of message signal.

PPM : pulse position Modulation is a process of Changing the position of high frequency periodic rectangular pulse in accordance with the amplitude of Message signal.

5. Define TIR ? (2M)

Ans : When the incident angle is increased beyond the critical angle, the light ray does not pass through the interface into the other medium. In this condition angle of reflection ϕ_2 is equal to the angle of Incidence ϕ_1. This action is called as Total Internal Reflection (TIR) of the beam.

PART – B

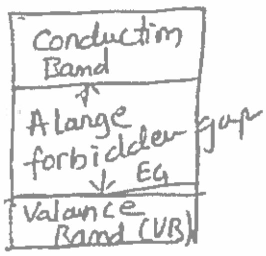
6(a) Explain Insulator, Semiconductor & Conductor with help of energy band structure. (2M + 2M + 2M = 6M)

Ans

Insulators: Insulators passes no free Charge Carriers and thus are non-conductive. Insulators are implemented in household items and electrical Circuits as protection.

Insulators possess a high resistivity and low Conductivity. Their atoms have tightly bound electrons that donot move throughout the material. Because the electrons are static and not freely roaming, a current cannot easily pass.

Eg: Rubber, Teflon, Cloth, Wood and fibaglass



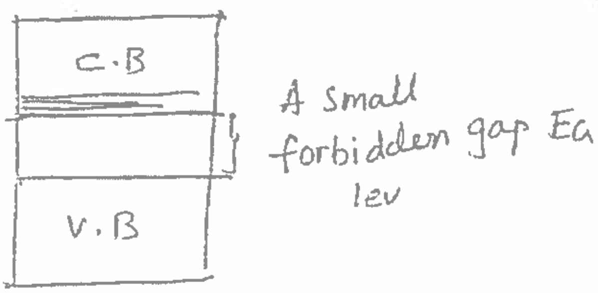
(a) Insulator

Semiconductor

In semiconductors the gap between ValanceBand and Conduction band is smaller

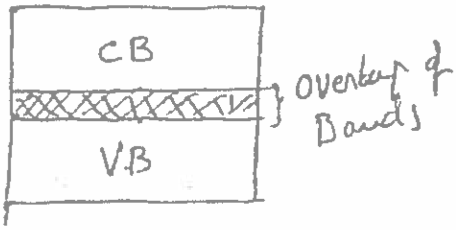
Ex: Ga, As, Si and Ge

At room temparature there is sufficient energy available for electrons to make a transistion from V.B to C.B. This allows some conduction to takes place.



Conductor : A conductor is defined as an object or type of material That allows the flow of Charge in one or more directions. Materials made of metal are Common electrical conductor, as metal have a high Conductance and low resistance.

Eg: Aluminium, Silver, Copper etc.



6(b) Differentiate between intrinsic and extrinsic semiconductor.

Ans

Intrinsic semiconductor	Extrinsic semiconductor
1. pure form of semiconductor	1. Impure form of semiconductor
2. It exhibits poor conductivity	2. It possesses comparatively better conductivity than intrinsic semiconductor
3. It is present in the middle of forbidden energy gap.	3. The presence of fermi level varies according to the type of extrinsic semiconductor
4. The conduction relies on temperature.	4. The conduction depends on the concentration of dopped impurity and temperature.
5. Equal amount of electron and holes are present in C.B & V.B	5. The majority presence of electrons and holes depends on the type of extrinsic semiconductor
6. It is not further classified	6. It is classified as p-type and n-type.

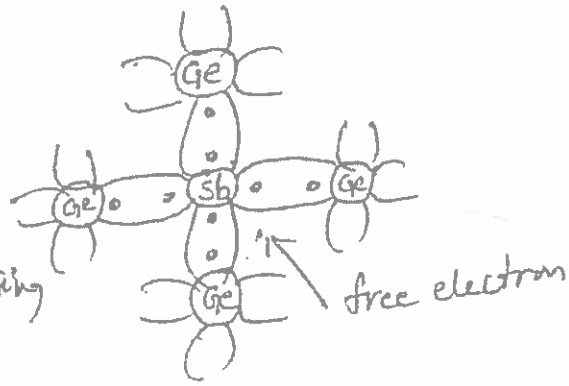
Marks: each difference 1 Mark

total – 6 Marks

7(a) Explain n-type semiconductor ? Content – 4M diagram – 2M

Ans: A small amount of pentavalent impurities such as Arsenic, antimony or phosphorus is added to the pure semiconductor (germanium or silicon crystal) to get N-type semiconductor.

Ge atom has four Valance electrons and antimony has five Valance electrons each antimony atom forms a Covalent bond with Surrounding four germanium atoms. Thus, four Valance electrons of antimony atom form Covalent bond with four Valance electrons of individual Germanium atom and fifth Valence electron is left free which is loosely bound to the antimony atom.



This loosely bound electron Can be easily excited from the V·B to the C·B by the application electric field or increasing the Thermal energy.

7(b) Derive the expression for current generated due to drifting of Charge Carriers in Semiconductors in the presence of electric field ?
[Content : 2M Equations: 4M = 6M]

Ans: When an electric field is applied accross the Semiconductor material, the Charge Carriers attain a Certain drift Velocity V_d. Which is equal to the product of the mobility of the Charge Carriers and the applied electric field intensity, E. The holes move towards the negative terminal of the battery and electrons move toward the positive terminal. This Combined effect of movement of the Charge Carriers Constitutes a Current known as the Drift Current.

Thus the Drift current is defined as the flow of electric current due to the motion of the charge carriers under the influence of an external electric field.

The equation for the drift current density J_n, due to free electrons given by $J_n = qn\mu_n E$ A/cm^2

and the drift current density J_p, due to holes is given by

$$J_p = qp\mu_p E \quad A/cm^2$$

Where n = number of free electrons per cubic centimeter

p = number of holes per cubic centimeter

μ_n = mobility of electrons in $cm^2/V\text{-}s$

μ_p = mobility of holes in $cm^2/V\text{-}s$

E = applied electric field intensity in V/cm

q = charge of an electron $= 1.6 \times 10^{-19}$ coulomb.

8(b) Explain ac characteristics of op-amp. [each characteristic 2M, 3X2 = 6M]

Ans: Slew Rate: It is defined as the maximum rate of change of output voltage with time.

The slew rate is specified in $V/\mu sec$. Thus

$$\text{Slew rate} = S = \left.\frac{dV_o}{dt}\right|_{max}$$

Transient Response Rise time:

When the o/p of the op-amp is suddenly changing like pulse type then the rise time of the response depends on the cut-off frequency f_H of the op-amp. Such a rise time is called cut-off frequency limited rise time or transient response rise time. It is inversely propotional to the cut-off frequency and given by

$$t_r = \frac{0.35}{F_H}$$

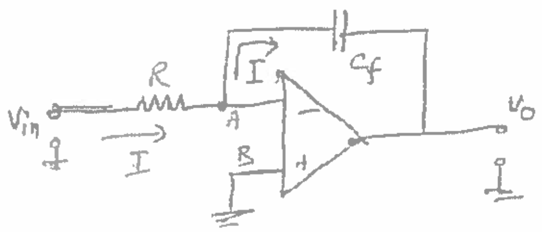
When t_r = rise time f_H = Cut-off frequency

frequency Response of op-amp

The plot showing the Variations in magnitude and phase angle of the gain due to the change in frequency is called frequency response of the op-amp.

8(a) Explain application of op-amp as Integrator & Differentiator [Integrator - 3M, Differ - 3M]

Ans Integrator: In an integrator circuit, the o/p Voltage is the integration of the I/p Voltage.



Consider the op-amp integrator ckt. The node B is grounded. The node A is also at the ground potential from the Concept of Virtual ground.

from the I/p side we can write

$$I = \frac{V_{in} - V_A}{R_1} = \frac{V_{in}}{R_1} \quad \text{--- (1)}$$

from o/p side we can write

$$I = C_f \frac{d(V_A - V_o)}{dt} = -C_f \frac{dV_o}{dt} \quad \text{--- (2)}$$

equating eq (1) & (2)

$$\frac{V_{in}}{R_1} = -C_f \frac{dV_o}{dt}$$

Integrating both sides

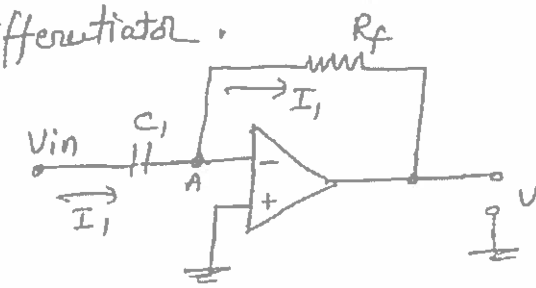
$$\int_0^t \frac{V_{in}}{R_1} dt = -C_f \int \frac{dV_o}{dt} dt$$

i.e $$\int_0^t \frac{V_{in}}{R_1} dt = -C_f V_o$$

$$\therefore V_o = -\frac{1}{R_1 C_f} \int_0^t V_{in}\, dt + V_o(0)$$

Where $V_o(0)$ is the Constant of Integration

Differentiator: The Circuit which produces the differentiation of the Input Voltage at its output is called Differentiator.



The node B is grounded. The node A is also at the ground potential hence $V_A = 0$

As I/p Current of op-amp is zero, either Current I_1 flows through the resistance R_f.

from the I/p side we can write

$$I_1 = C_1 \frac{d(V_{in} - V_A)}{dt} = C_1 \frac{dV_{in}}{dt} \quad \text{——} ①$$

from the o/p side

$$I = \frac{V_A - V_o}{R_f} = \frac{-V_o}{R_f} \quad \text{——} ②$$

Equating the two equations

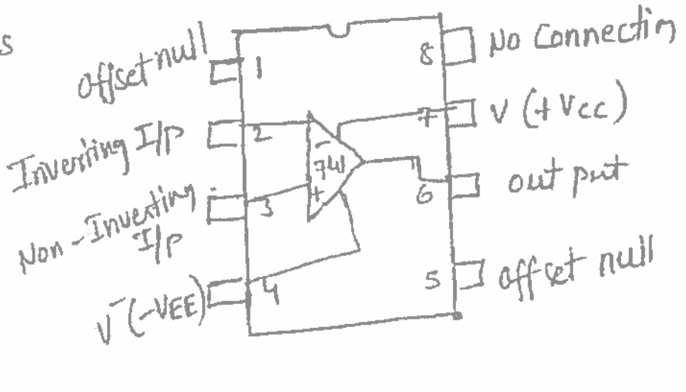
$$C_1 \frac{dV_{in}}{dt} = \frac{-V_o}{R_f}$$

$$\boxed{V_o = -C_1 R_f \frac{d\,V_{in}}{dt}}$$

The equation shows that the o/p is C_1R_f times the differentiation of the input and product C_1R_f is called time Constant of the differentiator.

9(a) Draw and Explain the pin Diagram IC741 op-amp. Pin Diagram – 3M, Description – 3M } 6M

Ans



Description of op-amp 741 IC pins

Pin 1 and 5. These two pins are used for offset process

Pin 2 : Inverting I/p terminal, i-e when a sinusoidal signal is applied to the Input pin 2:

Pin 3 : Non-inverting input terminal i.e when a sinusoidal signal is applied to the input pin 3, waveform of same phase o/p is obtained.

Pin 4 : $-V_{cc}$, i.e negative terminal of supply voltage is connected to this pin

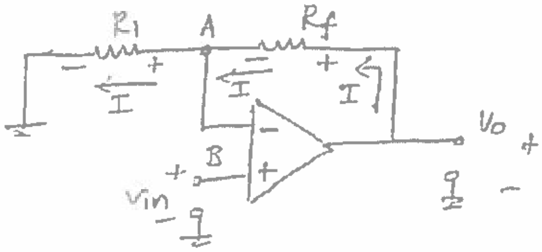
Pin 6 : o/p terminal.

Pin 7 : $+V_{cc}$ i.e +ve terminal of supply voltage is connected to this pin.

Pin 8 : No electrical connection is there in this pin : this pin is just for balance and the symmetric dual-input package look.

9(b) Derive the gain for non-inverting op-amp. Diagram — 2M, Derivation — 4M } 6M

Ans An amplifier which amplifies the input without producing any phase shift b/w I/p and o/p is called non-inverting amplifier.



Derivation of closed loop gain :

The node B is at potential Vin, hence the potential of point A is same as B which is Vin, from the concept of Virtual Ground.

$\therefore V_A = V_B = V_{in}$ —— ①

from the o/p side we can write

$$I = \frac{V_0 - V_A}{R_f}$$

$$\therefore I = \frac{V_0 - V_{in}}{R_f} \quad \text{—— ②}$$

At the inverting terminal

$$I = \frac{V_A - 0}{R_1}$$

$$\therefore I = \frac{V_{in}}{R_1} \quad \text{---} ②$$

equating 2 and 3

$$\therefore \frac{V_o - V_{in}}{R_f} = \frac{V_{in}}{R_1}$$

$$\therefore \frac{V_o}{R_f} = \frac{V_{in}}{R_f} + \frac{V_{in}}{R_1} = V_{in}\left[\frac{R_1 + R_f}{R_1 R_f}\right]$$

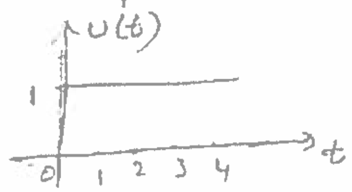
$$\frac{V_o}{V_{in}} = \frac{R_f(R_1+R_f)}{R_1 R_f} = \frac{R_1 + R_f}{R_1} =$$

$$\therefore A_{VF} = \frac{V_o}{V_{in}} = 1 + \frac{R_f}{R_1}$$

10(a) State and Explain the properties of Continuous signals.

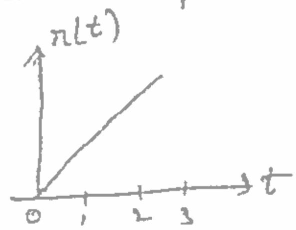
Ans: Unit step signal: The Unit Step function is defined as

$$u(t) = 1 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Unit Ramp Signal: The Unit ramp function is defined as

$$r(t) = t \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



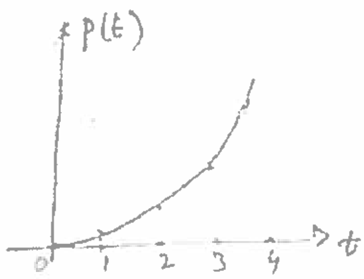
or $r(t) = t\,u(t)$

The ramp function can be obtained by Integrating the Unit step funct

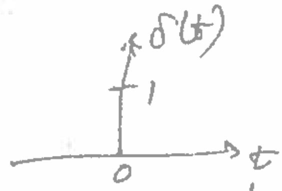
$$r(t) = \int u(t)\,dt = \int dt = t$$

Unit parabolic: $p(t) = t^2/2 \quad \text{for } t \geq 0$

$$= 0 \quad \text{for } t < 0$$



Impulse Signal : $\delta(t) = 0$ for $t \neq 0$



Sinusoidal signal : A Continuous-time Sinusoidal signal is given by

$$x(t) = A \sin(\Omega t + \theta)$$

A — Amplitude

$\Omega \rightarrow$ frequency in radians per second

$\theta \rightarrow$ is the phase angle in radians

10(b) List any four applications of FM System. (4M)

Ans: 1) It is mostly used in radio broadcasting.

2) It is used in radar, telemetry, Seismic prospecting.

3) It is used in music synthesis as well as in Video-transmission intstruments.

4) It is used medical applications like EEG.

11(a) Define Amplitude Modulation. Derive an expression for the AM wave.
[Def: 2M + Expression 4M = 6M]

Ans: Amplitude modulation is a process of Changing the amplitude of the high frequency analog Carrier in accordance with the amplitude of the message signal.

Expression for AM Wave

$$m(t) = A_m \cos 2\pi f_m t$$

$$c(t) = A_c \cos 2\pi f_c t$$

$$s(t) = A_c \cos 2\pi f_c t + A_c k_a m(t) \cos 2\pi f_c t$$

$$s(t) = A_c [1 + k_a m(t)] \cos 2\pi f_c t$$ $\longrightarrow$ time domain equation of AM wave

$$s(t) = A_c[1 + k_a A_m \cos 2\pi f_m t] \cos 2\pi f_c t$$

$\mu = k_a A_m$ modulation index

$$\therefore s(t) = A_c[1 + \mu \cos 2\pi f_m t] \cos 2\pi f_c t$$

$$= A_c \cos 2\pi f_c t + A_c \mu \cos 2\pi f_c t \cos 2\pi f_m t$$

$$= \underbrace{A_c \cos 2\pi f_c t}_{\text{Carrier}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t}_{\text{USB}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t}_{\text{LSB}}$$

11 b) Write about an Voltage distribution. (4M)

Ans

$$P_t = P_c + P_{USB} + P_{LSB}$$

$$P = I^2 R$$

$$s(t) = A_c \cos 2\pi f_c t + \frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t + \frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t$$

$$P_c = \frac{(A_c/\sqrt{2})^2}{R} \quad P_{USB} = \frac{(A_c \mu/2\sqrt{2})^2}{R} \quad P_{LSB} = \frac{(A_c \mu/2\sqrt{2})^2}{R}$$

$$\therefore P_t = \frac{A_c^2}{2R} + \frac{A_c^2 \mu^2}{8R} + \frac{A_c^2 \mu^2}{8R}$$

$[\because R = 1\Omega]$

$$P_t = \frac{A_c^2}{2}\left[1 + \frac{\mu^2}{2}\right] = P_c\left[1 + \frac{\mu^2}{2}\right]$$

12(a) State and prove Sampling theorem. [Statement 2M + proof 4M = 6M]

Ans: Statement : A Bandlimited signal of finite energy which has no frequency components higher than f_m Hz may be completely recovered from the knowledge of its samples taken at the rate of $2f_m$ samples per second.

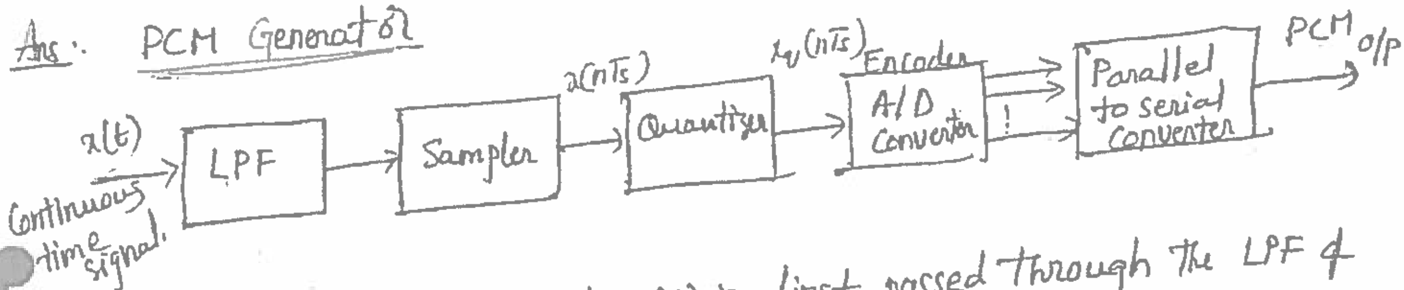
code words (combination of 1's and 0's) for each distinct symbol there is an unique code word. At receiver side source decoder is used it performs opposite operation of source encoder.

Channel Encoder and Decoder: The input of the channel encoder is Binary sequence. The communication channel adds noise and interference to the signal being transmitted. Hence errors are introduced in the binary sequence received at the receiver end. Thus, channel coding is done to avoid this type of errors.

Digital Modulators and Demodulators: After converting into binary information the pulses are to be transmitted by using digital modulation techniques like ASK, FSK, PSK ... etc. depends on application requirement.

13(b) With a neat diagram explain the Generation of PCM & DPCM.
[PCM 3M + DPCM 3M = 6M]

Ans: PCM Generator



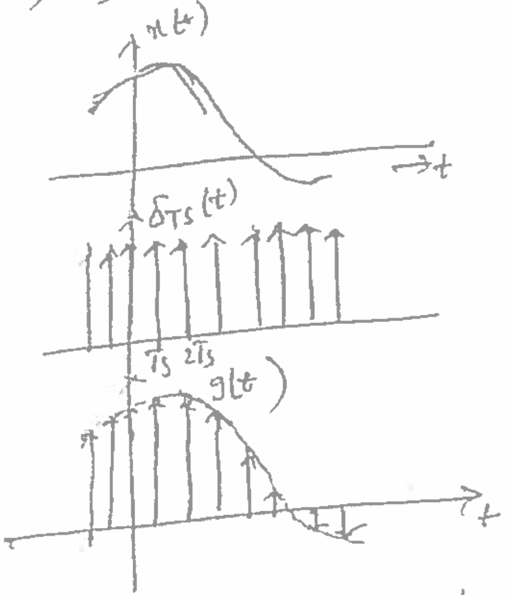
In PCM generator the signal x(t) is first passed through the LPF of cut-off frequency fm Hz. This LPF blocks all the frequency component which are lying above fm Hz. Samples this signal at the rate of fs. The o/p of sampler is denoted by x(nTs). A Quantizer compares input x(nTs) with its fixed digital levels. It assigns any one of the digital to x(nTs) with its fixed digital levels. The o/p of quantizer given to the I/p of encoder. This encoder converts input signal to 'v' digits binary word. Encoder o/p is given to the parallel to serial converter it converts parllel data into serial data it is suitable to transmission through channel.

Proof of Sampling theorem :

Let $x(t)$ is a Continuous signal, with maximum frequency f_m &

The Sampling of $x(t)$ at a rate f_s Hz, may be achieved by multiplying $x(t)$ by an impulse train $\delta_{T_s}(t)$

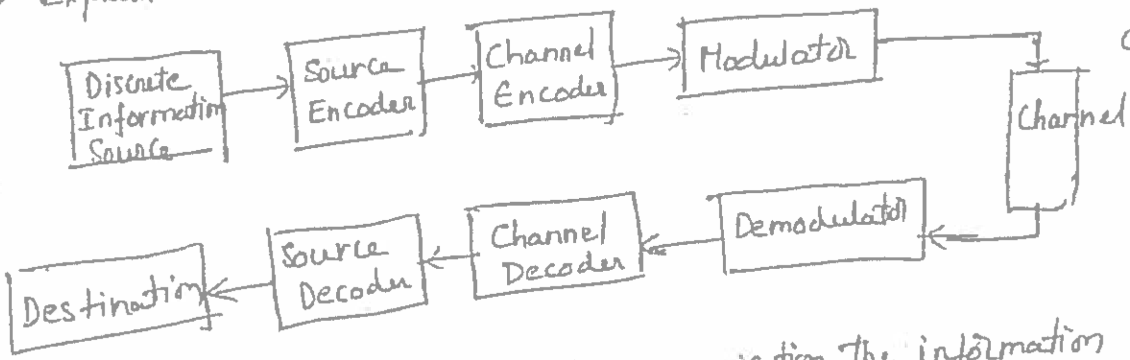
$\delta_{T_s}(t) \rightarrow$ impulse train Consist of unit impulses repeating periodically every T_s seconds where $T_s = 1/f_s$



13 (a) Explain the Basic elements of Digital Communication system ?.

diagram 2M
Content 4M } 6M

Ans

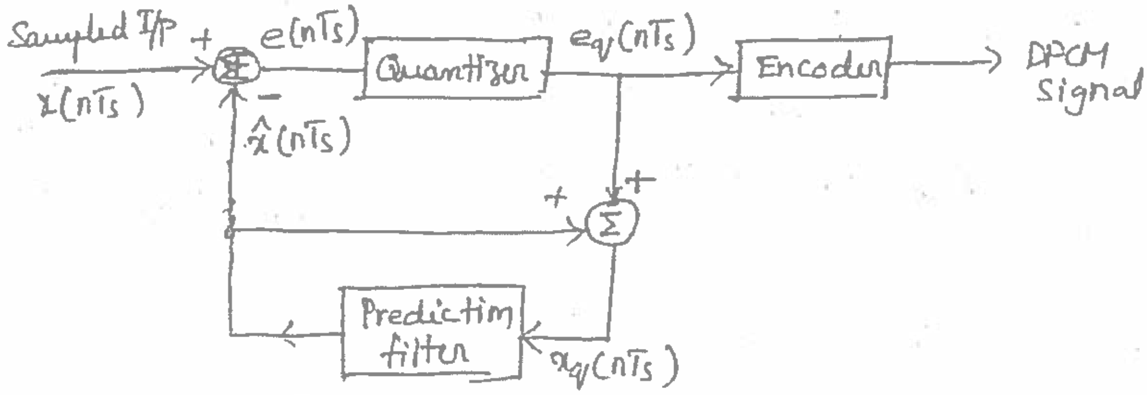


Discrete Information Source : In Digital Communication the information in Discrete w.r.t time. This information is obtained by process of sampling and Quantization. So the Discrete Information Source Can be letters, digits, Special Characters, code words . . .

Source Encoder and Decoder : The Symbols produced by The information Source are given to the Source encoder. The Symbols Can not be transmitted directly Source encoder Converts symbols into group of bits called

DPCM Generation :

The DPCM works on the principle of prediction. The value of the present sample is predicted from the past samples.



The sampled signal is denoted by $x(nT_s)$ and the predicted signal is denoted by $\hat{x}(nT_s)$. The Comparator finds out the difference between the actual sample value $x(nT_s)$ and predicted sample value $\hat{x}(nT_s)$. This is known as prediction error and it is denoted by $e(nT_s)$.

~~Thus, error is~~ The predicted value is produced by using a prediction filter. The Quantizer o/p signal $e_q(nT_s)$ and previous prediction is added and given as Input to the prediction filter. This signal is called $x_q(nT_s)$. This makes the prediction more and more close to the actual sampled signal. The quantized error signal $e_q(nT_s)$ is very small and can be encoded by using smaller number of bits. Thus no. of bits per sample are reduced in DPCM.

12(b) Describe the Basic principles of PCM system and PCM transmitter. [Principles of PCM system 2M + PCM Tx 4M = 6M]

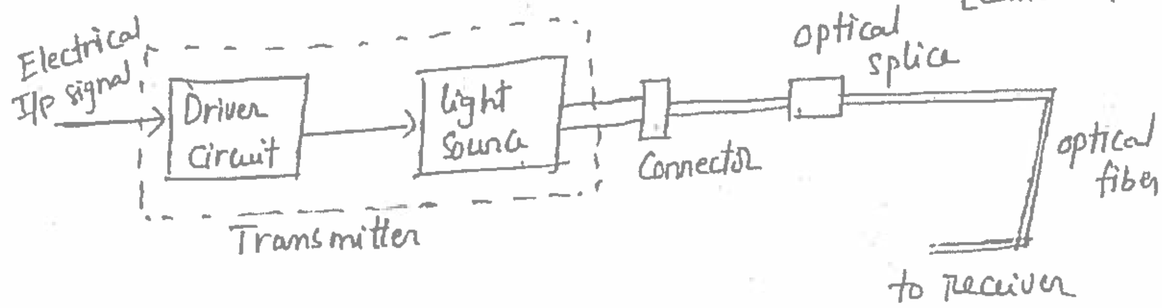
Ans Principles of PCM system

1) PCM is a digital pulse modulation system
2) PCM o/p is in the coded digital form.
3) PCM Consist of a PCM encoder and PCM decoder
4) PCM is not modulation in the Conventional sense.

PCM transmitter : refer 13 (b) Answer

14(a) Draw and Explain the working principle of an optical transmitter. [Content 4M + Diagram 2M = 6M]

Ans



Transmitter: The transmitter first converts the input voltage to current value which is used to drive the light source. Thus it interfaces the input circuit and light source.

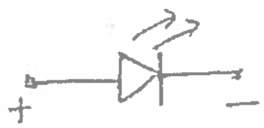
The light source is normally an infrared LED or LASER device which is driven by the current value from the V to I converter. It emits light which is proportional to the input voltage value is generated and given as input to the fiber.

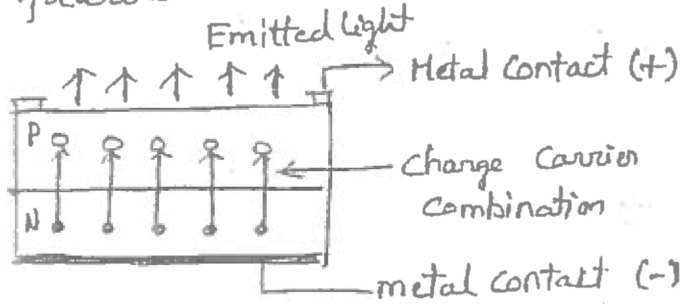
Optical Splice: for creating long haul communication link, it is necessary to join one fiber to other fibers permanently.

14(b) Explain about LED and its type. (content 4M + diagram 2M = 6M)

Ans: The Light Emitting Diode (LED) is a PN junction diode which emits light when forward biased, by a phenomenon called electroluminescence. In all semiconductor PN junctions, some of the energy will be radiated as heat and some in the form of photons. In Si and Ge the emitted light is insignificant. In other materials such as Gallium phosphide (GaP) or Gallium Arsenide phosphide (GaAsP), the number of photons of light energy emitted is sufficient to create a visible light source. Here the charge carrier recombination takes place when electrons from the N-side cross the junction and recombine with the holes on the P-side.

When LED is forward biased, The electrons and holes moves towards the junction and recombination takes place. As a result the e^- lying in The conduction bands of N-region fall into the holes lying in the VB of P-region. The difference of energy b/w The CB and VB is radiated in the form of light energy. The brightness of The emitted light is directly propotional to the forward current.





The color of the emitted light depends on the type of material used.

Gallium Arsenide (Ga As) ⟶ infrared radiation (invisible)

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Gallium Arsenide phosphide (Ga AsP) ⟶ red or yellow.

15(a) Explain the working principle of GSM? (GM)

Ans: Global System for Mobile Communication (GSM) is a digital mobile network that is widely used by mobile phone users in the world

The GSM. network has four Separate parts that work together to function as a whole. 1) Mobile Station (MS)

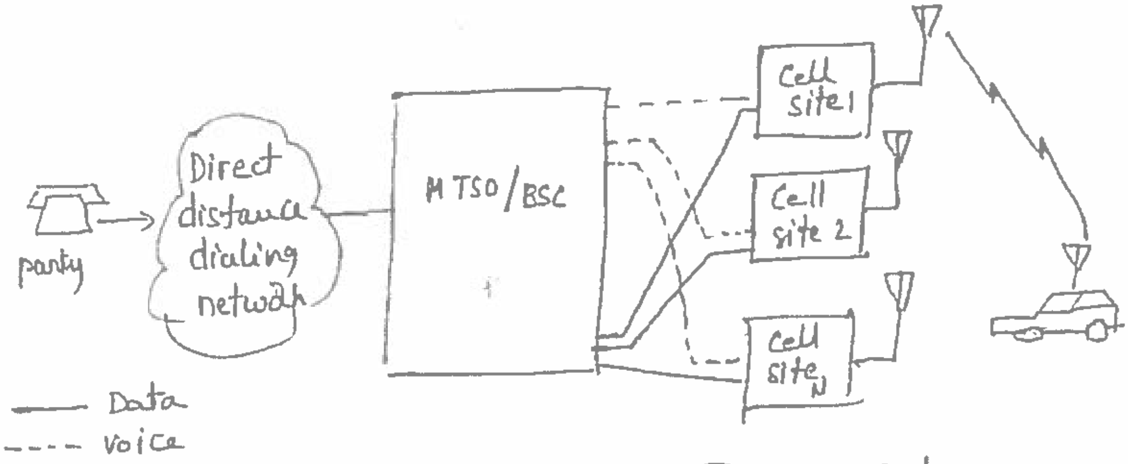
2) Base Station Subsystem (BSS)

3) Network Switching Subsystem (NSS)

4) Operation and support Subsystem (OSS)

15 (b) Explain Cellular Telephone System. [Diagram 3M + Content 3M = 6M]

Ans



A general View of Cellular Telephone System.

Antenna: Antenna pattern, antenna gain, antenna tilting and antenna height all affect the cellular system design. The antenna pattern can be omnidirectional, directional or any shape in both the vertical and the horizon planes. Antenna gain compensates for the transmitted power. Antenna gain at the mobile units would affect the system performance.

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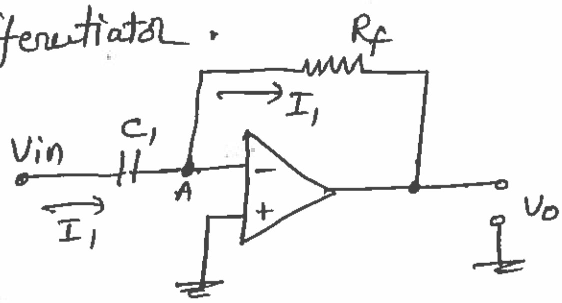
Data links: The data link are not directly affected by the cellular system, they are important in the system. Each data link can carry multiple channel data (10 kbps data transmitted per channel) from the cell site to the MTSO.

Differentiator: The circuit which produces the differentiation of the Input Voltage at its output is called Differentiator.



The node B is grounded. The node A is also at the ground potential hence $V_A = 0$

As I/p current of op-amp is zero, either current I_1 flows through the resistance R_f.

from the I/p side we can write

$$I_1 = C_1 \frac{d(V_{in} - V_A)}{dt} = C_1 \frac{dV_{in}}{dt} \quad \text{—— (1)}$$

from the o/p side

$$I = \frac{V_A - V_o}{R_f} = \frac{-V_o}{R_f} \quad \text{—— (2)}$$

Equating the two equations

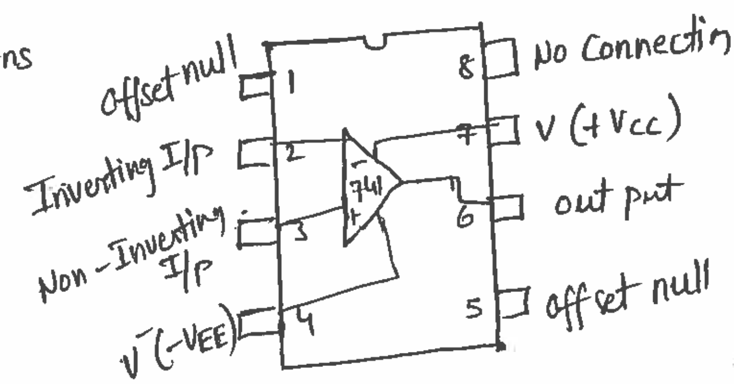
$$C_1 \frac{dV_{in}}{dt} = \frac{-V_o}{R_f}$$

$$\boxed{V_o = -C_1 R_f \frac{dV_{in}}{dt}}$$

The equation shows that the o/p is $C_1 R_f$ times the differentiation of the input and product $C_1 R_f$ is called time constant of the differentiator.

9(a) Draw and Explain the pin Diagram IC741 Op-amp. Pin Diagram – 3M, Description – 3M } 6M

Ans



Description of op-amp 741 IC pins

Pin 1 and 5: These two pins are used for offset process

Pin 2: Inverting I/p terminal, i.e when a sinusoidal signal is applied to the Input pin 2:

Pin 3: Non-inverting input terminal i.e when a sinusoidal signal is applied to the input pin 3, waveform of same phase o/p is obtained.

Pin 4: $-V_{cc}$, i.e negative terminal of supply voltage is connected to this pin

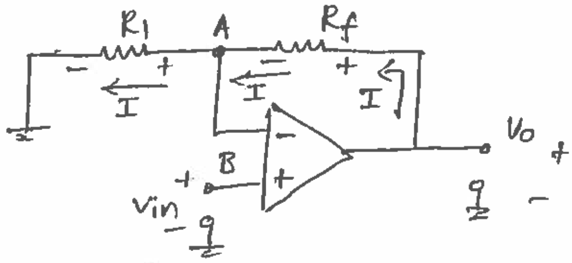
Pin 6: O/p terminal.

Pin 7: $+V_{cc}$ i.e +ve terminal of supply voltage is connected to this pin.

Pin 8: No electrical connection is there in this pin: this pin is just for balance and the symmetric dual-input package look.

9(b) Derive the gain for non-inverting op-amp. Diagram – 2M, Derivation – 4M } 6M

Ans An amplifier which amplifies the input without producing any phase shift b/w I/p and O/p is called non-inverting amplifier.



Derivation of closed loop gain:

The node B is at potential Vin, hence the potential of point A is same as B which is Vin, from the concept of virtual ground.

$$\therefore V_A = V_B = V_{in} \quad \text{——— (1)}$$

from the o/p side we can write

$$I = \frac{V_o - V_A}{R_f}$$

$$\therefore I = \frac{V_o - V_{in}}{R_f} \quad \text{——— (2)}$$

At the inverting terminal

$$I = \frac{V_A - 0}{R_1}$$

$$\therefore I = \frac{V_{in}}{R_1} \quad \text{——— (3)}$$

equating 2 and 3

$$\therefore \frac{V_0 - V_{in}}{R_f} = \frac{V_{in}}{R_1}$$

$$\therefore \frac{V_0}{R_f} = \frac{V_{in}}{R_f} + \frac{V_{in}}{R_1} = V_{in}\left[\frac{R_1 + R_f}{R_1 R_f}\right]$$

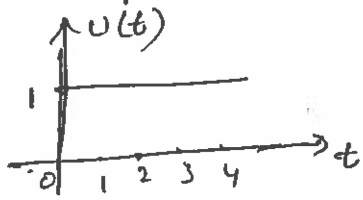
$$\frac{V_0}{V_{in}} = \frac{R_f(R_1 + R_f)}{R_1 R_f} = \frac{R_1 + R_f}{R_1} =$$

$$\therefore A_{VF} = \frac{V_0}{V_{in}} = 1 + \frac{R_f}{R_1}$$

10(a) State and Explain the properties of Continuous signals.

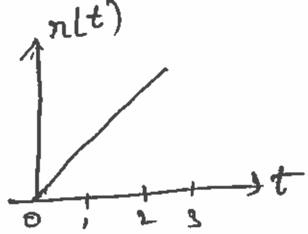
Ans: Unit step signal: The Unit step function is defined as

$$u(t) = 1 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Unit Ramp Signal: The Unit ramp function is defined as

$$r(t) = t \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



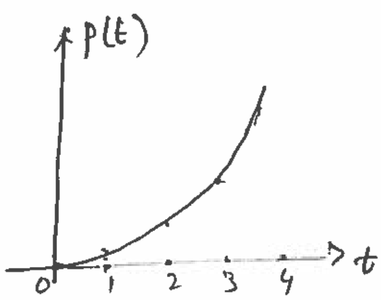
or $r(t) = t\,u(t)$

The ramp function can be obtained by Integrating the Unit step funct

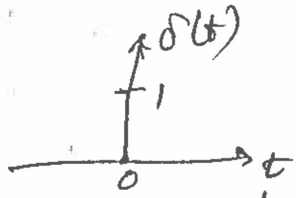
$$r(t) = \int u(t)\,dt = \int dt = t$$

Unit parabolic:

$$p(t) = t^2/2 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Impulse signal : $\delta(t) = 0$ for $t \neq 0$



Sinusoidal signal : A Continuous-time sinusoidal signal is given by

$$x(t) = A \sin(\Omega t + \theta)$$

A — Amplitude

$\Omega \rightarrow$ frequency in radians per second

$\theta \rightarrow$ is the phase angle in radians

10(b) List any four applications of FM System. (4M)

Ans: 1) It is mostly used in radio broadcasting.

2) It is used in radar, telemetry, seismic prospecting.

3) It is used in music synthesis as well as in Video-transmission intstruments.

4) It is used medical applications like EEG.

11(a) Define Amplitude Modulation. Derive an expression for the AM wave.
[Def: 2M + Expression 4M = 6M]

Ans: Amplitude modulation is a process of changing the amplitude of the high frequency analog carrier in accordance with the amplitude of the message signal.

Expression for AM Wave

$$m(t) = A_m \cos 2\pi f_m t$$

$$c(t) = A_c \cos 2\pi f_c t$$

$$s(t) = A_c \cos 2\pi f_c t + A_c k_a m(t) \cos 2\pi f_c t$$

$$s(t) = A_c [1 + k_a m(t)] \cos 2\pi f_c t$$ $\longrightarrow$ time domain equation of AM wave

$$s(t) = A_c\left[1 + k_a A_m \cos 2\pi f_m t\right] \cos 2\pi f_c t$$

$\mu = k_a A_m$ modulation index

$$\therefore \; s(t) = A_c\left[1 + \mu \cos 2\pi f_m t\right] \cos 2\pi f_c t$$

$$= A_c \cos 2\pi f_c t + A_c \mu \cos 2\pi f_c t \cos 2\pi f_m t$$

$$= \underbrace{A_c \cos 2\pi f_c t}_{\text{Carrier}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t}_{\text{USB}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t}_{\text{LSB}}$$

11 b) Write aboutan Voltage distribution. (4M)

$$P_t = P_c + P_{USB} + P_{LSB}$$

$$P = I^2 R$$

$$s(t) = A_c \cos 2\pi f_c t + \frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t + \frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t$$

$$P_c = \frac{(A_c/\sqrt{2})^2}{R}, \quad P_{USB} = \frac{(A_c \mu / 2\sqrt{2})^2}{R} \quad P_{LSB} = \frac{(A_c \mu / 2\sqrt{2})^2}{R}$$

$$\therefore \; P_t = \frac{A_c^2}{2R} + \frac{A_c^2 \mu^2}{8R} + \frac{A_c^2 \mu^2}{8R}$$

$[\because R = 1\Omega]$

$$P_t = \frac{A_c^2}{2}\left[1 + \frac{\mu^2}{2}\right] = P_c\left[1 + \frac{\mu^2}{2}\right]$$

12(a) State and prove Sampling theorem. [Statement 2M + proof 4M = 6M]

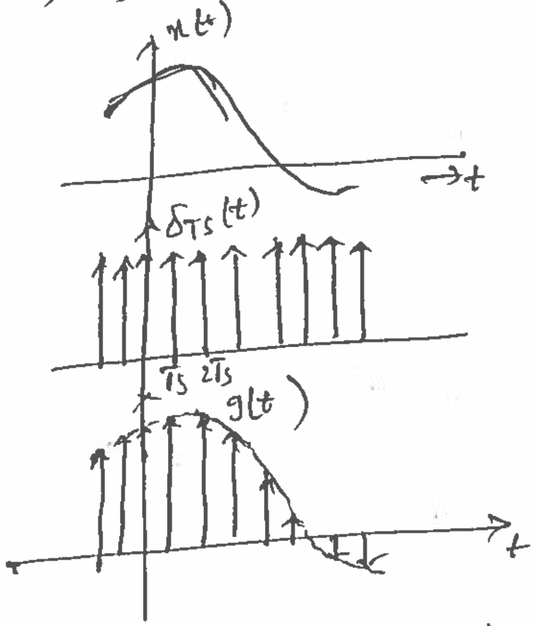
Ans: Statement : A Bandlimited signal of finite energy which has no frequency Components higher than f_m Hz may be Completely recovered from the knowledge of its samples taken at the rate of $2f_m$ samples per Second.

Proof of Sampling theorem :

Let $x(t)$ is a Continuous Signal, with maximum frequency f_m &

The Sampling of $x(t)$ at a rate f_s Hz, may be achieved by multiplying $x(t)$ by an Impulse train $\delta_{Ts}(t)$

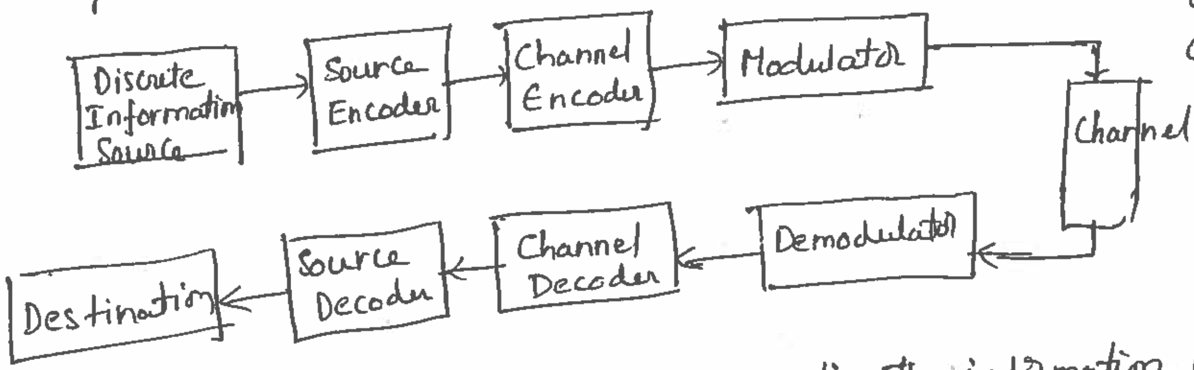
$\delta_{Ts}(t) \rightarrow$ impulse train Consist of unit impulses repeating periodically every T_s seconds where $T_s = 1/f_s$



13 (a) Explain the Basic elements of Digital Communication system ?

diagram 2M
Content 4M } 6M.

Ans



Discrete Information Source : In Digital Communication the information in Discrete w.r.t time. This information is obtained by process of Sampling and Quantization. So the Discrete information Source Can be letters, digits, Special Characters, Code words...

Source Encoder and Decoder: The Symbols produced by the information Source are given to the Source encoder. The Symbols Can not be transmitted directly Source encoder Converts symbols into group of bits called

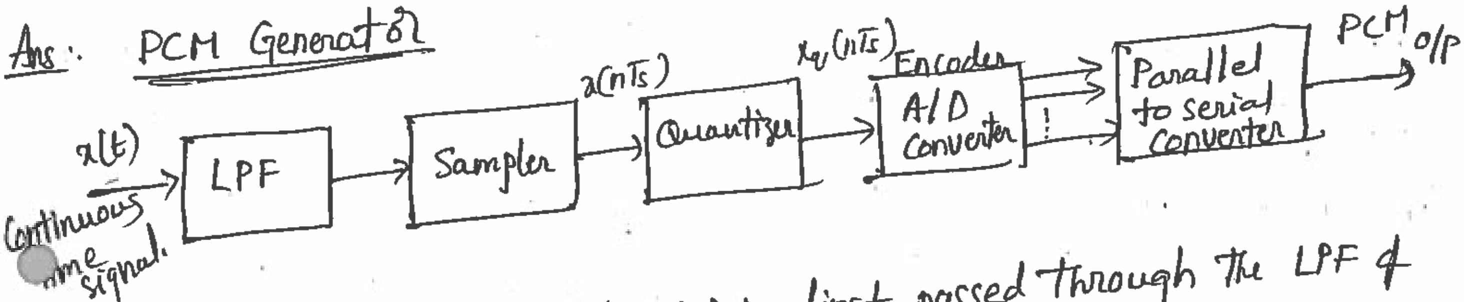
Code words (combination of 1's and 0's) for each distinct symbol there is an unique code word. At receiver side ~~channel~~ Source decoder is used it performs opposite operation of ~~channel~~ Source encoder.

<u>Channel Encoder and Decoder</u>: The input of the Channel encoder is Binary Sequence. The communication channel adds noise and interference to the signal being transmitted. Hence errors are introduced in the binary sequence received at the receiver end. Thus channel coding is done to avoid this type of errors.

<u>Digital Modulators and Demodulators</u>: After converting into binary information the pulses are to be transmitted by using digital modulation techniques like ASK, FSK, PSK ... etc. depends on application Requirement.

13(b) With a neat diagram explain the Generation of PCM & DPCM.
[PCM 3M + DPCM 3M = 6M]

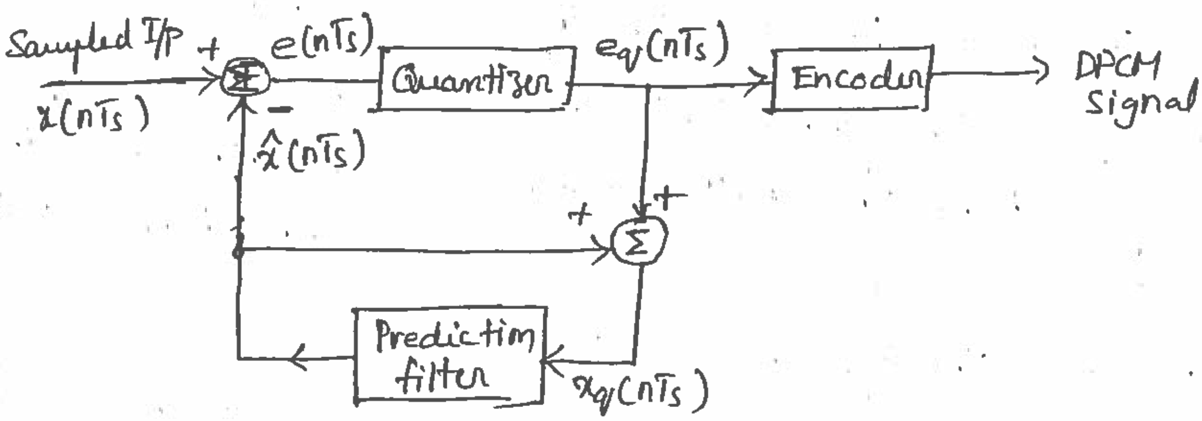
<u>Ans</u>: <u>PCM Generator</u>



In PCM generator the signal x(t) is first passed through the LPF of cut-off frequency fm Hz. This LPF blocks all the frequency component which are lying above fm Hz. Samples this signal at the rate of fs. The o/p of Sampler is denoted by x(nTs). A Quantizer compares input x(nTs) with its fixed digital levels. It assigns any one of the digital to x(nTs) with its fixed digital levels. The o/p of quantizer given to the I/p of encoder. This encoder converts input signal to 'v' digits binary word. encoder o/p is given to the parallel to serial converter it converts parllel data in to serial data it is suitable to transmission through channel.

DPCM Generation :

The DPCM works on the principle of prediction. The value of the present sample is predicted from the past samples.



The sampled signal is denoted by $x(nT_s)$ and the predicted signal is denoted by $\hat{x}(nT_s)$. The Comparator finds out the difference between the actual sample value $x(nT_s)$ and predicted sample value $\hat{x}(nT_s)$. This is known as prediction error and it is denoted by $e(nT_s)$.

~~Thus, error is~~ The predicted value is produced by using a prediction filter. The Quantizer o/p signal $e_q(nT_s)$ and previous prediction is added and given as input to the prediction filter. This signal is called $x_q(nT_s)$. This makes the prediction more and more close to the actual sampled signal. The quantized error signal $e_q(nT_s)$ is very small and can be encoded by using smaller number of bits. Thus no. of bits per sample are reduced in DPCM.

12(b) Describe the Basic principles of PCM system and PCM transmitter.

[Principles of PCM system 2M + PCM Tx 4M = 6M]

Ans **Principles of PCM system**

1) PCM is a digital pulse modulation system
2) PCM o/p is in the coded digital form.
3) PCM consist of a PCM encoder and PCM decoder
4) PCM is not modulation in the conventional sense.

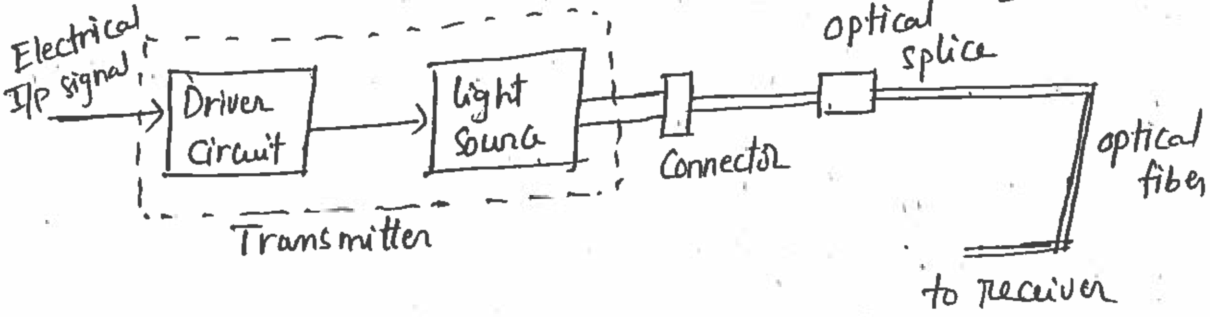
PCM transmitter : refer 13 (b) Answer

(8)

14(a) Draw and Explain The working principle of an optical transmitter.

[Content 4M + Diagram 2M = 6M]

Ans



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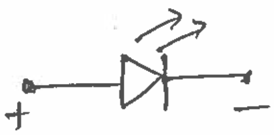

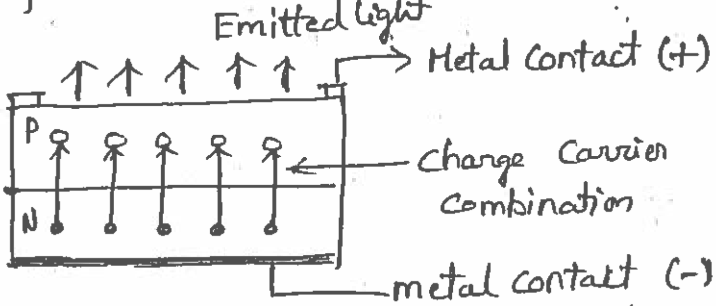

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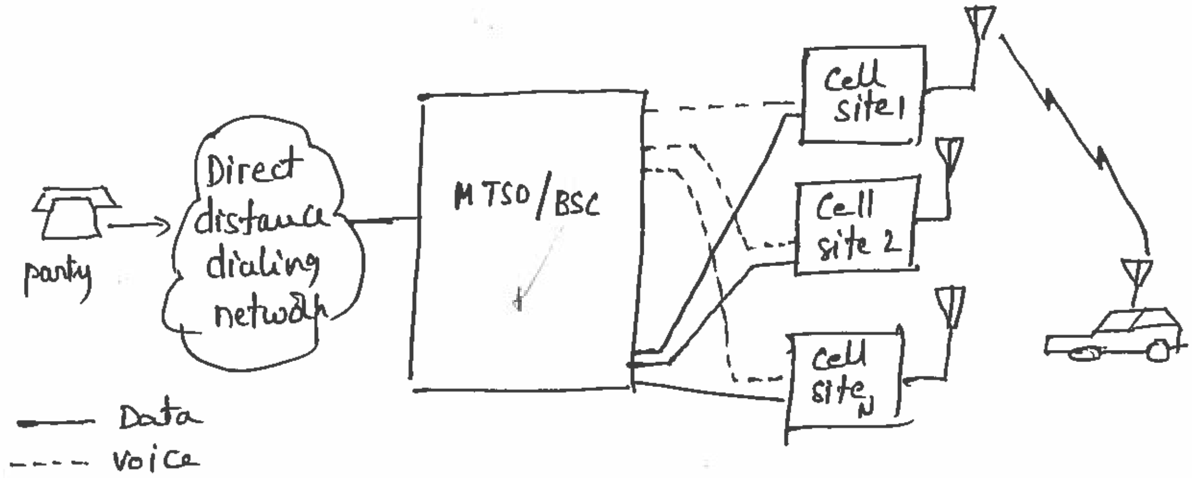
(4)

15 (b) Explain Cellular Telephone System. [Diagram 3M + Content 3M = 6M]

Ans



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NSRIT

Semester End Examination, Sept./Oct., 2021

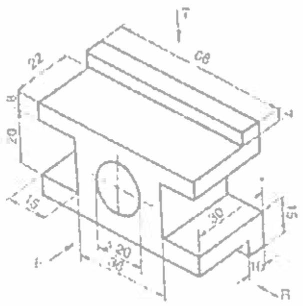
Degree	B. Tech. (U. G.)	**Program**	Common to EEE/ECE			**Academic Year**	2020 - 2021
Course Code	20ESX01	**Test Duration**	3 Hrs.	**Max. Marks**	70	**Semester**	II
Course	**ENGINEERING DRAWING**						

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

No.	Questions (1 through 2)	Learning Outcome (s)	DoK
1	Construct a scale to measure up to 50 m if 1cm represents 4 m, find its RF and mark a distance 37 m on it	20ESX01.2	L1
2	Draw a pentagon of side 30 mm	20ESX01.4	L3

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

No.	Questions (3 through 12)	Marks	Learning Outcome (s)	DoK
3 (a)	Draw a hyperbola having its two asymptotes passing through a point P at a distance of 30 mm from one asymptote and 36 mm from the other. Draw a normal and tangent at any convenient point	6M	20ESX01.1	L2
3 (b)	Construct a hexagon of side 25 mm by using general method	6M	20ESX01.1	L3
4 (a)	Draw the major axis of an ellipse is 110 mm long and the foci are at a distance of 15 mm from its ends. Draw the ellipse by concentric circles method	6M	20ESX01.1	L3
4 (b)	A 4 cm long line on map represents 1.5 metre length. Determine the RF and draw a scale long enough to measure up to 6 meters. Show a distance of 4.6 metres on it	6M	20ESX01.1	L2
5 (a)	A 70 mm long line PQ is inclined at 30° to the HP. The end P is 15 mm in front of the VP and 25 mm above the HP. Draw its projections	4M	20ESX01.2	L3
5 (b)	A line AB 75 mm long is inclined at 45° to the HP and 30° to VP. Its end A is in the HP and 40 mm infront of the VP. Draw its projections and determine traces	8M	20ESX01.2	L3
	OR			
6 (a)	Draw the following projection of points: I. A, 30 mm above HP and 20 mm infront of VP II. B, 20 mm above HP and 40 mm behind VP III. C, 20 mm below HP and 30 mm behind VP IV. D, is on both HP and VP	4M	20ESX01.2	L2
6 (b)	A 60 mm line AB, has an end P at 25 mm above the HP and 30 mm in front of VP. The line is inclined at 50° to HP and 40° to VP. Draw its projections	8M	20ESX01.2	L2
7 (a)	Draw the projections of a regular pentagon of 30 mm side, having one of its sides in the HP and its surface making an angle of 45° with the HP	6M	20ESX01.3	L2
7 (b)	Draw the projections of a circular lamina of 50 mm diameter having one of its sides in the VP and inclined at 30° to the VP	6M	20ESX01.3	L3
	OR			
8 (a)	Draw the projections of a 60° set square of 30 mm side and longer edge 120 mm one of its sides in the HP and its surface making an angle of 45° with the HP	6M	20ESX01.3	L2

8 (b)	Draw the projections of a regular hexagon of 30 mm side, having one of its sides in the HP and inclined at 60° to the V.P and its surface making an angle of 45° with the H.P	6M	20ESX01.3	L3
9 (a)	A square prism, side of base 30 mm and axis 50 mm long , has its axis inclined at 60° to HP its has an edge of its base in the HP and inclined at 45° to VP. Draw the projections	6M	20ESX01.4	L2
9 (b)	Draw the projection of a cone, base 75 mm diameter and axis 100 mm long, lying on HP. on one of its generators with axis parallel to the V.P	6M	20ESX01.4	L3
	OR			
10 (a)	A square prism, side of base 30 mm and axis 50 mm long , has its axis inclined at 60° to HP its has an edge of its base in the H.P and inclined at 45° to VP. Draw the projections	6M	20ESX01.4	L2
10 (b)	Draw the projections of a cone, base 65 mm diameter and axis 120 mm long, lying on the ground on one of its generators with the axis parallel to the VP	6M	20ESX01.4	L3
11	Draw top, front and side views of the isometric projection given in the figure 	12M	20ESX01.5	L4
	OR			
12	Draw an isometric view of a square prism having a base with a 40 mm side and a 60 mm long axis, resting on the HP. a) on its base with axis perpendicular to the HP, b) on its rectangular faces with axis perpendicular to the VP and c) on its rectangular face with axis parallel to the VP	12M	20ESX01.5	L4

Degree: B.Tech.

Scheme of Evaluation

AYear: 2020-21

Course Code: 20ESX01

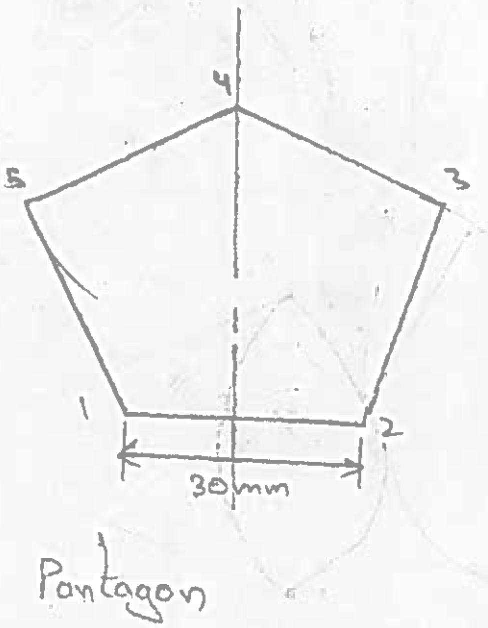
Course: Eng. Drawing

Year & Sem: I - II

Part-A

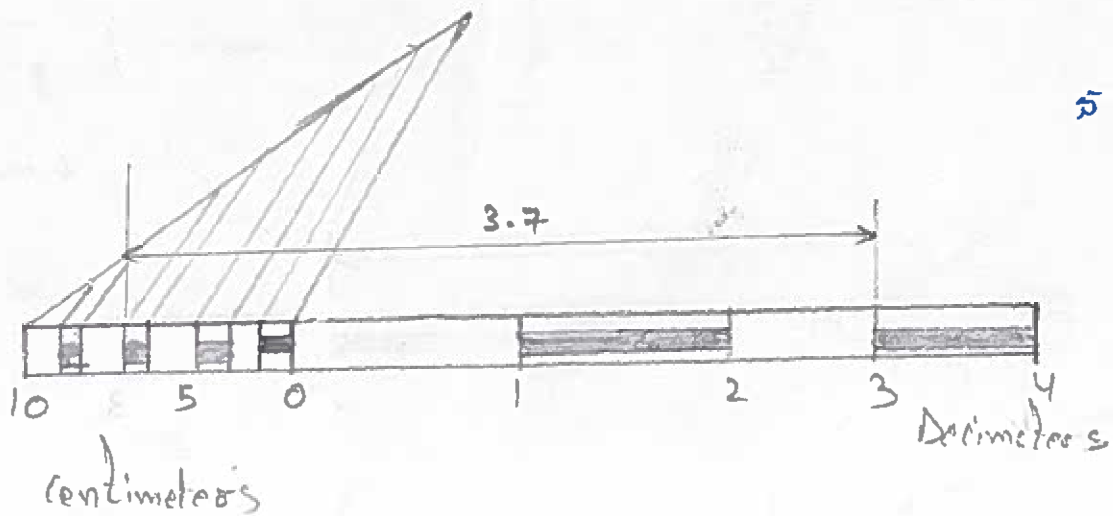
(2)

5 marks



(1)

5 marks



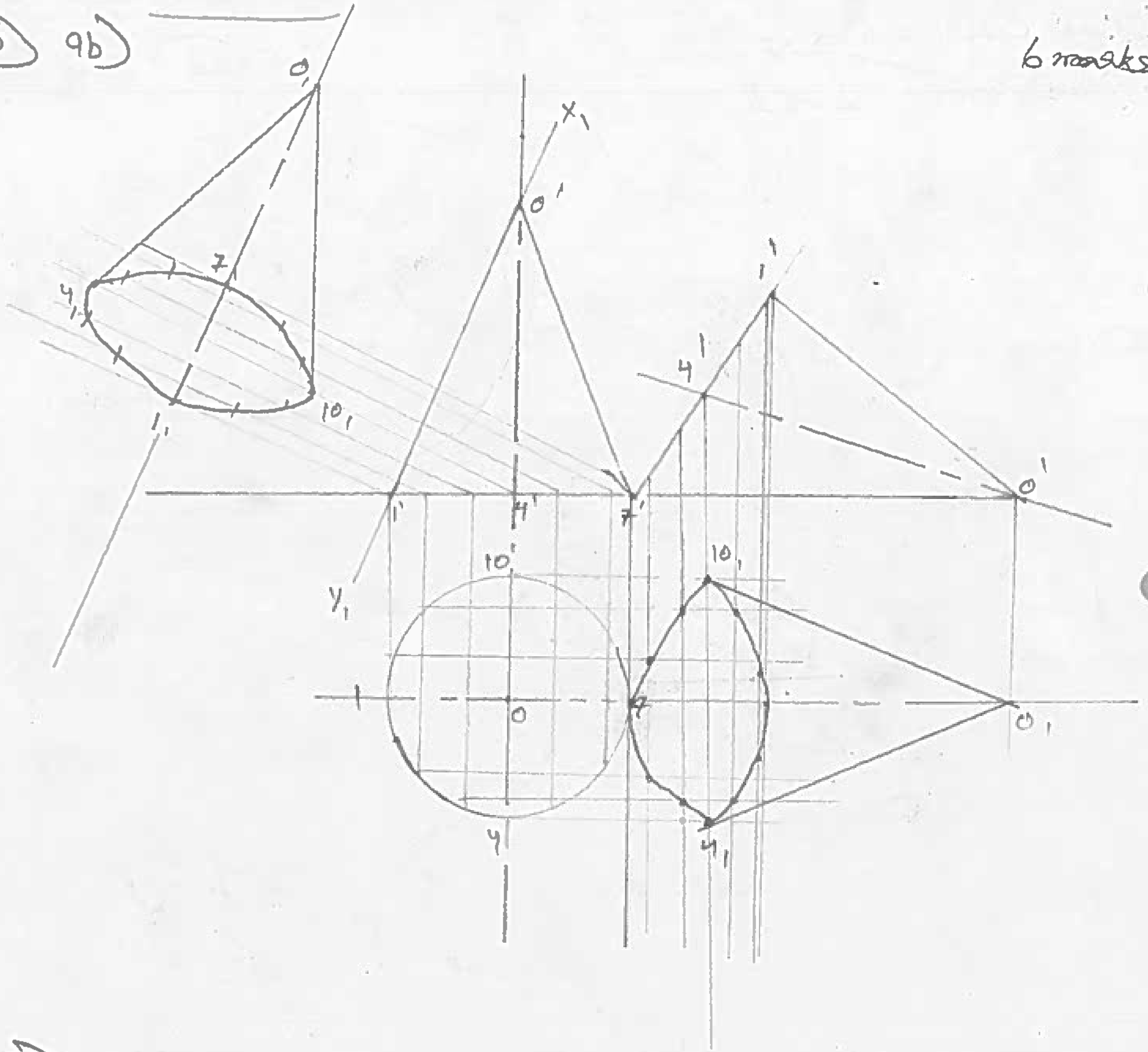
* Determine R.F of scale. Here it is $\frac{1}{4}$

* Length of Scale = R.F x max length $= \frac{1}{4} \times 50$
$= 12.5cm$

PART — B

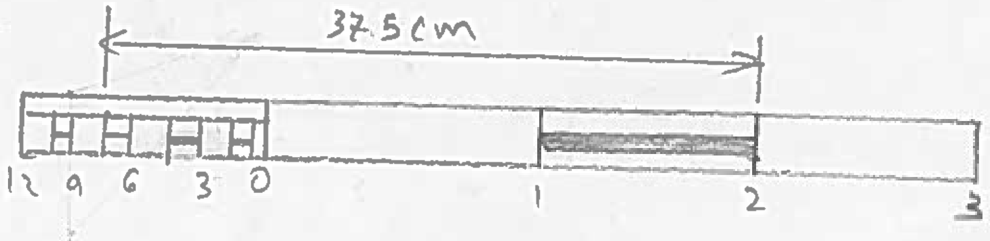
10b) 9b)

6 marks



4b)

6 mark

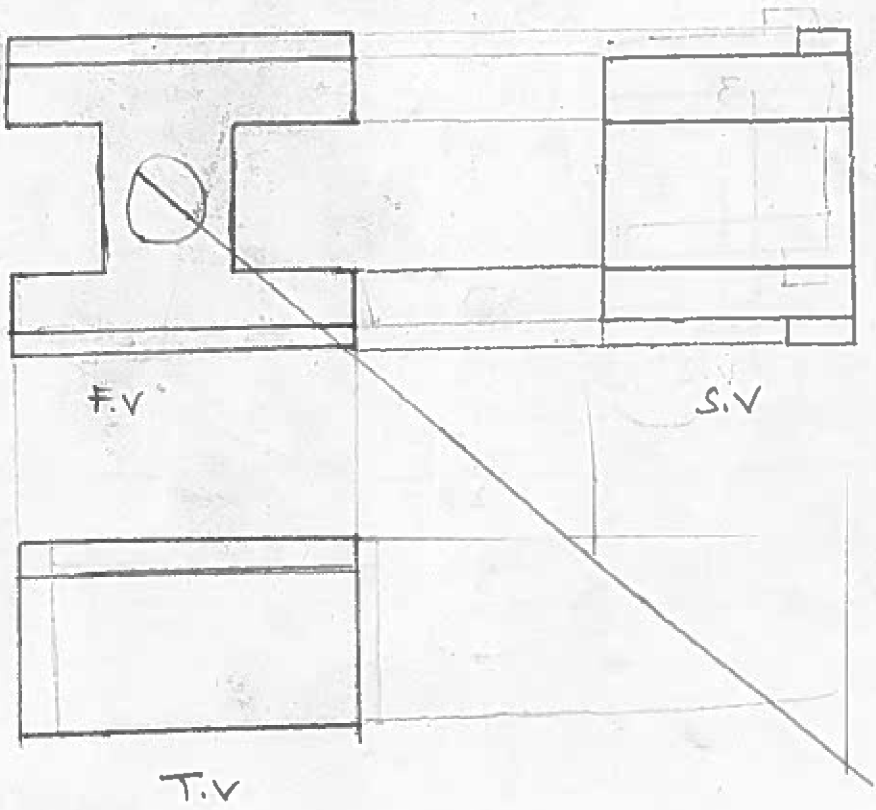


$$R.F = \frac{150}{4}$$

$$= 37.5\ cm$$

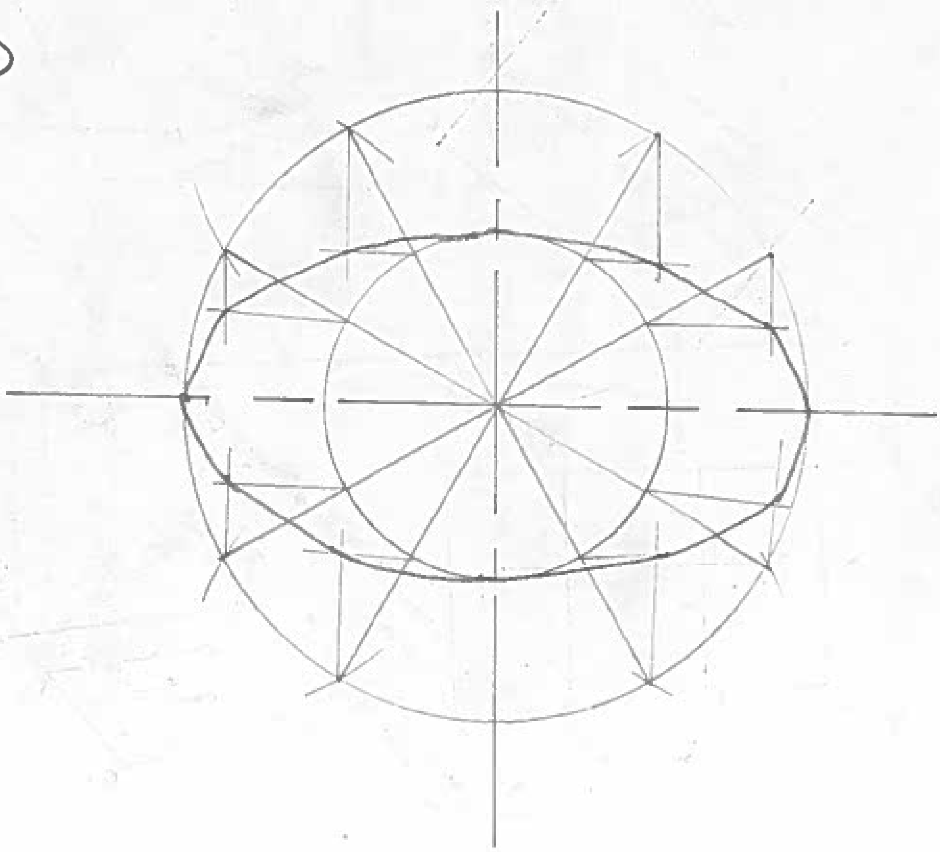
(11)

12 marks



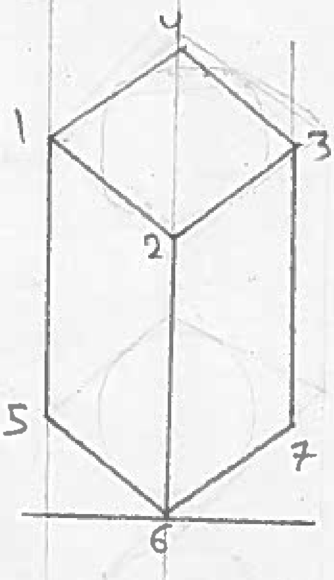
4a)

6 marks

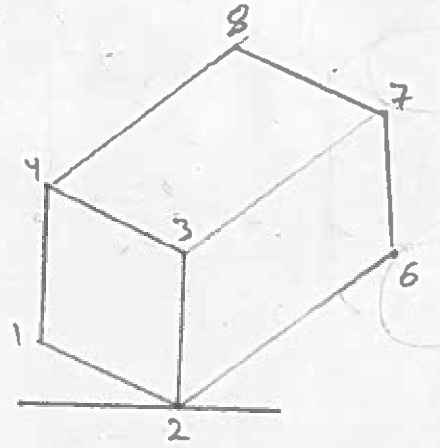


(12)

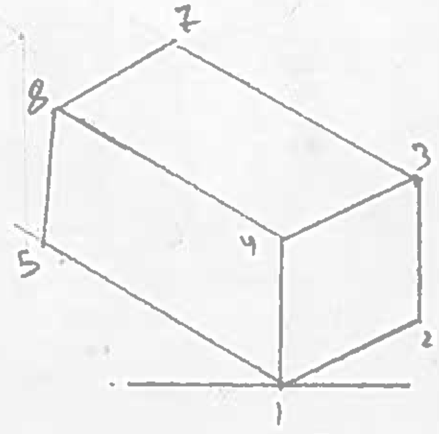
12 marks



(a)



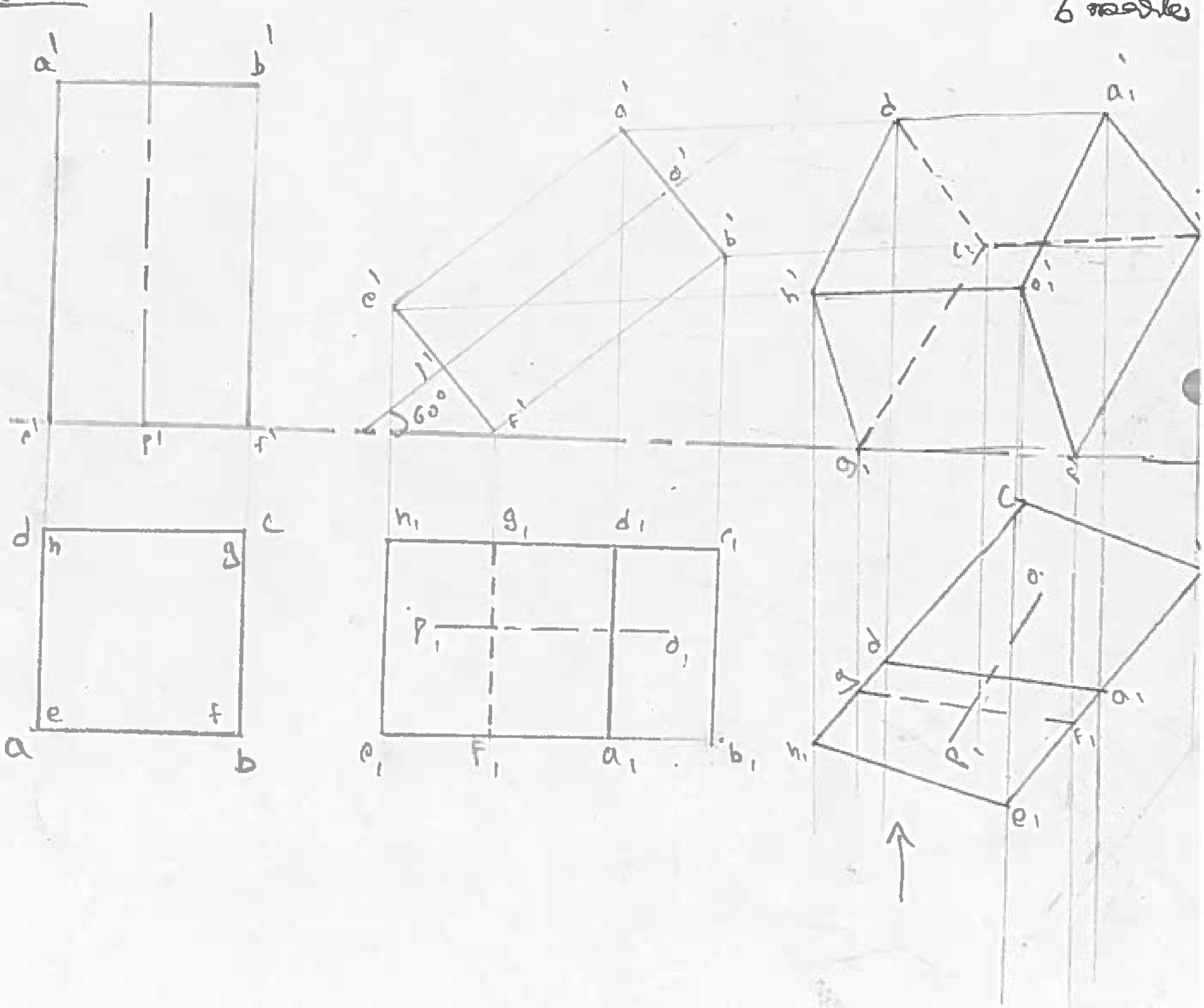
(b)



(c)

10 a) 9 a)

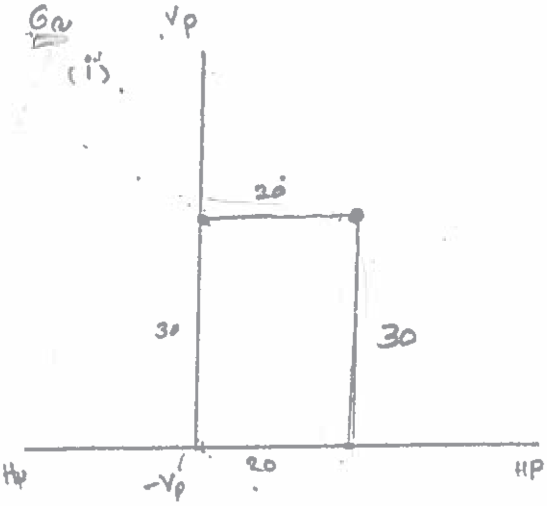
6 marks



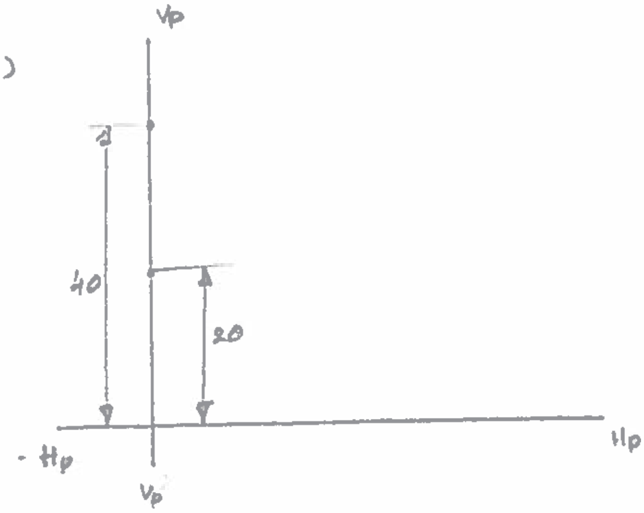
6a

4 marks

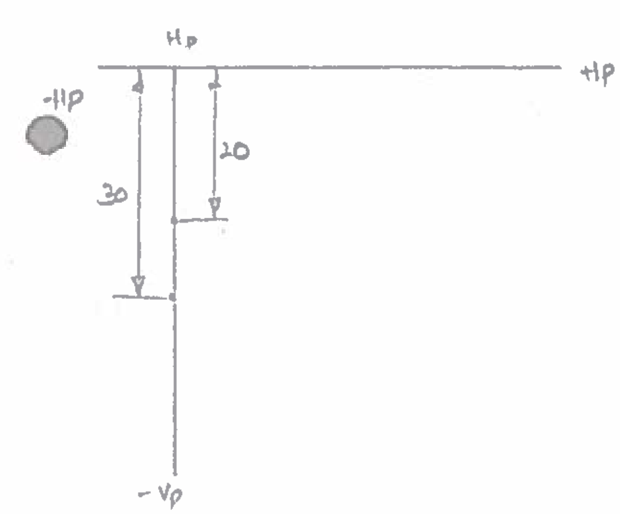
(i)



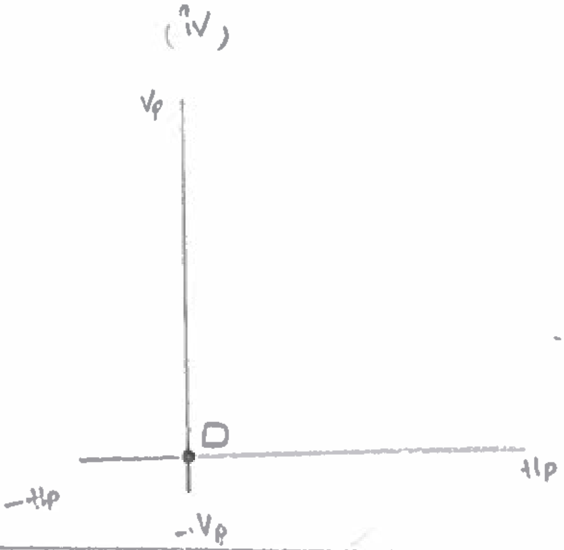
(ii)



(iii)

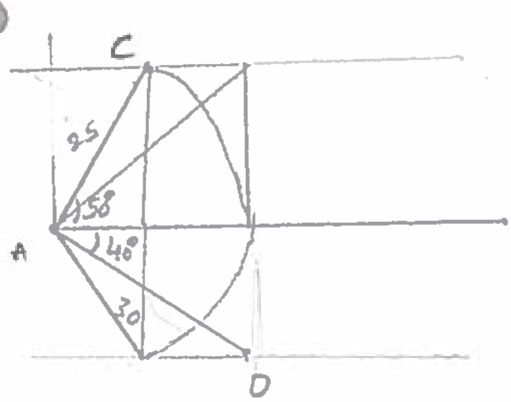


(iv)



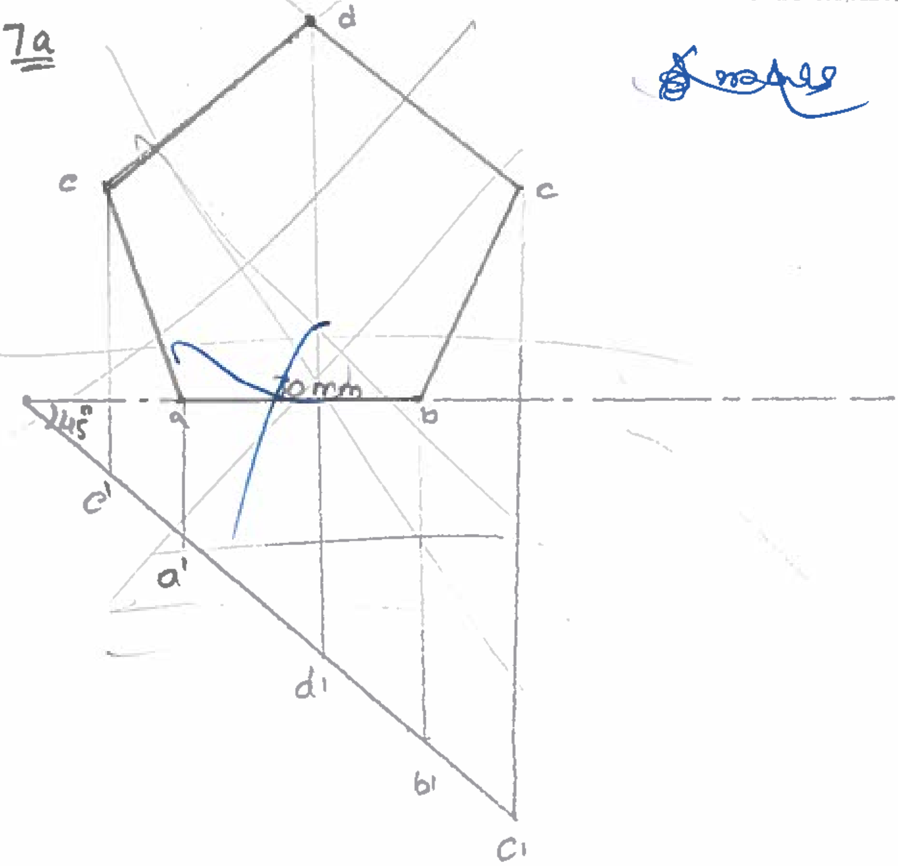
6b

8 marks



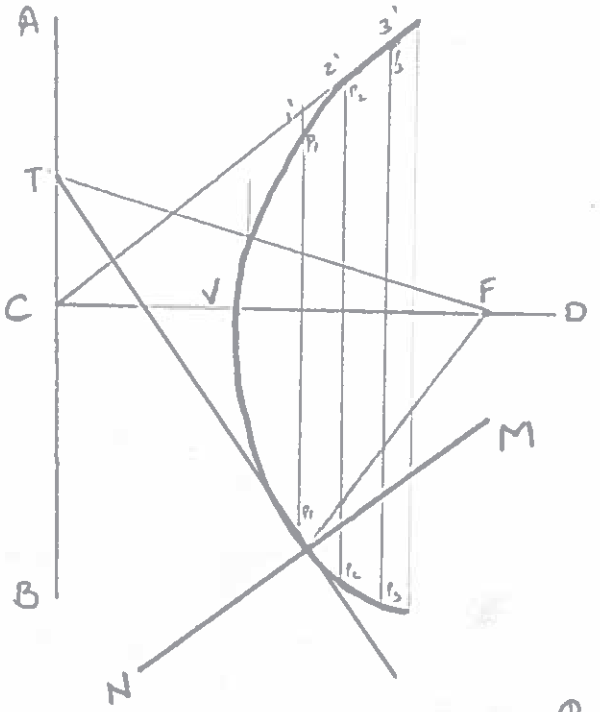
7a

8 marks



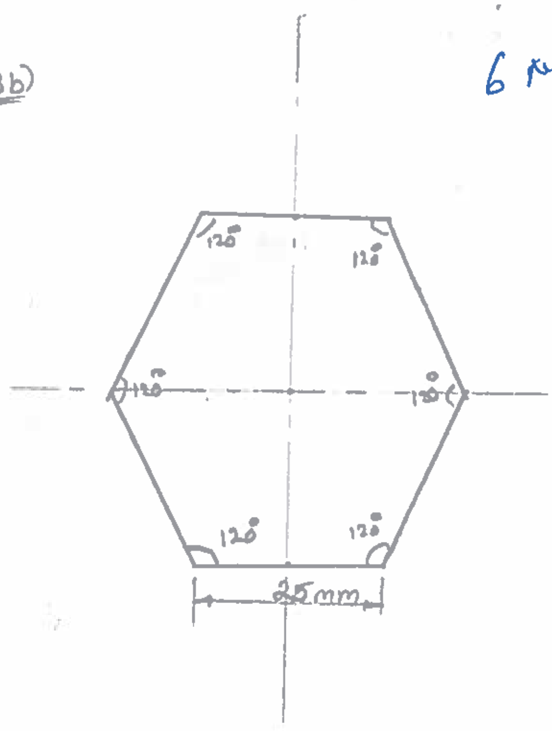
6 M

3(a)

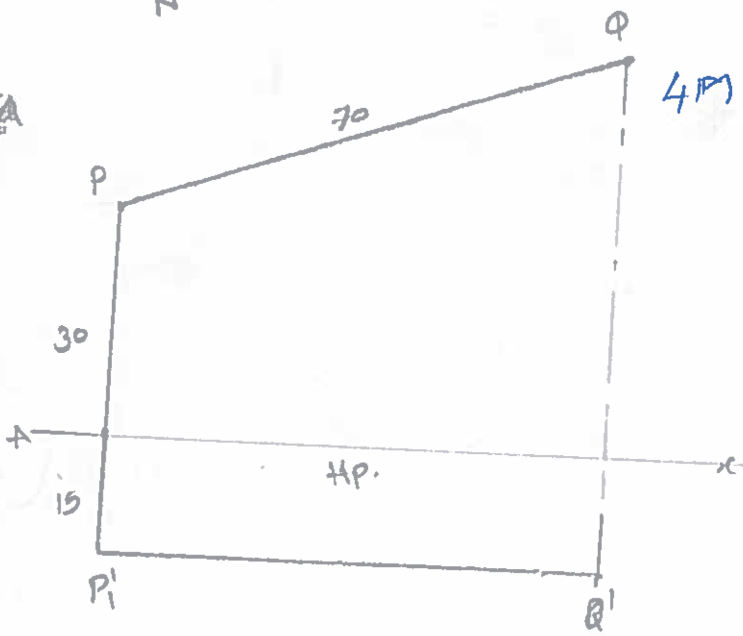


3b)

6 M



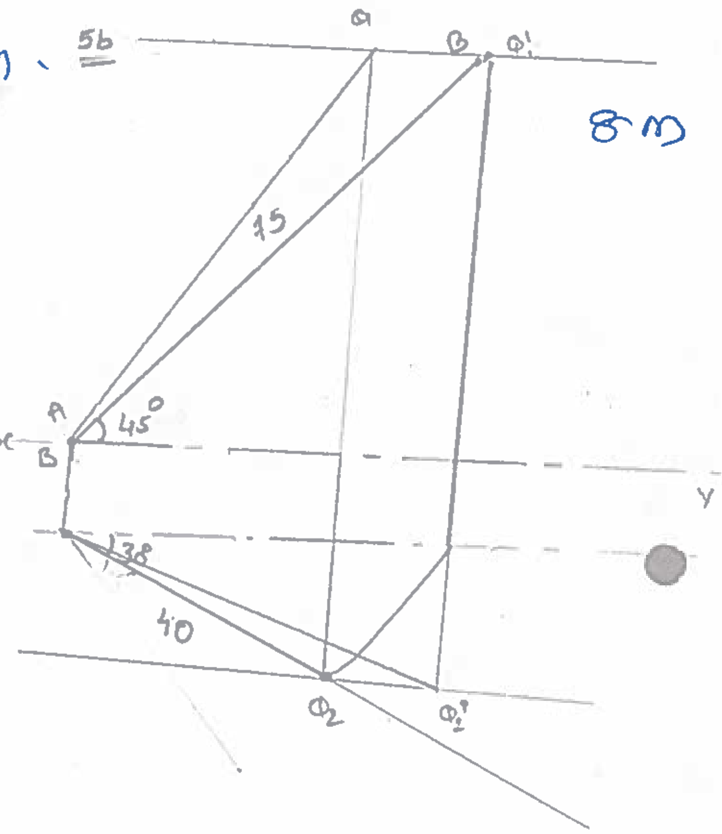
5A



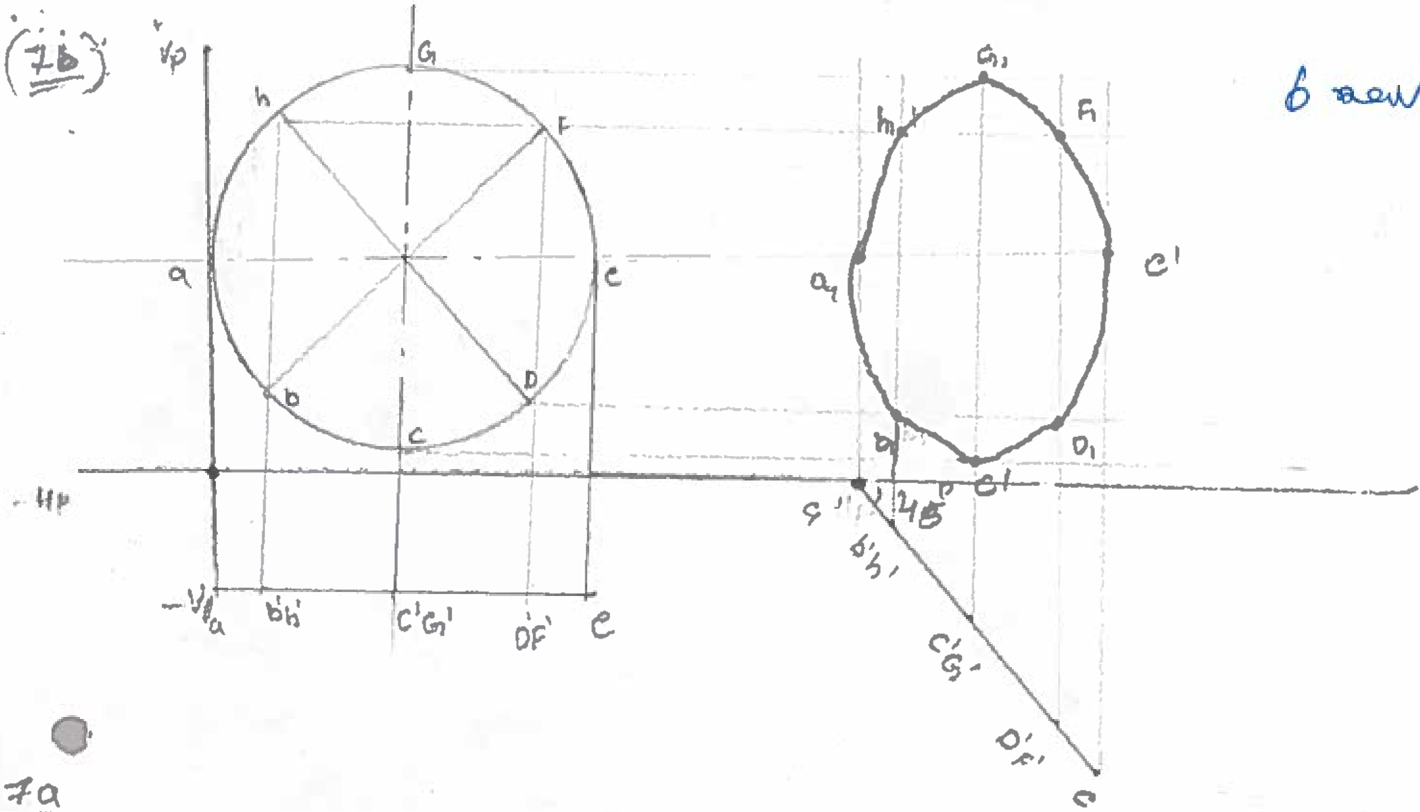
4M .

5b

8 M



(7b)



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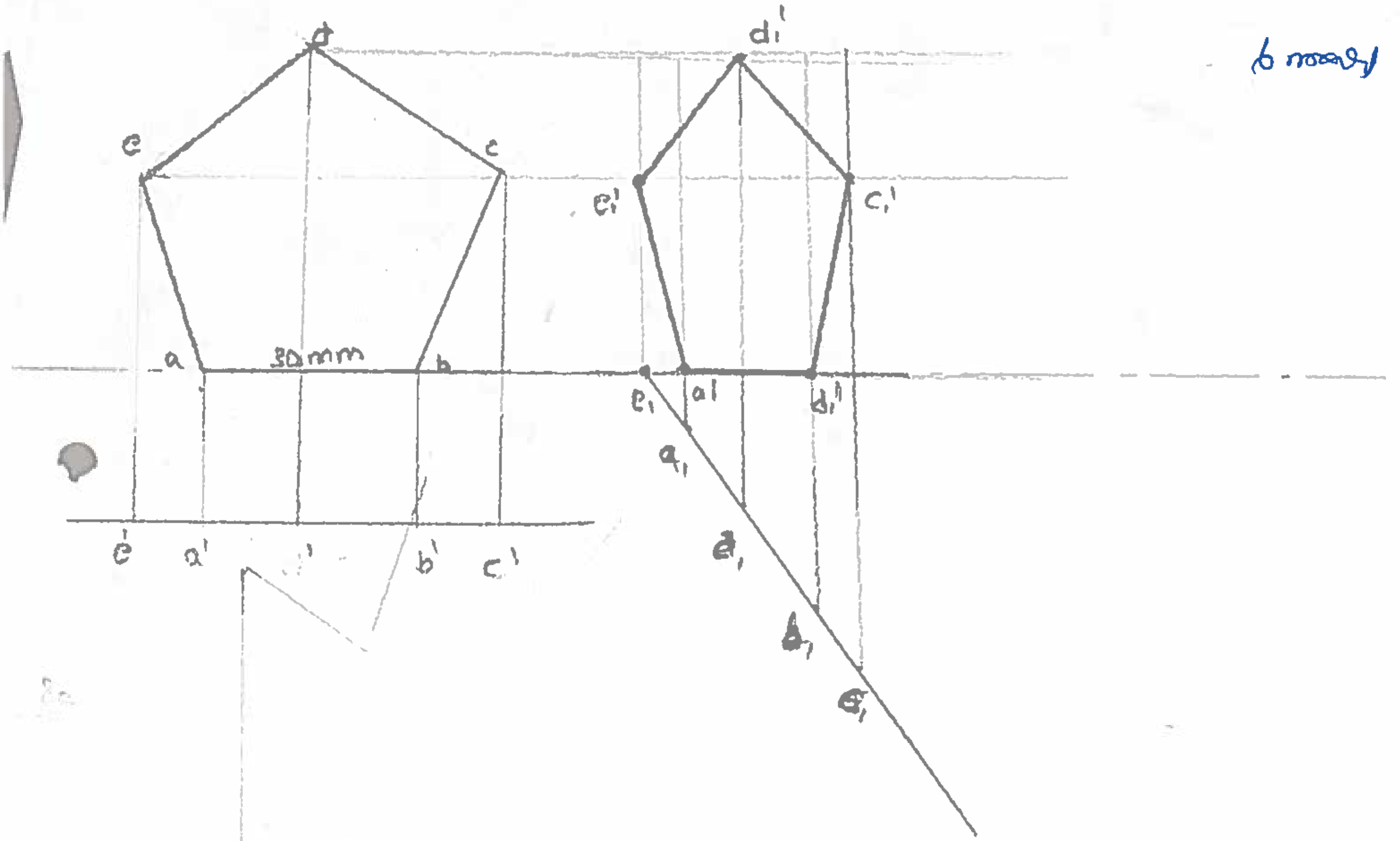

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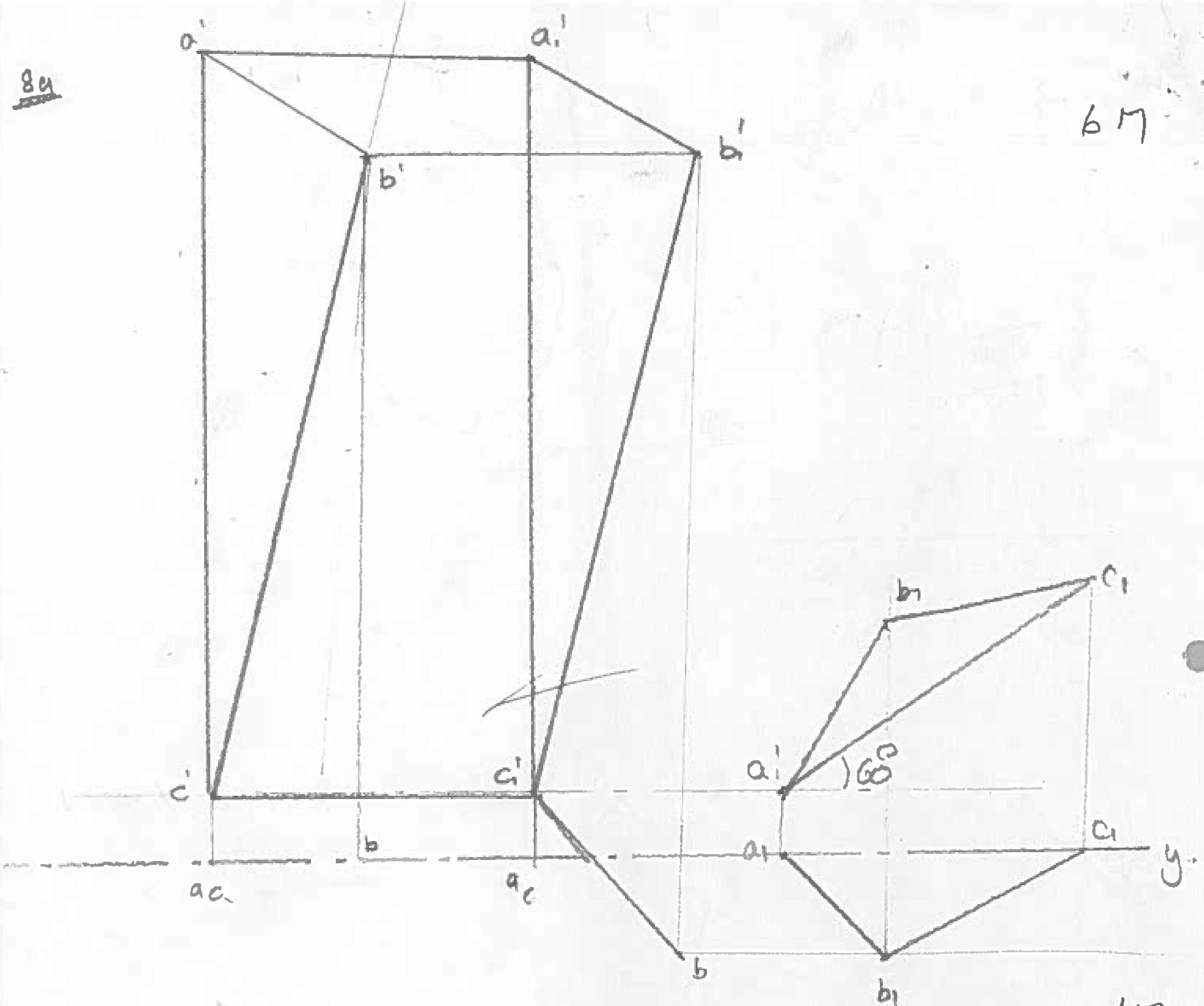

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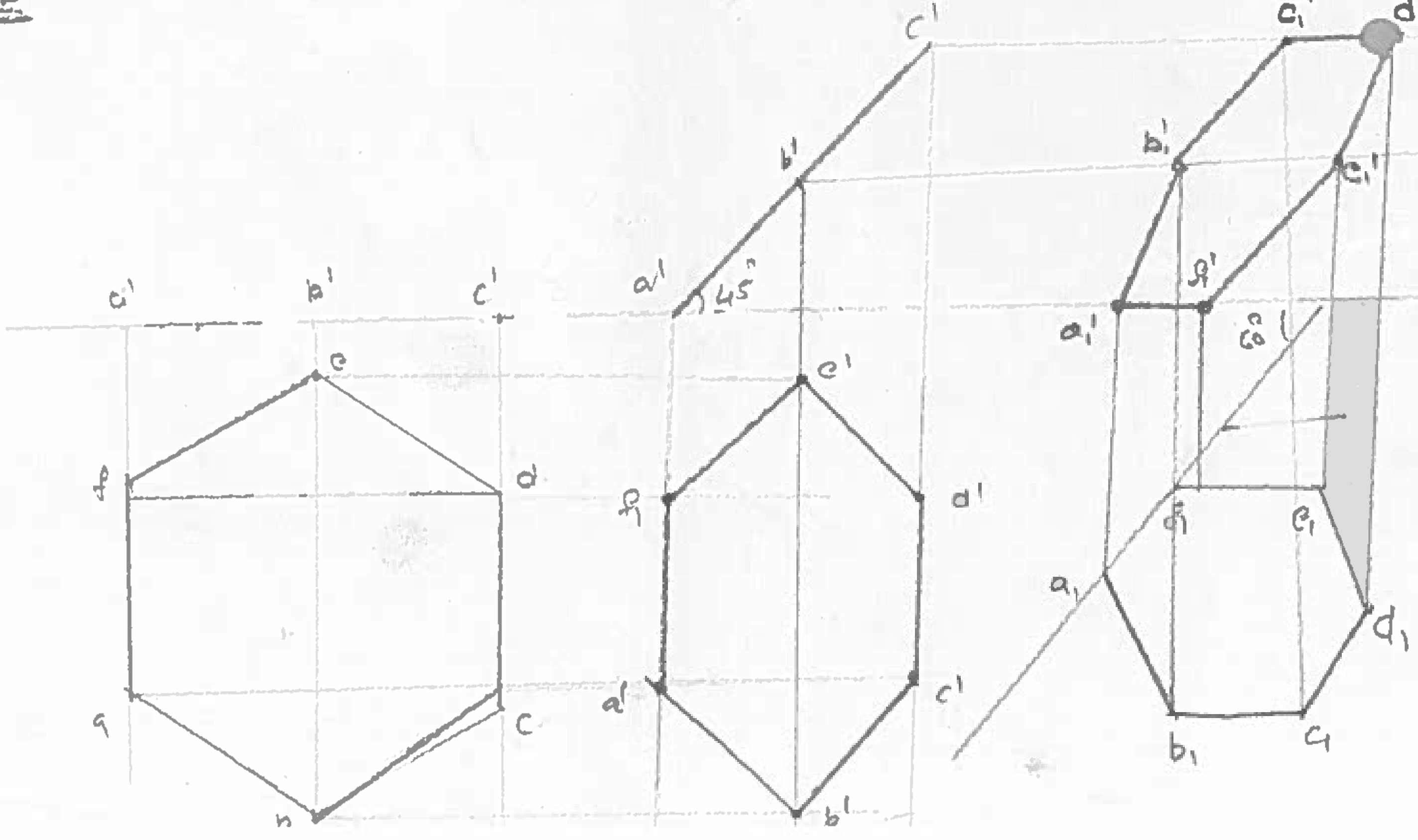

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SEMESTER Question Paper

Degree	B. Tech. (U. G.)	Program	CE	Test	I/II	Academic Year	2020 - 2021
Course Code	20CE201	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	BUILDING MATERIALS AND CONSTRUCTION COMPONENTS						

Key and Scheme of Evaluation

No.	Questions (1 through 5)	Marks
1	**List the advantages and disadvantages of stabilized mud blocks?** Advantage: i. Availability: Mud is readily available at most of the sites in India. Not talking about construction in metro cities like Mumbai etc. ii. mud brick construction will be good way to maintain temperature of your building at low level especially in Hot weather cities, Mud brick are ecological. Disadvantages: I. Strength largely depend in the stabilization process and degree of stabilization. Buildings that incorporate the use of clay are particularly vulnerable to deterioration and deserving of care and maintenance. ii. Picture of the associated works mud in our latitudes (for misinformation) with "poverty." Ironically, in other more technologically updated and effluent is currently considered a symbol of status.	Content 2M
2	**List any two objectives of seasoning of timber?** i.Maintain the size and shape of timber& Improve strength, hardness and stiffness of timber. ii. Make it suitable for receiving various treatments like paints, preservatives, varnishes etc.	Content 2M
3	**List any two advantages of cavity walls?** I. The moisture cannot enter from outer wall to inner wall, since there is no direct contact ii.Provide good insulation against sound& Protection against efflorescence.	Content 2M
4	**What are materials used damp proofing course?** **Following are the materials which are commonly used for the damp-proofing:** • Hot Bitumen. • Mastic Asphalt. • Bituminous Felts. • Metal Sheets. • Combination of Sheets and Felts. • Stones. • Bricks. • Mortar The mortar.	Any 2 or 3 Content 2M
5	Classify aggregates based on size? According to Particle Size: 1) Fine Aggregate (sand): Fine aggregate includes the particles that all passes through 4.75 mm sieve and retain on 0.075 mm sieve. EX: sand, crushed stone, ash or cinder and surki. 2) Coarse Aggregate (gravel): Coarse aggregate includes the particles that retain on 4.75 mm sieve.	Content 2M
No.	**Questions (6 through 11)**	
6	**Discuss the requirement of a good building stone? Explain the dressing of stones?** The following are the qualities or requirements of a good building ' stone. 1. Crushing strength: For a good building stone, the crushing strength should be greater than l000kg per cm2.	List 2m; content explanation 6m; dressing

	2. Appearance: Good building stone should be a uniform color, and free from clay holes, spots of other color bands etc capable of preserving the color for longtime. 3. Durability: A good building stone should be durable. The factors like heat and cold alternative wet and dry, dissolved gases in rain, high wind velocity etc affect the durability. 4. Fracture: For good building stone its fracture should be sharp, even and clear. 5. Hardness: The hardness greater than 17, treated as hard used in road works. It is between 14 to 17, medium hardness, less 14 said be poor hardness. 6. Percentage wear: For a good building stone, the percentage wear should be equal to or less then 3 percent. 7. Resistance to fire: A good building stone be fire proof. Sandstone, Argillaceous stone resists fire quite well 8. Specific gravity: For a good building stone the specific gravity should be greater then 8.7 or so. 9. Texture: A good building stone should have compact fine crystalline structure should be free from cavities, cracks or patches of stuff or loose material. Stones 10. Water absorption: For a good building stone, the percentage absorption by weight after 24 hours should not exceed 0.60. **Dressing of stones:** The stone dressing is a **process of surfacing and shaping of rocks available naturally.** The place where the rocks are abundantly available is called as a quarry. ... Once quarried, the stones are cut into the suitable size and surface finishes. This process is termed as dressing of stones **Following are the types of dressing** 1. Hammer Dressed or Quarry-faced Surface 2. Rough tooled surface 3. Tooled Surface 4. Cut stone Surface 5. Rubbed Surface	definition 3m; types 1m
	OR	
7 (a)	**What are the common ingredients of good brick earth?** Following are the constituents of good brick earth. Alumina: - It is the chief constituent of every kind of clay. A good brick earth should contain 20 to 30 percent of alumina. This constituent imparts plasticity to earth so that it can be moulded. If alumina is present in excess, raw bricks shrink and warp during drying and burning. Silica-A good brick earth should contain about 50 to 60 percent of silica. Silica exists in clay either as free or combined form. As free sand, it is mechanically mixed with clay and in combined form; it exists in chemical composition with alumina. Presence of silica prevents crackers shrinking and warping of raw bricks. It thus imparts uniform shape to the bricks. Durability of bricks depends on the proper proportion of silica in brick earth. Excess of silica destroys the cohesion between particles and bricks become brittle. Lime – A small quantity of lime is desirable in finely powdered state to prevents shrinkage of raw bricks. Excess of lime causes the brick to melt and hence, its shape is last due to the splitting of bricks. Oxide of iron- A small quantity of oxide of Iron to the extent of 5 to 6percent is desirable in good brick to imparts red color to bricks. Excess of oxide of iron makes the bricks dark blue or blackish. Magnesia- A small quantity of magnesia in brick earth imparts yellow tint to bricks, and decreases shrinkage. But excess of magnesia decreases shrink leads to the decay of bricks.	List 2m; Content explanation 4m

The ingredients like, lime, iron pyrites, alkalis, pebbles, organic matter should not present in good brick earth

7 (b) **State any five desirable properties of good bricks**

The following are the required properties of good bricks:

Color: Color should be uniform and bright.

Shape: Bricks should have plane faces. They should have sharp and true right-angled corners.

Size: Bricks should be of standard sizes as prescribed by codes.

Texture: They should possess fine, dense and uniform texture. They should not possess fissures, cavities, loose grit and unburnt lime.

Soundness: When struck with hammer or with another brick, it should produce metallic sound. (vi) Hardness: Finger scratching should not produce any impression on the brick. (vii) Strength: Crushing strength of brick should not be less than 3.5 N/mm2. A field test for strength is that when dropped from a height of 0.9 m to 1.0 mm on a hard ground, the brick should not break into pieces.

Water Absorption: After immersing the brick in water for 24 hours, water absorption should not be more than 20 per cent by weight. For class-I works this limit is 15 per cent.

Fire Resistance: Fire resistance of bricks is usually good. In fact, bricks are used to encase steel columns to protect them from fire.

Thermal Conductivity: Bricks should have low thermal conductivity, so that buildings built with them are cool in summer and warm in winter.

List 2m; Content explanation 4m

8 (a) **Explain with a neat sketch of any five defects in timber?**

Most common defects in timber are

1. Heart Shakes
2. Star Shakes
3. Cup Shakes
4. Radial Shakes
5. Rind Galls
6. Wind Cracks
7. Knots
8. Dead Wood

List 2m; Content explanation 3m; sketch 1m

OR

9 (a) **How many types of brick bond are there?**

Masonry may define as the construction of building units bonded together with mortar

Types of Bonds:

- Stretching Bond
- Heading Bond
- **English Bond**
- **Flemish Bond; 1. Double Flemish Bond, 2. Single Flemish Bond**

List 2m; Content explanation 4m

9 (b) **What is meant by fiber-reinforced concrete? What are the advantages and disadvantages of fiber-reinforced concrete?**

Fiber - reinforced concrete: Fiber Reinforced Concrete is a composite material consisting of fibrous material which increases its structural integrity. It includes mixtures of cement, mortar or concrete and discontinuous, discrete, uniformly dispersed suitable fibers. Fibers are usually used in concrete to control cracking due to plastic shrinkage and to drying shrinkage. They also reduce the permeability of concrete and thus reduce the bleeding of water

Definition Content 3m; Advantages& disadvantages 3m

Advantages of Fiber - reinforced concrete:

• Main role of fibers is to bridge the cracks that develop in concrete and increase the ductility of concrete elements

	• Improvement on Post-Cracking behavior of concrete • Imparts more resistance to Impact load • Lowers the permeability of concrete matrix and thus reduce the bleeding of water Disadvantages of Fiber - reinforced concrete: • Increase in specific gravity of the concrete, this means that the concrete will be heavier than normal concrete in case of some fibers • Higher cost because of its control issues (production issues) as well as the cost of raw material is high • Corrosion of steel fibers	
10 (a)	**What are the characteristics of lime and its uses?** plasticity Flexiable and easy workable Setting time Greter strength High resistance to moisture Uses of lime: Lime is very useful material that finds extensive applications in building construction, industry and agriculture • as a mortar (lime-mortar) mixed with sand or surkhi. • as a plaster • as a whitewash • as sand-lime bricks which are quite popular in many countries	Characteristics 4m; uses 2m
10 (b)	**What are the classifications of lime?** Limes classified as non-hydraulic or hydraulic. Non-hydraulic limes do not harden without air being present I. According to Chemical Composition (Quicklime, Hydrated lime, Hydraulic lime) ii. According to Use	Content 5m; Presentation 1m
	OR	
11	**What are the ingredients of ordinary cement? State the contributions of such in gradients** Ordinary Portland cement contains two basic ingredients, namely argillaceous and calcareous. In argillaceous materials, clay predominates and in calcareous materials, calcium carbonate predominates. A) Good ordinary cement contains following ingredients. 1. Lime (cao) .... 62% 2. silica (Sio2) .... 22% 3. Aluminca(Al2 u3) .... 5% 4. Calcium sulphate (CaSo4) .... 4% 5. Iron Oxide (Fe2 O3) .... 3% 6. Magnescia (Mgo) .... 2% 7. Sulphur .... 1% 8. Alkalies .... 1% B) Contribution of ingredients (Functions of ingredients)	Ingredient content 6m; functions 6m
12	**List various classification of flooring. Explain any four types of floors?** Floors are classified into two categories 1. Timber Floors 2. Composite Floors Types of floors: Marble flooring Granite flooring Bamboo flooring	2 m for the list; classification 5m, Types content 5m

	Hardwood flooring Tile flooring	
	OR	
13(a)	**Briefly explain the shoring and under pinning** Shoring definition and methods of shoring Under pinning definition and types and methods of underpinning	Each explanation 3m
13(b)	**List the classification of pitched roof. With neat sketch explain any two of them** A sloping roof is known as pitched roof A) The Pitched roofs classified into the following three categories: B) sketch	2marks for the list, explanation content 3m Sketch 1m
14(a)	**Define Fine Modulus of Aggregate? Explain the detailed test process to calculate the fine modulus of fine aggregate** The fineness modulus of aggregate is simply a measurement of the average size of the aggregate Test procedure: Take the sieves and arrange them in descending order with the largest sieve on top. If mechanical shaker is using then put the ordered sieves in position and pour the sample in the top sieve and then close it with sieve plate. Then switch on the machine and shaking of sieves should be done at least 5 minutes. If shaking is done by the hands, then pour the sample in the top sieve and close it then hold the top two sieves and shake it inwards and outwards, vertically and horizontally. After some time shake the 3rd and 4th sieves and finally last sieves. After sieving, record the sample weights retained on each sieve. Then find the cumulative weight retained. Finally determine the cumulative percentage retained on each sieve. Add the all-cumulative percentage values and divide with 100 then we will get the value of fineness modulus. fineness modulus of aggregate = (cumulative % retained) / 100	Content 3m ; Test process3m
14(b)	**What is the importance of specific gravity aggregate? Mention the testing process to determine its character** a) The specific gravity of aggregates indirectly measures its density; hence it is the most essential parameter of strength or quality of the aggregates. Higher the specific gravity, higher is the strength. b) Test procedure for sp. gravity ofcement/ Coarse aggregate	Content 3m ; Test process3m
	OR	
15(a)	**Briefly explain the importance of size, shape, and texture on coarse aggregates** Characteristics of aggregates: size, shape, and texture Particle Size and Gradation: has more influence on the performance of hardened concrete, asphalt, and base material performance than any other characteristic of aggregates. The size and distribution of particles directly impact properties of stiffness, strength, workability, permeability, stability, skid resistance, and more. It is no surprise that this is by far the most common and primary test to be performed on an aggregate sample. As most of these aggregate characterization tests, it is not difficult to perform properly and can be conducted effectively by technicians with minimal training. Once the proportions of the individual fractions are determined and plotted in graphical form as a gradation curve, the information can be used for more than just a report of grain sizes. The values can qualitatively group the aggregate with classification terms like gap-graded, open-graded, or uniformly-graded to describe particle distribution. This information can be used to adjust the proportions of the fractions to manipulate the qualities of the final mix	Content 6m

	designs. ASTM C136 and AASHTO T 27 spell out requirements of the sieve analysis test for aggregates. ASTM E11 lists specifications and tolerances for the test sieves.	
15(b)	**Explain Flakiness index and elongation index on coarse aggregate** Flat and Elongated Particles testing measures dimensional ratios of individual coarse aggregate particles. Particles with significantly greater length compared to their width will tend to fracture across the narrow aspect when loaded and can resist reorientation during compaction of asphalt paving mixtures. The fracturing of the particles also negatively affects the void content, stability, and binder distribution of asphalt. These dimensional characteristics also interfere with the placement and consolidation properties of freshly mixed concrete. In the ASTM D4791 test method, a proportional caliper is used to test and classify a representative sample of about 100 individual aggregate particles from each size fraction. Flakiness Index is looking for some of the same dimensional properties as the flat and elongated test but uses a slotted thickness gauge and a separate length gauge to classify the particles. This test method is based on procedures in British Standard BS 812 and is preferred by some state departments of transportation over the ASTM flat and elongated method. Individual particles from each size fraction are tried in the thickness and length gauges. Aggregate particles in this test are classified as flaky when their smallest dimension is less than 0.6 of their nominal size.	Each index Content 3 m

NSRIT

Semester End Examination, Sept./Oct., 2021

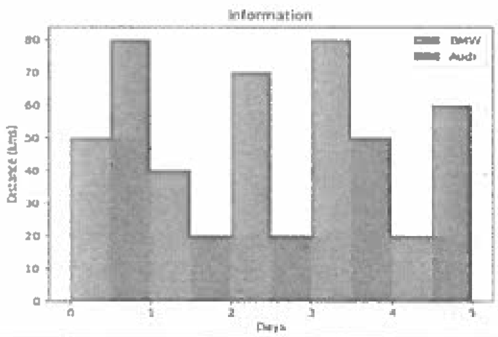
Degree	B. Tech. (U. G.)	**Program**	EEE			**Academic Year**	2020 - 2021
Course Code	20CS403	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	II
Course	PYTHON PROGRAMMING						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is type conversion?	20CS403.1	L2
2	Compare List and Tuple	20CS403.2	L3
3	Define Module. What is the use of Module?	20CS403.3	L2
4	How will you manipulate file pointer using seek?	20CS403.4	L1
5	List out geometry manager classes in tkinter module	20CS403.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Give short note on the following: i) Python variables ii) Keywords iii) Python Indentation	6M	20CS403.1	L2
6 (b)	Write a Python program that solve the quadratic equation ax**2 + bx + c = 0	6M	20CS403.1	L3
	OR			
7 (a)	Explain in detail about Program development cycle	6M	20CS403.1	L2
7 (b)	Write a Python program to demonstrate the application of identity operators & Membership operators.	6M	20CS403.1	L2
8 (a)	Explain how Accessing Character and Substring in Strings is done in Python with example	6M	20CS403.2	L2
8 (b)	Write a Python Program to Check if a Number is Positive, Negative or 0	6M	20CS403.2	L3
	OR			
9 (a)	Explain the various List methods available in Python	6M	20CS403.2	L2
9 (b)	Write a Python program to check if the number is an Armstrong number or not. (A positive integer is called an Armstrong number of order n if abcd... = $a^n + b^n + c^n + d^n + \ldots$ An Armstrong number of 3 digits, the sum of cubes of each digit is equal to the number itself. For example: 153 = 1*1*1 + 5*5*5 + 3*3*3 // 153 is an Armstrong number)	6M	20CS403.2	L2
10 (a)	Explain any 3 functions of the following modules: i. Cmath ii. Random	6M	20CS403.3	L2
10 (b)	Explain arbitrary and keyword argument in Python with example	6M	20CS403.3	L2
	OR			
11 (a)	What is Recursion? Explain the working of recursive function with an example	6M	20CS403.3	L2
11 (b)	What is PIP? How packages are installed using PIP?	6M	20CS403.3	L2

12 (a)	What is File? Explain the file handling functions in Python with example	6M	20CS403.4	L2
12 (b)	How to create a constructor and destructor in Python? Give an example	6M	20CS403.4	L2
	OR			
13 (a)	Demonstrate implementation of multilevel inheritance in Python, with a program	6M	20CS403.4	L2
13(b)	What is operator overloading in Python?	6M	20CS403.4	L2
14	Explain any 5 functions in Numpy module with example	12M	20CS403.5	L2
	OR			
15	Demonstrate the usage MatPlotlib library. Write a program for the following graph 	12M	20CS403.5	L3

SEMESTER Question Paper

Degree	B. Tech. (U. G.)	Program	EEE	Test	I/II	Academic Year	2020 - 2021
Course Code	20CS403	Test Duration	90 Min.	Max. Marks	40	Semester	II
Course	PYTHON PROGRAMMING						

Key and Scheme of Evaluation

No.	Questions (1 through 5)	Marks
1	**What is type conversion?** The process of converting the value of one data type to another data type is called typeconversion. Python has two types of type conversion. Implicit Type Conversion Explicit Type conversion	Definition-1M Types-0.5M Example-0.5M
2	**Compare List and Tuple** S. No / List / Tuple 1 / Lists are mutable / Tuples are immutable 2 / Lists consume more memory / Tuple consume less memory as compared to the list 3 / List is created using [] / Tuple is created using ()	Any 4 differences-2M
3	**Define Module. What is the use of Module?** In Python, Modules are simply files with the ".py" extension containing Python code that can be imported inside another Python Program.In simple terms, we can consider a module to be the same as a code library or a file that contains a set of functions that you want to include in your application. To incorporate the module into our program, we will use the import keyword, and to get only a few or specific methods or functions from a module, we use the from keyword. **Syntax:** module_name.function_name **Eg:** import math Print(math.pi)	Description-1M Importing module-1M
4	**How will you manipulate file pointer using seek?** **seek():** In Python, seek() function is used to **change the position of the File Handle** to a given specific position. File handle is like a cursor, which defines from where the data has to be read or written in the file. **Eg:** f = open("xyz.txt", "r") f.seek(20)	Description-1M Syntax-0.5M Example-0.5M
5	**List out geometry manager classes in tkinter module** The geometry manager is used to manage the geometry of the window and other frames. We can use it to handle the position and size of the window and frames. There are mainly three methods in Geometry Managers: The pack() method The grid() method The place() method	Description-0.5M List-1.5M
No.	Questions (6 through 11)	
6 (a)	**Give short note on the following:** **i) Python variables ii) Keywords iii) Python Indentation** i) Python variables **Definition** It is an identifier that is used to refer to memory location and used to hold value. In Python, we don't need to specify the type of variable Rules for creating variables in Python • It must start with a letter or the underscore character. • It can not start with a number. • It can only contain alpha-numeric characters and underscores (A-z, 0-9, and _)	Content 4m Grammar & Spellings 1m; presentation 1m

	• Variable names are case-sensitive **ii) Keywords** Python keywords are the fundamental building blocks of any program Python keywords are special reserved words that have specific meanings and purposes There are a number of ways you can identify valid Python keywords 1. Using help() Eg: >>> help("keywords") 2.import keyword module Eg: >>>import keyword >>> keyword. Kwlist **iii) Python Indentation** In python code blocks are identified by indentation rather than using symbols like curly braces. Without extra symbols programs are easier to read and also indentation clearly identifies which block of code a statement belongs to. Python does not support braces to indicate blocks of code for class or function definitions or flow control. Blocks of code is identified by line indentation. All the continues lines indented with same number of spaces would form a block. Python strictly follow indentation rules to indicate the blocks	

6 (b)	**Write a Python program that solve the quadratic equation ax**2 + bx + c = 0**
	Program:
	# import Complex math module
	import cmath
	a=float(input("Enter a value"))
	b=float(input("Enter b value"))
	c=float(input("Enter c value"))
	# calculate the discriminant
	d = (b**2) - (4*a*c)
	# find two solutions
	root1 = (-b-cmath.sqrt(d))/(2*a)
	root2 = (-b+cmath.sqrt(d))/(2*a)
	print('The solution are {0} and {1}'.format(root1,root2))
		Program-5M
		Output-0.5M
		Explaination-0.5M

	OR	
7 (a)	**Explain in detail about Program development cycle** Python's development cycle is dramatically shorter than that of traditional tools. In Python, there are nocompile or link steps -- Python programs simply import modules at runtime and use the objects they contain. The Program Development Cycle (PDC) has various states as follow Problem Definition Program Design Coding Debugging Testing Documentation Maintenance	List of states-1M Explanation of each state-5M
7 (b)	**Write a Python program to demonstrate the application of identity operators & Membership operators.** Operators are symbols that perform certain mathematical or logical operation to manipulate data values and produce a result. **Identity operators:** used to check if two values (variable) are located on the same object or same memory **Membership operators:** used to test whether a value or operand is found in the sequence such as list, string, set, or dictionary.	Description-2M Identity operator with example-2M Membership operator with example-2M
8 (a)	**Explain how Accessing Character and Substring in Strings is done in Python with example** A string is a sequence of zero or more characters. It is treated as a data structure. A string's length is the number of characters it contains. The length can be obtained using the len() function by passing the string as an argument to it. The positions of a string's characters are numbered from 0,	Definition-1M Accessing character-

	on the left, to the length of the string minus 1, on the right. **Accessing character & substring** 1. Using Subscript Operator 2. Using slice opertor	2.5M Accessing substring- 2.5M

8 (b)	**Write a Python Program to Check if a Number is Positive, Negative or 0**
	<u>Program</u>
	a=int(input ("enter a number:"))
	if(a>0):
	    print("The given number is a Positive number ")
	elif(a<0):
	    print ("The given number is Negative number")
	else:
	    print("The given number is equal to Zero")
		Program-5M
		Output-0.5M
		Explaination-
		0.5M

	OR	
9 (a)	**Explain the various List methods available in Python** List is used to store the group of values & we can manipulate them, in list the values are stores in index format starts with 0.List is mutable object so we can do the manipulations. Syntax: <list_name> = [value1,value2,value3,...,valuen] Example: data5=['TEC',10,56.4,'a'] # list with mixed data types **List Methods:** append()-Adds an element at the end of the list clear()-Removes all the elements from the list copy()-Returns a copy of the list count()-Returns the number of elements with the specified value extend()-Add the elements of a list (or any iterable), to the end of the current list index()-Returns the index of the first element with the specified value insert()-Adds an element at the specified position pop() Removes the element at the specified position remove() Removes the first item with the specified value	Description- 1M Any 5 methods with example-5M

9 (b)	**Write a Python program to check if the number is an Armstrong number or not. (A positive integer is called an Armstrong number of order n if abcd... = $a^n + b^n + c^n + d^n + ...$In case of An Armstrong number of 3 digits, the sum of cubes of each digit is equal to the number itself. For example: 153 = 1*1*1 + 5*5*5 + 3*3*3 // 153 is an Armstrong number)**
	<u>Program:</u>
	n=int(input("Enter Number"))
	sum=0
	temp=n
	while n!=0:
	    rem=n%10
	    sum=sum+(rem*rem*rem)
	    n=n//10
	if(sum==temp):
	    print("The given number is Armstrong number")
	else:
	    print("The given number is not a Armstrong number")
		Program-5M
		Output-0.5M
		Explaination-
		0.5M

10 (a)	**Explain any 3 functions of the following modules** **i. Cmath** Python has a built-in module that you can use for mathematical tasks for complex numbers.The methods in this module accepts int, float, and complex numbers. It even accepts Python objects that has a __complex__() or __float__() method.The methods in this module almost always return a complex number. If the return value can be expressed as a real number, the return value has an imaginary part of 0. The cmath module has a set of methods and constants.	cmath module functions with example-3M random module functions

	1.Sqrt() 2.log10() 3.cos() ii. **Random** Python Random module is an in-built module of Python which is used to generate random numbers. These are pseudo-random numbers means these are not truly random. This module can be used to perform random actions such as generating random numbers, print random a value for a list or string, etc. some common operations performed by this module. 1.random.randint() 2.random.random() 3.random.choice()	with example-3M
10 (b)	**Explain arbitrary and keyword argument in Python with example** **Arbitrary arguments** Sometimes, we do not know in advance the number of arguments that will be passed into a function. Python allows us to handle this kind of situation through function calls with an arbitrary number of arguments. In the function definition, we use an asterisk (*) before the parameter name to denote this kind of argument. **Keyword arguments** Python allows functions to be called using keyword arguments. When we call functions in this way, the order (position) of the arguments can be changed.	Arbitrary argument with example-3M Keyword argument with example-3M
	OR	

11 (a)	**What is Recursion? Explain the working of recursive function with an example**
	A function that calls itself is known as Recursive Function.
	Program:
	def fact(n):
	    if(n==0 or n==1):
	        return 1
	    else:
	        return n*fact(n-1)
	n=int(input("Enter number"))
	print("The factorial of a given number is:",fact(n))
	Output:
	Enter number5
	The factorial of a given number is: 120
		Definition-0.5M
		Program-5M
		Output-0.5M

11 (b)	**What is PIP? How packages are installed using PIP?** pip is a package-management system written in Python used to install and manage software packages.It connects to an online repository of public packages, called the Python Package Index. pip is the ease of its command-line interface, which makes installing Python software packages as easy as issuing a command: pip install some-package-name **Eg:** pip install Matplotlib pip uninstall some-package-name **Eg:** pip uninstall Matplotlib	Description-2M Commands with example-4M
12(a)	**What is File? Explain the file handling functions in Python with example** File: A file is some information or data which stays in the computer storage devices. ... Python gives you easy ways to manipulate these files. Generally we divide files in two categories, text file and binary file. **File handling functions** • Open() • Close() • Write() • Writelines() • Read() • Readlines()	Definition-1M List of functions-1M Any 4 functions with example-4M

	• tell()	
12(b)	**How to create a constructor and destructor in Python? Give an example** A constructor is a special type of method (function) which is used to initialize the instance members of the class. In C++ or Java, the constructor has the same name as its class, but it treats constructor differently in Python. It is used to create an object. Constructors can be of three types. 1.Default Constructor 2.Parameterized Constructor 3.Non-parameterized Constructor	Description-1M Types-1M Any constructor with example-3M
	OR	

13(a)	**Demonstrate implementation of multilevel inheritance in Python, with a program**
	Inheritance is a powerful feature in object oriented programming. It refers to defining a new class with little or no modification to an existing class. The new class is called derived (or child) class and the one from which it inherits is called the base (or parent) class.
	Multilevel Inheritance
	In multilevel inheritance, features of the base class and the derived class are further inherited into the new derived class. This is similar to a relationship representing a child and grandfather.
	# Python program to demonstrate multilevel inheritance
	# Base class
	class Grandfather:
	grandfathername =""
	def grandfather(self):
	print(self.grandfathername)
	# Intermediate class
	class Father(Grandfather):
	fathername = ""
	def father(self):
	print(self.fathername)
	# Derived class
	class Son(Father):
	def parent(self):
	print("GrandFather :", self.grandfathername)
	print("Father :", self.fathername)
	# Driver's code
	s1 = Son()
	s1.grandfathername = "Srinivas"
	s1.fathername = "Ankush"
	s1.parent()
		Description-2M
		Multilevel inheritance Program-4M

13(b)	**What is operator overloading in Python?** Python allows the same operator to have different meaning according to the context is called operator overloading. Python Special functions used for operator overloading: • __add__(self, other) • __sub__(self, other) • __mul__(self, other) • __floordiv__(self, other) • __lt__(self, other)) **Example** # Python Program illustrate how to overload an binary + operator class addoperator:	Description-2M Special functions-1M Program-3M

```
    def __init__(self, X):
        self.X = X
    # __add__ () method is magic function to perform addition of two objects
    def __add__(self, other):
        return self.X + other.X
obj1 = addoperator(234)
obj2 = addoperator(456)
print (obj1 + obj2)
obj3 = addoperator("Welcome ")
obj4 = addoperator("to NSRIT")
print (obj3 + obj4)
```

14

Explain any 5 functions in Numpy module with example

NumPy stands for numeric python which is a python package for the computation and processing of the multidimensional and single dimensional array elements.

It provides various functions which are capable of performing the numeric computations with a high speed.

NumPy Functions

1. **numpy.array()** : We can create a NumPy ndarray object by using the array() function. The array object in NumPy is called ndarray. It is basically a table of elements which are all of the same type and indexed by a tuple of positive integers.

Example:

```
import numpy as np
#Here create 1-D Array
arr=np.array([1, 2, 3, 4, 5])Herecreate1-DArray
#Here create 2-D Array
arr=np.array([[1, 2, 3],[4, 5, 6]])
```

2. numpy.sum()

This function is used to compute the sum of all elements. It is also possible to add rows and column elements of an array. The output will be in the form of an array object.

Example:

```
a=np.array([[1,4],[3,5]])
b=np.sum(a)
print(b)  #13
```

3. **numpy.append()**

The numpy append() function is used to merge two arrays. It returns a new array, and the original array remains unchanged.

The numpy.append() function is used to add or append new values to an existing numpy array. This function adds the new values at the end of the array.

Example:

```
import numpy as np
a=np.array([[10, 20], [40, 50], [70, 80]])
b=np.array([[11, 21], [42, 52], [73, 83]])
c=np.append(a,b)
print(c) #array([ 10,  20, 40, 50, 70, 80,11, 21,42, 52, 73, 83])
```

4. numpy.sort()

The NumPy ndarray object has a function called sort(), that will sort a specified array. Sorting means putting elements in an *ordered sequence*. *Ordered sequence* is any sequence that has an order corresponding to elements, like numeric or alphabetical.

Example

```
arr=np.array(['banana', 'cherry', 'apple'])
print(np.sort(arr))#['apple''banana''cherry']
```

5. numpy. arrange ()

It creates an array by using the evenly spaced values over the given interval.

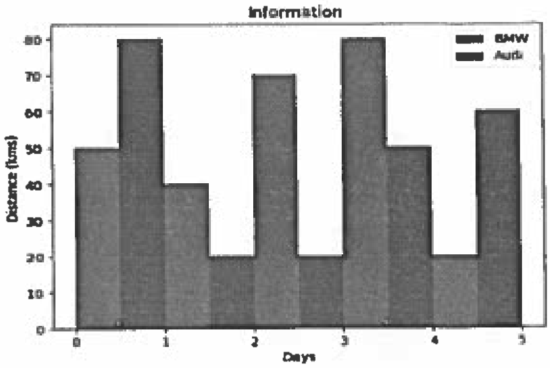
Example:

```
arr = np.arange(0,10,2,int) # [0 2 4 6 8]
```

Description-1M
List of functions-1M
Five functions with example-10M

OR

Demonstrate the usage MatPlotlib library. Write a program for the following graph



Matplotlib.pyplot is a plotting library used for 2D graphics in python programming language. It can be used in python scripts, shell, web application servers and other graphical user interface toolkits. Matploitlib is a Python Library used for plotting, this python library provides and objected-oriented APIs for integrating plots into applications.

Types of Plots

There are various plots which can be created using python matplotlib. Some of them are listed below:

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1. Bargraph
2. histogram
3. Scatter Plot
4. Area Plot
5. Pie Plot

Python Matplotlib – Histogram

Histograms are used to show a distribution whereas a bar chart is used to compare different entities. Histograms are useful when you have arrays or a very long list.

<u>Program</u>

```
import matplotlib.pyplot as plt
population_age
=[22,55,62,45,21,22,34,42,42,4,2,102,95,85,55,110,120,70,65,55,111,115,80,75,65,54,44,43,42,48]
bins = [0,10,20,30,40,50,60,70,80,90,100]
plt.hist(population_age, bins, histtype='bar', rwidth=0.8)
plt.xlabel('age groups')
plt.ylabel('Number of people')
plt.title('Histogram')
plt.show()
```

Description-3M
Types-2M
Program-9M
Expaination-1M

NSRIT

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2020 - 2021
Course Code	20EC201	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	II
Course	**Principles of Electronics & Communication Systems**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Fermi level	20EC201.1	L1
2	What is CMRR?	20EC201.2	L1
3	What is the need for modulation?	20EC201.3	L1
4	Define PAM and PPM	20EC201.4	L1
5	Define TIR	20EC201.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain Insulator, Semiconductor & conductor with help of energy band structure	6M	20EC201.1	L2
6 (b)	Differentiate between intrinsic and extrinsic semiconductor	6M	20EC201.1	L2
	OR			
7 (a)	Explain n-type semiconductor	6M	20EC403.1	L2
7 (b)	Derive the expression for current generated due to drifting of charge carriers in semiconductors in the presence of electric field	6M	20EC403.1	L2
8 (a)	Explain application of op-amp as integrator and differentiator	6M	20EC201.2	L2
8 (b)	Explain ac characteristics of op-amp	6M	20EC201.2	L2
	OR			
9 (a)	Draw and explain the pin diagram IC741 op-amp	6M	20EC201.2	L2
9 (b)	Derive the gain for non-inverting op-amp	6M	20EC201.2	L2
10 (a)	State and explain properties of continuous signals	8M	20EC201.3	L2
10 (b)	List any four applications of FM system	4M	20EC201.3	L2
	OR			
11 (a)	Define amplitude modulation. Derive an expression for the AM wave	8M	20EC201.3	L2
11 (b)	Write about am voltage distribution	4M	20EC201.3	L2
12 (a)	State and prove sampling theorem	6M	20EC201.4	L2
12 (b)	Describe the basic principles of PCM system and PCM transmitter	6M	20EC201.4	L2
	OR			
13 (a)	Explain the basic Elements of Digital Communication System	6M	20EC201.4	L2
13(b)	With a neat diagram explain the Generation of PCM & DPCM	6M	20EC201.4	L2
14(a)	Draw and explain the working principle of an Optical transmitter	6M	20EC201.5	L2
14 (b)	Explain about LED and its type	6M	20EC201.5	L2
	OR			
15(a)	Explain the working principle of GSM	6M	20EC201.6	L2
15(b)	Explain Cellular Telephone Systems	6M	20EC201.6	L2

Degree : B.Tech   Program : ECE   Semester : II

Course Code : 20EC201   Course : Principles of Electronics & Communication Systems

Key and Scheme of Evaluation

PART - A

1. Define Fermi level ?   (2M)

Ans. The Fermi level E_F indicates the probability of occupancy of an energy level by an electron.

2. What is CMRR ?   (2M)

Ans. It is defined as the ratio of the differential Voltage gain Ad to the Common mode Voltage gain Acm.

$$CMRR = A_d / A_{cm}$$

This parameter indicates the capability of the op-amp to reject noise.

3. What is the need for Modulation ?   (2M)

Ans : 1) To reduce the antenna height.
2) for Multiplexing of Signals.
3) To increase the range of Communication.
4) To reduce Noise and interference.

4. Define PAM and PPM.   (PAM - 1M   PPM - 1M)

Ans. PAM : Pulse Amplitude Modulation is a process of Changing the amplitude of high frequency periodic rectangular pulse in accordance with the amplitude of message signal.

PPM : pulse position Modulation is a process of Changing the position of high frequency periodic rectangular pulse in accordance with the amplitude of Message signal.

5. Define TIR ? (2M)

Ans : When the incident angle is increased beyond the critical angle, the light ray does not pass through the interface into the other medium. In this condition angle of reflection ϕ_2 is equal to the angle of incidence ϕ_1. This action is called as Total Internal Reflection (TIR) of the beam.

PART – B

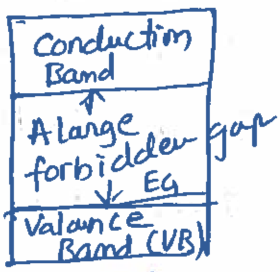
6(a) Explain Insulator, semiconductor & conductor with help of energy band structure. (2M + 2M + 2M = 6M)

Ans

Insulators: Insulators passes no free charge carriers and thus are non - conductive. Insulators are implemented in household items and electrical circuits as protection.

Insulators possess a high resistivity and low conductivity. Their atoms have tightly bound electrons that donot move throughout the material. Because the electrons are static and not freely roaming, a current cannot easily pass.

Eg: Rubber, Teflon, Cloth, wood and fiberglass



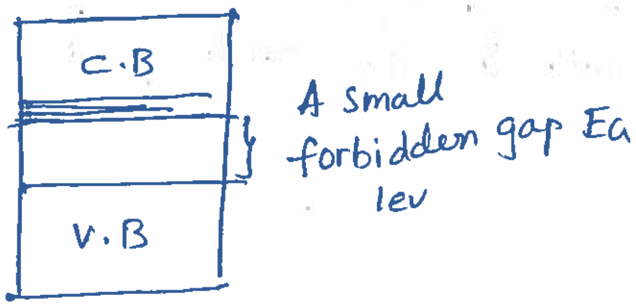
(a) Insulator

Semiconductor

In semiconductors the gap between ValanceBand and Conduction band is smaller

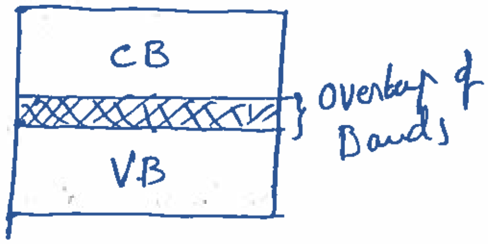
Ex: Ga, As, Si and Ge

At room temparature there is sufficient energy available for electrons to make a transistion from V.B to C.B. This allows some conduction to takes place.



Conductor: A conductor is defined as an object or type of material that allows the flow of charge in one or more directions. Materials made of metal are common electrical conductor, as metal have a high conductance and low resistance.

Eg: Aluminium, Silver, Copper etc.



6(b) Differentiate between intrinsic and extrinsic semiconductor.

Ans

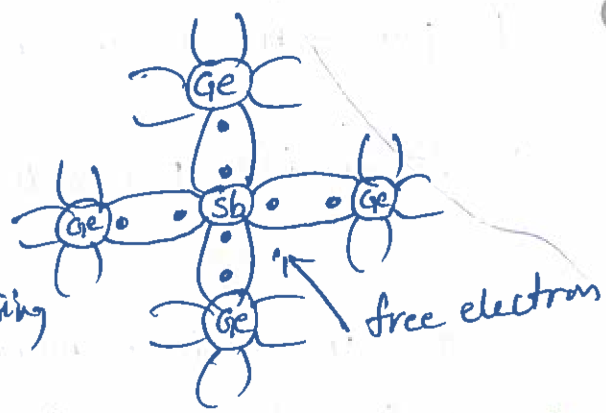
Intrinsic semiconductor	Extrinsic semiconductor
1. pure form of semiconductor	1. Impure form of semiconductor.
2. It exhibits poor conductivity	2. It possesses comparatively better conductivity than intrinsic semiconductor
3. It is present in the middle of forbidden energy gap.	3. The presence of fermi level varies according to the type of extrinsic semiconductor
4. The conduction relies on temperature.	4. The conduction depends on the concentration of dopped impurity and temperature.
5. Equal amount of electron and holes are present in CB & V.B	5. The majority presence of electrons and holes depends on the type of extrinsic semiconductor
6. It is not further classified	6. It is classified as p-type and n-type.

Marks: each difference 1 Mark
total – 6 Marks

7(a) Explain n-type semiconductor? Content - 4M diagram - 2M

Ans: A small amount of pentavalent impurities such as Arsenic, antimony or phosphorus is added to the pure Semiconductor (germanium or silicon crystal) to get N-type semiconductor.

Ge atom has four Valance electrons and antimony has five Valance electrons. each antimony atom forms a covalent bond with surrounding four germanium atoms. Thus, four Valance electrons of antimony atom form covalent bond with four Valance electrons of individual Germanium atom and fifth Valence electron is left free which is loosely bound to the antimony atom.



This loosely bound electron can be easily excited from the V.B to the C.B by the application electric field or increasing the Thermal energy.

7(b) Derive the expression for current generated due to drifting of Charge Carriers in Semiconductors in the presence of electric field?
[Content : 2M Equations: 4M = 6M]

Ans: When an electric field is applied accross the Semiconductor V_d.
material, the Charge carriers attain a certain drift Velocity
which is equal to the product of the mobility of the Charge
Carriers and the applied electric field intensity, E. The holes
move towards the negative terminal of the battery and electron
move toward the positive terminal. This combined effect
of movement of the Charge Carriers Constitutes a current known
as the Drift Current.

Thus the Drift Current is defined as the flow of electric current due to the motion of the Charge Carriers under the influence of an external electric field.

The equation for the drift Current density J_n, due to free electrons given by $J_n = q n \mu_n E$ A/cm^2

and the drift Current density J_p, due to holes is given by

$J_p = q p \mu_p E$ A/cm^2

Where n = number of free electrons per Cubic Centimeter
p = number of holes per Cubic Centimeter
μ_n = mobility of electrons in cm^2/V-s
μ_p = mobility of holes in cm^2/V-s
E = applied electric field Intensity in V/cm
q = Charge of an electron. $= 1.6 \times 10^{-19}$ Coulomb.

8(b) Explain ac Characteristics of op-amp. [each Characteristic 2M
$3 \times 2 = 6M$]

Ans: Slew Rate: It is defined as the maximum rate of Change of output Voltage with time.

The Slew rate is specified in V/μsec. Thus

$$\text{Slew rate} = S = \left.\frac{dV_o}{dt}\right|_{max}$$

Transient Response Rise time:

When the o/p of the op-amp is suddenly Changing like pulse type Then the rise time of the response depends on the cut-off frequency f_H of the op-amp. Such a rise time is called cut-off frequency limited rise time or transient response rise time. It is inversely proportional to the cut-off frequency and given by

$$t_r = \frac{0.35}{f_H}$$

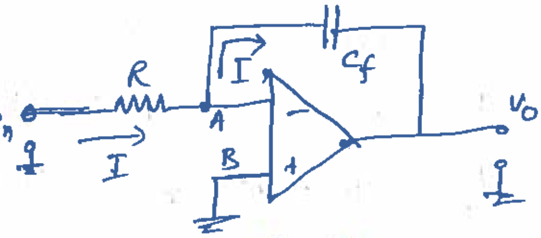
When t_r = rise time f_H = cut-off frequency

frequency Response of op-amp

The plot showing the variations in magnitude and phase angle of the gain due to the change in frequency is called frequency response of the op-amp.

8(a) Explain application of op-amp as Integrator & Differentiator [Integrator – 3M; Differ – 3M]

Ans Integrator: In an integrator circuit, the o/p voltage is the integration of the I/p voltage.



Consider the op-amp integrator ckt. The node B is grounded. The node A is also at the ground potential from the concept of virtual ground.

from the I/p side we can write

$$I = \frac{V_{in} - V_A}{R_1} = \frac{V_{in}}{R_1} \quad \text{—— (1)}$$

from o/p side we can write

$$I = C_f \frac{d(V_A - V_o)}{dt} = -C_f \frac{dV_o}{dt} \quad \text{—— (2)}$$

equating eq(1) & (2)

$$\frac{V_{in}}{R_1} = -C_f \frac{dV_o}{dt}$$

Integrating both sides

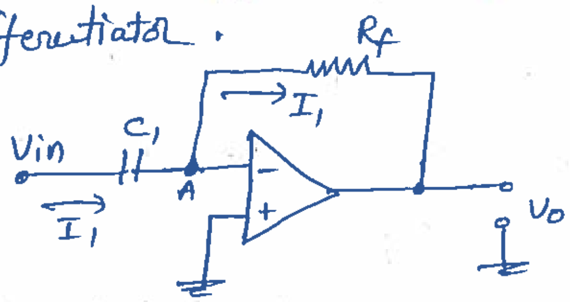
$$\int_0^t \frac{V_{in}}{R_1} dt = -C_f \int \frac{dV_o}{dt} dt$$

i.e $$\int_0^t \frac{V_{in}}{R_1} dt = -C_f V_o$$

$$\therefore V_o = \frac{-1}{R_1 C_f} \int_0^t V_{in}\, dt + V_o(0)$$

Where $V_o(0)$ is the constant of integration

Differentiator: The circuit which produces the differentiation of the input voltage at its output is called differentiator.



The node B is grounded. The node A is also at the ground potential hence $V_A = 0$

As I/p current of op-amp is zero, either current I_1 flows through the resistance R_f.

from the I/p side we can write

$$I_1 = C_1 \frac{d(V_{in} - V_A)}{dt} = C_1 \frac{dV_{in}}{dt} \quad \text{—— (1)}$$

from the o/p side

$$I = \frac{V_A - V_o}{R_f} = \frac{-V_o}{R_f} \quad \text{—— (2)}$$

Equating the two equations

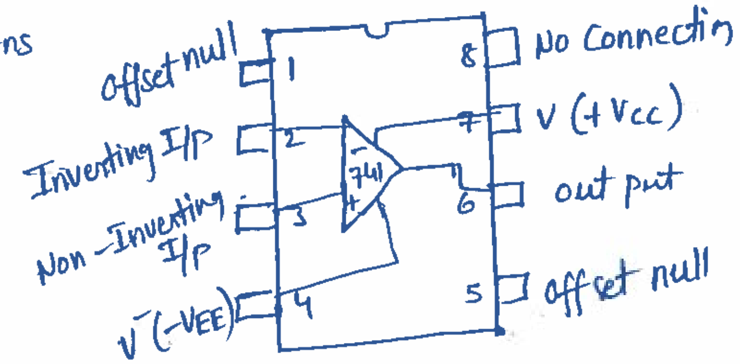
$$C_1 \frac{dV_{in}}{dt} = \frac{-V_o}{R_f}$$

$$\boxed{V_o = -C_1 R_f \frac{dV_{in}}{dt}}$$

The equation shows that the o/p is C_1R_f times the differentiation of the input and product C_1R_f is called time constant of the differentiator.

9(a) Draw and Explain the pin Diagram IC741 Op-amp. Pin Diagram – 3M, Description – 3M } 6M

Ans



Description of op-amp 741 IC pins

Pin 1 and 5: These two pins are used for offset process

Pin 2: Inverting I/p terminal, i.e when a sinusoidal signal is applied to the Input pin 2:

Pin 3: Non-inverting input terminal i.e when a sinusoidal signal is applied to the input pin 3, wave form of same phase o/p is obtained.

Pin 4: $-V_{cc}$, i.e negative terminal of supply voltage is connected to this pin

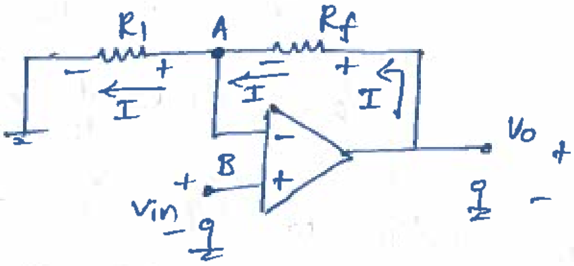
Pin 6: O/p terminal.

Pin 7: $+V_{cc}$ i.e +ve terminal of supply voltage is connected to this pin.

Pin 8: No electrical connection is there in this pin: this pin is just for balance and the symmetric dual-input package look.

9(b) Derive the gain for non-inverting op-amp. Diagram – 2M, Derivation – 4M } 6M

Ans An amplifier which amplifies the input without producing any phase shift b/w I/p and o/p is called non-inverting amplifier.



Derivation of closed loop gain:

The node B is at potential Vin, hence the potential of point A is same as B which is Vin, from the concept of virtual ground.

$$\therefore \; V_A = V_B = V_{in} \quad \text{——— (1)}$$

from the o/p side we can write

$$I = \frac{V_o - V_A}{R_f}$$

$$\therefore \; I = \frac{V_o - V_{in}}{R_f} \quad \text{——— (2)}$$

At the inverting terminal

$$I = \frac{V_A - 0}{R_1}$$

$$\therefore I = \frac{V_{in}}{R_1} \quad \text{———} \ (3)$$

equating 2 and 3

$$\therefore \frac{V_o - V_{in}}{R_f} = \frac{V_{in}}{R_1}$$

$$\therefore \frac{V_o}{R_f} = \frac{V_{in}}{R_f} + \frac{V_{in}}{R_1} = V_{in}\left[\frac{R_1 + R_f}{R_1 R_f}\right]$$

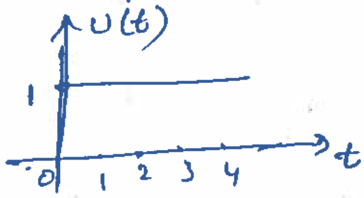
$$\frac{V_o}{V_{in}} = \frac{R_f(R_1+R_f)}{R_1 R_f} = \frac{R_1 + R_f}{R_1} =$$

$$\therefore A_{VF} = \frac{V_o}{V_{in}} = 1 + \frac{R_f}{R_1}$$

10(a) State and Explain the properties of Continuous signals.

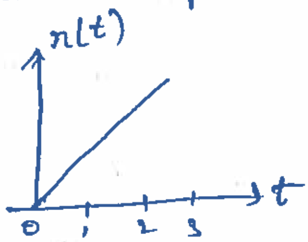
Ans: Unit step signal: The Unit Step function is defined as

$$u(t) = 1 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Unit Ramp Signal: The Unit ramp function is defined as

$$r(t) = t \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



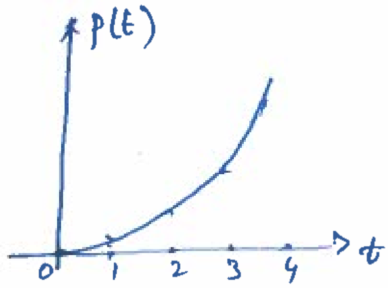
or $r(t) = t\,u(t)$

The ramp function can be obtained by Integrating the Unit step funct

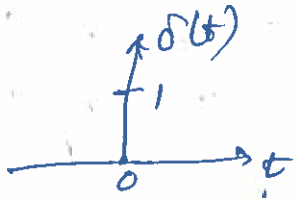
$$r(t) = \int u(t)\,dt = \int dt = t$$

Unit parabolic:

$$p(t) = t^2/2 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Impulse Signal : $\delta(t) = 0$ for $t \neq 0$



Sinusoidal signal : A Continuous-time sinusoidal signal is given by

$$x(t) = A \sin(\Omega t + \theta)$$

A — Amplitude
Ω → frequency in radians per second
θ → is the phase angle in radians

10(b) List any four applications of FM System. (4M)

Ans: 1) It is mostly used in radio broadcasting.
2) It is used in radar, telemetry, seismic prospecting.
3) It is used in music synthesis as well as in Video-transmission intstruments.
4) It is used medical applications like EEG.

11(a) Define Amplitude Modulation. Derive an expression for the AM wave.
[Def: 2M + Expression 4M = 6M]

Ans: Amplitude modulation is a process of Changing the amplitude of the high frequency analog Carrier in accordance with the amplitude of the message signal.

Expression for AM Wave

$$m(t) = A_m \cos 2\pi f_m t$$

$$c(t) = A_c \cos 2\pi f_c t$$

$$s(t) = A_c \cos 2\pi f_c t + A_c k_a m(t) \cos 2\pi f_c t$$

$$s(t) = A_c [1 + k_a m(t)] \cos 2\pi f_c t$$ → time domain equation of AM wave

$$s(t) = A_c[1 + k_a A_m \cos 2\pi f_m t] \cos 2\pi f_c t$$

$\mu = k_a A_m$ modulation index

$$\therefore s(t) = A_c[1 + \mu \cos 2\pi f_m t] \cos 2\pi f_c t$$

$$= A_c \cos 2\pi f_c t + A_c \mu \cos 2\pi f_c t \cos 2\pi f_m t$$

$$= \underbrace{A_c \cos 2\pi f_c t}_{\text{Carrier}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t}_{\text{USB}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t}_{\text{LSB}}$$

11 b) Write aboutan Voltage distribution. (4M)

Ans

$$P_t = P_c + P_{USB} + P_{LSB}$$

$$P = I^2 R$$

$$s(t) = A_c \cos 2\pi f_c t + \frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t + \frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t$$

$$P_c = \frac{(A_c/\sqrt{2})^2}{R} \quad P_{USB} = \frac{(A_c \mu / 2\sqrt{2})^2}{R} \quad P_{LSB} = \frac{(A_c \mu / 2\sqrt{2})^2}{R}$$

$$\therefore P_t = \frac{A_c^2}{2R} + \frac{A_c^2 \mu^2}{8R} + \frac{A_c^2 \mu^2}{8R}$$

$[\because R = 1\Omega]$

$$P_t = \frac{A_c^2}{2}\left[1 + \frac{\mu^2}{2}\right] = P_c\left[1 + \frac{\mu^2}{2}\right]$$

12(a) State and prove Sampling theorem. [Statement 2M + proof 4M = 6M]

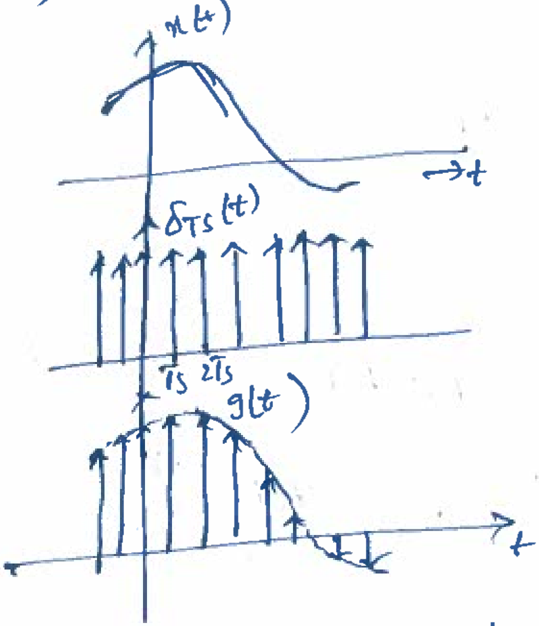
Ans: Statement: A Bandlimited signal of finite energy which has no frequency components higher than f_m Hz may be completely recovered from the knowledge of its samples taken at the rate of $2f_m$ samples per second.

Proof of Sampling theorem :

Let $x(t)$ is a Continuous Signal, with maximum frequency f_m e

The Sampling of $x(t)$ at a rate f_s Hz, may be achieved by multiplying $x(t)$ by an impulse train $\delta_{Ts}(t)$

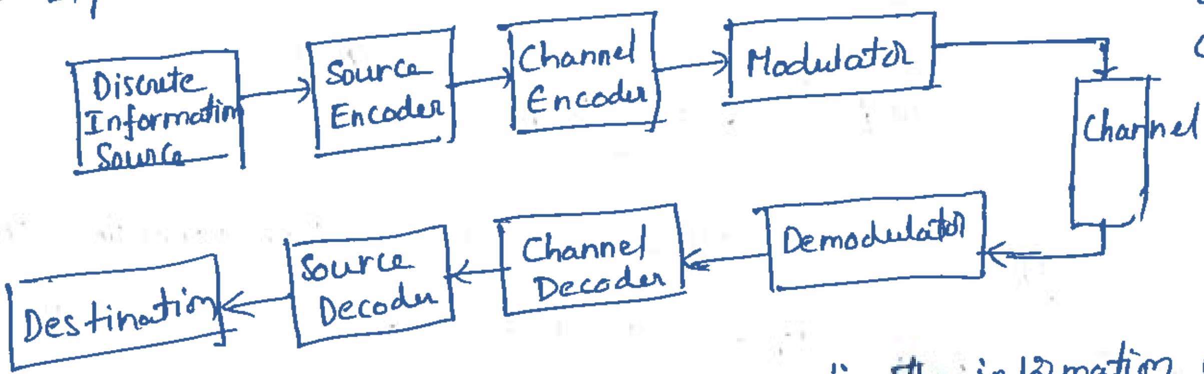
$\delta_{Ts}(t) \rightarrow$ impulse train Consist of unit impulses repeating periodically every T_s seconds where $T_s = 1/f_s$



13 (a) Explain the Basic elements of Digital Communication system ?

diagram 2M, Content 4M } 6M

Ans



Discrete Information Source : In Digital Communication the information in Discrete w.r.t time. This information is obtained by process of Sampling and Quantization. So the Discrete Information Source Can be letters, digits, Special Characters, code words . . .

Source Encoder and Decoder : The Symbols produced by the information Source are given to the Source encoder. The Symbols Can not be transmitted directly Source encoder Converts symbols into group of bits called

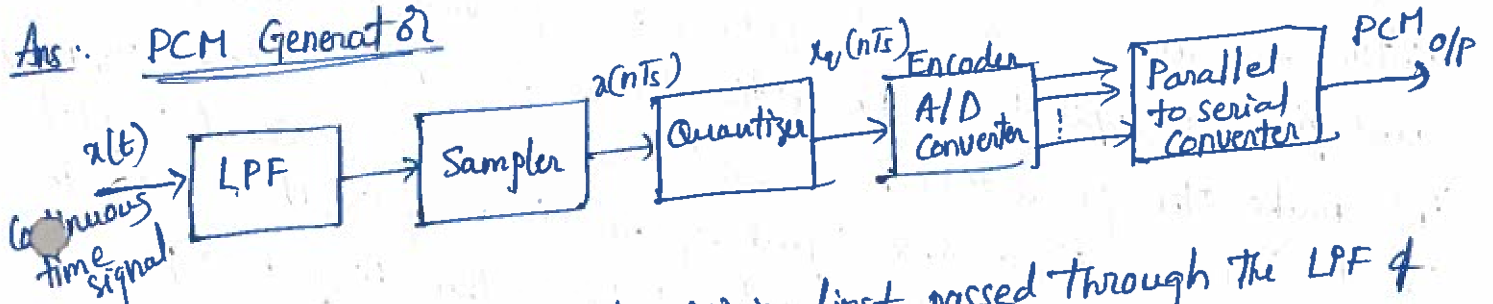
Code words (combination of 1's and 0's) for each distinct symbol there
is an unique code word. At receiver side ~~channel~~ Source decoder is used
it performs opposite operation of ~~channel~~ Source encoder.

Channel Encoder and Decoder: The input of the Channel encoder is
Binary Sequence. The communication channel adds noise and interference
to the signal being transmitted. Hence errors are introduced in the
binary sequence received at the receiver end. Thus, channel coding is
done to avoid this type of errors.

Digital Modulators and Demodulators: After converting into binary information
The pulses are to be transmitted by using digital modulation techniques
like ASK, FSK, PSK ... etc. depends on application Requirement.

13(b) With a neat diagram explain the Generation of PCM & DPCM.
[PCM 3M + DPCM 3M = 6M]

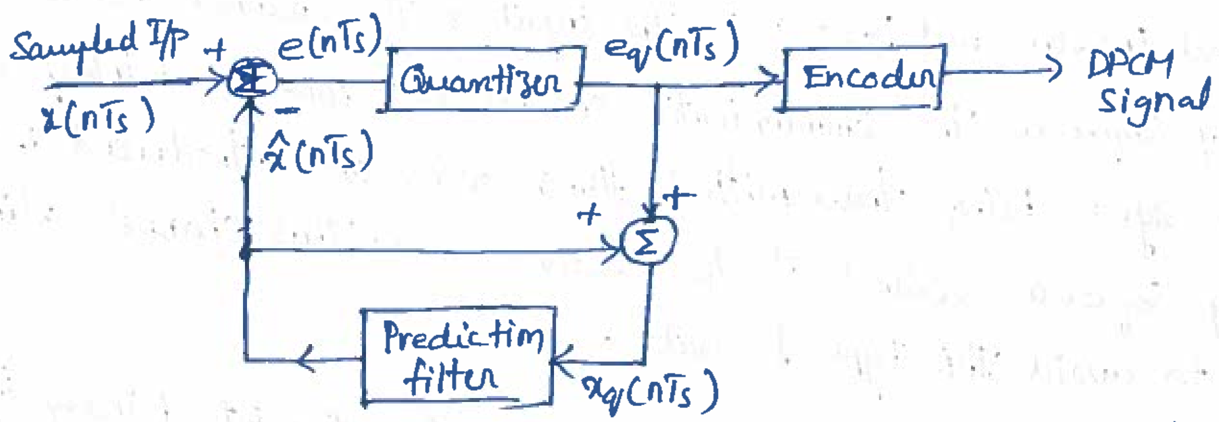
Ans: PCM Generator



In PCM generator the signal x(t) is first passed through the LPF of
cut-off frequency fm Hz. This LPF blocks all the frequency component which
are lying above fm Hz. Samples this signal at the rate of fs. The o/p of
Sampler is denoted by x(nTs). A Quantizer compares input x(nTs) with
its fixed digital levels. It assigns any one of the digital to x(nTs) with
its fixed digital levels. The o/p of quantizer given to the I/p of encoder.
This encoder converts input signal to 'v' digits binary word. encoder
o/p is given to the parallel to serial converter it converts parllel data
In to serial data it is suitable to transmission through channel.

DPCM Generation :

The DPCM works on the principle of prediction. The value of the present sample is predicted from the past samples.



The sampled signal is denoted by $x(nT_s)$ and the predicted signal is denoted by $\hat{x}(nT_s)$. The comparator finds out the difference between the actual sample value $x(nT_s)$ and predicted sample value $\hat{x}(nT_s)$. This is known as prediction error and it is denoted by $e(nT_s)$.

~~Thus, error is~~ The predicted value is produced by using a prediction filter. The Quantizer o/p signal $e_q(nT_s)$ and previous prediction is added and given as input to the prediction filter. This signal is called $x_q(nT_s)$. This makes the prediction more and more close to the actual sampled signal. The quantized error signal $e_q(nT_s)$ is very small and can be encoded by using smaller number of bits. Thus no. of bits per sample are reduced in DPCM.

12(b) Describe the Basic principles of PCM system and PCM transmitter.

[Principles of PCM system 2M + PCM Tx 4M = 6M]

Ans Principles of PCM system

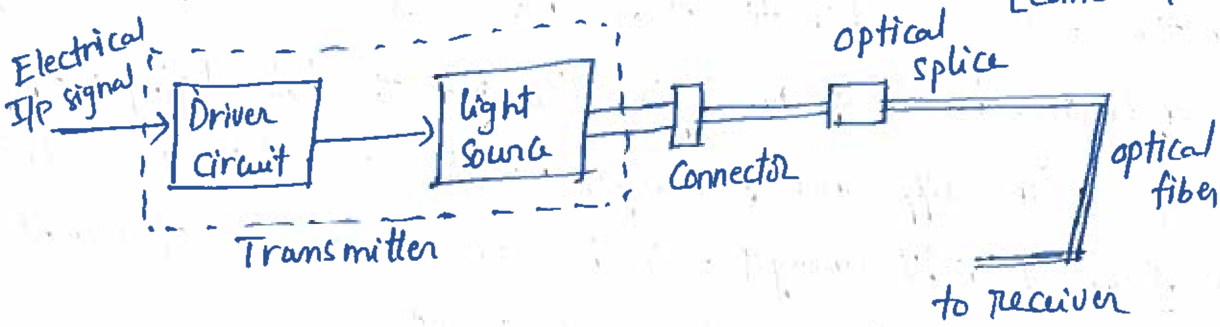
1) PCM is a digital pulse modulation system
2) PCM o/p is in the coded digital form.
3) PCM consist of a PCM encoder and PCM decoder
4) PCM is not modulation in the conventional sense.

PCM transmitter : refer 13 (b) Answer

(8)

14(a) Draw and Explain the working principle of an optical transmitter. [Content 4M + Diagram 2M = 6M]

Ans



Transmitter: The transmitter first converts the input voltage to current value which is used to drive the light source. Thus it interfaces the input circuit and light source.

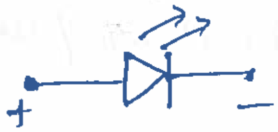
· The light source is normally an infrared LED or LASER device which is driven by the current value from the V to I converter. It emits light which is proportional to the input voltage value is generated and given as input to the fiber.

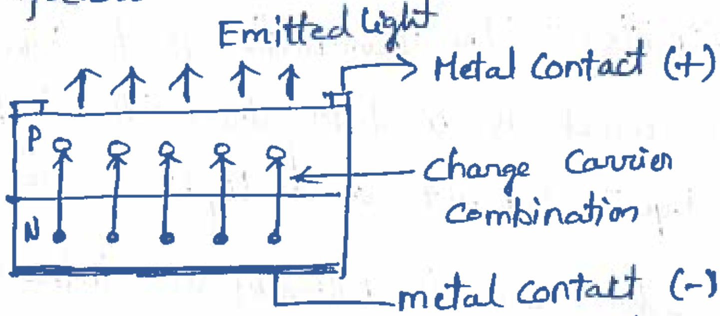
optical Splice: for creating long haul communication link, it is necessary to join one fiber to other fibers permanently.

14(b) Explain about LED and its type. (Content 4M + diagram 2M = 6M)

Ans · The Light Emitting Diode (LED) is a PN junction diode which emits light when forward biased, by a phenomenon called electroluminescence. In all semiconductor PN junctions, some of the energy will be radiated as heat and some in the form of photons. In Si and Ge the emitted light is insignificant. In other materials such as Gallium phosphide (GaP) or Gallium Arsenide phosphide (GaAsP), the number of photons of light energy emitted is sufficient to create a visible light source. Here the charge carrier recombination takes place when electrons from the N-side cross the junction and recombine with the holes on the p-side.

When LED is forward biased, The electrons and holes moves towards the junction and recombination takes place. As a result the e^- lying in The Conduction bands of N-region fall into the holes lying in the VB of P-region. The difference of energy b/w The CB and VB is radiated in the form of light energy. The brightness of The emitted light is directly proportional to the forward Current.





The Color of the emitted light depends on the type of material used.

Gallium Arsenide (Ga As) ⟶ infrared radiation (invisible)

Gallium Phosphide (GaP) ⟶ red & green

Gallium Arsenide phosphide (Ga AsP) ⟶ red & yellow.

15(a) Explain the working principle of GSM? (6M)

Ans: Global System for Mobile Communication (GSM) is a digital mobile network that is widely used by mobile phone users in the world

The GSM network has four Separate parts that work together to function as a whole. 1) Mobile Station (MS)

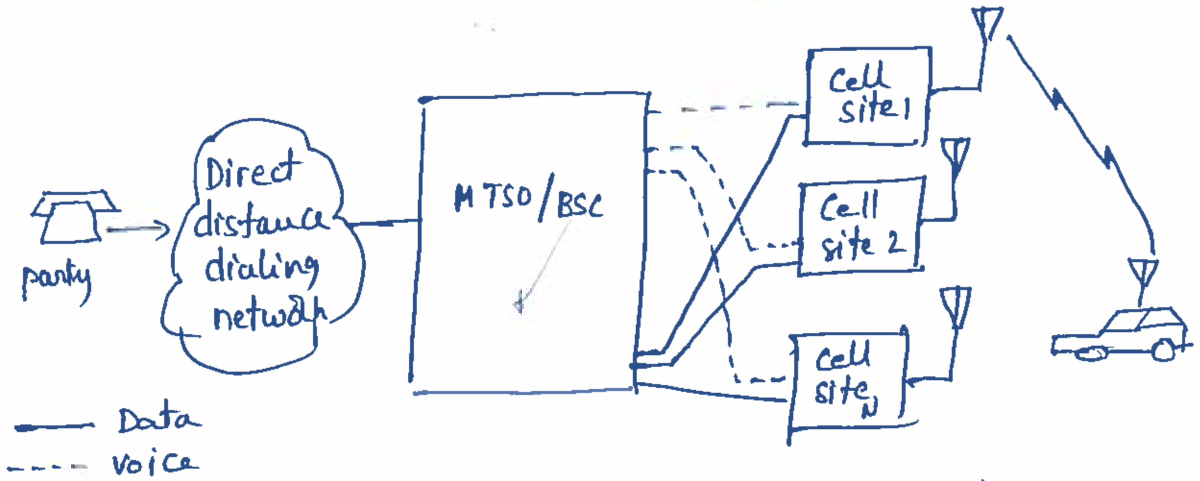
2) Base Station Subsystem (BSS)

3) Network Switching Subsystem (NSS)

4) Operation and support Subsystem (OSS)

15 (b) Explain Cellular Telephone System. [Diagram 3M + Content 3M = 6M]

Ans



A general View of Cellular Telephone System.

Antenna : Antenna pattern, antenna gain, antenna tilting and antenna
height all affect the cellular system design. The antenna pattern
Can be omnidirectional, directional & any shape in both the vertical and
The horizon planes. Antenna gain Compensates for the transmitted power.
Antenna gain at the mobile units would affect the system performance.

Switching Equipment : The Capacity of switching equipment in cellular
Systems is not based on the number of switch ports but on the Capacity
of the processor associated with the switches.

Data links : The Data link are not directly affected by the cellular system,
They are important in the system. Each data link can carry multiple
Channel data (10 kbps data transmitted per channel) from the cell site to
to the MTSO.

NSRIT

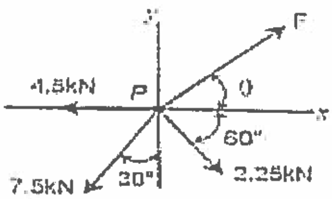
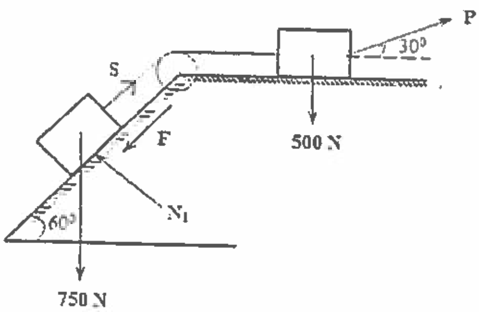
Semester End Examination, October, 2021

Degree	B. Tech. (U. G.)	Program	CE, EEE & ME	Academic Year	2020 - 2021
Course Code	20ESX04	Test Duration	3 Hrs. Max. Marks 70	Semester	II
Course	ENGINEERING MECHANICS				

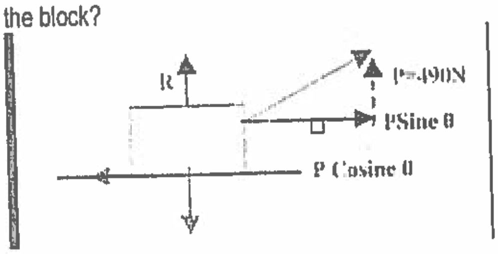
Part A (Short Answer Questions 5 x 2 = 10 Marks)

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Lami's Theorm,	20ESX04.1	L1
2	Write any four advantages and limitations of friction	20ESX04.2	L1
3	Differentiate between moment of inertia and polar moment of inertia	20ESX04.3	L2
4	Define and mention units for velocity of projection	20ESX04.4	L1
5	Write Impulse Momentum Method.	20ESX04.5	L1

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

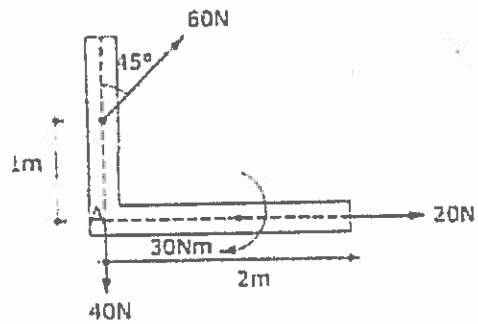
No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Derive and explain about Parallelogram Law	6M	20ESX04.1	L2
6 (b)	State and prove Triangular law of forces	6M	20ESX04.1	L3
	OR			
7 (a)	State and Explain the concept of Equilibrium	4M	20ESX04.1	L2
7 (b)	Determine the magnitude and angle and F so that particle shown in figure, is in Equilibrium 	8M	20ESX04.1	L2
8 (a)	What is the value of P in the system shown in the figure to cause the motion to impend? Assume the pulley is smooth and coefficient of friction between the other two contact surfaces is 0.20 	8M	20ESX04.2	L3
8 (b)	Define the following (i) Law of transmissibility (ii) Converse of the Law of Polygon of Forces	4M	20ESX04.2	L2
	OR			

9 (a) A pull of 490 N inclined at 30° to the horizontal is necessary to move a block of wood on a horizontal table. If the coefficient of friction between two bodies in contact is 0.2. What is the mass of the block? — 7M — 20ESX04.2 — L2



9 (b) Differentiate between the angle of repose and angle of friction — 5 M — 20ESX04.2 — L3

10 (a) Locate the centroid of L – section shown in figure — 7M — 20ESX04.3 — L3



10(b) Explain briefly about Centre of Gravity using Varigmon's theorem — 5M — 20ESX04.3 — L2

OR

11 (a) Determine the centroid of a triangle having base width b and height h — 6M — 20ESX04.3 — L3

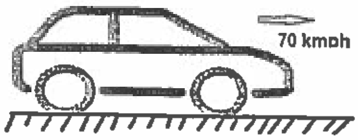
11(b) Locate the centroid of the following figure — 6M — 20ESX04.3 — L2

12 (a) A man weight 100 Newton entered a lift, which moves with an acceleration of 5 m/sec². Find the force exerted by the man on the floor of lift when — 5M — 20ESX04.4 — L3

a) Lift is moving downward
b) Lift is moving upward

12(b) A motorist travelling at a speed of 80 kmph, suddenly applies brakes and halts after 70 m. Determine — 7M — 20ESX04.4 — L3

a) The time required to stop the car
b) The coefficient of friction between the tyres and the road



OR

13(a)	A Particle is projected vertically upwards from the ground with an initial velocity of 10 m/sec. find a) The time taken to reach the maximum height b) The maximum height reached c) Time required for descending d) Velocity when it strikes the ground. Consider the upward motion of the particle	6M	20ESX04.4	L3
13(b)	A small Steel ball is shot vertically upwards from the top of a building 45m above the ground with an initial velocity of 28 m/sec a) In what time, it will reach the maximum height. b) How high above the building will the ball rise	6M	20ESX04.4	L3
14	Find the Power of a locomotive, drawing a train whose weight including that of engine is 500 kN up an incline 1 in 135 at a steady speed of 60 kmph, the frictional resistance being 6 N/kN. While the train is ascending the incline, the steam is shut off. Find how far it will move before coming to rest, assuming that the resistance to motion remains the same	12M	20ESX04.5	L3
	OR			
15	Derive the Work Energy equation for translation about Fixed Axis	12M	20ESX04.5	L3

Engineering Mechanics

20 ES X04.

Semester End Examinations

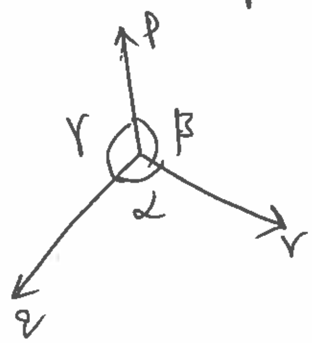
oct, 2021.

I. Part-A Short Answers $5 \times 2 = 10M$

1. Define lami's theorem (2m)

Ans: If three forces acting at a point are in a equilibrium then each force is directly proportional to the sine of the angle of b/w the other two forces.

$$\boxed{\frac{R}{\sin\gamma} = \frac{P}{\sin\alpha} = \frac{Q}{\sin\beta}}$$



2. Advantages & limitations of friction.

Ans: Friction help to walk, turn and stop (2m)
produces unwanted heat
Reduces Efficiency of machines
Eco-hazard.

3. Difference b/w moment of inertia and polar moment of Inertia. (2m)

Ans:

Moment of Inertia	Polar Moment of Inertia
① Moment of Inertia used to measure an object's ability to oppose angular acceleration	it is measurement of an object's ability to oppose torsion
② $I = r^2 dm$	$J = r^2 dA$
③ $kg\,m^2$	m^4.
④ Depends on the mass of a Body	geometry of the body.

④ Velocity of projection. (2M)

Ans: velocity with which the projectile is projected. velocity of projection largely determined the various aspects of the projectile such as time of flight, range, etc.

Trajectory _ it is the path followed by the projectile after it has been projected.

⑤ Impulse momentum. (2M)

Impulse applied to an object produces an equivalent vector change in its linear momentum, also in the result direction

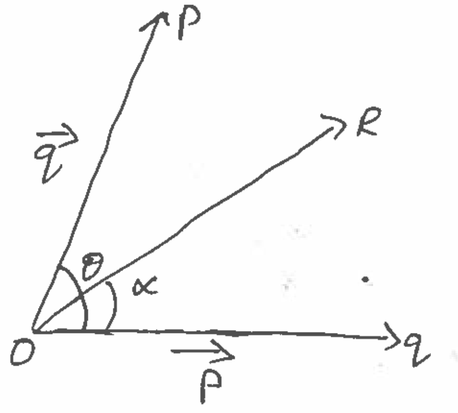
$$\Delta p = F \Delta t$$

Δp = change in momentum

F = applied force ; Δt = elapsed time

6(a) Derive and explain about parallelogram Law?

Ans:- parallelogram law:-



Magnitude :-

$$R = \sqrt{p^2+q^2+2pq\cos\theta}$$

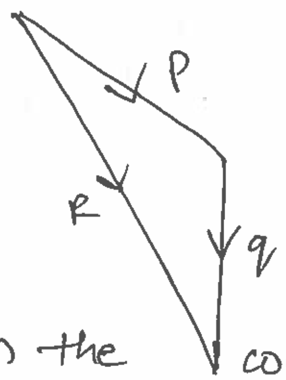
Direction :-

$$\alpha = \tan^{-1}\left[\frac{q\sin\theta}{p+q\cos\theta}\right]$$

6(b) State and prove Triangular law of forces.

A. Law of triangular forces:-

"If two types forces p and q are represented by the two sides of a triangle both in magnitude and the direction which are in same order then the closing side of the triangle gives the resultant of these two. forces both in magnitude and direction but in opposite order"

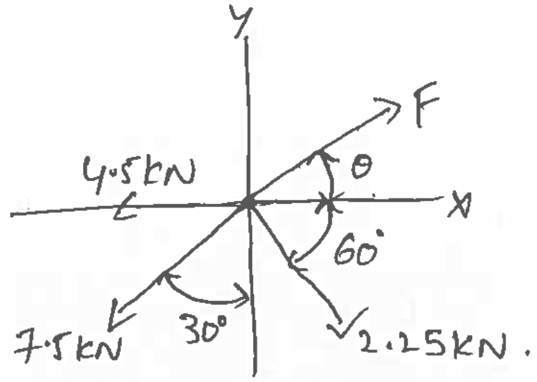


7(a) State, and explain the concept of Equilibrium.

A. concept of equilibrium:-

when some external forces (which may be parallel or concurrent) are acting on a stationary body then the body may start moving or will start rotating any point but doesn't start moving and also doesn't start's rotating about any points then the body is said to be in equilibrium.

7(b) Determine the magnitude and angle and F so that particle shown in figure, is in equilibrium.



$$R = \sqrt{(\Sigma H)^2 + (\Sigma V)^2}$$

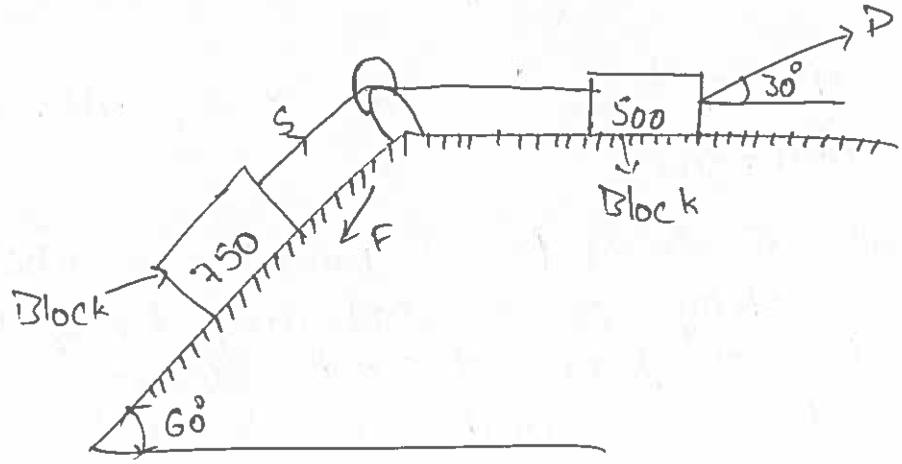
$\Sigma H = P_1 \cos\theta_1 - P_2 \cos\theta_2 - P_3 \cos\theta_3 + P_4 \cos\theta_4.$

$\Sigma V = P_1 \sin\theta_1 + P_2 \sin\theta_2 - P_3 \sin\theta_3 - P_4 \sin\theta_4.$

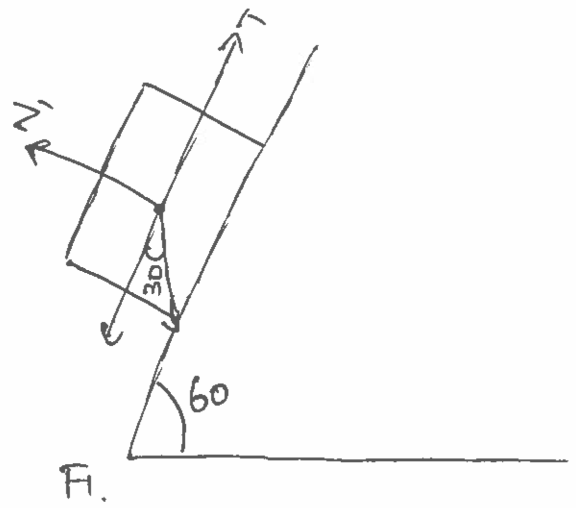
Magnitude:- $R = \sqrt{(\Sigma H)^2 + (\Sigma V)^2}$

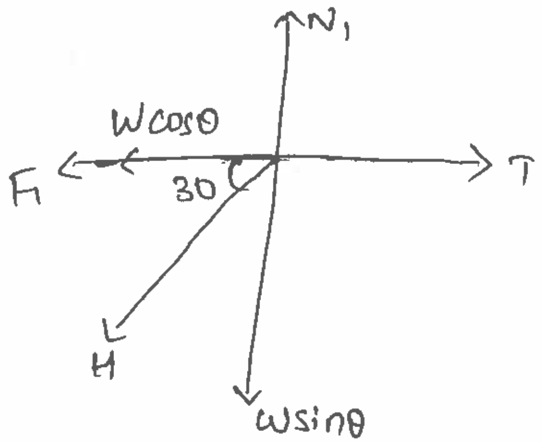
Direction :- $\tan\theta = \frac{\Sigma V}{\Sigma H}.$

8(a) What is the value of P in the figure to cause the motion to impend? Assume the pulley is smooth and coefficient of friction between the other two contact surface is 0.20.



A. Equilibrium of 750 N Block:





$\therefore\ \Sigma x = 0.$

$\Rightarrow\ T - W\cos 30 - F_1 = 0.$

$T = 750\cos 30 + F_1$

$T = 649.51 + \mu_1 N_1$

$T = 649.51 + 0.2 N_1$

$\therefore\ \Sigma y = 0.$

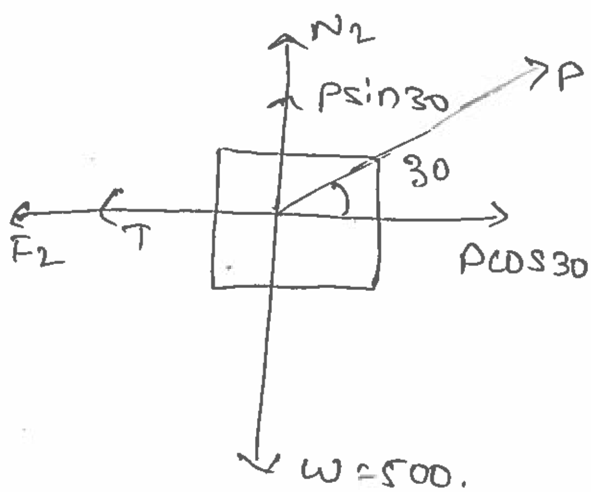
$= N_1 - W\sin 30 = 0.$

$N_1 = 750\sin 30.$

$\boxed{N_1 = 375}$

$\therefore\ T = 649.51 + (0.2)(375)$

$\boxed{T = 724.51 N}$



$\Sigma H = 0$

$\Rightarrow P\cos 30 - F_2 - T = 0$

$= P\cos 30 - 724.5 - F_2 = 0.$

$= (0.86)P - 724.5 - F_2 = 0.$

$= (0.866)P = 724.5 + (0.2)N_2$

$\Sigma V = 0.$

$\Rightarrow N_2 + P\sin 30 = 500.$

$P/2 = 500 - N_2$

$P = 1000 - 2N_2$

$N_2 = -P/2 + 500.$

$\Rightarrow (0.866)P = 724.5 + (0.2)(-P/2 + 500)$

$(0.866)P = 724.5 - (0.1)P + 100$

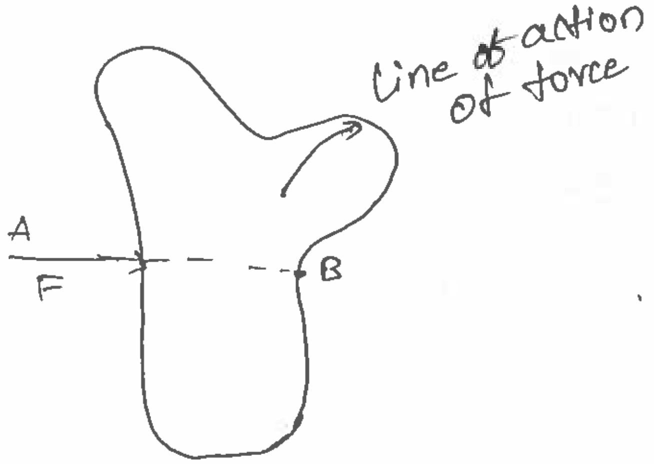
$(0.966)P = 824.5$

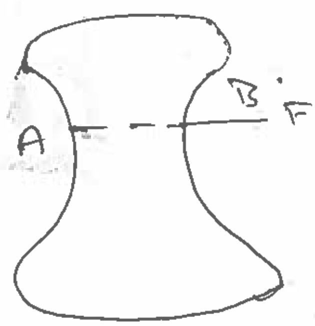
$P = \frac{824.5}{0.966}$

$\boxed{P = 853.5 IN}$

8b) i) Law of transmissibility!.

As per law of transmissibility or forces the force can be replaced from one point to other within the same point in the line without any effect to the actual system.

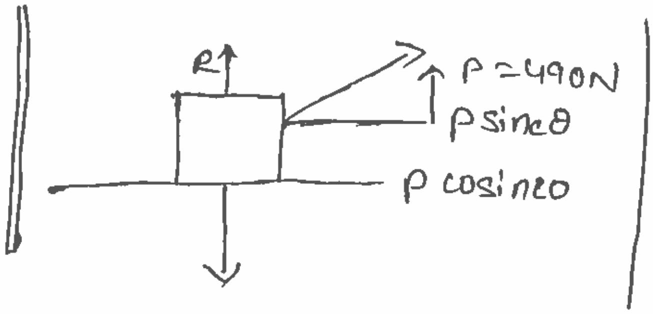




ii) Laws of polygon forces:-

"If 'n' forces are represents at the 'n' sides of the polygon both in magnitude and direction and which are in same order then the closing side of the polygon gives the resultant of the 'n' forces both in magnitude and direction but in opposite order".

9a)



$\Sigma H = 0.$

$\Rightarrow 180 \cos 30 - F_1 = 0.$

$\boxed{F_1 = 155.88N}$

$\Rightarrow \mu N = 155.88N$

$\Sigma V = 0$

$= N + 180 \sin 30 - W = 0.$

$\boxed{N + 90 = W}$

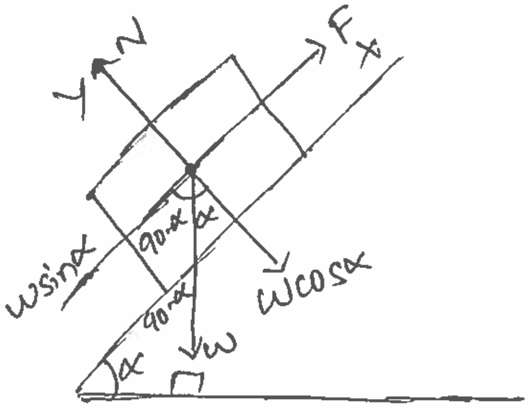
$N = W - 90.$

$\Rightarrow \mu(W-90) = 155.88$

$\therefore \mu(W-90) = 155.88 \longleftarrow (1).$

9b. Angle of Repose (θ)!

The Angle of repose is defined as the maximum inclinati-n of a plane at which body remains in equilibrium over the incline plane by assitance of friaction only.



$\Rightarrow \Sigma X = 0.$

$F - W\sin\alpha = 0.$

$\Rightarrow \Sigma y = 0.$

$$N - W\cos\alpha = 0.$$

$$N = W\cos\alpha$$

$$\frac{F}{N} = \frac{W\sin\alpha}{W\cos\alpha}$$

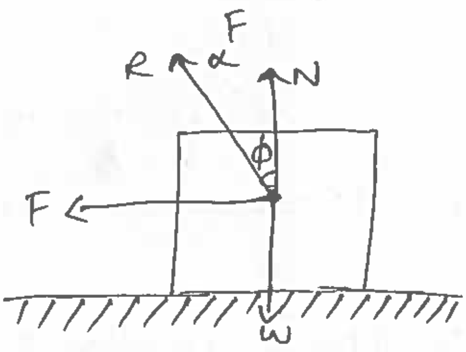
$$\tan\alpha = F/N = \mu = \tan\phi$$

$$\mu = \tan\phi$$

$$\boxed{\alpha = \phi}$$

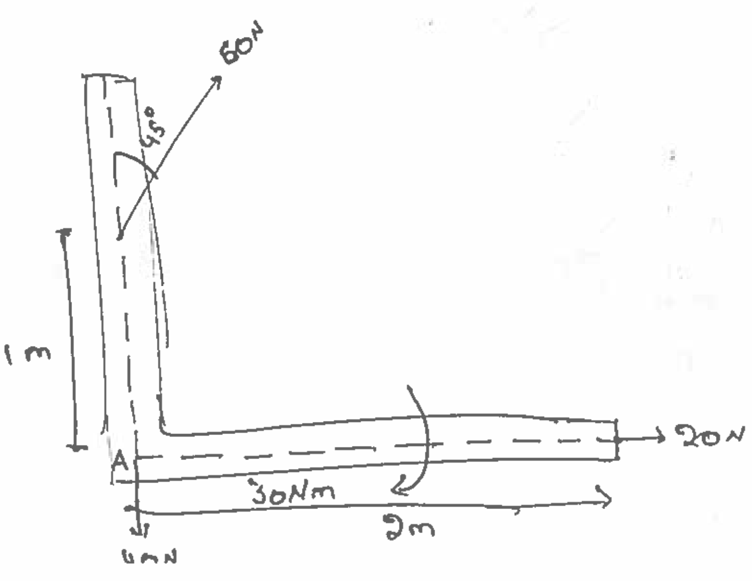
* Angle of friction (ϕ):

It is defined as the angle made by the Resultant of the Normal reaction N and limiting frictional force F with the Normal reaction 'N' and it is denoted by 'ϕ'



$$\boxed{\tan\phi = F/N = \mu}$$

10(a)



$$C = \left(\frac{A_1 x_1 + A_2 x_2 + \cdots}{A_1 + A_2}, \frac{A_1 y_1 + A_2 y_2 + \cdots}{A_1 + A_2} \right)$$

① $A_1 = 100 \times 20.$

$= 2000 \text{ mm}^2$

$x_1 = 50$

$y_1 = 10.$

② $A_2 = 80 \times 30 = 2400 \text{ mm}^2$

$x_2 = 15$

$y_2 = 20 + 40 = 60.$

$$C = \left(\frac{2000(50) + 2400(15)}{2000 + 2400}, \frac{2000(10) + 2400(60)}{2000 + 2400} \right)$$

$$= \left(\frac{100000 + 56000}{4400}, \frac{20{,}000 + 144000}{4400} \right)$$

$$= \left(\frac{136000}{4400}, \frac{164{,}000}{4400} \right)$$

$$C = (30.90, 37.27)$$

10b) Centre of Gravity

$W = W_1 + W_2 + \cdots$

$= \Sigma W_i$

$W\bar{x} = \Sigma w_i x_i$

$W\bar{y} = \Sigma w_i y_i$

$W\bar{z} = \Sigma w_i z_i$

mass $m\bar{x} = \Sigma m_i x_i$

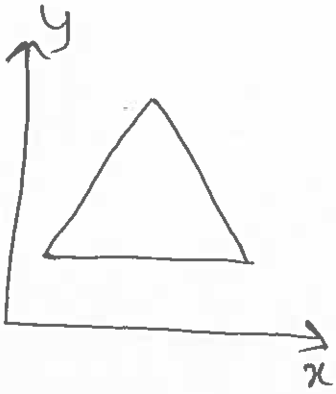
$m\bar{y} = \Sigma m_i y_i$

$M\bar{z} = \Sigma m_i z_i.$

same of volume also

$\boxed{W = V i r}$.

11a)

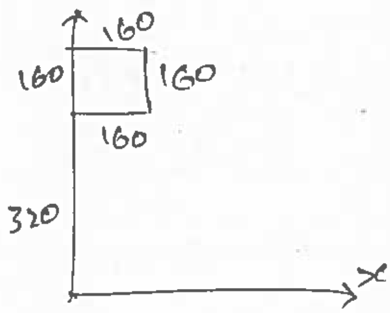


$$\text{Centroid} = \left(\frac{A_1 x_1 + A_2 x_2}{A_1 + A_2}, \frac{A_1 y_1 + A_2 y_2}{A_1 + A_2} \right)$$

$$\text{Centroid} = \left(b/3, \; b/3 \right)$$

11b)

$A_2 = 16 \times 16 = 25600 mm^2$



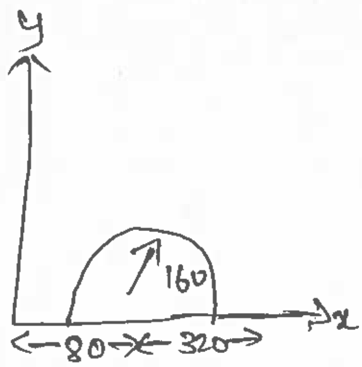
$x_2 = 80 mm$

$y_2 = 400 m$.

$$A_3 = \frac{\pi r^2}{2} = \frac{3.14 \times (160)^2}{2}$$

$$= \frac{3.14 \times 25600}{2}.$$

$$= 40,192 \, mm^2$$



$$x_3 = \frac{80 + 320}{2} = 80 + 320/2$$

$$= 80 + 160.$$

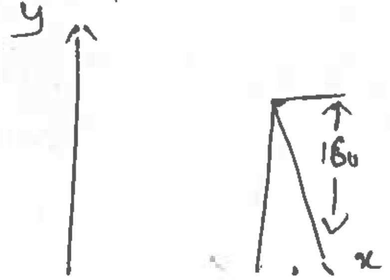
$$= 240 mm.$$

$$y_3 = \frac{4r}{3\pi} = \frac{4 \times 160}{3 \times 3.14} = \frac{690.}{9.42} = 67.94$$

$$A_4 = \frac{1}{2} \times b \times h$$

$$= 1/2 \times 240 \times 160$$

$$= 38400/2$$



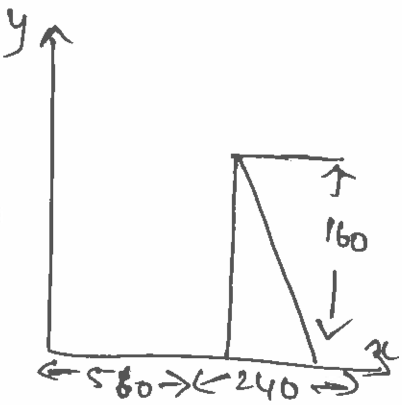
$= 19{,}200\ mm^2$

$x_4 = 560 + b/3 = 560 + 240/3$

$= 560 + 80.$

$= 640\ mm.$

$y_4 = \frac{160}{3} = 53.33\ mm.$



12a). $\Sigma V = 0.$

$R - W + f = 0.$

$R - w + w/g\ (a) = 0.$

$\boxed{R = -w/g\ (a) + w}$

② $\Sigma V = 0.$

$R - w - F = 0.$

$R - w - w/g\ (a) = 0.$

$\boxed{R = w + w/g\ (a).}$

12b)

i) $v^2 - u^2 = 2as$

$v = u + at$

$v = 0.,\ u = \frac{70 \times 1000}{\cancel{36 \times 36}\ 60 \times 60.} = \frac{70000}{3600} = 19.4\ m/s.$

$v^2 - u^2 = 2as.$

$v^2 - u^2 = 2as.$

$0 - (19.4)^2 = 2a \times 50.$

$-(19.4)^2 = 100a$

$\frac{-376.36}{100} = a.$

$\boxed{a = -3.78}\ m/sec^2.$

$v = u + at.$

$0 = 19.4 + (-3.76)t$

$+19.4 = +3.76t.$

$\frac{19.4}{} = t.$

$\boxed{t = 5.15}$

ii) $\Sigma H = 0.$

$w/g\,(a) - \mu N = 0$ —— (1)

$\Sigma V = 0.$

$-w + N = 0.$

$\boxed{= N = w}$

$= w/g\,(a) = \mu w = 0.$

$\cancel{w}/g\,(a) = \mu \cancel{w}$

$\boxed{\mu = a/g}$

$\mu = \frac{-3.78}{9.81}$

$\boxed{\mu = 0.38}$

13a) i) Time takens:

$v = u + at$

$0 = u + (-g)t$

$0 = u - gt.$

$+u = +gt$

$\boxed{t = u/g}$

$\boxed{\begin{matrix} u = -g = -981 \\ u = g = 981 \end{matrix}}$

(iii) $v = u + at$

$0 = u + gt$

$0 - u = gt$

$\boxed{t = -u/g}$

ii) $v^2 - u^2 = 2as.$

$0 - u^2 = 2 \times (-g) \times h$

$+u^2 = +2gh$

$\boxed{h = u^2/2g}$

iv) $v^2 = \cancel{2}\cancel{g}\left(\frac{u^2}{\cancel{2g}}\right)$

$v^2 = u^2$

$\boxed{v = u}$

13b) . i) $v = u + at$

$0 = 18 + (-g)t$.

$0 = 18 - 9.81\,t$

$+18 = -9.81t$

$t = \frac{18}{9.81} = 1.8.$

$\boxed{t = 1.8\ sec}$

ii) $v^2 - u^2 = 2as.$

$-(18)^2 = 2(-g)h$

$+(18)^2 = +2 \times 9.81 \times h$

$\frac{324}{19.62} = h$

iii) Total $h = 16.5 + 25.$

$\boxed{h = 41.5m}$

iv) $v^2 - u^2 = 2as.$

$v^2 - 0 = 2 \times 9.81 \times 41.5.$

$v^2 = 28.5\ m/s.$

$v^2 = 28.5 M/s.$

v) $v = u + at.$

$28.5 = 0 + (9.81)t$

$\frac{28.5}{9.81} = t.$

$\boxed{t = 2.90\ sec}$

Total time $= 2.90 + 1.8$

$t = 4.70 Sec.$

14) $N = W = 420.$

Speed $= 56 km/ph$

$M =$ Resistance $= 5N/kN$

velocity $= \frac{56 \times 1000}{60 \times 60} = 15.5 M/sec.$

$F = MN$

$F = 5 \times 420.$

$F = 2100 = 2.1N.$

$\Sigma H = 0.$

$P - W\sin\theta - F = 0.$

$P - 420 \times \frac{1}{120} - 2.1 = 0$

$P - \frac{420}{120} - 2.1 = 0.$

$P - 3.5 - 2.1 = 0.$

$P = 3.5 + 2.1$

$\boxed{P = 5.6\ kw}$

power of the locomotive = Work done of p/section

$= P \times V$

$= 5.6 \times 15.5$

$$\boxed{P = 86.80 \text{ kN}}$$

Energy = KE.

$P \times S = \frac{1}{2} mv^2$

$5.6 \times S = \frac{1}{2} \times \frac{W}{g} (15.5)^2$

$5.6 \times S = \frac{1}{2} \times \frac{420}{9.81} \times 240.25$

$5.6 \times S = \frac{1}{2} \times 42.81 \times 240.25$

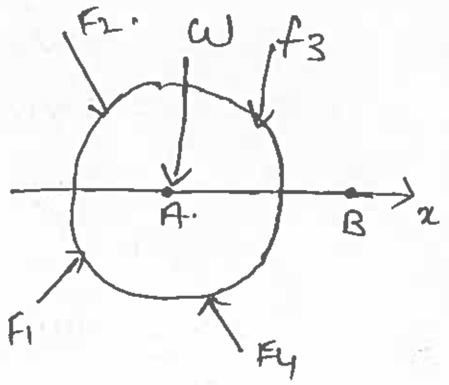
$S = \frac{42.81 \times 240.25}{5.6 \times 2}$

$S = \frac{10285.10}{11.2}$

$$\boxed{S = 918.31 \text{ m}}$$

15) Work energy equation for Translation.

Consider the body subject to a system of forces F_1, F_2 and moving with an acceleration a in x-direction, let its intial velocity of a be 'u' and final velocity when it moves distance $AB = S$ be 'v' Then the resultant of system of forces must be in 'x' direction.



let

$$R = \Sigma f_x.$$

from newton's second law of motion

$$R = \frac{W}{g} a$$

multiplying both sides by elementary distance ds, we get.

$$R\,ds = \frac{W}{g}\,a\,ds \qquad \left[a = v\frac{dv}{ds}\right]$$

$$= \frac{W}{g}\,v\,\frac{dv}{ds}\,ds$$

$$= \frac{W}{g}\,v\,dv.$$

Integrating both sides for the motion from.

A to B we get

$$\int_0^s R\,ds = \int_u^v \frac{W}{g}\,v\,dv.$$

$$Rs = \frac{W}{g}\left[\frac{v^2}{2}\right]_u^v$$

$$= \frac{W}{2g}(v^2 - u^2)$$

Now, Rs is the work done by the forces acting on the body $\frac{W}{2g}v^2$ is finial kinetic energy and $\frac{Wu^2}{2g}$ is ~~init~~ initial kinetic energy.

Hence, we can say, work done in a motion is equal to change in kinetic energy that is.

Work done = Final kinetic energy - Initial kinetic energy.

and it is called work energy equation.

This work energy principle may be stated as the work done by a system of forces acting on a body during a dis placement is equal to the change in kinetic energy of the body during the same displacement.

$$F = m\frac{dv}{dt}$$

$$= m\frac{dv}{dx}\cdot\frac{dx}{dt}$$

$$= mv\frac{dv}{dx}$$

Integrating both sides with respect to x between limits

$$F\,dx = mv\,dv$$

$$\int F\,dx = \int mv\,dv$$

$$= \frac{1}{2}mv^2\Big|$$

$$= \frac{1}{2}m(v^2-u^2)$$

Now we see that work done by the force acting on the body. $\frac{1}{2}mv^2$ is defined as the energy or $\frac{1}{2}mv^2$ is kinetic energy.

Work done by a force on a body results in equal change of kinetic energy.

Work done = Final kinetic energy − Initial kinetic energy

This is called work energy equation.

This work energy principle may be stated as the work done by a system of forces acting on a rigid body during a displacement is equal to the change in kinetic energy of the body during the same displacement.

NSRIT

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	**Program**	ECE			**Academic Year**	2020 - 2021
Course Code	20EC201	**Test Duration**	3 Hrs.	Max. Marks	70	**Semester**	II
Course	**Principles of Electronics & Communication Systems**						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define Fermi level	20EC201.1	L1
2	What is CMRR?	20EC201.2	L1
3	What is the need for modulation?	20EC201.3	L1
4	Define PAM and PPM	20EC201.4	L1
5	Define TIR	20EC201.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Explain Insulator, Semiconductor & conductor with help of energy band structure	6M	20EC201.1	L2
6 (b)	Differentiate between intrinsic and extrinsic semiconductor	6M	20EC201.1	L2
	OR			
7 (a)	Explain n-type semiconductor	6M	20EC403.1	L2
7 (b)	Derive the expression for current generated due to drifting of charge carriers in semiconductors in the presence of electric field	6M	20EC403.1	L2
8 (a)	Explain application of op-amp as integrator and differentiator	6M	20EC201.2	L2
8 (b)	Explain ac characteristics of op-amp	6M	20EC201.2	L2
	OR			
9 (a)	Draw and explain the pin diagram IC741 op-amp	6M	20EC201.2	L2
9 (b)	Derive the gain for non-inverting op-amp	6M	20EC201.2	L2
10 (a)	State and explain properties of continuous signals	8M	20EC201.3	L2
10 (b)	List any four applications of FM system	4M	20EC201.3	L2
	OR			
11 (a)	Define amplitude modulation. Derive an expression for the AM wave	8M	20EC201.3	L2
11 (b)	Write about am voltage distribution	4M	20EC201.3	L2
12 (a)	State and prove sampling theorem	6M	20EC201.4	L2
12 (b)	Describe the basic principles of PCM system and PCM transmitter	6M	20EC201.4	L2
	OR			
13 (a)	Explain the basic Elements of Digital Communication System	6M	20EC201.4	L2
13(b)	With a neat diagram explain the Generation of PCM & DPCM	6M	20EC201.4	L2
14(a)	Draw and explain the working principle of an Optical transmitter	6M	20EC201.5	L2
14 (b)	Explain about LED and its type	6M	20EC201.5	L2
	OR			
15(a)	Explain the working principle of GSM	6M	20EC201.6	L2
15(b)	Explain Cellular Telephone Systems	6M	20EC201.6	L2

Degree : B.Tech   Program: ECE   Semester : II

Course Code: 20EC201   Course: Principles of Electronics & Communication Systems

Key and Scheme of Evaluation

PART A

1. Define Fermi level?   (2M)

Ans. The Fermi level E_F indicates the probability of occupancy of an energy level by an electron.

2. What is CMRR?   (2M)

Ans. It is defined as the ratio of the differential Voltage gain Ad to the Common mode Voltage gain A_{cm}.

$$CMRR = A_d / A_{cm}$$

This parameter indicates the Capability of the op-amp to reject noise.

3. What is the need for Modulation?   (2M)

Ans: 1) To reduce the antenna height.
2) for Multiplexing of Signals.
3) To increase the range of Communication.
4) To reduce Noise and interference.

4. Define PAM and PPM.   (PAM - 1M   PPM - 1M)

Ans. PAM: Pulse Amplitude Modulation is a process of Changing the amplitude of high frequency periodic rectangular pulse in accordance with the amplitude of message signal.

PPM: pulse position Modulation is a process of Changing the position of high frequency periodic rectangular pulse in accordance with the amplitude of Message signal.

5. Define TIR ? (2M)

Ans : When the incident angle is increased beyond the critical angle, the light ray does not pass through the interface into the other medium. In this condition angle of reflection ϕ_2 is equal to the angle of incidence ϕ_1. This action is called as Total Internal Reflection (TIR) of the beam.

PART - B

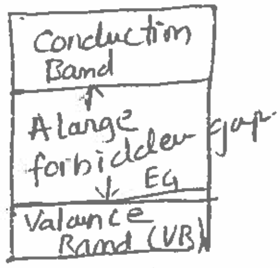
6(a) Explain Insulator, semiconductor & conductor with help of energy band structure. (2M + 2M + 2M = 6M)

Ans

Insulators: Insulators passes no free charge carriers and thus are non - conductive. Insulators are implemented in household items and electrical circuits as protection.

Insulators possess a high resistivity and low conductivity. Their atoms have tightly bound electrons that donot move throughout the material. Because the electrons are static and not freely roaming, a current cannot easily pass.

Eg: Rubber, Teflon, Cloth, wood and fibreglass



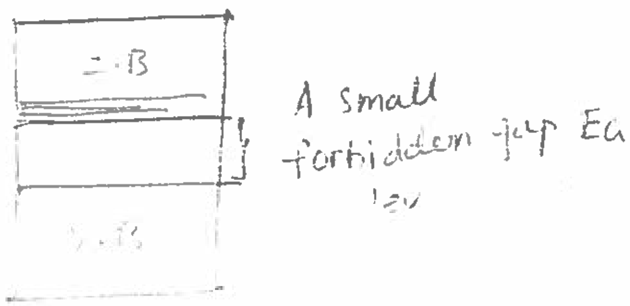
(a) Insulator

Semiconductor

In semiconductors the gap between Valance Band and Conduction band is smaller

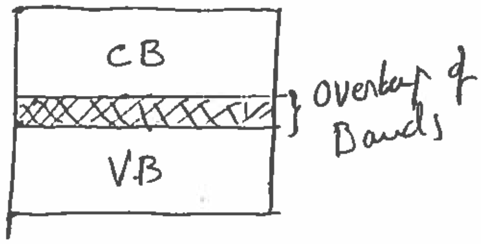
Ex: Ga, As, Si and Ge

At room temperature there is sufficient energy available for electrons to make a transition from V.B to C.B. This allows some conduction to takes place.



Conductor : A conductor is defined as an object or type of material that allows the flow of charge in one or more directions. Materials made of metal are common electrical conductor, as metal have a high conductance and low resistance.

Eg: Aluminium, Silver, Copper etc.



6(b) Differentiate between intrinsic and extrinsic semiconductor.

Ans

Intrinsic semiconductor	Extrinsic semiconductor
1. pure form of semiconductor	1. Impure form of semiconductor.
2. It exhibits poor conductivity	2. It possesses comparatively better conductivity than intrinsic semiconductor
3. It is present in the middle of forbidden energy gap.	3. The presence of fermi level varies according to the type of extrinsic semiconductor
4. The conduction relies on temperature.	4. The conduction depends on the concentration of dopped impurity and temperature.
5. Equal amount of electron and holes are present in C.B & V.B	5. The majority presence of electrons and holes depends on the type of extrinsic semiconductor
6. It is not further classified	6. It is classified as p-type and n-type.

Marks: each difference 1 Mark

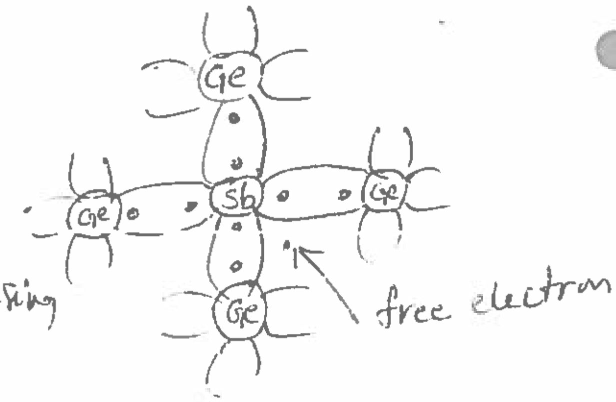
total - 6 Marks

7(a) Explain n-type semiconductor? Content - 4M diagram - 2M

Ans: A small amount of pentavalent impurities such as Arsenic, antimony or phosphorus is added to the pure Semiconductor (germanium or silicon Crystal) to get N-type semiconductor.

Ge atom has four Valance electrons and antimony has five Valance electron each antimony atom forms a Covalent bond with surrounding four germanium atoms. Thus, four Valance electrons of antimony atom form Covalent bond with four Valance electrons of individual Germanium atom and fifth Valence electron is left free which is loosely bound to the antimony atom.

This loosely bound electron Can be easily excited from the V.B to the C.B by the application electric field or increasing the Thermal energy.



7(b) Derive the expression for current generated due to drifting of Charge Carriers in Semiconductors in the presence of electric field?
[Content : 2M Equations: 4M = 6M]

Ans: When an electric field is applied accross the Semiconductor material, the Charge Carriers attain a certain drift Velocity V_d. Which is equal to the product of the mobility of the Charge Carriers and the applied electric field intensity, E. The holes move towards the negative terminal of the battery and electrons move toward the positive terminal. This Combined effect of movement of the Charge Carriers Constitutes a Current known is the Drift Current.

Thus the Drift Current is defined as the flow of electric current due to the motion of the Charge Carriers under the influence of an external electric field.

The equation for the drift Current density J_n, due to free electrons given by $J_n = qn\mu_n E$ A/cm²

and the drift Current density J_p, due to holes is given by

$$J_p = qp\mu_p E \ \ A/cm^2$$

Where n = number of free electrons per Cubic Centimeter

p = number of holes per Cubic Centimeter

μ_n = mobility of electrons in cm²/V-s

μ_p = mobility of holes in cm²/V-s

E = applied electric field Intensity in V/cm

q = Charge of an electron = 1.6×10^{-19} Coulomb.

3(b) Explain ac Characteristics of op-amp. [each Characteristic 2M, 3X2 = 6M]

Ans: slew Rate: It is defined as the maximum rate of Change of output Voltage with time.

The slew rate is specified in V/μsec. Thus

$$\text{Slew rate} = S = \left.\frac{dV_o}{dt}\right|_{max}$$

Transient Response Rise time:

When the o/p of the op-amp is suddenly changing like pulse type Then the rise time of the response depends on the cut-off frequency f_H of the op-amp. Such a rise time is called cut-off frequency limited rise time or transient response rise time. It is inversly propotional to the cut-off frequency and given by

$$t_r = \frac{0.35}{F_H}$$

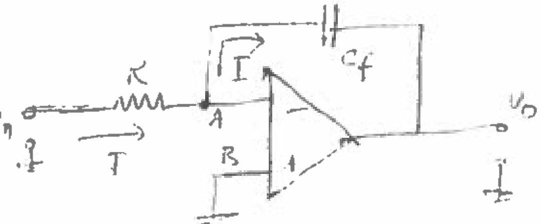
When t_r = rise time f_H = Cut-off frequency

frequency Response of op-amp

The plot showing the variations in magnitude and phase angle of the gain due to the change in frequency is called frequency response of the op-amp.

3(a) Explain application of op-amp as Integrator & Differentiator [Integrator – 3M, Differ – 3M]

Ans Integrator: In an integrator circuit, the o/p voltage is the integration of the I/p voltage.



Consider the op-amp integrator ckt. The node B is grounded. The node A is also at the ground potential from the concept of virtual ground.

from the I/p side we can write

$$I = \frac{V_{in} - V_A}{R_1} = \frac{V_{in}}{R_1} \quad \text{--- (1)}$$

from o/p side we can write

$$I = C_f \frac{d(V_A - V_o)}{dt} = -C_f \frac{dV_o}{dt} \quad \text{--- (2)}$$

equating eq (1) & (2)

$$\frac{V_{in}}{R_1} = -C_f \frac{dV_o}{dt}$$

Integrating both sides

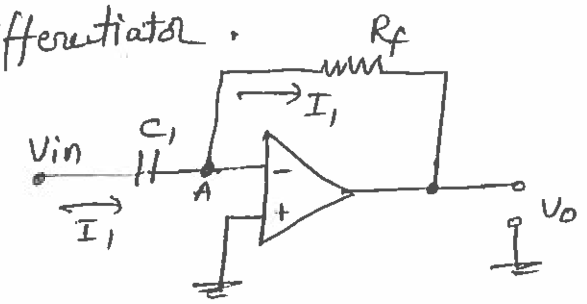
$$\int_0^t \frac{V_{in}}{R_1} dt = -C_f \int \frac{dV_o}{dt} dt$$

i.e. $$\int_0^t \frac{V_{in}}{R_1} dt = -C_f V_o$$

$$V_o = -\frac{1}{R_1 C_f} \int_0^t V_{in}\, dt + V_o(0)$$

where $V_o(0)$ is the constant of integration

Differentiator: The circuit which produces the differentiation of the input voltage at its output is called Differentiator.



The node B is grounded. The node A is also at the ground potential hence $V_A = 0$

As I/p current of op-amp is zero, either current I_1 flows through the resistance R_f.

from the I/p side we can write

$$I_1 = C_1 \frac{d(V_{in} - V_A)}{dt} = C_1 \frac{dV_{in}}{dt} \quad \text{—— (1)}$$

from the O/p side

$$I = \frac{V_A - V_o}{R_f} = \frac{-V_o}{R_f} \quad \text{—— (2)}$$

Equating the two equations

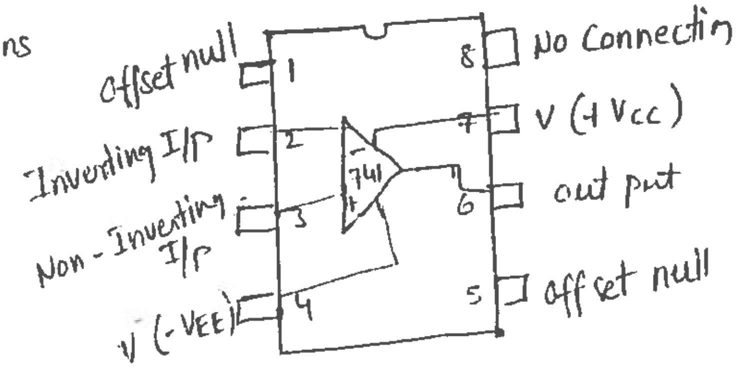
$$C_1 \frac{dV_{in}}{dt} = \frac{-V_o}{R_f}$$

$$\boxed{V_o = -C_1 R_f \frac{dV_{in}}{dt}}$$

The equation shows that the O/p is $C_1 R_f$ times the differentiation of the input and product $C_1 R_f$ is called time constant of the differentiator.

9(a) Draw and Explain the pin Diagram IC741 Op-amp. Pin Diagram – 3M, Description – 3M } 6M

Ans



Description of op-amp 741 IC pins

Pin 1 and 5: These two pins are used for offset process

Pin 2: Inverting I/p terminal, i.e when a Sinusoidal signal is applied to the Input pin 2:

Pin 3: Non-inverting input terminal i.e when a Sinusoidal signal is applied to the input pin 3, wave form of Same phase o/p is obtained.

Pin 4: $-V_{cc}$, i.e negative terminal of Supply Voltage is connected to this pin

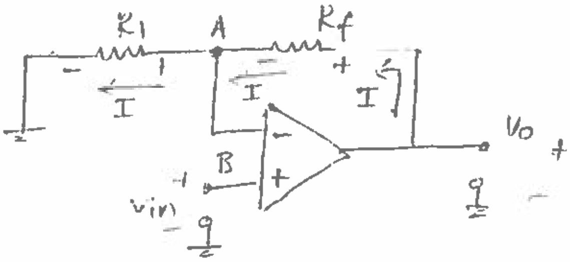
Pin 6: O/p terminal.

Pin 7: $+V_{cc}$ i.e +ve terminal of Supply Voltage is connected to this pin.

Pin 8: No electrical Connection is there in this pin: this pin is just for having the Symmetric dual-input package look.

7(b) Derive the gain for non-inverting op-amp. Diagram - 2M, Derivation - 4M } 6M

Ans An amplifier which amplifies the input without producing any phase shift b/w I/p and o/p is called non-inverting amplifier.



Derivation of closed loop gain:

The node B is at potential Vin, hence the potential of point A is same as B which is Vin, from the Concept of Virtual Ground.

$\therefore V_A = V_B = V_{in}$ —— (1)

from the o/p side we can write

$$I = \frac{V_o - V_A}{R_f}$$

$$I = \frac{V_o - V_{in}}{R_f} \quad \text{—— (2)}$$

At the inverting terminal

$$I = \frac{V_A - 0}{R_1}$$

$$\therefore I = \frac{V_{in}}{R_1} \quad \text{——③}$$

equating 2 and 3

$$\therefore \frac{V_0 - V_{in}}{R_f} = \frac{V_{in}}{R_1}$$

$$\therefore \frac{V_0}{R_f} = \frac{V_{in}}{R_f} + \frac{V_{in}}{R_1} = V_{in}\left[\frac{R_1 + R_f}{R_1 R_f}\right]$$

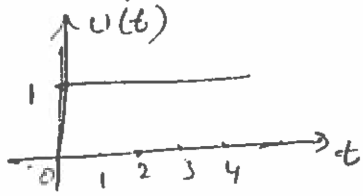
$$\frac{V_0}{V_{in}} = \frac{R_f(R_1 + R_f)}{R_1 R_f} = \frac{R_1 + R_f}{R_1} =$$

$$\therefore A_{VF} = \frac{V_0}{V_{in}} = 1 + \frac{R_f}{R_1}$$

10(a) State and Explain the properties of Continuous signals.

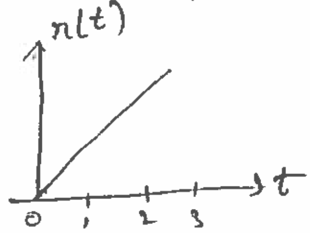
Ans: Unit step signal: The Unit Step function is defined as

$$u(t) = 1 \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



Unit Ramp Signal: The Unit ramp function is defined as

$$r(t) = t \quad \text{for } t \geq 0$$
$$= 0 \quad \text{for } t < 0$$



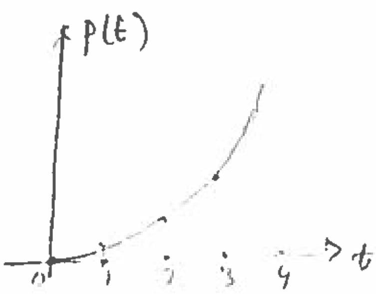
or $r(t) = t\,u(t)$

The ramp function can be obtained by Integrating the Unit step funct

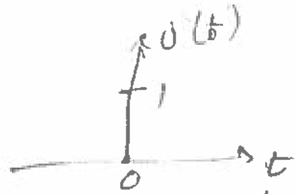
$$r(t) = \int u(t)\,dt = \int dt = t$$

Unit parabolic: $p(t) = t^2/2 \quad \text{for } t \geq 0$

$$= 0 \quad \text{for } t < 0$$



Impulse Signal : $\delta(t) = 0$ for $t \neq 0$



Sinusoidal signal : A Continuous - time Sinusoidal Signal is given by

$$x(t) = A \sin(\Omega t + \theta)$$

A — Amplitude

Ω → frequency in radians per second

θ → is the phase angle in radians

10(b) List any four applications of FM System. (4M)

Ans: 1) It is mostly used in radio broadcasting.

2) It is used in radar, telemetry, Seismic prospecting.

3) It is used in music Synthesis as well as in Video transmission intstruments.

4) It is used medical applications like EEG.

11(a) Define Amplitude Modulation. Derive an expression for The AM Wave.
[Def : 2M + Expression 4M = 6M]

Ans : Amplitude modulation is a process of Changing The amplitude of The high frequency analog Carrier in accordance with The amplitude of The message signal.

Expression for AM Wave

$$m(t) = A_m \cos 2\pi f_m t$$

$$c(t) = A_c \cos 2\pi f_c t$$

$$s(t) = A_c \cos 2\pi f_c t + A_c k_a m(t) \cos 2\pi f_c t$$

$$s(t) = A_c [1 + k_a m(t)] \cos 2\pi f_c t$$ → time domain equation of AM wave

$$s(t) = A_c\left[1 + k_a A_m \cos 2\pi f_m t\right] \cos 2\pi f_c t$$

$\mu = k_a A_m$ modulation index

$$\therefore\ s(t) = A_c\left[1 + \mu \cos 2\pi f_m t\right] \cos 2\pi f_c t$$

$$= A_c \cos 2\pi f_c t + A_c \mu \cos 2\pi f_c t \cos 2\pi f_m t$$

$$= \underbrace{A_c \cos 2\pi f_c t}_{\text{Carrier}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t}_{\text{USB}} + \underbrace{\frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t}_{\text{LSB}}$$

11 b) Write about am Voltage distribution. (4M)

$$P_t = P_c + P_{USB} + P_{LSB}$$

$$P = I^2 R$$

$$s(t) = A_c \cos 2\pi f_c t + \frac{A_c \mu}{2} \cos 2\pi (f_c + f_m)t + \frac{A_c \mu}{2} \cos 2\pi (f_c - f_m)t$$

$$P_c = \frac{(A_c/\sqrt{2})^2}{R}, \quad P_{USB} = \frac{\left(A_c \mu / 2\sqrt{2}\right)^2}{R} \quad P_{LSB} = \frac{\left(A_c \mu / 2\sqrt{2}\right)^2}{R}$$

$$\therefore\ P_t = \frac{A_c^2}{2R} + \frac{A_c^2 \mu^2}{8R} + \frac{A_c^2 \mu^2}{8R}$$

$$[\because R = 1\Omega]$$

$$P_t = \frac{A_c^2}{2}\left[1 + \frac{\mu^2}{2}\right] = P_c\left[1 + \frac{\mu^2}{2}\right]$$

12(a) State and prove Sampling theorem. [Statement 2M + proof 4M = 6M]

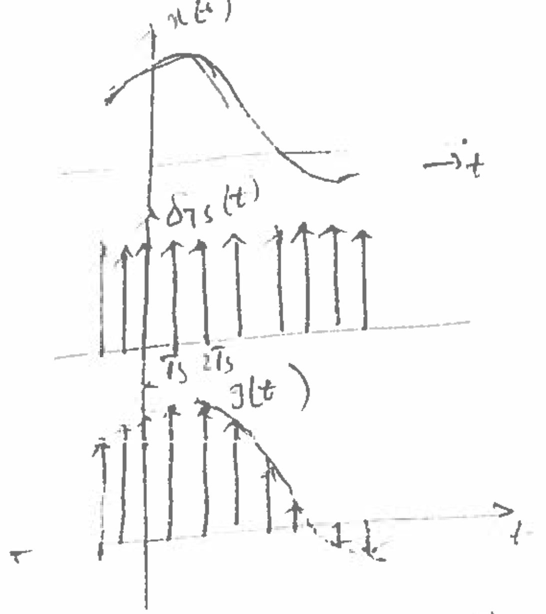
Ans: Statement : A Bandlimited signal of finite energy which has no frequency Components higher than f_m Hz may be Completely recovered from the knowledge of its samples taken at the rate of $2f_m$ samples per Second.

Proof of Sampling theorem :

Let $x(t)$ is a Continuous signal, with maximum frequency f_m

The Sampling of $x(t)$ at a rate f_s Hz, may be achieved by multiplying $x(t)$ by an Impulse train $\delta_{Ts}(t)$

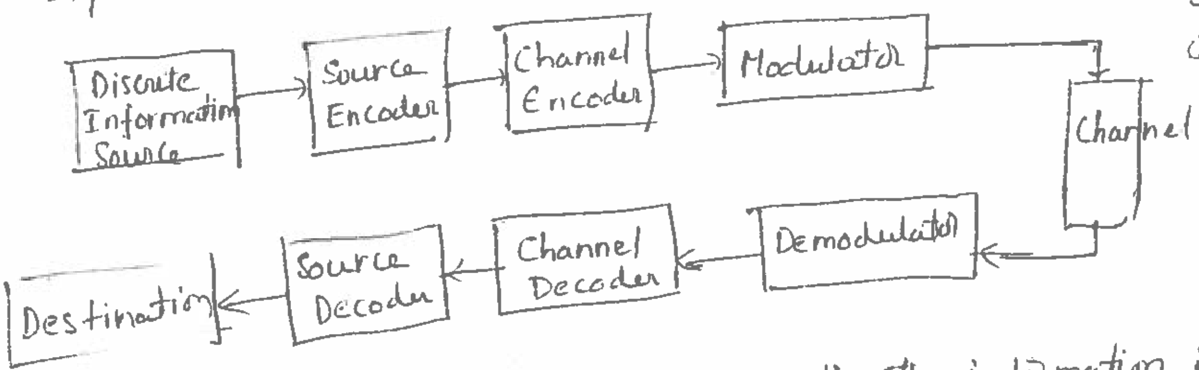
$\delta_{Ts}(t) \rightarrow$ impulse train Consist of unit impulses repeating periodically every T_s seconds where $T_s = 1/f_s$



13(a) Explain the Basic elements of Digital Communication system.

Diagram 2M
Content 4M } 6M

Ans



Discrete Information Source : In Digital Communication the information in Discrete w.r.t time. This information is obtained by process of Sampling and quantization. So The Discrete Information Source can be letters, digits, Special Characters, Code words...

Source Encoder and Decoder: The symbols produced by the information Source are given to the Source encoder. The Symbols Can not be transmitted directly. Source encoder converts symbols into group of bits called

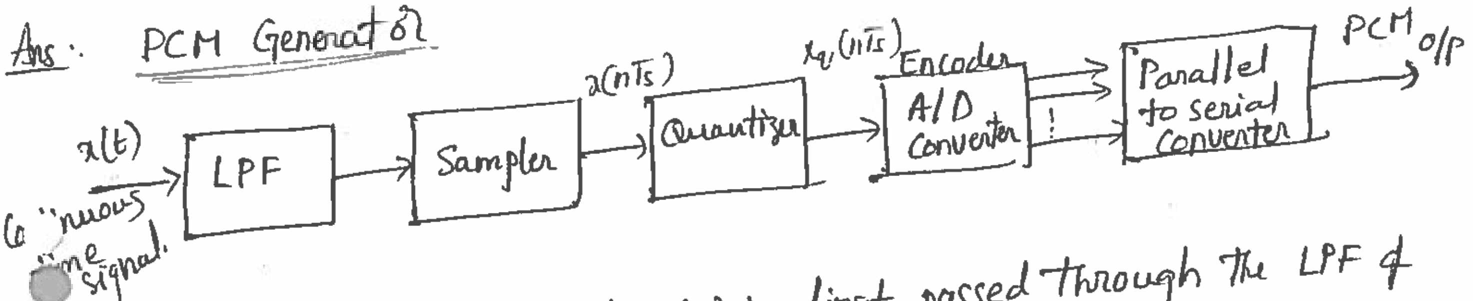
Code words (combination of 1's and 0's) for each distinct symbol there is an unique code word. At receiver side ~~channel~~ source decoder is used it performs opposite operation of ~~channel~~ source encoder.

Channel Encoder and Decoder: The input of the channel encoder is Binary sequence the communication channel adds noise and interference to the signal being transmitted. Hence errors are introduced in the binary sequence received at the receiver end. Thus channel coding is done to avoid this type of errors.

Digital Modulators and Demodulators: After converting into binary information The pulses are to be transmitted by using digital modulation techniques like ASK, FSK, PSK ... etc. depends on application Requirement.

13(b) With a neat diagram explain the Generation of PCM & DPCM. [PCM 3M + DPCM 3M = 6M]

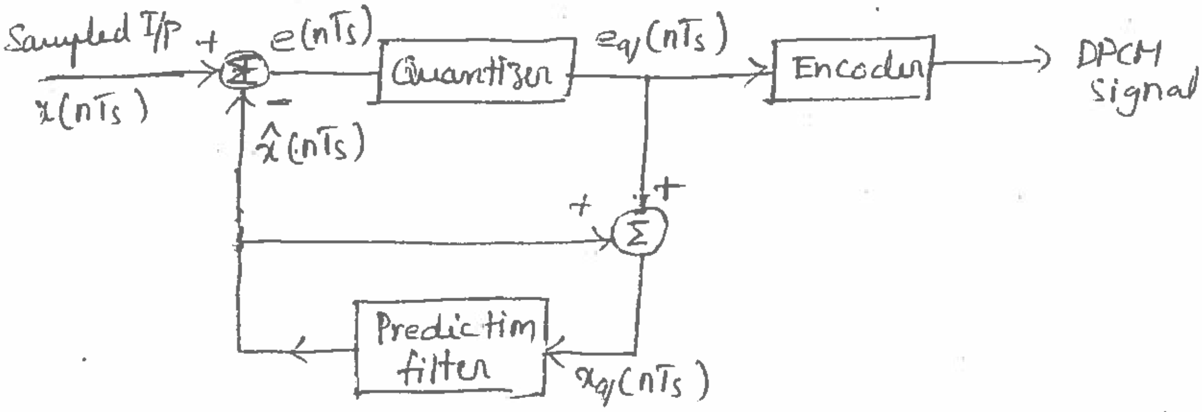
Ans: PCM Generator



In PCM generator the signal x(t) is first passed through the LPF of cut-off frequency fm Hz. This LPF blocks all the frequency component which are lying above fm Hz. Samples this signal at the rate of fs. The o/p of sampler is denoted by x(nTs). A quantizer compares input x(nTs) with its fixed digital levels. It assigns any one of the digital to x(nTs) with its fixed digital levels. The o/p of quantizer given to the I/p of encoder. This encoder converts input signal to 'v' digits binary word. Encoder o/p is given to the parallel to serial converter it converts parallel data into serial data it is suitable to transmission through channel.

DPCM Generation :

The DPCM works on the principle of prediction. The value of the present sample is predicted from the past samples.



The sampled signal is denoted by $x(nT_s)$ and the predicted signal is denoted by $\hat{x}(nT_s)$. The comparator finds out the difference between the actual sample value $x(nT_s)$ and predicted sample value $\hat{x}(nT_s)$. This is known as prediction error and it is denoted by $e(nT_s)$.

~~Thus, error is~~ The predicted value is produced by using a prediction filter. The quantizer o/p signal $e_q(nT_s)$ and previous prediction is added and given as input to the prediction filter. This signal is called $x_q(nT_s)$. This makes the prediction more and more close to the actual sampled signal. The quantized error signal $e_q(nT_s)$ is very small and can be encoded by using smaller number of bits. Thus no. of bits per sample are reduced in DPCM.

12(b) Describe the Basic principles of PCM system and PCM transmitter.

[Principles of PCM system 2M + PCM Tx 4M = 6M]

Ans **Principles of PCM system**

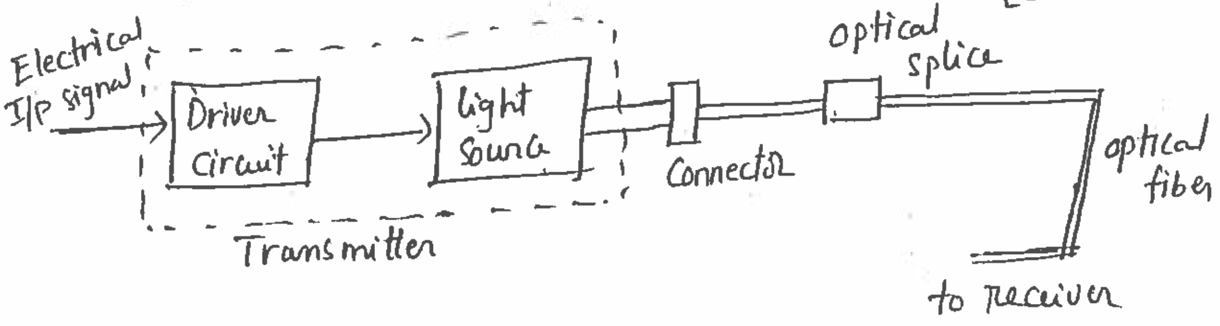
1) PCM is a digital pulse modulation system
2) PCM o/p is in the coded digital form.
3) PCM consist of a PCM encoder and PCM decoder
4) PCM is not modulation in the conventional sense.

PCM transmitter : Refer 13 (b) Answer

(8)

14(a) Draw and Explain The working principle of an optical transmitter. [Content 4M + Diagram 2M = 6M]

Ans



Transmitter: The transmitter first converts the input Voltage to Current Value which is used to drive The light Source. Thus it interfaces The input Circuit and light Source.

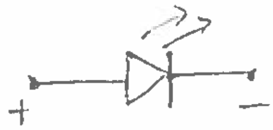
The light Source is normally an infrared LED or LASER device which is driven by the current Value from the V to I Converter. It emits light which is proportional to the input Voltage Value is generated and given as input to the fiber.

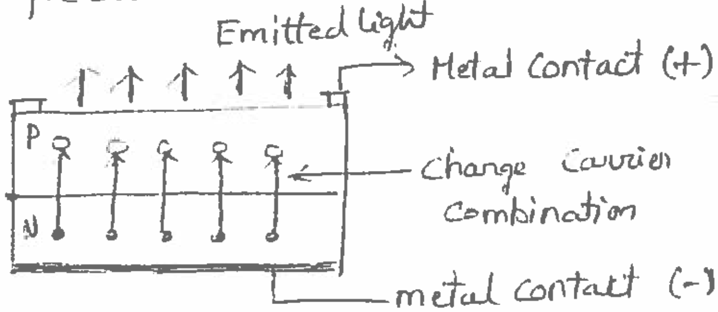
optical Splice: for Creating long haul Communication link, it is necessary to join one fiber to other fibers permanently.

14(b) Explain about LED and its type. (content 4M + diagram 2M = 6M)

Ans: The Light Emitting Diode (LED) is a PN junction diode which emits light When forward biased, by a phenomenon called electroluminescence. In all Semiconductor PN junctions, Some of the energy will be radiated as heat and some in the form of photons. In Si and Ge the emitted light is insignificant. In other materials such as Gallium phosphide (GaP) or Gallium Arsenide phosphide (GaAsP), The number of photons of light energy emitted is Sufficient to Create a Visible light Source. Here the Charge carrier recombination takes place when electrons from The N-Side Cross the junction and recombine with The holes on the p-side.

When LED is forward biased, The electrons and holes moves towards the junction and recombination takes place. As a result the e^- lying in The Conduction bands of N-region fall into the holes lying in the VB of P-region. The difference of energy b/w The CB and VB is radiated in the form of light energy. The brightness of The emitted light is directly proportional to the forward current.





The color of the emitted light depends on the type of material used.

Gallium Arsenide (Ga As) ⟶ infrared radiation (invisible)

Gallium Phosphide (GaP) ⟶ red or green

Gallium Arsenide phosphide (Ga AsP) ⟶ red or yellow

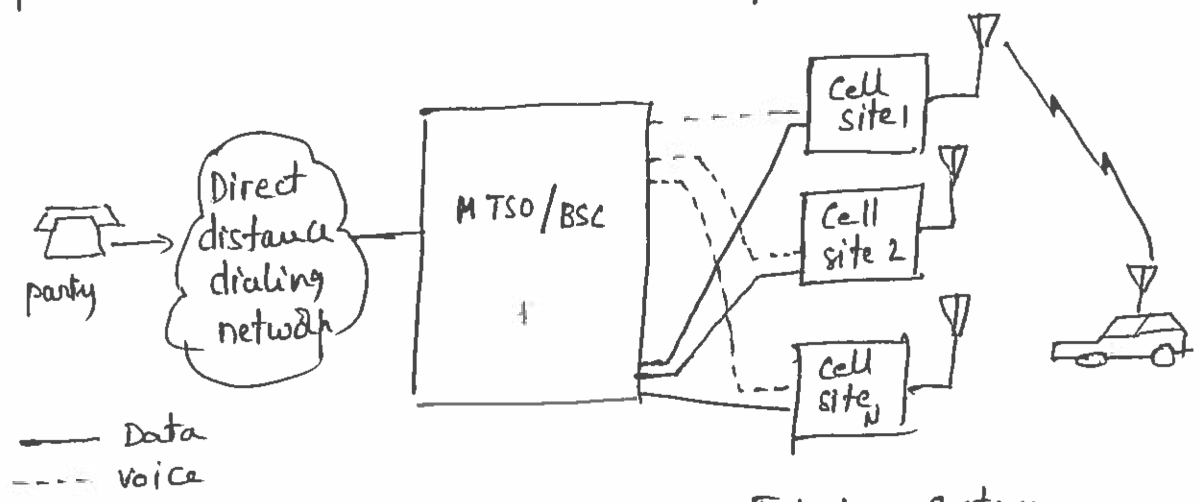
15(a) Explain the working principle of GSM? (6M)

Ans: Global System for Mobile Communication (GSM) is a digital mobile network that is widely used by mobile phone users in the world

The GSM. network has four Separate parts that work together to function as a whole.
1) Mobile Station (MS)
2) Base Station Subsystem (BSS)
3) Network Switching Subsystem (NSS)
4) Operation and Support Subsystem (OSS)

15 (b) Explain Cellular Telephone System. [Diagram 3M + Content 3M = 6M]

Ans.



A general View of Cellular Telephone System.

Antenna: Antenna pattern, antenna gain, antenna tilting and antenna height all affect the cellular System design. The antenna pattern can be omnidirectional, directional or any shape in both the vertical and the horizon planes. Antenna gain compensates for the transmitted power. Antenna gain at the mobile units would affect the system performance.

Switching Equipment: The capacity of switching equipment in cellular systems is not based on the number of switch ports but on the capacity of the processor associated with the switches.

Data links: The data link are not directly affected by the cellular system, they are important in the system. Each data link can carry multiple channel data (10 kbps data transmitted per channel) from the cell site to the MTSO.

NSRIT

Semester End Examination, Sept./Oct., 2021

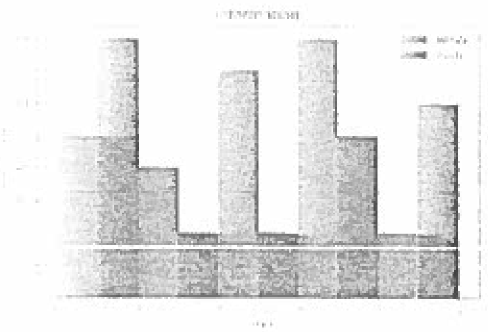
Degree	B. Tech. (U. G.)	Program	EEE			Academic Year	2020 - 2021
Course Code	20CS403	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	PYTHON PROGRAMMING						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	What is type conversion?	20CS403.1	L2
2	Compare List and Tuple	20CS403.2	L3
3	Define Module. What is the use of Module?	20CS403.3	L2
4	How will you manipulate file pointer using seek?	20CS403.4	L1
5	List out geometry manager classes in tkinter module	20CS403.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Give short note on the following: i) Python variables ii) Keywords iii) Python Indentation	6M	20CS403.1	L2
6 (b)	Write a Python program that solve the quadratic equation ax**2 + bx + c = 0	6M	20CS403.1	L3
	OR			
7 (a)	Explain in detail about Program development cycle	6M	20CS403.1	L2
7 (b)	Write a Python program to demonstrate the application of identity operators & Membership operators.	6M	20CS403.1	L2
8 (a)	Explain how Accessing Character and Substring in Strings is done in Python with example	6M	20CS403.2	L2
8 (b)	Write a Python Program to Check if a Number is Positive, Negative or 0	6M	20CS403.2	L3
	OR			
9 (a)	Explain the various List methods available in Python	6M	20CS403.2	L2
9 (b)	Write a Python program to check if the number is an Armstrong number or not. (A positive integer is called an Armstrong number of order n if abcd... = $a^n + b^n + c^n + d^n + \dots$ An Armstrong number of 3 digits, the sum of cubes of each digit is equal to the number itself. For example: 153 = 1*1*1 + 5*5*5 + 3*3*3 // 153 is an Armstrong number)	6M	20CS403.2	L2
10 (a)	Explain any 3 functions of the following modules: i. Cmath ii. Random	6M	20CS403.3	L2
10 (b)	Explain arbitrary and keyword argument in Python with example	6M	20CS403.3	L2
	OR			
11 (a)	What is Recursion? Explain the working of recursive function with an example	6M	20CS403.3	L2
11 (b)	What is PIP? How packages are installed using PIP?	6M	20CS403.3	L2

12 (a)	What is File? Explain the file handling functions in Python with example	6M	20CS403.4	L2
12 (b)	How to create a constructor and destructor in Python? Give an example	6M	20CS403.4	L2
	OR			
13 (a)	Demonstrate implementation of multilevel inheritance in Python, with a program	6M	20CS403.4	L2
13(b)	What is operator overloading in Python?	6M	20CS403.4	L2
14	Explain any 5 functions in Numpy module with example	12M	20CS403.5	L2
	OR			
15	Demonstrate the usage MatPlotlib library. Write a program for the following graph 	12M	20CS403.5	L3

SEMESTER Question Paper

Degree	B. Tech. (U. G.)	Program	EEE	Test	I/II	Academic Year	2020 - 2021
Course Code	20CS403	Test Duration	90 Min.	Max. Marks	40	Semester	II
Course	PYTHON PROGRAMMING						

Key and Scheme of Evaluation

No.	Questions (1 through 5)	Marks
1	**What is type conversion?** The process of converting the value of one data type to another data type is called typeconversion. Python has two types of type conversion. Implicit Type Conversion Explicit Type conversion	Definition-1M Types-0.5M Example-0.5M
2	**Compare List and Tuple** (see table below)	Any 4 differences-2M

S. No	List	Tuple
1	Lists are mutable	Tuples are immutable
2	Lists consume more memory	Tuple consume less memory as compared to the list
3	List is created using []	Tuple is created using ()

No.	Questions (1 through 5)	Marks
3	**Define Module. What is the use of Module?** In Python, Modules are simply files with the ".py" extension containing Python code that can be imported inside another Python Program.In simple terms, we can consider a module to be the same as a code library or a file that contains a set of functions that you want to include in your application. To incorporate the module into our program, we will use the import keyword, and to get only a few or specific methods or functions from a module, we use the from keyword. **Syntax:** module_name.function_name **Eg:** import math Print(math.pi)	Description-1M Importing module-1M
4	**How will you manipulate file pointer using seek?** seek(): In Python, seek() function is used to **change the position of the File Handle** to a given specific position. File handle is like a cursor, which defines from where the data has to be read or written in the file. Eg: f = open("xyz.txt", "r") f.seek(20)	Description-1M Syntax-0.5M Example-0.5M
5	**List out geometry manager classes in tkinter module** The geometry manager is used to manage the geometry of the window and other frames. We can use it to handle the position and size of the window and frames. There are mainly three methods in Geometry Managers: The pack() method The grid() method The place() method	Description-0.5M List-1.5M

No.	Questions (6 through 11)	Marks
6 (a)	**Give short note on the following:** **i) Python variables ii) Keywords iii) Python Indentation** i) Python variables Definition It is an identifier that is used to refer to memory location and used to hold value. In Python, we don't need to specify the type of variable Rules for creating variables in Python • It must start with a letter or the underscore character. • It can not start with a number. • It can only contain alpha-numeric characters and underscores (A-z, 0-9, and _)	Content 4m Grammar & Spellings 1m; presentation 1m

- Variable names are case-sensitive

ii) Keywords

Python keywords are the fundamental building blocks of any program
Python keywords are special reserved words that have specific meanings and purposes
There are a number of ways you can identify valid Python keywords
1. Using help()
Eg: >>> help("keywords")
2.import keyword module
Eg: >>>import keyword
>>> keyword. Kwlist

iii) Python Indentation

In python code blocks are identified by indentation rather than using symbols like curly braces. Without extra symbols programs are easier to read and also indentation clearly identifies which block of code a statement belongs to. Python does not support braces to indicate blocks of code for class or function definitions or flow control. Blocks of code is identified by line indentation. All the continues lines indented with same number of spaces would form a block. Python strictly follow indentation rules to indicate the blocks

6 (b) **Write a Python program that solve the quadratic equation ax**2 + bx + c = 0**

Program:

```
# import Complex math module
import cmath
a=float(input("Enter a value"))
b=float(input("Enter b value"))
c=float(input("Enter c value"))
# calculate the discriminant
d = (b**2) - (4*a*c)
# find two solutions
root1 = (-b-cmath.sqrt(d))/(2*a)
root2 = (-b+cmath.sqrt(d))/(2*a)
print('The solution are {0} and {1}'.format(root1,root2))
```

Program-5M
Output-0.5M
Explaination-0.5M

OR

7 (a) **Explain in detail about Program development cycle**

Python's development cycle is dramatically shorter than that of traditional tools. In Python, there are nocompile or link steps -- Python programs simply import modules at runtime and use the objects they contain.

The Program Development Cycle (PDC) has various states as follow
Problem Definition
Program Design
Coding
Debugging
Testing
Documentation
Maintenance

List of states-1M
Explanation of each state-5M

7 (b) **Write a Python program to demonstrate the application of identity operators & Membership operators.**

Operators are symbols that perform certain mathematical or logical operation to manipulate data values and produce a result.

Identity operators: used to check if two values (variable) are located on the same object or same memory

Membership operators: used to test whether a value or operand is found in the sequence such as list, string, set, or dictionary.

Description-2M
Identity operator with example-2M
Membership operator with example-2M

8 (a) **Explain how Accessing Character and Substring in Strings is done in Python with example**

A string is a sequence of zero or more characters. It is treated as a data structure. A string's length is the number of characters it contains. The length can be obtained using the len() function by passing the string as an argument to it. The positions of a string's characters are numbered from 0,

Definition-1M
Accessing character-

on the left, to the length of the string minus 1, on the right. | 2.5M

Accessing character & substring

1. Using Subscript Operator
2. Using slice opertor

Accessing substring-2.5M

8 (b) **Write a Python Program to Check if a Number is Positive, Negative or 0**

Program

```
a=int(input ("enter a number:"))
if(a>0):
    print("The given number is a Positive number ")
elif(a<0):
    print ("The given number is Negative number")
else:
    print("The given number is equal to Zero")
```

Program-5M
Output-0.5M
Explaination-0.5M

OR

9 (a) **Explain the various List methods available in Python**

List is used to store the group of values & we can manipulate them, in list the values are stores in index format starts with 0.List is mutable object so we can do the manipulations.

Syntax: <list_name> = [value1,value2,value3,...,valuen]

Example: data5=['TEC',10,56.4,'a'] # list with mixed data types

List Methods:

append()-Adds an element at the end of the list
clear()-Removes all the elements from the list
copy()-Returns a copy of the list
count()-Returns the number of elements with the specified value
extend()-Add the elements of a list (or any iterable), to the end of the current list
index()-Returns the index of the first element with the specified value
insert()-Adds an element at the specified position
pop() Removes the element at the specified position
remove() Removes the first item with the specified value

Description-1M
Any 5 methods with example-5M

9 (b) **Write a Python program to check if the number is an Armstrong number or not. (A positive integer is called an Armstrong number of order n if abcd... = $a^n + b^n + c^n + d^n + \ldots$ In case of An Armstrong number of 3 digits, the sum of cubes of each digit is equal to the number itself. For example: 153 = 1*1*1 + 5*5*5 + 3*3*3 // 153 is an Armstrong number)**

Program:

```
n=int(input("Enter Number"))
sum=0
temp=n
while n!=0:
   rem=n%10
   sum=sum+(rem*rem*rem)
   n=n//10
if(sum==temp):
   print("The given number is Armstrong number")
else:
   print("The given number is not a Armstrong number")
```

Program-5M
Output-0.5M
Explaination-0.5M

10 (a) **Explain any 3 functions of the following modules**

i. Cmath

Python has a built-in module that you can use for mathematical tasks for complex numbers.The methods in this module accepts int, float, and complex numbers. It even accepts Python objects that has a __complex__() or __float__() method.The methods in this module almost always return a complex number. If the return value can be expressed as a real number, the return value has an imaginary part of 0.

The cmath module has a set of methods and constants.

cmath module functions with example-3M
random module functions

1.Sqrt()
2.log10()
3.cos()

with example-3M

ii. **Random**

Python Random module is an in-built module of Python which is used to generate random numbers. These are pseudo-random numbers means these are not truly random. This module can be used to perform random actions such as generating random numbers, print random a value for a list or string, etc.

some common operations performed by this module.

1.random.randint()
2.random.random()
3.random.choice()

10 (b) **Explain arbitrary and keyword argument in Python with example**

Arbitrary arguments

Sometimes, we do not know in advance the number of arguments that will be passed into a function. Python allows us to handle this kind of situation through function calls with an arbitrary number of arguments. In the function definition, we use an asterisk (*) before the parameter name to denote this kind of argument.

Keyword arguments

Python allows functions to be called using keyword arguments. When we call functions in this way, the order (position) of the arguments can be changed.

Arbitrary argument with example-3M
Keyword argument with example-3M

OR

11 (a) **What is Recursion? Explain the working of recursive function with an example**

A function that calls itself is known as Recursive Function.

Program:

```
def fact(n):
    if(n==0 or n==1):
        return 1
    else:
        return n*fact(n-1)
n=int(input("Enter number"))
print("The factorial of a given number is:",fact(n))
```

Output:

```
Enter number5
The factorial of a given number is: 120
```

Definition-0.5M
Program-5M
Output-0.5M

11 (b) **What is PIP? How packages are installed using PIP?**

pip is a package-management system written in Python used to install and manage software packages.It connects to an online repository of public packages, called the Python Package Index. pip is the ease of its command-line interface, which makes installing Python software packages as easy as issuing a command:

pip install some-package-name

Eg: pip install Matplotlib

pip uninstall some-package-name

Eg: pip uninstall Matplotlib

Description-2M
Commands with example-4M

12(a) **What is File? Explain the file handling functions in Python with example**

File: A file is some information or data which stays in the computer storage devices. ... Python gives you easy ways to manipulate these files. Generally we divide files in two categories, text file and binary file.

File handling functions

- Open()
- Close()
- Write()
- Writelines()
- Read()
- Readlines()

Definition-1M
List of functions-1M
Any 4 functions with example-4M

- tell()

12(b)	**How to create a constructor and destructor in Python? Give an example** A constructor is a special type of method (function) which is used to initialize the instance members of the class. In C++ or Java, the constructor has the same name as its class, but it treats constructor differently in Python. It is used to create an object. Constructors can be of three types. 1.Default Constructor 2.Parameterized Constructor 3.Non-parameterized Constructor	Description-1M Types-1M Any constructor with example-3M
	OR	

13(a)	**Demonstrate implementation of multilevel inheritance in Python, with a program**
	Inheritance is a powerful feature in object oriented programming. It refers to defining a new class with little or no modification to an existing class. The new class is called derived (or child) class and the one from which it inherits is called the base (or parent) class.
	Multilevel Inheritance
	In multilevel inheritance, features of the base class and the derived class are further inherited into the new derived class. This is similar to a relationship representing a child and grandfather.
	# Python program to demonstrate multilevel inheritance
	# Base class
	class Grandfather:
	grandfathername =""
	def grandfather(self):
	print(self.grandfathername)
	# Intermediate class
	class Father(Grandfather):
	fathername = ""
	def father(self):
	print(self.fathername)
	# Derived class
	class Son(Father):
	def parent(self):
	print("GrandFather :", self.grandfathername)
	print("Father :", self.fathername)
	# Driver's code
	s1 = Son()
	s1.grandfathername = "Srinivas"
	s1.fathername = "Ankush"
	s1.parent()
		Description-2M
		Multilevel inheritance
		Program-4M

13(b)	**What is operator overloading in Python?** Python allows the same operator to have different meaning according to the context is called operator overloading. Python Special functions used for operator overloading: • __add__(self, other) • __sub__(self, other) • __mul__(self, other) • __floordiv__(self, other) • __lt__(self, other)) **Example** # Python Program illustrate how to overload an binary + operator class addoperator:	Description-2M Special functions-1M Program-3M

```
    def __init__(self, X):
        self.X = X
    # __add__ () method is magic function to perform addition of two objects
    def __add__(self, other):
        return self.X + other.X
obj1 = addoperator(234)
obj2 = addoperator(456)
print (obj1 + obj2)
obj3 = addoperator("Welcome ")
obj4 = addoperator("to NSRIT")
print (obj3 + obj4)
```

14

Explain any 5 functions in Numpy module with example

NumPy stands for numeric python which is a python package for the computation and processing of the multidimensional and single dimensional array elements.

It provides various functions which are capable of performing the numeric computations with a high speed.

<u>**NumPy Functions**</u>

1. <u>**numpy.array()**</u> : We can create a NumPy ndarray object by using the array() function. The array object in NumPy is called ndarray. It is basically a table of elements which are all of the same type and indexed by a tuple of positive integers.

Example:

```
import numpy as np
#Here create 1-D Array
arr=np.array([1, 2, 3, 4, 5])Herecreate1-DArray
#Here create 2-D Array
arr=np.array([[1, 2, 3],[4, 5, 6]])
```

<u>**2. numpy.sum()**</u>

This function is used to compute the sum of all elements. It is also possible to add rows and column elements of an array. The output will be in the form of an array object.

Example:

```
a=np.array([[1,4],[3,5]])
b=np.sum(a)
print(b)  #13
```

3. <u>**numpy.append()**</u>

The numpy append() function is used to merge two arrays. It returns a new array, and the original array remains unchanged.

The numpy.append() function is used to add or append new values to an existing numpy array. This function adds the new values at the end of the array.

<u>**Example:**</u>

```
import numpy as np
a=np.array([[10, 20], [40, 50], [70, 80]])
b=np.array([[11, 21], [42, 52], [73, 83]])
c=np.append(a,b)
print(c) #array([ 10,  20, 40, 50, 70, 80,11, 21,42, 52, 73, 83])
```

4. <u>**numpy.sort()**</u>

The NumPy ndarray object has a function called sort(), that will sort a specified array. Sorting means putting elements in an *ordered sequence*. *Ordered sequence* is any sequence that has an order corresponding to elements, like numeric or alphabetical.

<u>**Example**</u>

```
arr=np.array(['banana', 'cherry', 'apple'])
print(np.sort(arr))#['apple''banana''cherry']
```

5. <u>numpy. arrange ()</u>

It creates an array by using the evenly spaced values over the given interval.

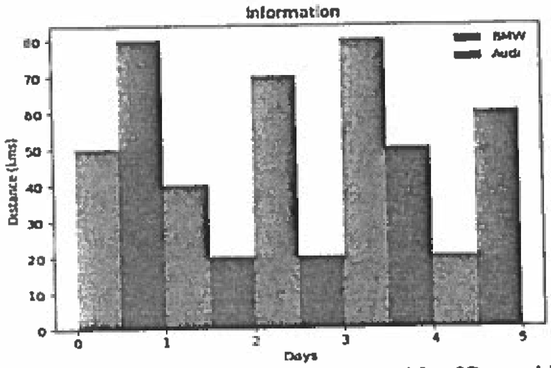
<u>**Example:**</u>

```
arr = np.arange(0,10,2,int) # [0 2 4 6 8]
```

Description-1M
List of functions-1M
Five functions with example-10M

OR

Demonstrate the usage MatPlotlib library. Write a program for the following graph



Matplotlib.pyplot is a plotting library used for 2D graphics in python programming language. It can be used in python scripts, shell, web application servers and other graphical user interface toolkits. Matploitlib is a Python Library used for plotting, this python library provides and objected-oriented APIs for integrating plots into applications.

Types of Plots

15 There are various plots which can be created using python matplotlib. Some of them are listed below:

1. Bargraph
2. histogram
3. Scatter Plot
4. Area Plot
5. Pie Plot

Python Matplotlib – Histogram

Histograms are used to show a distribution whereas a bar chart is used to compare different entities. Histograms are useful when you have arrays or a very long list.

<u>Program</u>

```
import matplotlib.pyplot as plt
population_age
=[22,55,62,45,21,22,34,42,42,4,2,102,95,85,55,110,120,70,65,55,111,115,80,75,65,54,44,43,42,48]
bins = [0,10,20,30,40,50,60,70,80,90,100]
plt.hist(population_age, bins, histtype='bar', rwidth=0.8)
plt.xlabel('age groups')
plt.ylabel('Number of people')
plt.title('Histogram')
plt.show()
```

Description-3M
Types-2M
Program-9M
Expaination-1M

NSRIT

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	Program	EEE/CSE			Academic Year	2020 - 2021
Course Code	20BSX23	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	APPLIED CHEMISTRY						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Define coordination polymerization	20BSX23.1	L1
2	What is conductivity cell?	20BSX23.2	L1
3	Write Schrodinger equation	20BSX23.3	L2
4	What is an electromagnetic radiation?	20BSX23.4	L1
5	How does a molecular switch work?	20BSX23.5	L2

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Discuss about the mechanism of free radical addition polymerization and copolymerization with suitable examples	6M	20BSX23.1	L2
6 (b)	Differentiate thermoplastics and thermosets	6M	20BSX23.1	L4
	OR			
7 (a)	Write the preparation, properties and applications of Bakelite and BUNA-N	8M	20BSX23.1	L2
7 (b)	Distinguish the properties and a pplications of Nylon 6:6 and carbon fibers(GCF)	4M	20BSX23.1	L4
8 (a)	Explain the construction & working of Ag|AgCl electrode	5M	20BSX23.2	L2
8 (b)	What is potentiometry? Explain how potentiometry method helps to determine the end point in oxidation-reduction titration	7M	20BSX23.2	L2
	OR			
9 (a)	Explain construction, working and applications of lead acid battery	6M	20BSX23.2	L2
9 (b)	Derive the Nernst equation and write its applications	6M	20BSX23.2	L2
10 (a)	Describe the energy level diagrams of O2 and CO molecule. Write their magnetic nature and bond order	7M	20BSX23.3	L1
10 (b)	What is molecular orbital theory? Describe the molecular orbitals of butadiene and benzene	5M	20BSX23.3	L1
	OR			
11 (a)	State the crystal field theory? List out the magnetic properties of coordination compounds	6M	20BSX23.3	L1
11 (b)	Define the following terms: a. conductors b .semiconductors c. Insulators	6M	20BSX23.3	L1
12 (a)	Write a short note on Beer-Lambert's Law	5M	20BSX23.4	L2
12 (b)	Explain the principle and instrumentation of FTIR spectroscopy with a neat diagram	7M	20BSX23.4	L2
	OR			
13 (a)	Explain the principle and instrumentation of HPLC	6M	20BSX23.4	L2
13 (b)	Explain the determination of end point in acid-base titration using pH meter	6M	20BSX23.4	L2
14 (a)	What is the basic lock and key principle?	5M	20BSX23.5	L1
14 (b)	Discuss about the supramolecular reactivity and catalysis	7M	20BSX23.5	L2
	OR			
15 (a)	List out the applications of Catenands and rotaxanes	6M	20BSX23.5	L1
15 (b)	Explain computational chemistry and molecular docking	6M	20BSX23.5	L2

Semester - II End Examinations, Sep/Oct - 2021

Key & Scheme of Evaluation

: Applied chemistry CODE: 20BS X23 MAX MARKS - 70M

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

Define coordination polymerisation.

ns:- It is also known as Ziegler - Natta polymerization. It is defined as " In the presence of a Combination of transition - metal halide like $TiCl_4$ with an organo metallic compounds like $(C_2H_5)_3Al$, stereo specific polymerization occurs. 2M

. What is Conductivity cell.

ns:- A Conductivity cell is Composed of two pt electrodes which are coated with pt black. the electrodes have area of cross section equal to 'A' and are separated by distance 'l'. Hence, the solution Confined b/w the two electrodes is a Column of length l and area of cross section A. 2M

3. Write Schrodinger wave equation.

ns:- A particle whose motion is described by three space Co-ordinates x, y, z. Then the Schrodinger equation is 2M

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

4. What is an electro magnetic radiation.

Ans:- It is a form of radiant energy which has both the particle as well as wave nature & It has both electric & magnetic components.

5. How does a molecular switch work?

Ans:- A molecular switch is a molecule that can be reversibly shifted b/w two or more stable states the molecule may be shifted b/w the states in response to environmental stimuli such as changes in pH, light, temperature, an electric current, micro environment (or) in presence of ions & other ligands. — 2M

PART-B (Long Answer Questions 5x12 = 60 Marks)

6 (a). Discuss about the mechanism of free radical addition polymerization & Copolymerization with suitable examples. — 6M

Ans: Free radical polymerization mechanism:- It is a type of polymerization in which the reactive centre of a polymer chain contains a free radical & this polymerization reaction is intiated by initiators which undergo homolytic — 1M

This mechanism involved in 3 steps.

1. Initiation
2. propagation
3. Termination

Initiation :- Initiators are unstable compounds & undergo homolytic fission to produce free radicals which react with π electrons of the monomer to produce monomer free radical. 1M

$$\underset{\text{Initiator}}{I - I} \xrightarrow{\text{heat}} \underset{\text{free radical}}{\dot{I} + \dot{I}}$$

$$\underset{\text{free radical}}{\dot{I}} + \underset{\text{monomer}}{CH_2 = \underset{|}{\overset{}{C}}H\!-\!X} \longrightarrow \underset{\text{Monomer free radical}}{I - CH_2 - \dot{C}H(X)}$$

2. **propagation :-** The monomer free radical reacts with a number of monomers to form chain growth with free radical site at the end of the chain producing a living polymer. 1M

$$\underset{\text{Monomer free radical}}{I - CH_2 - \dot{C}H(X)} + \underset{\text{Monomer}}{CH_2 = CH(X)} \longrightarrow \underset{\text{Growing chain}}{I - CH_2 - CH(X) - CH_2 - \dot{C}H(X)}$$

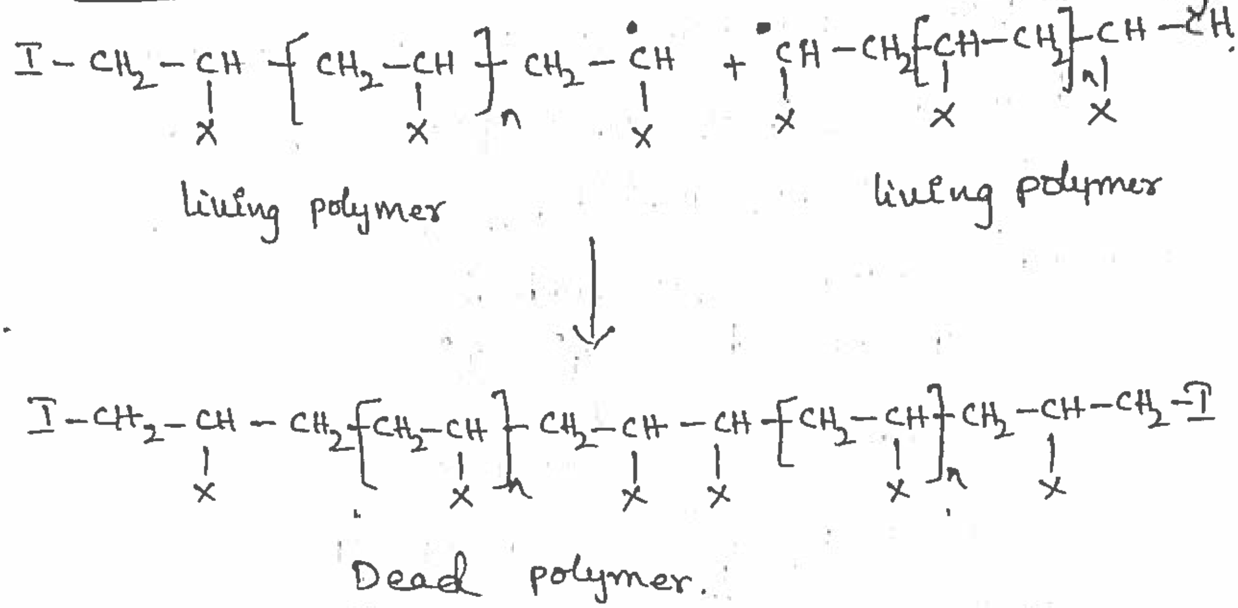
$$I - CH_2 - CH(X) - CH_2 - \dot{C}H(X) + \underset{\text{no. of monomers}}{n\,(CH_2 = CH(X))}$$

$$\downarrow$$

3. Termination :- (To stop chain growth) :-

The propagation polymer chain stops growing and terminates to produce dead polymer. & This process can be carried out by coupling & disproportionation.

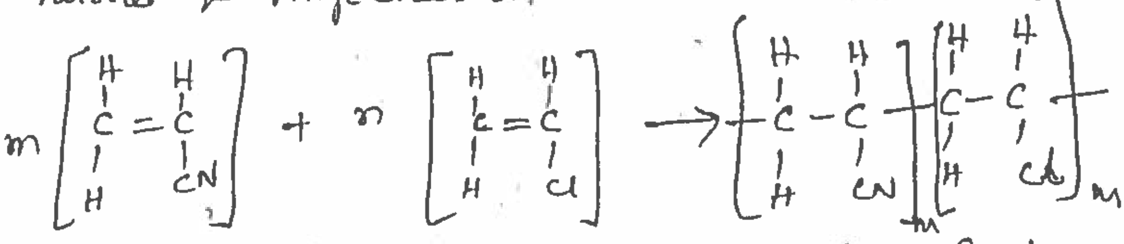
By coupling



Copolymerization :- It is a reaction in which two or more different monomers combine & add together to form a copolymer & this reaction is called as Co-polymerization. 1M

Eg :- Acrylo Nitrile Vinyl chloride, Buna-S rubber.

Acrylo nitrile vinyl chloride is prepared by copolymerization of nitrile & vinyl chloride.



Acrylo Vinyl Acrylonitrile

b) Differentiate Thermoplastics and Thermosets — 6M

S:-

Thermoplastics	Thermosets
* These are formed by additional polymerization	* These are formed by Condensation polymerization
* These are linear (or) branched link structures	* These are cross linked (or) 3D structures
* These can be remoulded, reshaped & reused	* These can't be moulded again & again
* They soften on heating & hardened on cooling	* They can't soften on heating
* These have low melting points	* These have high melting points
* These are soft, weak & less brittle	* These are strong, hard & more brittle
Eg:- polythene	Eg:- Bakelite

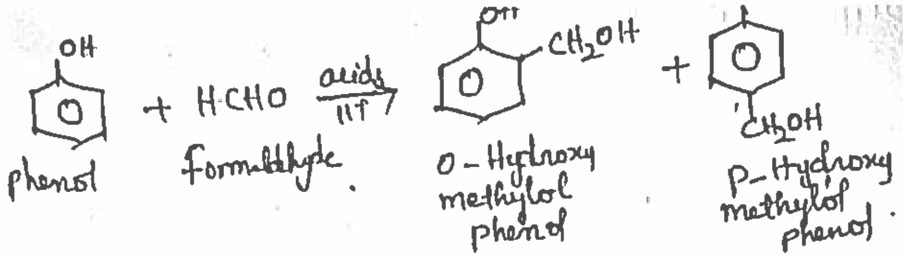
6M

(or)

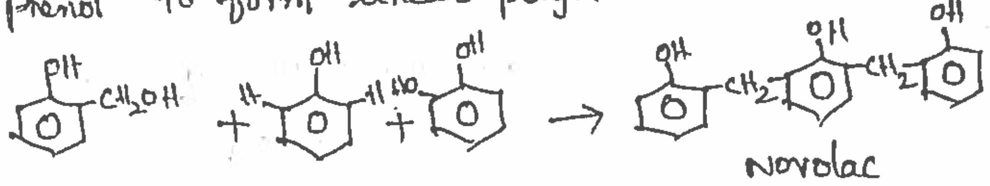
a) Write the preparation, properties and applications of Bakelite and BUNA-N — 8M

s:- preparation of Bakelite :- It is prepared by Condensation of phenol with formaldehyde in presence of acids or alkaline as catalyst. It is formed by different steps

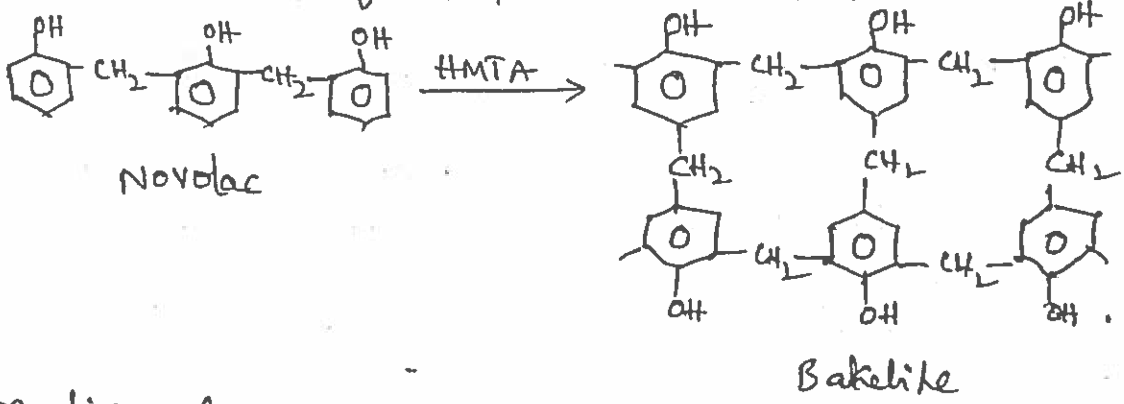
Step-I :-



Step-II :- O-hydroxy methylol phenol which reacts with phenol to form linear polymer called as Novolac



Step-III :- During moulding, catalyst hexamethylene tetra amine is added for providing excess of formaldehyde which converts soluble & fusible novolac resin into hard infusible, cross linked polymer Bakelite



Properties of Bakelite :-

* It is hard, rigid and strong
* It is a scratch resistance and water resistance
* It has good chemical resistance, resistance to acids, salts & many organic solvents

1M

Applications of Bakelite :-

* It is used widely for making electrical insulator parts like switches, switch boards, heater handles etc.
* It is used for moulded articles like ...

1M

Preparation of Buna-N :- It is prepared by the co-polymerization of buta di ene with acrylo nitrile, in presence of sodium as a catalyst. 2

$$m\,[CH_2=CH-CH=CH_2] + n\,[CH_2=\underset{CN}{CH}] \longrightarrow \left[CH_2-CH=CH-CH_2\right]_m \left[CH_2-\underset{CN}{CH}\right]_n$$

1,3-butadiene   Acrylo Nitrile   Buna-N Rubber.

Properties of Buna-N :-

* Because of presence of -CN group structure. It possess excellent resistant to heat, sun light, oils, acids & salts and less resistance to alkali than natural rubber. 1M
* It is an excellent insulator

Applications of Buna-N :-

* It is used for making conveyor belts, high altidude air craft components & automobile parts because of it's strength and light weight 1M
* It is used for making tank linings and pipes for chemical industries

7(b) Distinguish b/w properties & applications of Nylon 6:6 & Carbon fibre — 4M

Ans :- properties of Nylon 6:6 & Carbon fibre

Nylon 6:6	Carbon fibre
* It is translucent, whitish, horny, high melting point	* These are quite costly
* It possess high temperature stability & good scratch	* These are resistant to moisture, acids, bases
	* These are reinforcing agents

2M

Applications of Nylon 6:6 & Carbon fibres

Nylon 6:6	Carbon fibres
* It is used for fibre * It is used for moulding Purpose of gears, bearings * It is used for making filament ropes, bristles.	* These are used in structural components such as wing, body, stabilizer etc. in air crafts & helicopters. * These are used as reinforcing material with epoxy or poly ester resins to form Composites

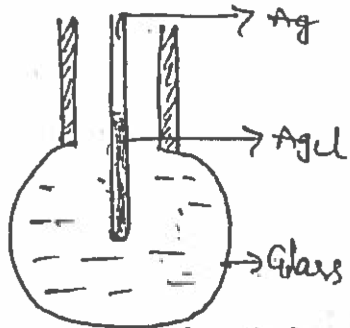
2M

8(a) Explain the Construction & working of Ag|AgCl electrode — 5M

Ans:- Construction:- A glass electrode is a type of ion-selective electrode and Consists of a thin-walled glass bulb attached to a glass tube. A very low melting point and high electrical conductivity glass are used for the Construction of this bulb. The glass tube Contains a dilute solution of Constant pH of HCl (0.1N) solution.

A silver - silver chloride electrode (Ag-AgCl electrode) or pt wire is immersed as reference electrode in the HCl solution

3M



Glass electrode

Working:- It is based upon the observation that when a glass surface is in contact with a solution surface is in contact with a solution, there exist a potential difference b/w glass surface & the solution, the magnitude of which depends upon the H^+ ion concentration of the solution and the nature of glass. The glass electrode may be represented as

$$Ag, AgCl\ (s)\ /\ 0.1N\ HCl\ /\ glass\ /\ H^+ = unknown$$

The electrode potential of the glass electrode depends upon the concn of H^+ ions contained in the experimental solution & are given by

$$E_G = E^0_G + \frac{2.303RT}{F} \log H^+$$

$$= E^0_G - 0.059 \log H^+$$

$$E_G = E^0_G + 0.0591\,pH$$ 2M

E^0_G = standard Electrode potential

3(b) What is potentiometry? Explain how potentiometry method helps to determine the end point in oxidation - reduction titration — 7M

Ans:- potentiometry :- The potential of an electrolyte can be determined by dipping an electrode into the electrolytic solution.

The measurement of potential with an appropriate indicator electrode in conjugation with a reference electrode can be used to find the endpoint in various titrations

principle :- The potential changes with the change in concn of solution being titrated and near the end point there is a sharp change in 2M

End point determination method :- potentiometric titrations are used to determine the end point in various acid - base titrations, redox titrations, precipitation titrations, complexation equilibrium etc.

- Acid - Base Titrations for these titrations, generally hydrogen electrode is used in conjugation with N - calomel electrode as a reference electrode.
- A known volume of acid to be titrated is taken in a beaker having an automatic stirrer.
- Hydrogen & calomel electrodes are dipped inside the beaker, & electrodes are connected to the potentiometer. that records EMF of the solution.
- The base is added gradually from the burette into the beaker and EMF is measured after each addition of the base.
- The values of EMF are plotted against the volume of base added & curve is obtained.

3M

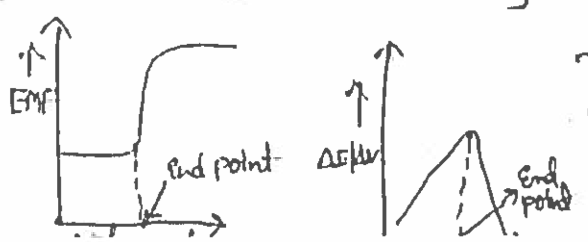
- The potential of hydrogen electrodes is given by $E = E^0 - 0.0591 \log_a H^+$ at $25^\circ c$

where E^0 = standard Electrode potential.

(or)

$$E = E^0 + 0.0591 \, pH \quad [\because pH = -\log(H^+)]$$

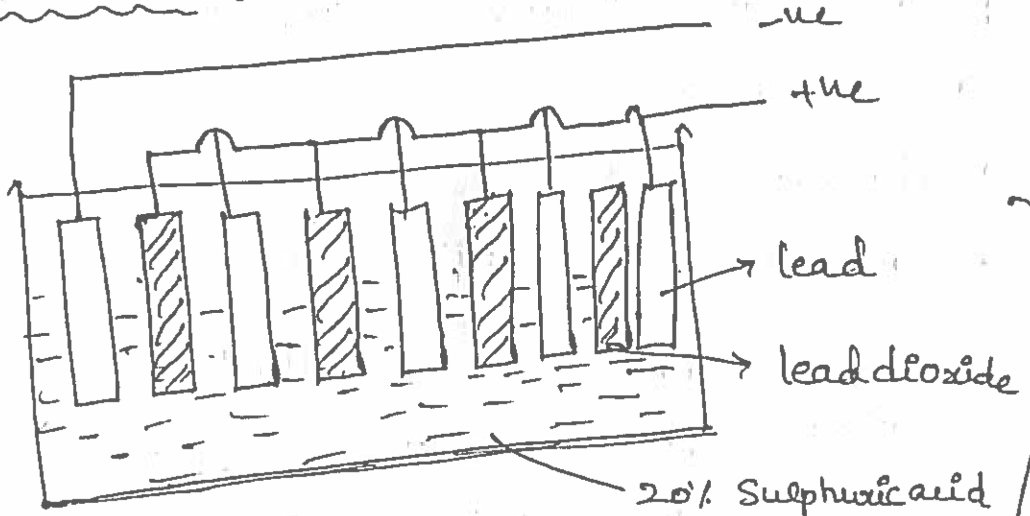
- The end point is given by the point where the EMF increases



2M

(a) · Explain Construction, working and applications of lead acid battery. — 6 M

Ans:- Construction :-



Lead acid battery

2M

If a number of cells are connected in series, the arrangement is called as a battery. The lead storage battery is one of the most Common batteries that are used in the automobiles A 12V lead storage battery is generally used, which Consists of six cells, each providing 2V. Each cell Consist of six cells ie lead anode & a grid of lead packed with lead oxide as the cathode. These electrodes are arranged alternatively, separated by a thin wooden piece & suspended in dil. H_2SO_4 (38%), which acts as an electrolyte. Hence it is called lead acid battery.

Working :- The cell represented as

$Pb/PbSO_4(s), H_2SO_4/PbSO_4(s), Pb$

In this process of discharging, i.e., when the battery produces current, the reactions at the electrodes are as follows.

Discharging reactions :-

At anode :- $Pb \rightarrow Pb^{2+} + 2e^-$

$Pb^{2+} + SO_4^{2-} \rightarrow PbSO_4 \downarrow$

At cathode :- $PbO_2(s) + 4H^+ + 2e^- \rightarrow Pb^{2+} + 2H_2O$

$Pb^{2+} + SO_4^{2-} \rightarrow PbSO_4$

Therefore overall reaction is:

$$Pb(s) + PbO_2 + 4H_2SO_4(aq) \rightarrow 2PbSO_4(s) + 2H_2O + Energy$$

During discharging the battery, H_2SO_4 is consumed & as a result, the density of H_2SO_4 falls. when it falls below 1.20 g/cc, the battery needs recharging. In discharging, the cell acts as voltaic cell where oxidation of lead occurs

Recharging :- During recharging, the cell is operated like electrolytic cell i.e Electrical energy is supplied to it from an external source.

Reactions:- $PbSO_4 + 2e^- \rightarrow Pb + SO_4^{2-}$ (Reaction at cathode)

$PbSO_4 + 2H_2O \rightarrow PbO_2 + 2H_2SO_4 + 2e^-$ (Reaction at anode)

$$2PbSO_4 + 2H_2O + Energy \rightarrow Pb + PbO_2 + 2H_2SO_4$$

Applications:- * It is used for supplying current to mines, laboratories, hospitals, automobiles. 21

(c). Derive the Nernest equation and write its applications - 6 M

A :- Derivation :- Nernest found that the single electrode potential varies with the change in concentration of ions and temperature i.e hence the E.M.F of the cell also varies

He derived mathematical relationship b/w the standard electrode potential, temperature & the concentration of ions. This relationship is known as the Nernest Equation.

Consider the redox reaction: $M^{n+} + ne^- \rightleftharpoons M$

In the above equation the free energy change (G) and its equilibrium constant (K) are related by following equation.

$$\Delta G = RT \ln k + RT \ln \frac{\text{product}}{\text{Reactant}}$$

$$\Delta G = \Delta G^\circ + RT \ln \frac{\text{product}}{\text{Reactant}}$$

where ΔG° is standard free energy change.

free energy change is equivalent to the electrical energy $-nFE$.

$$\therefore \Delta G = -nFE$$ 4M

where n = valency, F = Faraday (96500C), R = 8.314 Jouls.

T = Temperature (K), E = Electrode potential.

$$-nFE = -nFE^\circ + RT \ln \frac{[M]}{[M^{n+}]}$$ $\therefore$ Concentration of M is unity.

$$-nFE = -nFE^\circ - RT \ln [M^{n+}]$$

$$= -nFE^\circ - RT\, 2.303 \log_{10} [M^{n+}]$$

Dividing equation by $-nF$

$$E = E^\circ + \frac{2.303RT}{nF} \log_{10} [M^{n+}]$$

Applications :- It is used ... .

Constant of A cell Reaction.

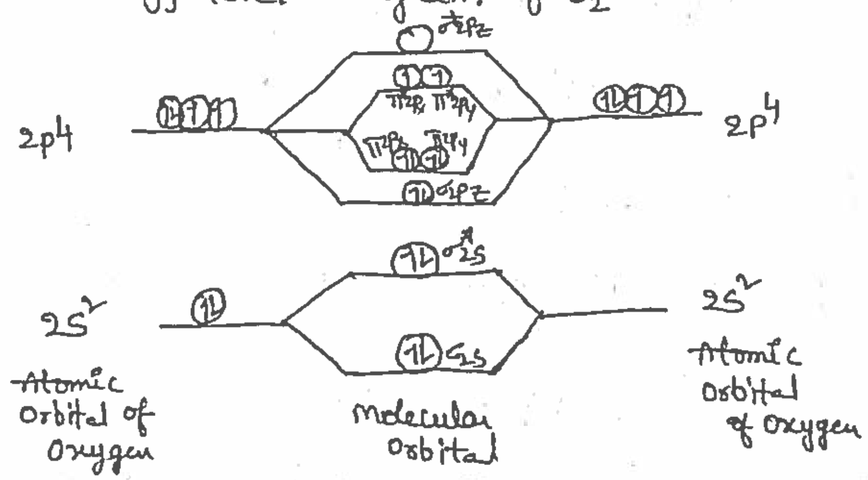
↳ It is used in determining the pH of an electrolytic solution.

↳ It is used in evaluating the heat of reaction inside A cell.

↳ It is also used in calculation of solubility product & cell membrane resting potentials

10(a). Describe the energy level diagram of O_2 and CO molecule. Write their magnetic nature & bond order. — 7M

Ans :- Energy level diagram of O_2



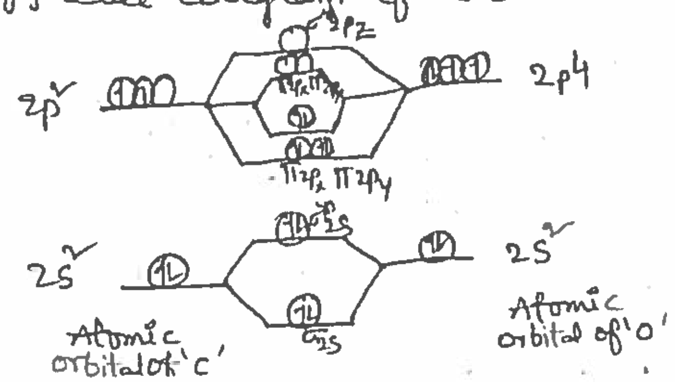
2M

Bond order = $\frac{N_b - N_a}{2} = \frac{8-4}{2} = \frac{4}{2} = 2$

Mangnetic nature = Para magnetic

1.5M

Energy level diagram of CO



2M

$N_b - N_a$ - 8-2 - 2?

b) What is molecular orbital theory ? Describe the molecular orbitals of buta di ene & benzene . — 5M

:- Molecular orbital theory :- It was proposed by Hund and Mulliken in 1932.

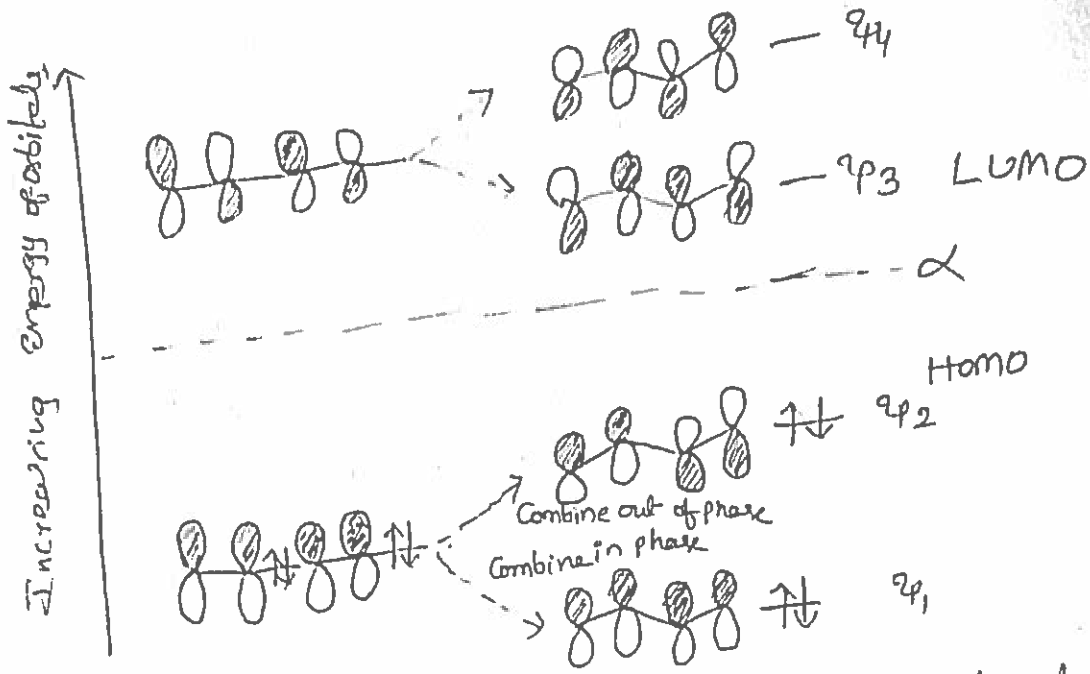
postulates :- New orbitals formed as the molecular orbitals are formed by the overlap of atomic orbitals of the combining atoms.

* The atomic orbitals lose their identity (individual) after the formation of the molecular orbitals.
* The no. of molecular orbitals is equal to the no. of atomic orbitals.
* The electron in an atomic orbital is influenced by just one positive nucleus of the atom i.e., it is monocentric.
* where as depending upon the no. of. atoms in the molecule the electron of a molecular orbital is under the influence of more than one nuclei, i.e., It is polycentric.
* Similar to the atomic orbitals, the filling of electrons in the molecular orbitals also follows the Aufbau, Hund's rule & pauli's exclusive principle.

3M

* Atomic orbitals with comparable energies as well as proper orientations combine to form MO's

Molecular orbitals of ...

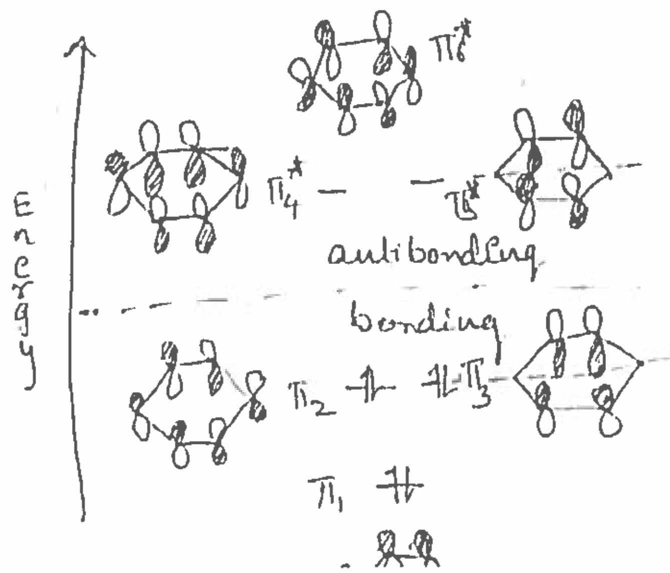


The molecular orbital of butadiene as a result of combining the π-molecular orbitals of two ethene molecules.

The highest occupied molecular orbital (or) Homo is π_2 (or) ψ_2 in 1,3-Butadiene. On contrast, the anti bonding π^* orbitals contain no electrons. The lowest unoccupied orbital (or) LUMO is π_3 (or) ψ_3.

1M

Molecular orbitals of benzene :-



1M

11(a) State the crystal field theory? List out the magnetic properties of coordination compounds. — 6M

Ans:- Crystal field Theory :- It explains many important properties of transition metal complexes including their colours, magnetism, structures & stability

* The central metal (cation) is surrounded by ligands which contain 1 or more lone pair of electrons
* Ligands are treated as point charges
* The ionic ligands like F^-, Cl^-, CN^- etc are regarded as negative point charges and neutral ligands like H_2O, NH_3 etc. regarded as dipoles (i.e. dipolar)
* In a metal complex, the negative end of the neutral ligand is oriented towards the central metal cation.
* There is no interaction b/w metal orbitals & ligand orbitals
* The bonding b/w the metal cation & ligand is purely electrostatic or coloumbic attraction b/w cation & negatively charged anion or negative end of neutral dipole molecule.
* All the d-orbitals on the metal have same energy i.e degenerate in the free atom. However, when a complex is formed, the ligands destroy the degeneracy of these orbitals i.e the orbitals now have different energies

4M

Magnetic properties :- Dia magnetic substances repel the magnetic lines of forces and decrease the flux

* Magnetic (para magnetic) compounds attract the magnetic lines of forces and increase the flux.

* $\mu_{LS} = \sqrt{L(L+1)+4S(S+1)}$ B.M

usually for certain first row transition metal ions, the orbital magnetic contribution is neglected

So $\mu = \sqrt{4S(S+1)}$ B.M

$S = \frac{n}{2}$, n = no. of unpaired electrons

$\mu = \sqrt{4\frac{n}{2}(\frac{n}{2}+1)} = \sqrt{n(n+2)}$ B.M it is

called "Spin-only" magnetic moment

* The experimental value is more than spin-only value but always lesser than μ_{LS} value

* The reason for this is due to L-L Covalent interaction The orbital angular momentum is partially quenched

11(b) Define the following terms.

a. Conductors b. Semi Conductors C Insulators - 6 m

a. Conductors : A Conductor is a material that allows the flow of charge when applied with a voltage. 2

b. Semi Conductor :- A semi conductor is a material whose conductivity lies between conductor and insulator. 2 M

Insulators:- An insulator is a material that does not allow the flow of current. } 2M.

2 (a). Write a short note on Beer - Lambert's law - 5M.

Ans:- Beer - Lambert law:- It explains the relation b/w Concentration & absorbance of solution & it can be expressed by using Beer - Lambert law.

↳ It is Combined form of Beer's law & Lambert's law.

"It states when a beam of monochromatic light is passed through a solution, "The decrease in intensity of radiation with thickness of the absorbing material is directly proportional to the intensity of incident radiation as well as the concentration of solution".

↳ If a mono chromatic light of intensity (I) passes through a solution of molar concentration (C) & the length of the path is x cm.

and then the mathematical form of Beer - Lambert's law is } 3M

$$\boxed{A = \log \frac{I_0}{I} = \varepsilon C x}$$

I_0 = Intensity of incident light

ε = molar absorption coefficient

A = absorbance } 2M

12(b) Explain the principle and instrumentation of FTIR Spectroscopy with neat diagram - 7M

Ans:- FT-IR Spectroscopy :- FT-IR means Fourier Transform Infrared is the Preferred method of IR spectroscopy.

principle :- Based on the principle of interferometry and interferogram will be obtained, which is a complex signal occurs in wave like pattern. Inter-ferogram signal is plotted b/w intensity Vs time. 2M

Instrumentation:- In this spectroscope the source of light is Nemest filament consisting of spindle of rare earth oxides about 1 inch long x 0.1 inch diameter.

↳ In this a parallel beam of radiation is directed from the source to the interferometer device.

↳ This interferometer device that separates a beam of light into two beams (rays) by a beam splitter.

↳ So as to reflect 50% of radiation falling on it.

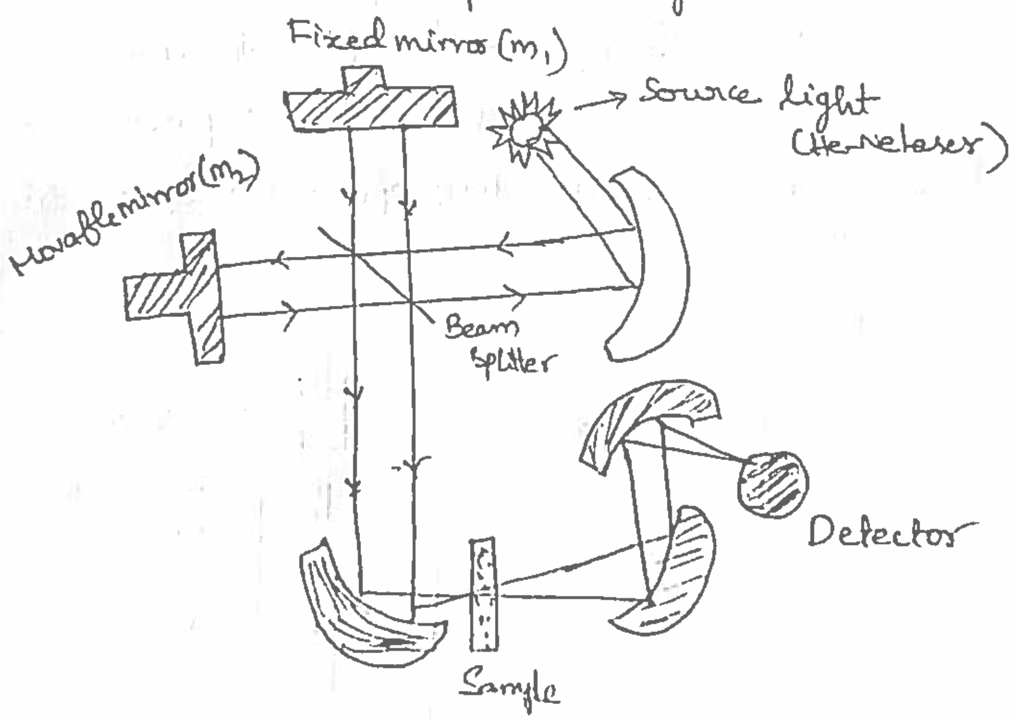
↳ one beam reflects off of a flat mirror (m_1) which is fixed in place.

↳ other beam reflects off of a flat mirror (m_2) which is on a mechanism which allows this mirror to move a very short distance away from the beam splitter. 3M

The two beams reflect off of their respective mirrors and are recombined when they meet back at the beam splitter.

Because the path that one beam travels in a fixed length and the other is constantly changing as its mirror moves, the signal which exists the interf-erometer is the result of two beams interfering with each other.

→ The resulting signal is called an interferogram which has the unique property that every data point which makes up the signal.



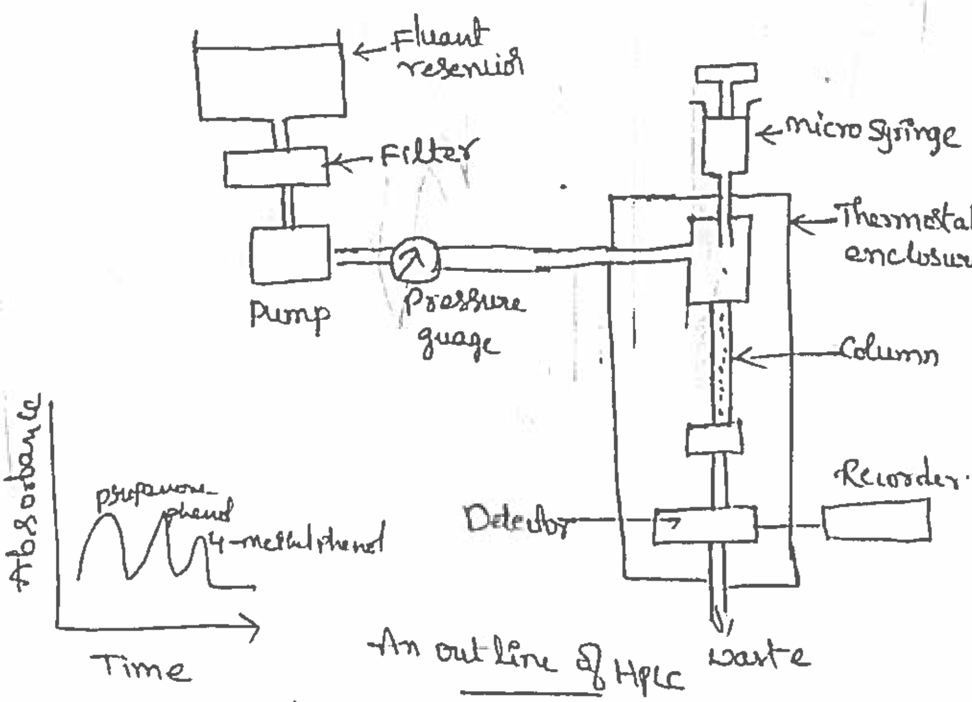
2 M

13(a) Explain the principle and instrumentation of HPLC —6M

Ans:- HPLC is a technique for separation, identification and quantification of components in a mixture.

Principle :- In this it relies on pumps to pass a pressurized liquid solvent containing the sample mixture through a column filled with a solid adsorbant material.

"Each component in the sample interacts slightly differently with the adsorbent material, causing different flow rates for the different components & leading to the separation of the components as they flow out of the column. 2



An out line of HPLC

2M

Output signal

Instrumentation :- However because liquids are more viscous than gases, so the pressure used to make them pass through a column is greater than in GLC between 20 and 200 atm.

↳ Such high pressure requires a strong column, which is often about 25 cm length.

↳ This principle is much the same as in GLC.

↳ In this process the molecules coming off the column are detected by an uv spectrophotometer & the output appears as a series of peaks very much like the GLC charts.

↳ Because of its accuracy HPLC has become very widely used in analysis & research.

2m

13(b) Explain the determination of end point in acid-base titration using pH meter. — 6m.

Ans:- pH metry involves the measurement of pH with the addition of the reactants similar to potentiometric titrations, the change in pH is noted with addition of reagent from the burette & the end point can be determined graphically by plotting the pH against the volume of the titrant added.

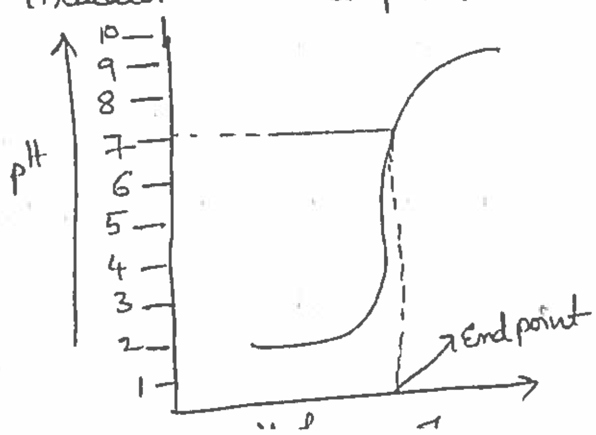
↳ The pH is measured with the temp ...

↳ pH metric Titrations :- Determination of end point in acid base titrations :-

* The acid to be titrated is taken in a beaker.
* The combined glass electrode connected to the pH meter is dipped in the beaker & the pH is noted.
* Strong base is filled in the burette and is added gradually to this solution which constant stirring & the pH is obtained after each addition. 4
* The pH is then plotted against the volume of the base added.
* The volume at pH 7 (neutral) gives the end point of the titration.

Advantages of pH metry :-

This method gives accurate results for end point without the use of indicator & eliminates errors because of difference in the observation of colour changes of indicator at end point.



2M

1(a) What is the basic lock and Key principle? — 5M

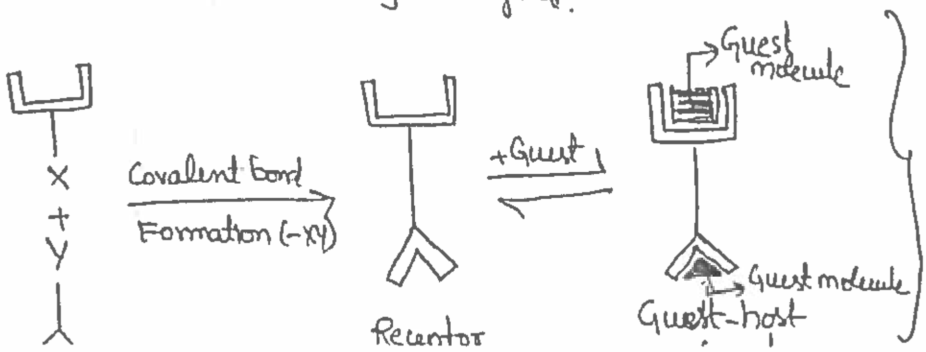
Ans :- Basic lock and key principle:- In the supra-molecular systems, the attractive forces operate efficiently when the receptor (i.e. host) provides a suitable cavity or site that properly matches both electronically and sterically (i.e. properties related with the shape and size) with the substrate (i.e. guest).

Thus, in the guest—host molecular assembly, the molecular components must maintain the proper complementarity both electronically & sterically.

This is why, the components of the supra molecular assembly can recognise each other through the interply of supramolecular noncovalent forces. This leads to molecular recognition & molecular recognition leads to the supra molecular assemblies which are also described as supermolecules.

To recognise the substrate (i.e. guest), the receptor (i.e host) must be suitably designed.

3M



2M

14(b). Discuss about the ... — 7M
and catalysis

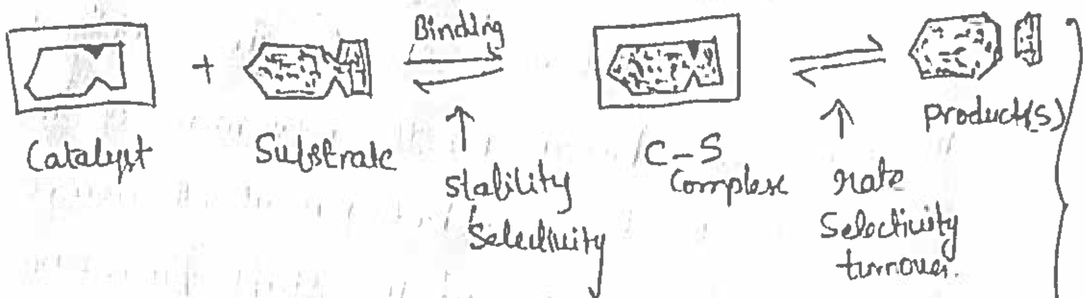
Ans:- Supra molecular reactivity & catalysis:-

The design of highly efficient and selective reagents & catalysis is one of the major goals of research in chemistry, the science of matter & of its transformations.

The particularly remarkable features displayed in this respect by the natural catalysts, the enzymes has provided major stimulus & inspiration for the development of novel catalysts by either manipulating the natural versions or by trying to device entirely artificial catalysts that would nevertheless display similar high efficiencies & selectivities.

Since enzymatic reactions involve binding of and reaction with precisely defined substrate they have the characteristics of a supramolecular process. on the otherhand, reactivity and catalysis represent major features of the functional properties of supra molecular systems

Molecular receptors bearing appropriate reactive groups in addition to binding sites may Complex a substrate (with given stability selectivity & kinetic features) and with it and release the products, thus regenerating the reagent for a new cycle.



3m

Schematic representation of supramolecular catalysis

Supramolecular reactivity & catalysis thus involve two main steps: binding, which selects the substrate and transformation of the bound species into products within the supermolecule formed. Both steps take part in the molecular recognition of the productive substrate and require the correct molecular information in the reactive receptor.

(OR)

15(a) List out the applications of Catenands & rotaxanes – 6M

Ans:- Applications of catenands:-

1. Catenanes have been used to create molecular switches, molecular motors, fabrication of molecular electronic devices, molecular sensors & chemical sensors.
2. Catenanes are catalysts & sensors to polymers.
3. Another important application of catenans is

3m

Applications of rotaxanes :-

1. Molecular machines :- The potential use of rotaxanes in molecular electronics on logic molecular switching elements and molecular shuttles. these molecular machines are based on the movement of the macrocycle on the dumbbell molecule. The macro cycle can rotate around the shaft from one site to another, rotate like wheel, and axle to function as molecular switch.

2. Ultrastable dyes :- Rotaxane's potential application in long lasting dyes based on enhanced stability of the inner portion of the dumbbell shaped molecule for example cyclo dextrin protected azo dyes.

3. Nanorecording :- Rotaxane is deposited as Long--muir - Blodgett film on ITO - Coated glass a memory dot.

3

15 (b) Explain computational chemistry & molecular docking - 6M

Ans :- Computational chemistry :- It is a branch of chemistry that used computer & simulations computer programs (software) to help in solving chemical problems.

the early 2000s the development of efficient computer based algorithms, into a science of its own, which today has reached a high level of maturity & sophistication.

→ The use of computational chemistry was for predicting the structures and properties of biomolecules. 3n

↳ It is widely used to design new drugs & materials.

↳ Examples of such properties are structures(expected position of atoms)., absolute & relative energies(electronic charge density, dipoles, vibrational frequencies & other spectroscopic quantities.

↳ Computational chemistry method ranges from very approximate to highly accurate.

Molecular docking:- In the field of molecular modelling, docking is a method which predicts the preferred orientation of one molecule to a second when bound to each other to form a stable complex.

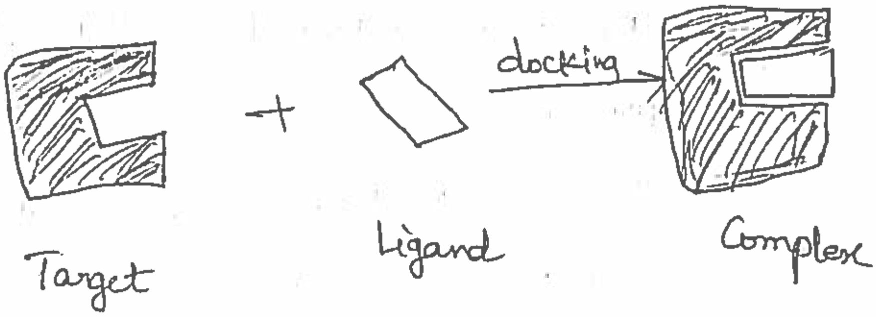
↳ MD used to predict the strength of association & binding affinity b/w two molecules using for example, scoring functions.

↳ The association b/w biologically relevant molecules such as proteins, peptides, nucleic acids, carbohydrates & lipids play a central role in signal

↳ one can think of molecular docking as "lock-and-key" in which one wants to find the correct relative orientation of the 'key' which will open up the 'lock'. Here, the protein can be thought of as the lock & the ligand can be thought as a key.

↳ However, since both the ligand & the protein are flexible, a "hand-in-glove" analogy is more appropriate than lock-and-key.

3



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ANSWER KEY & SCHEME OF EVALUATION

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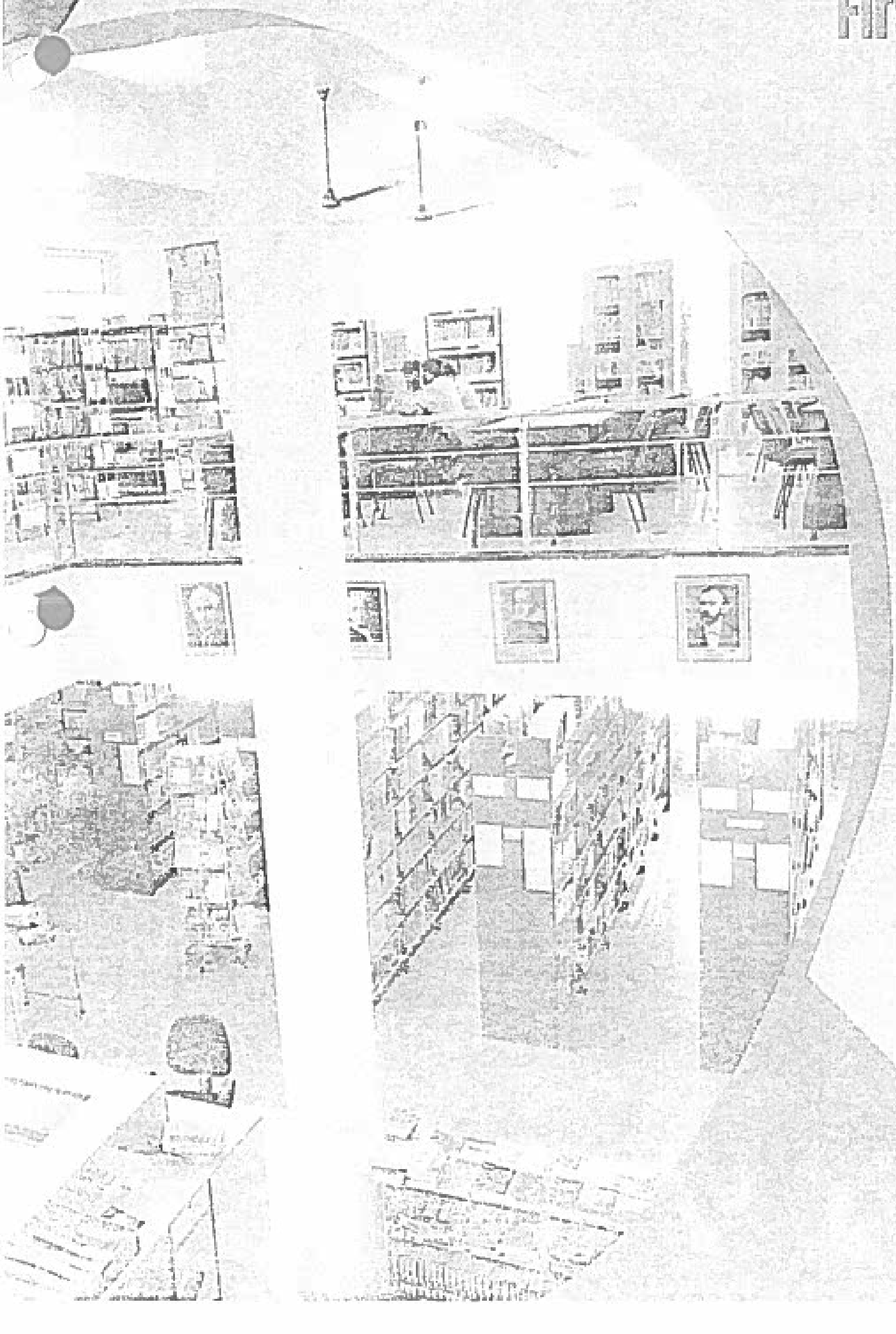


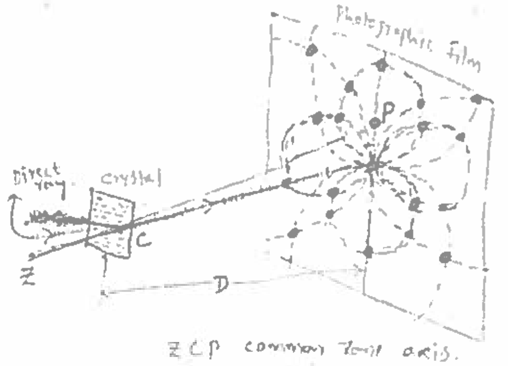
15(b)

The Laue method (For single crystal) :

→ This method is used to study the orientation of the crystal & verify crystal symmetry

→ Here, a single crystal specimen is held fixed and is illuminated with "white X-radiation" (X-radiation with continuous wavelength).

→ Rays diffracted through the crystal is made to fall on a photographic film as shown in fig.



ZCP common zone axis.

→ Since the crystal is fixed in position, the angles of diffraction 'θ' are also fixed

Reflection spots at various distances R from the direct beam

$$R = D \tan 2\theta$$

→ Each spot is due to all the orders of reflection n = 1, 2, 3, .. superimposed from a single plane

→ Interpretation of the spots in the film leads to know about the crystal structure of the specimen

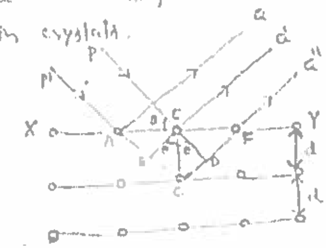
Drawback
→ Since λ values for different reflections are not known, the determination of unit cell dimensions is not possible

(9)

Diffraction of x-rays by crystal planes : Bragg's law

→ Since the interatomic spacing is only 2-3 Å (order), x-rays whose wavelength is in the same range can be used for crystal diffraction studies.

→ crystals act as 3-d space grating ; and the diffraction pattern thus produced reveals the internal arrangement of atoms in crystals.



Bragg's law

→ consider a crystal made up of equidistant parallel planes of atoms with the interplanar spacing 'd'

→ wave front of monochromatic x-ray beam of wavelength λ at an angle 'θ' incidents on these atomic planes

→ Each atom scatters x-rays in all directions

→ In certain directions these scattered radiations are in phase i.e they interfere constructively

→ let x-rays PE & P'A inclined at an angle 'θ' with the top of the crystal plane XY

→ They scattered along AQ, EQ' an angle 'θ'

→ another x-ray from P'C scattered along CQ''

→ Draw normals EB & ED to AC & CF

→ the path difference between two incident rays is

$$\Delta = BC + CD$$

In Δ^{le} BEC $\quad \sin\theta = \frac{BC}{EC} = \frac{BC}{d} \quad \Rightarrow \quad BC = d\sin\theta$

and in Δ^{le} DEC $\quad \sin\theta = \frac{CD}{EC} \quad \Rightarrow \quad CD = d\sin\theta$

Hence the path difference $\Delta = 2d\sin\theta$

→ If the scattered waves are in phase, then the path difference must be an integral multiple of wavelength

i.e $\boxed{2d\sin\theta = n\lambda}$ $\quad n = 0, 1, 2, 3, \ldots$ order of diffraction

→ This condition is known as Bragg's law.

Since the max possible value for $\sin\theta$ is '1'

$\frac{n\lambda}{2d} \leq 1$, this sets "limitation on wavelength"

i.e "λ should not exceed twice the interplanar spacing for diffraction to occur."

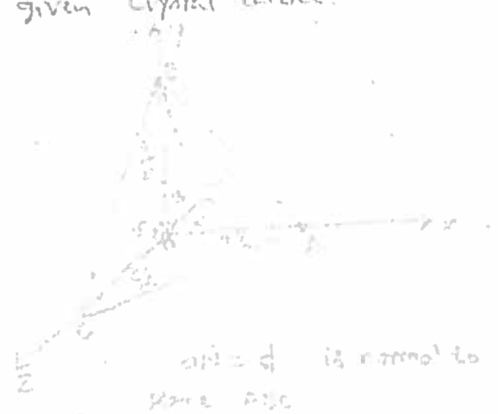
Distance of seperation between successive (hkl) planes

Now we shall derive an expression for the spacing between two parallel planes in a given crystal lattice.

In the fig 'O' the origin is taken at a lattice point.

Let any set of planes represented by miller indices (h k l).

Imagine the reference plane is passing through the origin O and the next plane cutting the intercepts $\frac{a}{h}, \frac{b}{k}, \frac{c}{l}$ on X, Y, Z axes



→ A normal 'ON' is drawn to the plane ABC

→ length of this normal d is called "Inter planar seperation"

α, β, γ be the angles made by ON with X, Y, Z axes

then
$$\cos\alpha = \frac{ON}{OA} = \frac{d}{a/h}$$

$$\cos\beta = \frac{ON}{OB} = \frac{d}{b/k}$$

$$\cos\gamma = \frac{ON}{OC} = \frac{d}{c/l}$$

But from the law of direction cosines $\cos^2\alpha + \cos^2\beta + \cos^2\gamma = 1$

$$\therefore \left[\frac{d}{a/h}\right]^2 + \left[\frac{d}{b/k}\right]^2 + \left[\frac{d}{c/l}\right]^2 = 1$$

$$d^2\left[\frac{h^2}{a^2} + \frac{k^2}{b^2} + \frac{l^2}{c^2}\right] = 1$$

Interplanar seperation $\boxed{d = \frac{1}{\sqrt{\frac{h^2}{a^2} + \frac{k^2}{b^2} + \frac{l^2}{c^2}}}}$

In a cubic system $a = b = c$

$$\boxed{d = \frac{a}{\sqrt{h^2 + k^2 + l^2}}}$$

where $P = 1,2,3,4$ … etc. for fundamental, first over tone, second over toneetc.,
Y = Young's modulus of the crystal and ρ = density of the crystal.

- The variable condenser C_1 is adjusted such that the frequency of the applied AC voltage is equal to the natural frequency of the quartz crystal, and thus resonance takesplace.
- The vibrating crystal produces longitudinal ultrasonic waves of largeamplitude.

Advantages

- Ultrasonic frequencies as high as 5 x 108Hz or 500 MHz can be obtained with thisarrangement.
- The output of this oscillator is veryhigh.
- It is not affected by temperature andhumidity.

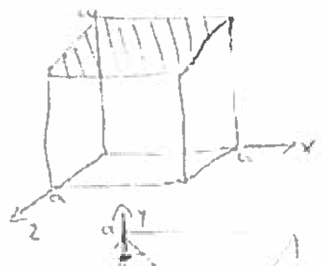
Disadvantages

- The cost of piezo electric quartz is veryhigh
- The cutting and shaping of quartz crystal are verycomplex.

14(a) Miller indices form a notation system in crystallography for lattice planes in crystal lattices. In particular, a family of lattice planes of a given Bravais lattice is determined by three integers h, k, and l, the Miller indices.

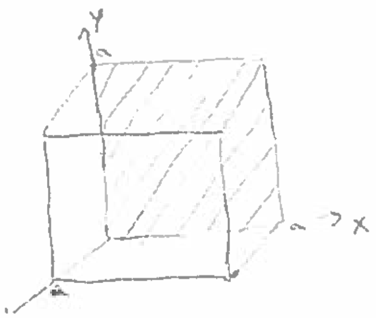
planes

(0 1 0)

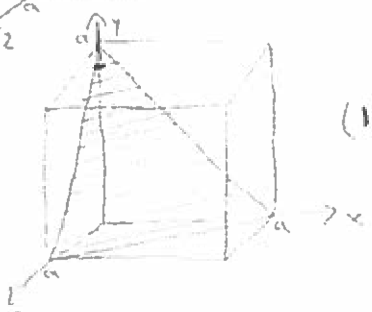


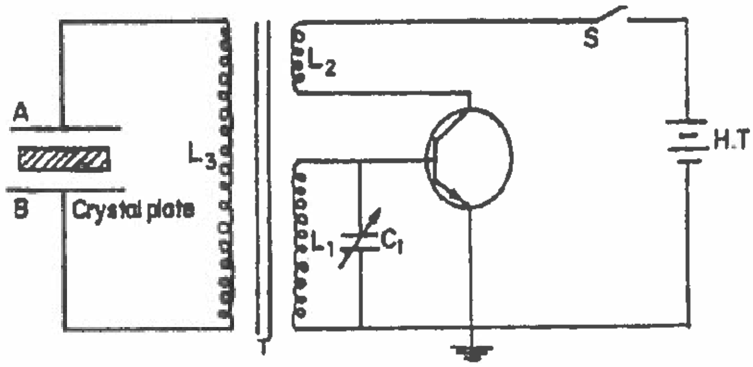
3×2 = 6

(1 1 0)



(1 1 1)





- The quartz crystal is placed between two metal plates A andB.
- The plates are connected to the primary (L_3) of a transformer which is inductively coupled tothe electronicsoscillator.
- The electronic oscillator circuit is a base tuned oscillatorcircuit.
- The coils L_1 and L_2 of oscillator circuit are taken from the secondary of a transformerT.
- The collector coil L_2 is inductively coupled to base coilL_1.
- The coil L_1 and variable capacitor C_1 form the *tank circuit* of theoscillator.

Working

- When H.T. battery is switched on, the oscillator produces high frequency alternatingvoltages withafrequency. $f = \frac{1}{2\pi\sqrt{L_1C_1}}$
- Due to the transformer action, an oscillatory e.m.f. is induced in the coil L_3. This high frequency alternating voltages are fed on the plates A andB.
- Inverse Piezo-electric effect takes place and the crystal contracts and expands alternatively. The crystal is set into mechanicalvibrations.
- The frequency of the vibration is givenby

$$f = \frac{P}{2l}\sqrt{\frac{Y}{\rho}}$$

Echo		
	- If the time interval between the direct sound and the reflected sound is less than 1/15 of a second, the reflected sound reaches the audience later than the direct sound.	▪ properly covering the long distance walls, high ceilings with suitable sound absorbing materials.
Echelon Effect		
	- new sound produced by *repetitive echoes* - regular reflecting surface like stair case may create this effect.	▪ Cover such regular reflecting surfaces properly.
Focusing		
	- Reflected sound by the ceiling and wall is focused at a particular area of the hall. - **Plane surface** : reflect and distribute the sound evenly.	▪ cover the curved surfaces with proper sound absorbing materials ▪ radius of curvature of concave ceiling

(1) 13(a) Inverse piezo electriceffect

- If mechanical pressure is applied to one pair of opposite faces of certain crystals like quartz, equal and opposite electrical charges appear across its other faces. This effect is called as piezo-electric effect.
- The converse of piezo electric effect is alsotrue.
- If an electric field is applied to one pair of faces, the corresponding changes in the dimensions of the other pair of faces of the crystal are produced. This effect is known as inverse piezo electriceffect.

Construction

The circuit diagram is shown in Figure

$$E_4 = \frac{P}{3\varepsilon_0}$$

The resultant internal field or Lorentz field can be written as

$$E_i = E_1 + E_2 + E_3 + E_4$$

$$E_i = (E + \frac{P}{\varepsilon_o}) - \frac{P}{\varepsilon_o} + 0 + \frac{P}{3\varepsilon_o}$$

$$E_i = E + \frac{P}{3\varepsilon_o}$$

This is the expression for internal field of a solid. This is also called Lorentz field.

11(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.

FACTORS	DEFINITION	REMEDIES
Reverberation Time	- Time taken by the sound wave to fall below the minimum audibility level after the source is stopped - Reverberation Time is too high: overlapping of successive sound - ReverberationTime is too low : produced sound will disappear - for the good audibility , reverberation time should be kept at an optimum value.	by installing sound absorbing materials like ▪ arranging full capacity of audience ▪ completely covering the floor with carpets ▪ decorating the walls with drawing boards, picture boards
Loudness	- degree of sensation produced in the ear. - uniform distribution of loudness must be maintained - due to high absorption or low reflecting surfaces near the sound source	***If loudness is low:*** ▪ speakers may be placed at regular distances ▪ lowering the ceiling and placing reflecting surfaces at necessary places. ***If loudness is high:*** ▪ sound absorbents can be placed at noisy places

A. 11(a) Local field or internal field in a dielectric is the space and time average of the electric field intensity acting on a particular molecule in the dielectric material.

Consider a dielectric be placed between the plates of a parallel plate capacitor and let there be an imaginary spherical cavity around the atom A inside the dielectric.

It is also assumed that the radius of the cavity is large compared to the radius of the atom.

The internal field at the atom site 'A' can be made up of four components E_1, E_2, E_3 and E_4.

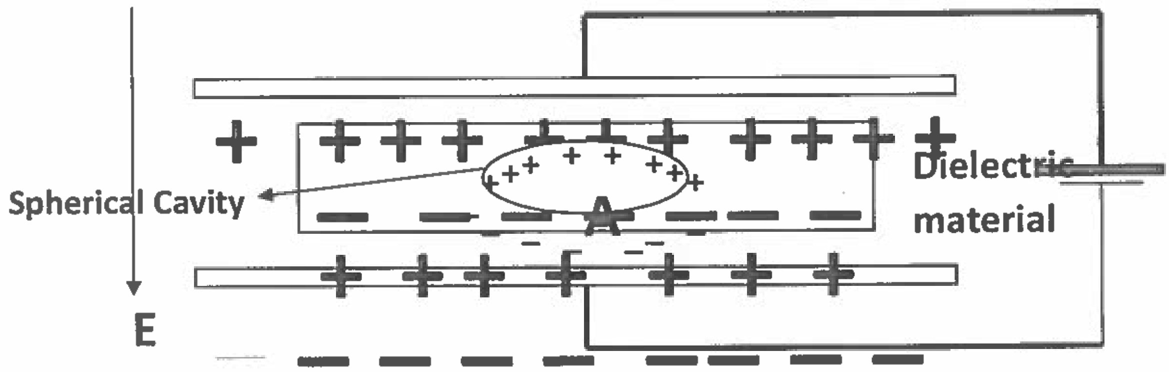
Field E_1:

E_1 is the field intensity at A due to the charge density on the plates, from the field theory,

$$E_1 = \frac{D}{\varepsilon_0} \text{ and } D = \varepsilon_0 E + P$$

$$E_1 = \frac{\varepsilon_0 E + P}{\varepsilon_0}$$

$$E_1 = E + \frac{P}{\varepsilon_0} \quad \text{........ (1)}$$



Field E_2:

E_2 is the field intensity at A due to the charge density induced on the two sides of the dielectric.

$$E_2 = \frac{-P}{\varepsilon_0} \quad \text{....... (2)}$$

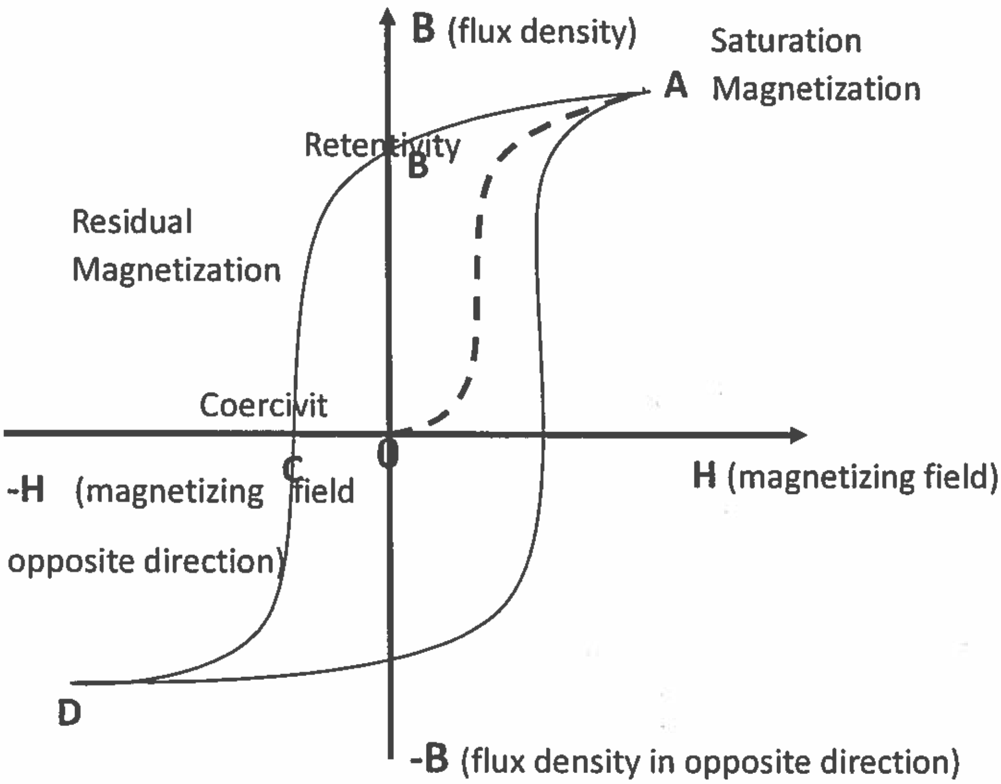
Field E_3:

E_3 is the field intensity at A due to the other atoms contained in the cavity, we are assuming a cubic structure, so $E_3 = 0$.

Field E_4:

E_4 is the field intensity due to polarizing charges on the surface of the spherical cavity and calculated by Lorentz.

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.

i. ***Retentivity***: It is the property of magnetic material in which the magnetic flux density remaining, when the applied field is reduced from saturation to zero. The value of magnetic flux density at point b on the hysteresis curve shows retentivity.

ii. ***Coercivity***: It is the property of magnetic material in which the residual magnetic flux density becomes zero at certain value of reverse magnetic field applied to the material.

10(b) Two equal and opposite magnetic poles separated by finite distance is known as magnetic diploe.

The magnetic flux density (B) is directly proportional to the magnetic filed intensity (H).

$$B \propto H \Rightarrow B = \mu H$$

where μ is proportionality constant and is known as permeability of the medium.

Since ($n_1 \approx n_2$) therefore $n_1 + n_2 \approx 2n_1$

$$NA = \sqrt{(2n_1)(n_1 - n_2)}$$

$$NA = \sqrt{(2n_1^2)(\frac{n_1 - n_2}{n_1})}$$

$$NA = \sqrt{n_1^2 2\Delta} \quad \text{Where } \Delta = \frac{n_1 - n_2}{n_1}$$

$$NA = n_1\sqrt{2\Delta}$$

Where Δ is a fractional difference between the refractive indices of core and cladding, it is known as fractional refractive index change. It is expressed as $\Delta = \frac{n_1 - n_2}{n_1}$.

9(b) Given acceptance angle=30^0

Numerical aperture=Sin 30^0

=0.5

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

$$\sin\theta_a = \frac{\sqrt{n_1^2 - n_2^2}}{n_0}$$

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.

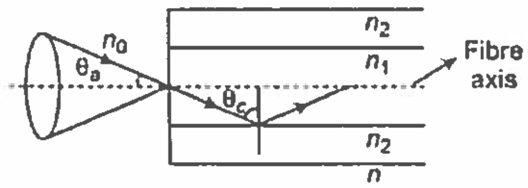
The angle $\theta_a$is called the acceptance angle of the fiber. The acceptance angle may be defined as the maximum angle of incidence that light ray makes with the axis of the fiber to get the total internal reflection. It is also called Acceptance cone half angle.

Acceptance Cone:

The light rays contained within the cone having a full angle $2\theta_a$ are accepted and transmitted along the fiber. Therefore the cone is called the acceptance cone.

If the diameter of the core is large, the acceptance angle is large.

Rotating the Acceptance angle about the fiber axis describes the acceptance cone of the fiber.



Acceptance angle of fibre

Numerical Aperture:

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. It is the measure of the amount of light that can be accepted by a fiber. It depends only on Refractive indices of core and cladding and not on fiber dimensions. It is always < 1 and ranges from 0.13 to 0.50. A larger numerical aperture implies that a fiber will accept a large amount of light from the source.

Numerically it is equal to sine of the acceptance angle.

$$NA = \sin\theta_a$$

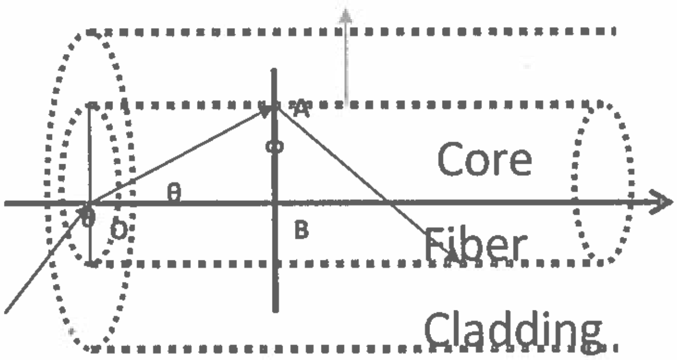
$$\sin\theta_a = \frac{\sqrt{n_1^2 - n_2^2}}{n_0}$$

For air medium $n_0=1$, then

$$NA = \sqrt{n_1^2 - n_2^2}$$

Generally n_1 is slightly greater than n_2.

$$NA = \sqrt{(n_1 + n_2)(n_1 - n_2)}$$



From Δ^{le} OAB, $\quad \phi + \theta_r = 90^0$

$$\theta_r = 90^0 - \phi$$

$$\sin\theta_r = \sin(90^0 - \phi)$$

$$\sin\theta_r = \cos\phi \qquad \text{.......... (2)}$$

Substituting equation (2) in equation (1), we get

$$\sin\theta_i = \frac{n_1}{n_0}\cos\phi \qquad \text{........... (3)}$$

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

4. The neon atoms in Ne_3 state de-excited spontaneously to Ne_2 state by emitting an electromagnetic radiation of wavelength 6000 Å.

The transition of 11500 Å and 33900Å reduces 6328 Å transition. In order to get only 6328Å output, the laser tube windows are made up of glass or quartz that absorbs strongly 11500Å and 33900Å.

The emitted photons during the transition from Ne_6 to Ne_3 travel through the gas mixture. If this photon is moving parallel to the axis of the tube, it is reflected back and forth by the reflectors until it stimulates and an excited Ne atom and causes it to emit a fresh photon in phase with the stimulating photon.

This process is continued and a laser beam builds up in the tube. When the beam becomes sufficiently intense, a portion of it escapes from the partially reflecting end. The wavelength of the laser beam is 6328Å.

8(b) **Population inversion:** If the number of atoms is more in higher energy level than the number atoms in the lower energy level is called as population inversion. The condition for population inversion is $N_2 > N_1$.

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 θ_c the light will stay within the fiber.

According to Snell's law

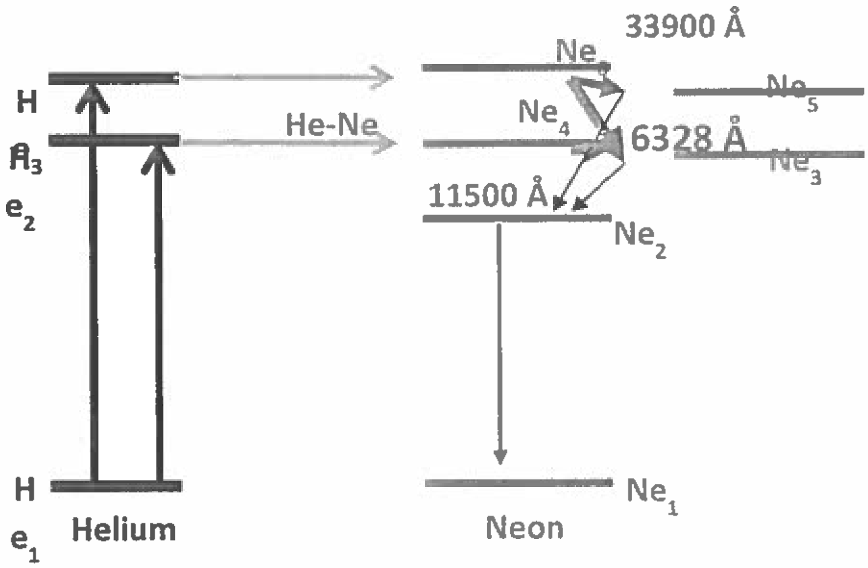
$$n_0 \sin\theta_i = n_1 \sin\theta_r$$

$$\sin\theta_i = \frac{n_1}{n_0}\sin\theta_r \quad .... (1)$$

If θ_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.

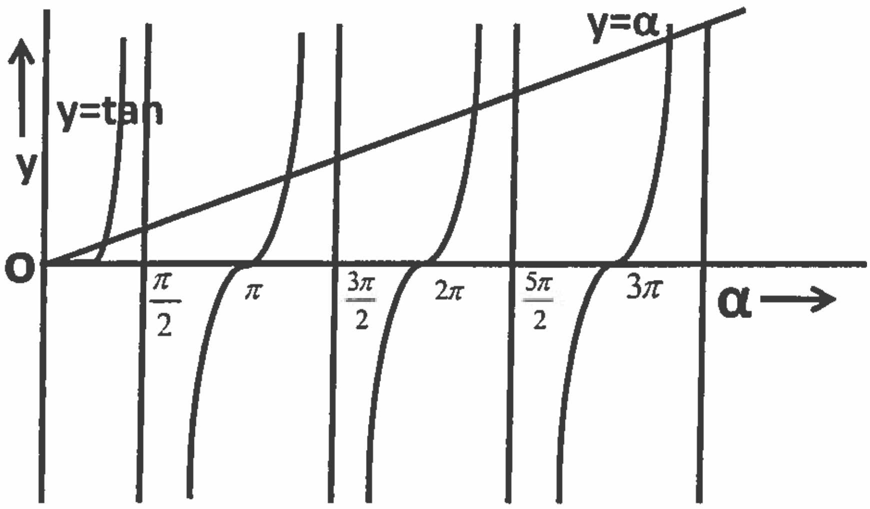
In helium, there are three active energy levels namely He_1, He_2 and He_3 where as in neon atom there are six active energy levels namely Ne_1, Ne_2, Ne_3, Ne_4, $Ne_5$and Ne_6. When an electric discharge passed through the gas the electrons which are accelerated down the tube, collide with the helium and neon atoms. Helium atoms are excited very efficiently by electron impact into the higher energy levels He_2 and He_3, but the neon atoms are remains in the ground state. The He_2, He_3 levels are metastable states of helium. The toms stay longer time in these states. The lifetimes of He_2 and He_3 levels are 10^{-4} and $5x10^{-5}$ seconds respectively.

Now these helium atoms in the metastable states are in elastically collided with the neon atoms which are in the ground state and excite the neon atoms to their metastable states Ne_4 and Ne_6, while the helium atoms return to their ground states. The energy level Ne_4 is coinciding with He_2 and Ne_6 is coinciding with He_3. Therefore the states Ne_4 and Ne_6 of neon atoms act as metastable states. So the neon atoms stay longer time in these states. As the energy exchange continues the population inversion will be achieved in metastable states Ne_4 and Ne_6.



The possible transitions are

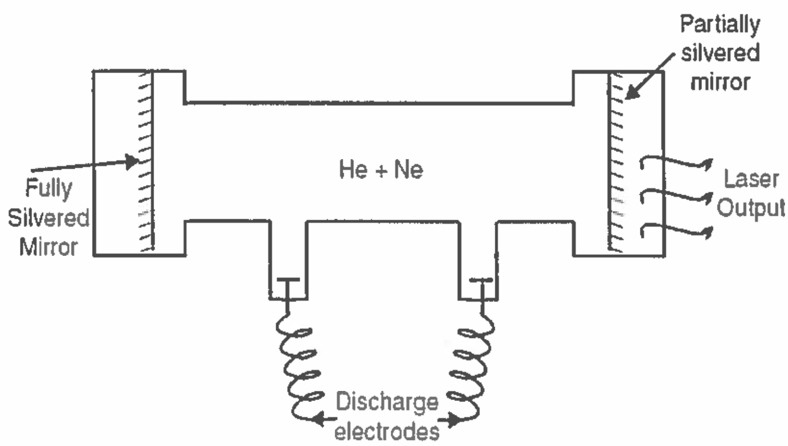
1. Some of the neon atoms de-excite from Ne_6 and Ne_5. In this transition the electromagnetic radiation of wavelength of 33900 Å will be emitted.
2. The other neon atoms de-excite from Ne_6 and Ne_3. During this transition a photon of wavelength 6328 Å is emitted. This is the important and major wavelength in this laser.
3. The neon atoms in the Ne_4 state are de-excited to Ne_3 then an electromagnetic radiation of wavelength 11500 Å is emitted.



8(a) He-Ne laser was first gas laser is based on the four level systems, so it is called four level lasers. It was built by Ali Javan, William R. Bennet Jr and Donald R. Herriott at Bell laboratories in December 1960. He-Ne laser produces a continuous output power of the order of few mWatts, so it is a continuous laser.

Construction:

Helium-Neon laser consists of a long narrow quartz discharge tube of a diameter of 1 to 1.5cm and length 80 to 100 cm is filled the mixture of helium and neon gases, in approximately a 10:1 ratio. The helium atoms are at a pressure of 1mm of Hg and the neon atoms are at pressure of 0.01 mm of Hg. The laser action takes place in the energy levels of the neon atom. Helium atom helps to achieve the population inversion by imparting their energy to the Ne atoms. The energy or pump source of the laser is provided by a high voltage electrical discharge passed through the gas between electrodes (anode and cathode) within the tube. The tube has got two parallel mirrors. One is completely reflecting and the other partially reflecting in order to amplify the output laser beam.



Working:

$$I = R^2$$

$$I = A^2\left(\frac{\sin^2\alpha}{\alpha^2}\right)$$

$$I = I_0\left(\frac{\sin\alpha}{\alpha}\right)^2 \quad .... (5) \qquad \text{Where } A^2 = I_0$$

Therefore the intensity I depends upon the value of α

Case i: Condition for minima

The intensity will be minimum, when

$\sin\alpha = 0 \Rightarrow \alpha = \pm m\pi$ (But $\alpha \neq 0$)

If $\alpha = \pm m\pi$

$$\Rightarrow \frac{\pi a \sin\theta}{\lambda} = \pm m\pi$$

$\therefore a\sin\theta = \pm m\lambda$ Where m = 1, 2, 3 .....

This is the condition for minimum intensity

The first order minima occur at $\theta = \pm\sin^{-1}\left(\frac{\lambda}{a}\right)$

The second order minim occur at $\theta = \pm\sin^{-1}\left(\frac{2\lambda}{a}\right)$ and so on

Case ii: Condition for maxima

The condition for secondary maxima can be obtained by differentiating equation (5) with respect to α and equating to zero.

$$\frac{dI}{d\alpha} = 0$$

$$\frac{d}{d\alpha}\left[\frac{I_0\sin^2\alpha}{\alpha^2}\right] = 0$$

$$I_0\frac{2\sin\alpha}{\alpha}\times\frac{\alpha\cos\alpha - \sin\alpha}{\alpha^2} = 0$$

$$\sin\alpha = 0 \text{ or } \alpha\cos\alpha - \sin\alpha = 0$$

But the condition for minima is $\sin\alpha = 0$

So, the condition for the maxima is $\alpha\cos\alpha - \sin\alpha = 0$

$$\Rightarrow \alpha\cos\alpha = \sin\alpha$$

$$\alpha = \tan\alpha \quad ..... (6)$$

This equation is called transcendental equation. If we draw the graph between $y = \alpha$ and $y = \tan\alpha$ then the points of intersection of these two curves gives the maximum intensity.

$\Rightarrow$ BC=AB $\sin\theta$

BC= a $\sin\theta$

The phase difference corresponding to this path difference is given as

$$\delta = \frac{2\pi}{\lambda} \times path\ difference$$

$$\delta = \frac{2\pi}{\lambda} \times a\sin\theta \ \ldots\ldots (1)$$

Let us consider that the width of the slit is divided into 'n' equal parts and the amplitude of the wave from each part is $A^{'}$.

The phase difference between any two consecutive waves from these parts would be $1/n$ [total phase]

$$\frac{\delta}{n} = \frac{\frac{2\pi}{\lambda} a\sin\theta}{n} = \text{d say} \ \ldots.. (2)$$

Using the method of vector addition of amplitudes, the resultant amplitudes R is given by

$$R = \frac{A^{'}\sin(\frac{nd}{2})}{\sin(\frac{d}{2})}$$

$$R = \frac{A^{'}\sin(\frac{\pi a\sin\theta}{\lambda})}{\sin(\frac{\pi a\sin\theta}{\lambda n})}$$

$$R = \frac{A^{'}\sin\alpha}{\sin(\frac{\alpha}{n})} \ \ldots\ldots (3) \qquad \text{Where } \alpha = \frac{\pi a\sin\theta}{\lambda}$$

When n is very large the α/n is small, $\sin(\frac{\alpha}{n}) = \frac{\alpha}{n}$

$$R = \frac{A^{'}\sin\alpha}{\frac{\alpha}{n}}$$

$$R = \frac{nA^{'}\sin\alpha}{\alpha}$$

$$R = \frac{A\sin\alpha}{\alpha} \ \ldots.. (4) \qquad \text{Where } (nA'=A)$$

The resultant amplitude $R = A\left(\frac{\sin\alpha}{\alpha}\right)$

The intensity at P is

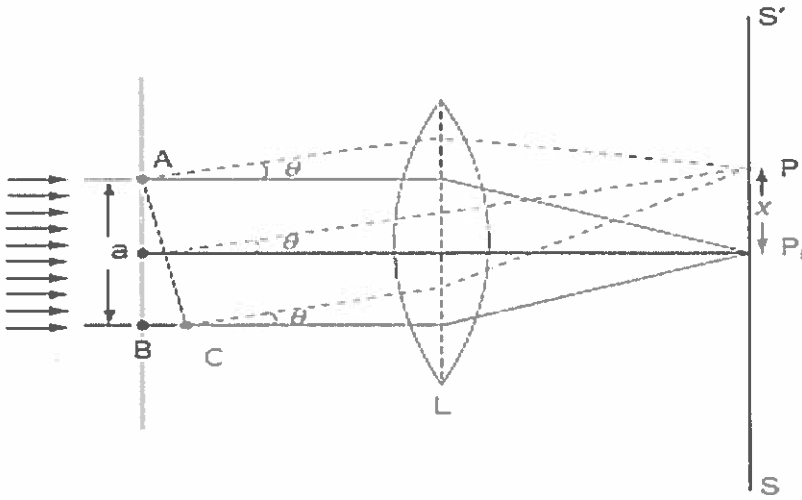
$\lambda = 5900\ \mathring{A}$ $\qquad D_5 = 0.2\,cm$ $\qquad D_{10} = 0.5\,cm$

$$R = \frac{D_m^2 - D_n^2}{4\lambda(m-n)} = \frac{D_{10}^2 - D_5^2}{4\lambda(10-5)}$$

$$= \frac{(0.5)^2 - (0.2)^2}{4 \times 5900 \times 10^{-8} \times 5} = 178\ cm$$

6(b)

A. 7(a)



Let AB represents a slit having width 'a' when the plane wave front of monochromatic wavelength 'λ' strike the slit, diffraction of slit occurs.

The diffracted waves from different parts of the slit traveling normally to the slit will converge at the point P_0, where maximum intensity is observed.

The diffracted waves from AB inclined at an angle θ from the direction PP_0. Let these waves are focused at point 'P' on the screen. The P is of minimum or subsidiary maximum intensity depending upon the path difference between the secondary waves originating from the corresponding points of the wave front.

To find out the intensity at P draw a perpendicular AC. The path difference between secondary wavelets from A and B in direction θ is equal to BC.

We know from $\Delta^{le}\ ABC$

$$\sin\theta = \frac{BC}{AB}$$

$$r^2 = 2Rt - t^2$$

$$r^2 \approx 2Rt$$ (Since t is small t^2 is very small)

$$r^2 = 2Rt$$

$$t = \frac{r^2}{2R} \quad \ldots\ldots (5)$$

For bright rings, $2 \times \frac{r^2}{2R} = (2n-1)\frac{\lambda}{2}$

$$\frac{r^2}{R} = (2n-1)\frac{\lambda}{2}$$

$$r^2 = \frac{(2n-1)\lambda R}{2}$$

If D is the diameter of the ring, $r = \frac{D}{2}$

$$\frac{D^2}{4} = \frac{(2n-1)\lambda R}{2}$$

$$D^2 = 2\lambda R(2n-1)$$

$$D = \sqrt{2\lambda R(2n-1)}$$

$$D_n \propto \sqrt{(2n-1)}$$

Therefore the diameter of the bright ring is proportional to the square root of the odd natural numbers.

For dark rings, $2 \times \frac{r^2}{2R} = n\lambda$

$$\frac{r^2}{R} = n\lambda$$

$$r^2 = n\lambda R$$

If D is the diameter of the ring, $r = \frac{D}{2}$

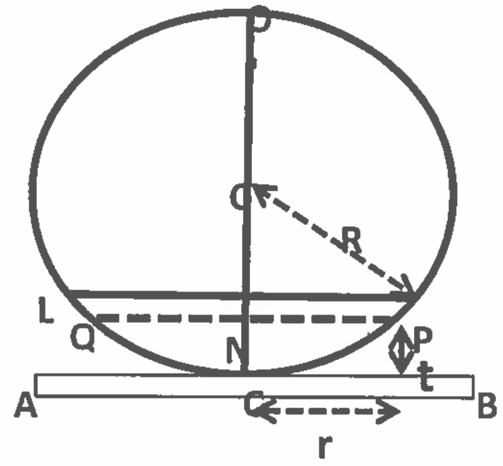
$$\frac{D^2}{4} = n\lambda R$$

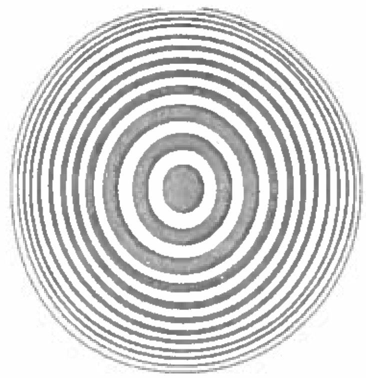
$$D^2 = 4n\lambda R$$

$$D = 2\sqrt{n\lambda R}$$

$$D_n \propto \sqrt{n}$$

Therefore, the diameter of the dark ring is proportional to the square root of natural numbers.





Let LOL' be the lens placed on a glass plate AB.

Let R be the radius of curvature of lens and r be the radius of Newton's ring corresponding to the constant film thickness t.

The rings are observed in the reflected light, an additional path $\lambda/2$ is introduced.

The effective path difference between the rays

$$\delta = 2\mu t\cos r + \frac{\lambda}{2} \quad \ldots\ldots (1)$$

For air film $\mu=1$ and for normal incidence r = 0

$$\delta = 2t + \frac{\lambda}{2} \quad \ldots\ldots (2)$$

At the point of contact $t = 0$, $\delta = \frac{\lambda}{2}$, this is the condition for minimum intensity. Hence the central spot is dark.

The condition for bright ring is

$$\delta = 2t + \frac{\lambda}{2} = n\lambda$$

$$\Rightarrow 2t = (2n-1)\frac{\lambda}{2} \quad \ldots\ldots. (3) \qquad \text{Where n=0, 1, 2, 3 ....}$$

The condition for dark ring is

$$\delta = 2t + \frac{\lambda}{2} = (2n+1)\frac{\lambda}{2}$$

$$\Rightarrow 2t = n\lambda \quad \ldots\ldots. (4) \qquad \text{Where n=0, 1, 2, 3 ....}$$

Let us consider the curved surface of the lens as an arc of a circle whose center is at C.

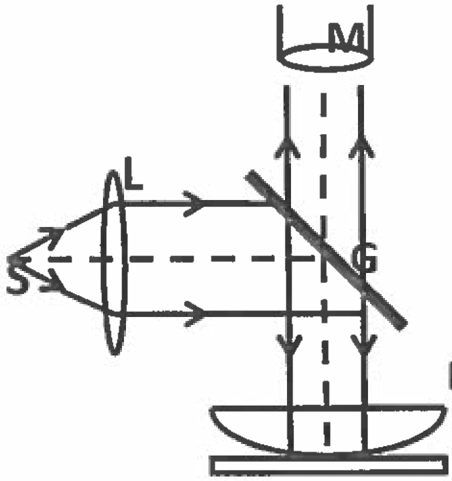
$$NP \times NQ = NO \times NO'$$

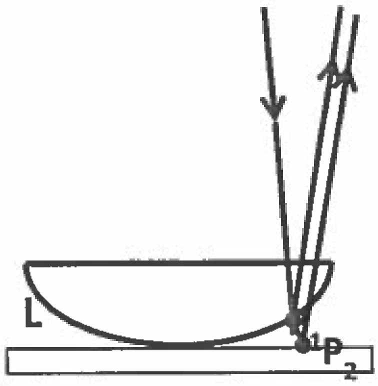
$$r \times r = t \times (2R - t)$$

around the point of contact. This phenomenon was first observed by Newton, the rings are called Newton's rings.

***Experimental arrangements*:**

A plano-convex lens L of large radius of curvature and is placed on a plane glass plate. The light from monochromatic source is incident on a glass plate, which is placed at an angle of 45° with vertical. The glass plate reflects normally a part of incident light towards the air film enclosed by the lens L and the glass plate P. A part of the incident light is reflected by the curved surface of the lens L and remaining is transmitted which is reflected back from the plane surface of the glass plate P. These two reflected rays (P_1 and P_2) are interfering and produce an interference pattern in the form of bright and dark circular rings. These rings can be viewed in a microscope M focused on the film.





Theory:

(1)

All the bright fringes have the same intensity.	The intensity of bright fringes usually decreases with increase of order.
All the dark fringes have zero intensity	The intensity of dark fringes is not zero.

(2)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

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 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

(3)**Electric susceptibility:** The polarization vector P is proportional to the applied electric field.

$$P \propto E$$

$$P = \chi E$$

Where χ is constant is called electric susceptibility

(4) Sabine's formula is for Reverberation time and is given by

T*=0.167*V*/Σ*as

Where V=volume of the hall

a= absorption coefficient

s=surface area

(5) A unit cell is the smallest portion of a crystal lattice that shows the three-dimensional pattern of the entire crystal. A crystal can be thought of as the same unit cell repeated over and over in three dimensions.

A. 6(a) When a plano-convex lens of long focal length with its convex surface is placed on a plane glass plate. At the point of contact where the lens touches the glass plate the thickness of the air film is zero and when moved gradually towards the edge of the lens, the thickness of the air film is increases. If a monochromatic light is allowed to fall normally and the film is viewed in reflected light, alternate dark and bright concentric circular rings are observed

Intensity expression $I = I_0 \left(\frac{\ldots}{\alpha}\right)$ [illegible]
(Qualitative) 2M

Condition for principal maxima 2M
Secondary maxima 2M

Intensity distribution curve

(8)(a) Introduction & Principle of He-Ne LASER 2M
Construction 4M
Working 4M
(b) Concept of Population Inversion 2M

(OR)

(9)(a) Expression for Acceptance angle $\sin\theta_{max} = \frac{\sqrt{n_1^2 - n_2^2}}{n_0}$ 5M

Expression for Numerical Aperture $NA = n_1\sqrt{2\Delta}$ 5M

(b) Acceptance angle $\theta_A = \theta_{max} = 30^\circ$ 1M

$NA = \sin\theta_A = \sin 30^\circ = 0.5$ 1M

(10)(a) Basic definition of ferromagnetism 2M
Hysteresis curve with description 4M
Concept of Retentivity & coercivity 4M
(b) Magnetic dipole definition 1M
Permeability definition 1M

(OR)

(11)(a) Diagram with description 2M
Each component of Internal field 4x2=8M
$(E_0, E_1, E_2 \,\&\, E_3)$
No derivation needed, only explanation & expression

(b) Any two applications of dielectrics (each 1M) 2x1=2M

(12) Various factors affecting the acoustics of a Hall (Minimum of 8–10 points) 12 M

(OR)

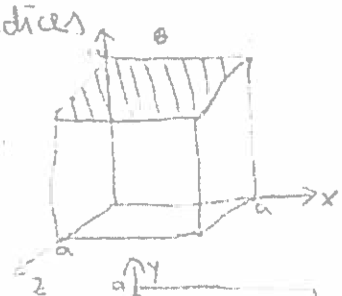
(13)(a) Principle & circuit construction of Piezo electric method 4 M

working 4 M.

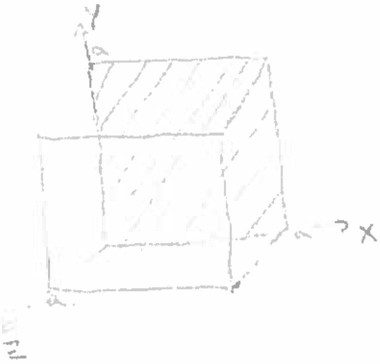
(b) Brief note on NDT applications 4 M.

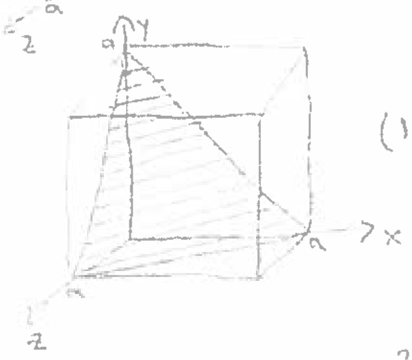
(14)(a) Definition of Miller indices 2 M.

(b) planes (0 1 0) $3 \times 2 = 6$ M



(1 1 0)





(1 1 1)

(b) concept & diagram 2 M.
(OR) Inter planar spacing 2 M

(15)(a) Bragg's X-ray diffraction Path difference & diagram 2 M

Bragg's condition 2 M.

(b) Introduction & diagram 2 M

Explanation 2 M

Condition ($R = D\tan 2\theta$) 2 M

Limitation of Laue Method 2 M.

HOD, S&H

3/09/21

NSRIT
(Autonomous)

20BSX31 : Engineering Physics. Sep 2021

Scheme of valuation

Part – A. $5 \times 2 = 10$ marks.

(1) Any two differences between interference and diffraction (each 1M) $2 \times 1 = 2M$

(2) Basic definitions of Einstein coefficients 2M

(3) Basic definition of Electric Susceptibility, $\chi = \frac{P}{E}$. 2M

(4) Sabine's formula $T = \frac{0.165\,V}{\Sigma as}$ 2M.

(5) Basic definition of unit cell. 2M.

Part – B.

(6)(a) Basic concept of Newton's rings 2M

Expression for diameter of dark ring ($D^2 = 4n\lambda R$) 4M

Expression for diameter of bright ring $D^2 = 2\lambda R(2n-1)$ 4M

(b) $\lambda = 5900\,Å$ $D_5 = 0.2\,cm$ $D_{10} = 0.5\,cm$ 1M

$$R = \frac{D_m^2 - D_n^2}{4\lambda(m-n)} = \frac{D_{10}^2 - D_5^2}{4\lambda(10-5)}$$ 1M

$$= \frac{(0.5)^2 - (0.2)^2}{4 \times 5900 \times 10^{-8} \times 5} = 178\ cm$$

(OR)

(7)(a) Single slit diagram with description (Path difference) 4M

NSRIT

Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	Program	CE/ME			Academic Year	2020 - 2021
Course Code	20BSX31	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	Engineering Physics						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Distinguish between interference and diffraction	20BSX31.1	L2
2	What are Einstein's coefficients	20BSX31.2	L1
3	Define electric susceptibility	20BSX31.3	L1
4	Write the Sabine's formula for reverberation time	20BSX31.4	L1
5	Define Unit Cell	20BSX31.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	How Newton's rings are formed? Obtain the expressions for diameters dark rings and bright rings	10M	20BSX31.1	L2
6 (b)	Newton's rings are observed in the reflected light of wavelength 5900A°. The diameter of fifth ring and tenth dark ring is 0.2 cm and 0.5 cm. Find the radius of curvature of the lens used	2M	20BSX31.1	L2
	OR			
7 (a)	Give the theory of Fraunhofer diffraction due to a single slit and hence obtain the condition for primary and secondary maxima. Using this obtain intensity distribution curve	12M	20BSX31.1	L2
8 (a)	Explain the principle, construction and working of a He-Ne laser	10M	20BSX31.2	L2
8 (b)	Explain how population inversion is obtained	2M	20BSX31.2	L2
	OR			
9 (a)	Derive expression for acceptance angle and numerical aperture of optical fiber	10M	20BSX31.2	L2
9 (b)	Acceptance angle of an optical fiber is 30°. Calculate Numerical Aperture	2M	20BSX31.2	L2
10 (a)	What is ferromagnetism? Explain hysteresis curve	10M	20BSX31.3	L2
10 (b)	Define magnetic dipole and permeability	2M	20BSX31.3	L2
	OR			
11 (a)	Write a note on internal field in dielectrics	10M	20BSX31.3	L2
11 (b)	Write any two important applications of dielectric materials	2M	20BSX31.3	L2
12	What are the factors affecting the acoustics of a hall	12M	20BSX31.4	L2
	OR			
13 (a)	Explain piezoelectric method to produce ultrasonics	8M	20BSX31.4	L2
13 (b)	Write a brief note on applications of NDT	4M	20BSX31.4	L2
14 (a)	Define Miller indices. Sketch planes in simple cubic structure (010), (110) and (111)	8M	20BSX31.5	L2
14(b)	Derive an expression for the inter planar spacing in case of cubic structure	4M	20BSX31.5	L2
	OR			
15 (a)	State and Explain Braggs law of X-ray diffraction	4M	20BSX31.5	L2
15 (b)	Describe in detail Laue method to determine crystal structure	8M	20BSX31.5	L2

Semiconductor in its pure form is known as intrinsic semiconductor.

Extrinsic semiconductors results when intrinsic semiconductors are doped with appropriate impurities.

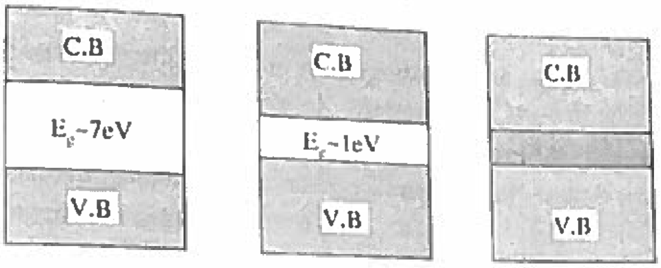
15(b) Sign of charge carriers can be determined.

Mobility of charge carriers can be determined.

Magnetic field can be measured.

Carrier density can be estimated.

of a solid. The electrons in the innermost shells, which are completely filled, do not take any part in the conduction process. The completely filled bands and the completely empty bands do not contribute to the electrical conduction. The valence band and the conduction band energies are important for the electrical properties of a solid.



(a) Insulators (b) Semiconductors (c) Conductors

Fig. 8.6 Energy band diagram of insulators, semiconductors and conductors

Insulators :

The energy band structure of an insulator is shown in Fig. 8.6(a). In insulator, the conduction band is completely empty, the valence band is completely filled and there is a large energy gap, $E_g > 2$ eV, 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 ($k_B T$) is much less than the band gap energy. The diamond is a perfect insulator having a band gap of 5.5 eV.

Semiconductors :

The energy band diagram of semiconductor is shown in Fig. 8.6(b). 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 gap energies 1.1 eV and 0.7 eV respectively. At absolute zero temperature, all the semiconductors act as insulators. The conductivity of semiconductor increases with increase in temperature. So the semiconductors have a negative temperature coefficient of resistance.

Conductors :

The energy band diagram of a conductor is shown in Fig. 8.6(c). In conductors, the valence band and the 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.

15(a)

8.5 Distinction between Conductors, Semiconductors and Insulators

The electrical properties of a solid depends upon its energy band structure and the way in which the energy bands are occupied by the electrons. Depending on the nature of band occupation by electrons and on the width of the forbidden band, the solids can be classified as insulators, semiconductors and conductors. The metals are good conductors of electricity while the insulators are bad conductors of electricity. The electrical conductivity of semiconductor lies between that of a metal and insulator. The energy band theory of solids can explain the electrical conductivity

$$\frac{\sin \alpha a}{\alpha a}+\frac{\cos \alpha a}{P}=\frac{\cos k(a+b)}{P}$$

$$\frac{\sin \alpha a}{\alpha a}+0=0$$

$$\sin \alpha a=0$$

$$\alpha a=n\pi$$

$$\alpha=\frac{n\pi}{a}$$

$$\alpha^2=\frac{n^2\pi^2}{a^2}$$

$$\frac{2mE}{\hbar^2}=\frac{n^2\pi^2}{a^2} \quad \text{(From equation (3))}$$

$$E=\frac{n^2\pi^2\hbar^2}{2ma^2} \quad \text{or} \quad E_n=\frac{n^2\pi^2\hbar^2}{2ma^2}$$

Electron is trapped in potential.

4. When *P=0, b=0* then

$$\cos \alpha a=\cos ka$$

$$\alpha a=ka$$

$$\alpha=k$$

$$\alpha^2=k^2$$

$$\frac{2mE}{\hbar^2}=k^2$$

$$E=\frac{\hbar^2k^2}{2m}$$

$$E=\frac{P^2}{2m} \quad (\because p=\hbar k \text{, momentum of an electron})$$

So the electron is free.

$$P = \frac{mV_0\, ba}{\text{h}^2}$$

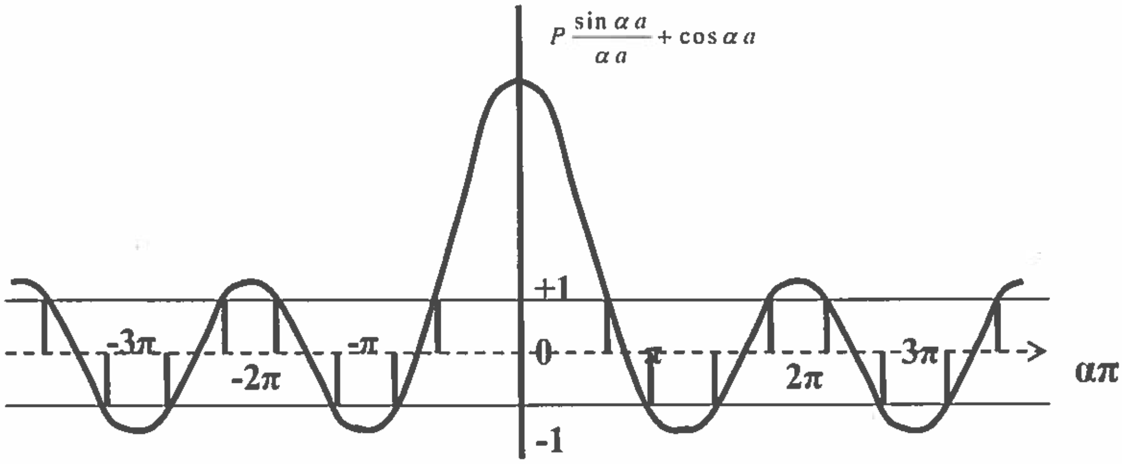
$$P = \frac{2mV_0\, ba}{2\text{h}^2}$$

$$P = \frac{\beta^2\, ba}{2} \qquad (\text{Q}\ \beta^2 = \frac{2mV_0}{\text{h}^2})$$

The physical significance of this quantity is that if P is increases the area of the potential barrier is increased and given electron is bound more strongly to a particular potential well.

From equation (3), $\alpha^2 = \frac{2mE}{\text{h}^2} \qquad \Rightarrow E = \frac{\alpha^2 \text{h}^2}{2m}$

$$\Rightarrow E = \frac{\alpha^2 h^2}{8\pi^2 m} \qquad (\text{Q}\ \text{h} = \frac{h}{2\pi})$$

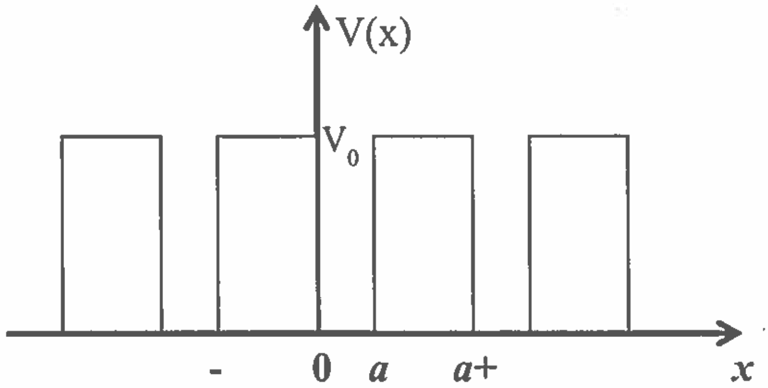


From the figure, the following conclusions are drawn,

1. The energy spectrum of the electrons consists of a large number of allowed energy bands separated by forbidden bands.
2. The width of allowed energy bands increases with increase of energy values (i.e., with increasing the values of αa) and forbidden energy regions become narrower.
3. The width of allowed band decreases with the increasing value of P i.e., with increasing binding energy of electrons.

As P→∞ then $\frac{1}{P} \rightarrow 0$

$$P\frac{\sin \alpha a}{\alpha a} + \cos \alpha a = \cos k(a+b)$$



In the regions where $0 < x < a$, the potential energy is assumed to be *zero* and in the region $-b < x < 0$, the potential is V_0.i.e., the potential rectangular potential well of period $(a+b)$ suggested by *Kronig-Penny*.

$V(x) = 0$, for the region $0 < x < a$

$V(x) = V_0$, for the region $-b < x < 0$

The Schrodinger time independent wave equation for the two regions can be written as

$$\frac{d^2\psi}{dx^2} + \frac{2m}{h^2} E\psi = 0 \quad \text{for } 0 < x < a \quad \ldots\ldots. (1)$$

$$\frac{d^2\psi}{dx^2} + \frac{2m}{h^2}(E - V_0)\psi = 0 \quad \text{for } -b < x < 0 \quad \ldots\ldots. (2)$$

Assuming the energy E of the electron is less than V_0, we define two real quantities αand β as

$$\alpha^2 = \frac{2mE}{h^2} \text{ and } \beta^2 = \frac{2m(V_0 - E)}{h^2} \quad \ldots\ldots (3)$$

Where α and β are real quantities.

Therefore the equations (1) and (2) becomes

$$\frac{d^2\psi}{dx^2} + \alpha^2\psi = 0 \quad \text{for } 0 < x < a \quad \ldots\ldots (4)$$

$$\frac{d^2\psi}{dx^2} - \beta^2\psi = 0 \quad \text{for } -b < x < 0 \quad \ldots.. (5)$$

By applying boundary conditions the value of these constants are evaluated.

$$P\frac{\sin\alpha a}{\alpha a} + \cos\alpha a = \cos ka$$

where *P is calledscattering power of the potential barrier* and is given by

14(a)

8.4 Origin of Energy Band Formation in Solids

Solids are usually strong and slightly elastic structures. The individual atoms are held together in solids by interatomic forces or bonds. The bonding is strongly dependent on the electronic structure of the atoms concerned. The attraction between the atoms brings them closer until the individual electron clouds begin to overlap. A strong repulsive force arises according with Pauli's exclusion principle. When the attraction force and the repulsive force between any two atoms are equal, the two atoms occupy a stable position with a minimum potential energy. The spacing between the atoms under this condition is called equilibrium spacing. In solids many atoms are brought together so that the split energy levels form a set of bands of very closely spaced levels with forbidden energy gaps between them as illustrated in the Fig. 8.5.

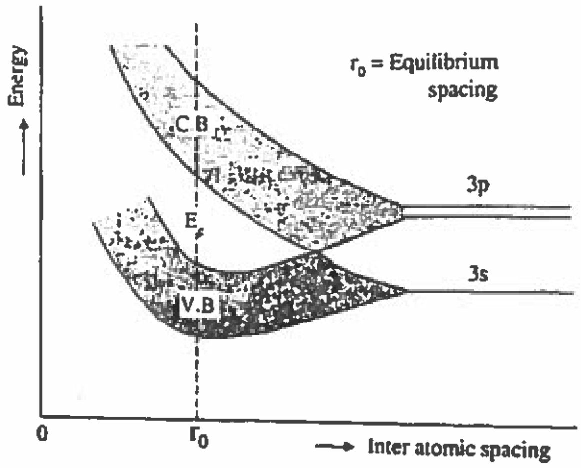
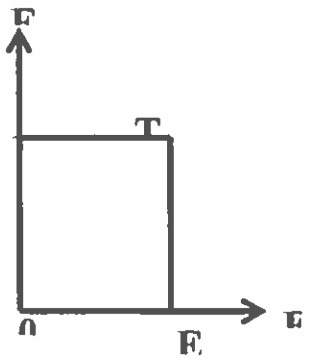


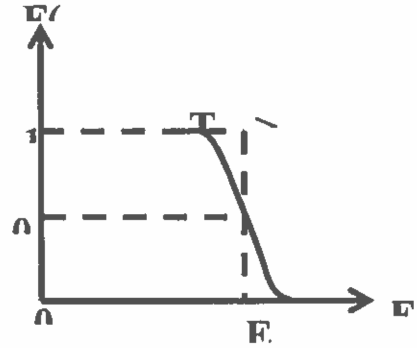
Fig. 8.5 Spreading of energy levels into energy bands in sodium

The electrons first occupy the lower energy band and are of no importance in determining many of the electrical properties of solids. Instead, the electrons in the higher energy bands of solids are important in determining many of the physical properties of solids. Hence we are interested in those two allowed energy bands called valence and conduction bands. The gap between these two allowed bands is called forbidden energy gap or band gap.

14(b) The essential feature of the behavior of an electron studied by considering a periodic potential well in one dimensional which was first discussed by *Kronig-Penny* in 1931. It is assumed that the potential energy of electron, when it moves in one dimensional perfect crystal lattice, it is represented in the form of rectangular wells and barriers of width '*b*'. The periodicity of the potentials energy is (*a*+*b*).

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.

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.

1. This theory could not explain the differences between the conductors, semiconductors and insulators.
2. It also fails to explain the positive value of Hall co-efficient.
3. Quantum free electron theory always predicts a spherical Fermi surface is often non spherical.
4. On the basis of quantum free electron theory, it has been show that the electrical conductivity is proportional to the electron concentration. It is surprising that the divalent metals (like Be, Cd, Zn) even trivalent metals (like Al, In) are consistently less conductive than monovalent metals (like Cu, Ag, Au) despite the fact that former have higher concentration of electrons.

13(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

$$E = \frac{1}{2}mv^2$$

$$E = \frac{1}{2}\frac{m^2v^2}{m}$$

$$E = \frac{P^2}{2m}$$

$$\Rightarrow P = \sqrt{2mE}$$

$$\therefore \lambda = \frac{h}{\sqrt{2mE}}$$

Characteristics of matter waves:

1. Lighter the particle, greater the wavelength associated with it.
2. Smaller the velocity of the particle, longer the wavelength associated with it.
3. When $\upsilon = 0$ then $\lambda = \infty$ and $\upsilon = \infty$ then $\lambda = 0$. This means that only with moving particle matter wave is associated.
4. Whether the particle is charged or not, matter wave is associated with it. This reveals that these waves are not electromagnetic but a new kind of waves.
5. The velocity of matter waves is greater than the velocity of light.
6. No single phenomena exhibit both particle nature and wave nature simultaneously.

12(b)

$\lambda = 0.21\,nm$ $\quad\quad v = ?$ $\quad\quad$ 1 M

$$\lambda = \frac{h}{mv} = \frac{6.625 \times 10^{-34}}{9.1 \times 10^{-31} \times v} = 0.21$$

$$v = \frac{6.625 \times 10^{-34}}{9.1 \times 10^{-31} \times 0.21 \times 10^{-9}} = 0.5232 \times 10^{6}\ m/s$$

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.

$\varepsilon_r = \dfrac{\varepsilon}{\varepsilon_0}$ Where ε_0 permittivity of free space

Electric diploe: A system or arrangement of two equal and opposite charges is separated by a small distance is called electric dipole.

12(a) According to de-Broglie's hypothesis (1924), a moving particle behaves as a wave and as a particle. *The moving particle is associated with a wave which is known as de-Broglie's wave or matter wave.* They are seen with particles like electrons, protons, neutrons etc. The wavelength of the matter wave is given by

$$\lambda = \frac{h}{mv} = \frac{h}{p}$$

where m is the mass of the material particle, v is the velocity and p is the momentum of the particle.

Consider the Planck's theory of radiation, the energy of a photon is given by

$$E = h\upsilon = \frac{hc}{\lambda} \quad \ldots\ldots (1)$$

where c is the velocity of light in vacuum and λ is its wavelength.

According to Einstein energy mass relation

$$E = mc^2 \quad \ldots\ldots (2)$$

From equation (1) and (2)

$$mc^2 = \frac{hc}{\lambda}$$

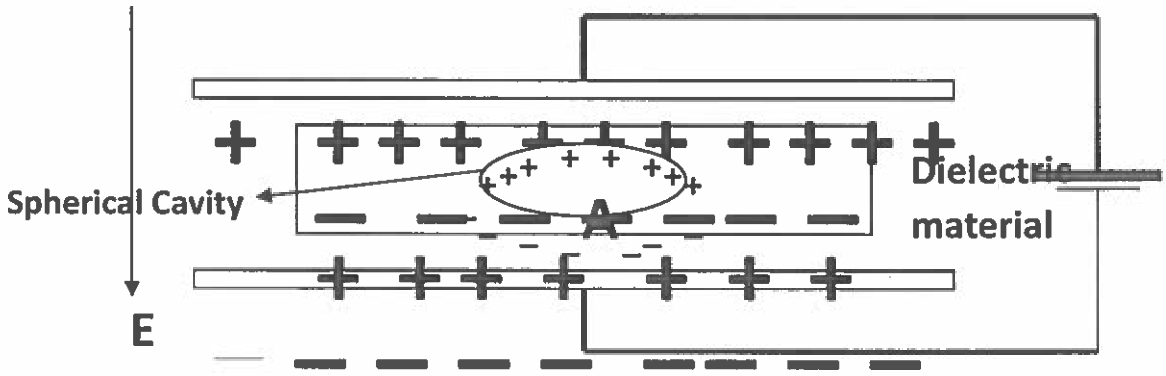
$$\Rightarrow \lambda = \frac{hc}{mc^2}$$

$$\Rightarrow \lambda = \frac{h}{mc}$$

$$\Rightarrow \lambda = \frac{h}{p} \qquad \text{(where } mc = p \text{ momentum)}$$

This is known as *de-Broglie's* equation.

If E is the kinetic energy of the material particle, then



Field E_2:

E_2 is the field intensity at A due to the charge density induced on the two sides of the dielectric.

$$E_2 = \frac{-P}{\varepsilon_0} \quad .... (2)$$

Field E_3:

E_3 is the field intensity at A due to the other atoms contained in the cavity, we are assuming a cubic structure, so $E_3 = 0$.

Field E_4:

E_4 is the field intensity due to polarizing charges on the surface of the spherical cavity and calculated by Lorentz.

$$E_4 = \frac{P}{3\varepsilon_0}$$

The resultant internal field or Lorentz field can be written as

$$E_i = E_1 + E_2 + E_3 + E_4$$

$$E_i = (E + \frac{P}{\varepsilon_o}) - \frac{P}{\varepsilon_o} + 0 + \frac{P}{3\varepsilon_o}$$

$$E_i = E + \frac{P}{3\varepsilon_o}$$

This is the expression for internal field of a solid. This is also called Lorentz field.

11(b) The dielectric constant is the ratio between the permittivity of the medium to the permittivity of free space.

v. The ferrimagnetic materials are also exhibit hysteresis property similar to ferromagnetic materials. The hysteresis curve of ferrites is normally has a square shape.

Eg: Fe_2O_4, $NiFe_2O_4$, $PbFe_{12}O_{19}$, $BaFe_{12}O_{19}$ etc

10(b)

$\chi_p = 3.7 \times 10^{-3}$, $\mu_r = ?$

$\mu_r = 1 + \chi_p = 1 + 3.7 \times 10^{-3} = 1.0037$

A. 11(a) Local field or internal field in a dielectric is the space and time average of the electric field intensity acting on a particular molecule in the dielectric material.

Consider a dielectric be placed between the plates of a parallel plate capacitor and let there be an imaginary spherical cavity around the atom A inside the dielectric.

It is also assumed that the radius of the cavity is large compared to the radius of the atom.

The internal field at the atom site 'A' can be made up of four components E_1, E_2, E_3 and E_4.

Field E_1:

E_1 is the field intensity at A due to the charge density on the plates, from the field theory,

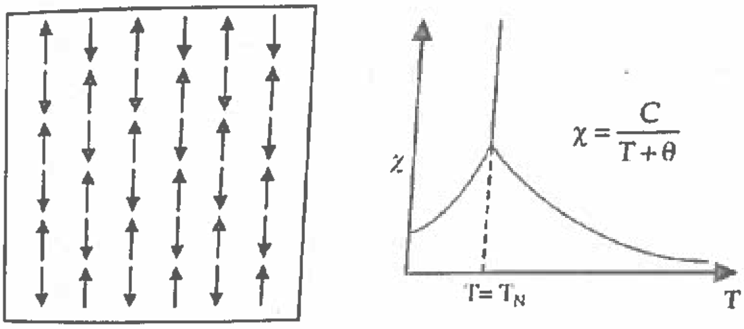
$$E_1 = \frac{D}{\varepsilon_0} \text{ and } D = \varepsilon_0 E + P$$

$$E_1 = \frac{\varepsilon_0 E + P}{\varepsilon_0}$$

$$E_1 = E + \frac{P}{\varepsilon_0} \quad \text{........ (1)}$$

iii. The variation of susceptibility with temperature is given by the relation, $\chi_{af} = \frac{C}{T+T_N}$,

$T > T_N$, where C is Curie constant and T_N is Neel temperature.



iv. They attain maximum susceptibility at Neel temperature, T_N. above T_N these materials become paramagnetic.

v. Anti-ferromagnetic materials show very little external magnetism.

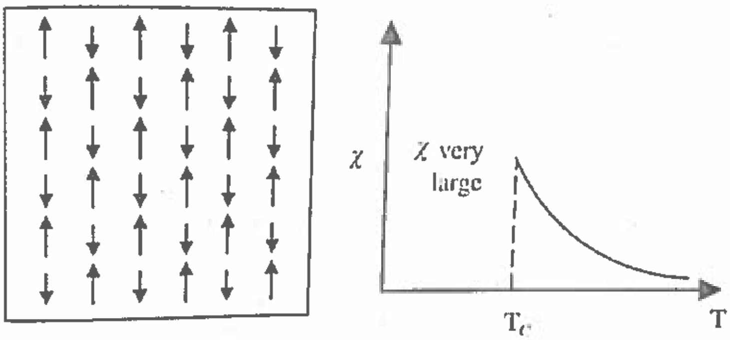
Eg: MnO, NiO, MnS, MnTe, CoO, $MnCl_2$, $FeCl_2$ etc.

5. **Properties of Ferrimagnetic materials (Ferrites):**

i. In these materials the atomic dipoles are arranged antiparallel to one another but the moments in one direction have a larger magnitude so that the net magnetization exists.

ii. The magnetic susceptibility is large and positive.

iii. They also show Curie-Weiss behavior. The susceptibility varies with temperature is given by the relation, $\chi_{ferri} = \frac{C}{T \pm T_N}$ where C is the Curie constant and T_N is Neel temperature.



iv. The ferrimagnetic materials behave like ferromagnetic materials below the Neel temperature and are paramagnetic above Neel temperature.

ii. These materials possess permanent magnetic moments even when applied field is zero i.e., they possess spontaneous magnetization.

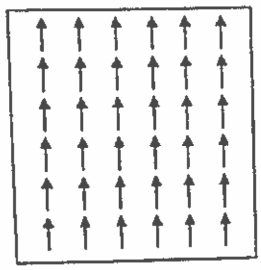
iii. The magnetic susceptibility and relative permeability are positive and exhibit very high values.

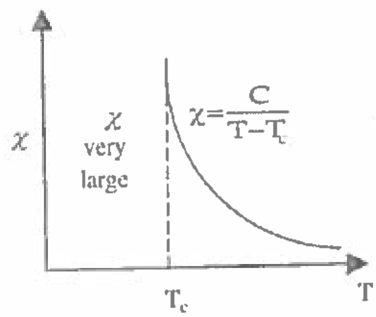
iv. These materials having permanent magnetic dipoles are orderly oriented.

v. Because of nonlinear relationship between B and H, the permeability of ferromagnetic material does not have a constant value.

vi. These materials possess all the properties of paramagnetic materials with much greater intensity.

vii. Above a certain temperature, ferromagnetic materials behaves paramagnetic and the susceptibility varies with temperature





$$\chi_F = \frac{C}{T-T_c} \qquad T > T_c$$

Where C is Curie constant and T_c is curie constant. This relation is called Curie-Weiss law. The Curie temperature is depends on the material.

Eg: Fe, Ni, Co, Gd, Fe_2O_3, $ZnFe_2O_3$, $MnFe_2O_3$ etc.

4. **Properties of Anti-ferromagnetic materials:**

i. These materials, the atomic dipoles are arranged antiparallel to one another net magnetic moment is zero.

ii. These are crystalline materials which exhibit small positive susceptibilities of the order of 10^{-3} to 10^{-5}.

Eg: metals (Cu, Au, Bi, Sb, Hg), semiconductors (Si, Ge), rare gas elements (He, Ne, Ar), benzene, Naphthalene, Nacl, air, water, H_2 etc.

2. **Properties of paramagnetic materials:**

i. The paramagnetic materials are feebly magnetized in the direction of the magnetizing field.

ii. When a paramagnetic rod is suspended freely in a uniform magnetic field, it aligns itself in the direction of magnetic field.

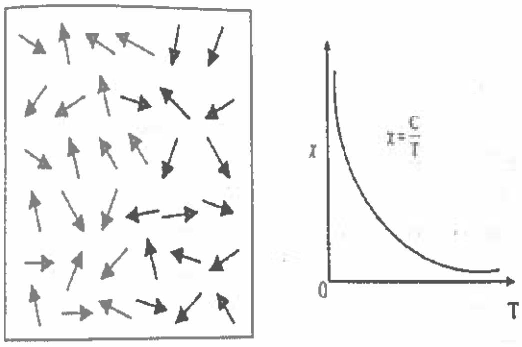
iii. The magnetic susceptibility is small and positive, is of the order of 10^{-3}

iv. In a non-uniform magnetic field, the paramagnetic substances are attracted towards the stronger parts of the magnetic field from the weaker part of the field.

v. As soon as the magnetizing field is removed, the paramagnetic materials lose their magnetization.

vi. The paramagnetic susceptibility varies inversely with temperature,

$$\chi_p = \frac{C}{T}$$



Where C is the Curie constant, this relation is called curies law.

Eg: metals (Al, Ca, Ti, Pt, Cr, Mn), salts of the transition elements, rare earths and actinide series containing elements (Cr^{3+}, Dy^{3+}, U^{4+}), compounds ($FeCl_2$, $CuCl_2$, $MnCl_2$, MnO, NiO, CoO) etc.

3. **Properties of ferromagnetic materials**:

i. These materials get strongly magnetized in the direction of the field.

9(b)

Numerical aperture $= 0.02$

cladding refractive index $n_2 = 1.59$

core refractive index $n_1 = ?$

$NA = \sqrt{n_1^2 - n_2^2}$ (Assume $n_0 = 1$)

$n_1^2 - n_2^2 = (NA)^2$

$n_1^2 = (NA)^2 + n_2^2 = 0.0004 + 2.5281$

$n_1^2 = 2.5285 \Rightarrow n_1 = \sqrt{2.5285} = 1.590125$

A. 10(a) **Classification of Magnetic Materials:**

1. **Properties of Diamagnetic materials:**

i. The materials which are weakly magnetized in a direction opposite to that of the applied magnetic field are called diamagnetic materials.

ii. When a diamagnetic material placed in a non-uniform field, then it tends to move towards the weaker part from the stronger part of the field.

iii. A diamagnetic liquid in a U shaped tube is depressed, when subjected to a magnetic field.

iv. The lines of force do not prefer to pass through the specimen, since the ability of a material to permit the passage of magnetic lines of force through it is less.

v. There is no permanent dipole moment, so the magnetic effects are very small.

vi. The magnetic susceptibility is negative. It is independent of temperature and magnetic field strength.

vii. The relative permeability μ_r for diamagnetic substances is less than one.

where n_1 = refractive index at the center of the core

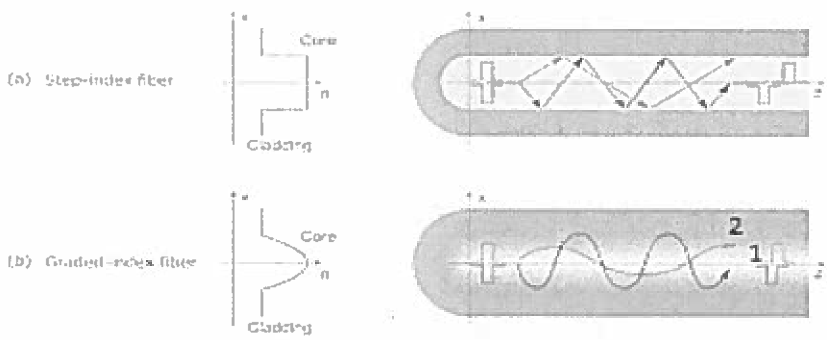
a= radius of the core

$\Delta = (n_1-n_2)/n_1$

P= grading profile index number.

Transmission of signal in graded index fibers:

Let us consider a signal pulse travelling through graded index fiber in two different paths 1 and 2. The pulse 1, travelling along the axis of the fiber in shorter route, it travels through the higher refractive index medium. The other ray 2, travelling away from the axis undergo and refraction and bend and covers longer distance in less refractive index and hence both the pulses reach the other end simultaneously. Hence the intermodal dispersion is overcome by using graded index fiber.

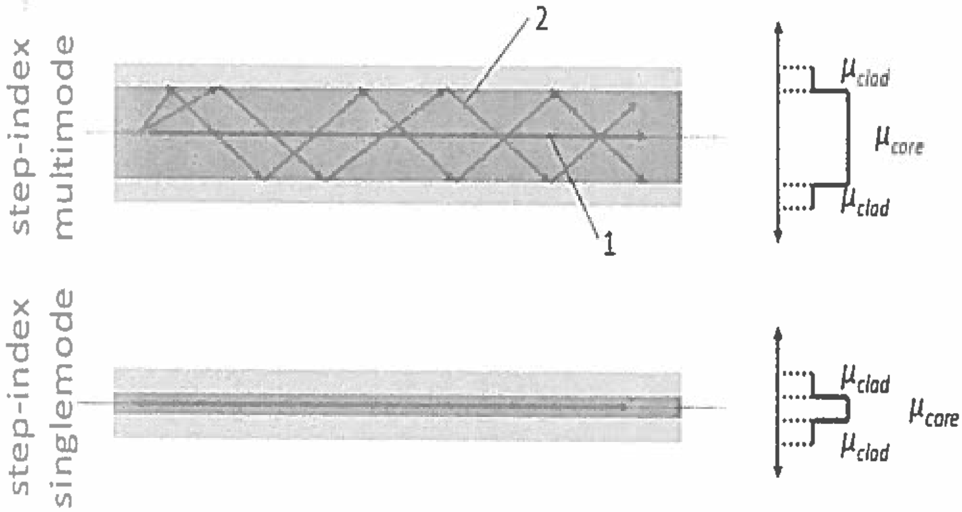


Polychromatic radiation Monochromatic radiation

9(a) **Step index fiber:** In this fiber the entire core has uniform refractive index n_1 slightly greater than the refractive index of the cladding n_2. Since the index profile is in the form of a step, these fibers are called **Step index fibers.**

Transmission of signal in step index fibers:

Generally, the signal is sent through the fiber in digital form i.e., in the form of pulses representing 0s and 1s. Let us now consider the propagation of one such pulse through multimode fiber. The same pulsed signal travels in different paths (represented by multimode). Hence at the receiving end only ray 1 travels along the fiber axis reaches first while the ray taking longer (zigzag) path 2 reach after time delay. Hence the pulsed signal received at the other end is broadened. This is called **intermodal dispersion.** This imposes limited on the separation between pulses thereby reducing the transmission rate and capacity. This difficulty is overcome by manufacturing of graded index fiber.



2.Graded index fiber:

In graded index multimode fiber, the refractive index of the core varies radially as shown in fig. It has maximum refractive index at its center, which gradually falls with increase of radius and at the core-cladding interface matches with the refractive index of the cladding. The variation of refracting index of the core (n) with radius (x), measured from the center of the core, is given by

$$n(x)= n_1 [1 -2 \Delta (x/a)^p]^2$$

7(b)

$N = \frac{15000}{2.54} = 5905$ lines/cm ; $\lambda = 600\,nm = 600 \times 10^{-9}\,m = 6000\,Å$

$n \leq \frac{1}{N\lambda}$

$n \leq \frac{1}{5905 \times 6000 \times 10^{-8}}$

$n_{max} = \underline{3}$

8(a) 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.

Communications: Optical fiber systems, CD/DVD/USB/HDD writing and reading.

Medicine: Blood less surgery, endoscopic studies.

Holography: Generation (recording) and reconstruction of holograms.

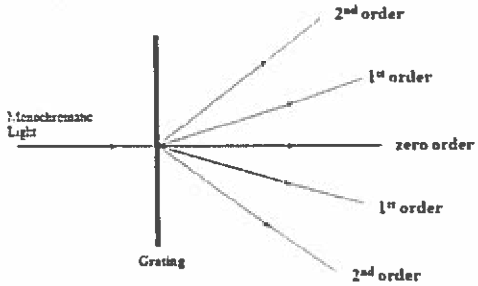
Commercial: Bar code readers, Printing

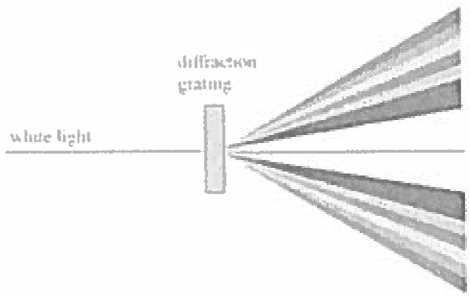
8(b)

This theory was postulated by Bohr	This theory was postulated by Einstein
Incoherent radiation	Coherent radiation

The second order maxima obtained for $n=2$ then $(a+d)\sin\theta_2 = 2\lambda$ and so on.

If we are using while light then the central maximum will also be white. However, for $n\neq 0$, in which order different colors are diffracted at different angles. The angles of diffraction are different for different wavelengths and therefore various spectral components appear at different positions. By measuring the angles of diffraction for various colors, we can determine the values of wavelength.





***Determination of wavelength of light*:**

The position of the principal maxima in a grating

$(a+d)\sin\theta = n\lambda$ Where n=0, 1, 2...

$$\lambda = \frac{(a+d)\sin\theta}{n}$$

$$\lambda = \frac{\sin\theta}{Nn}$$

Where $N = \dfrac{1}{(a+d)}$ is the number of grating elements or lines per unit width of grating.

A. 7(b) One of the most important applications of diffraction is the diffraction grating. It consists of a very large number of obstacles of equal widths arranged parallel and at equal distances from one another. Usually the width of the obstacles is the same as the width of the slits.

An arrangement consists of large number of equidistant parallel slits on a plane glass plate is called as diffraction grating. The corresponding diffraction pattern is known as grating spectrum.

Diffraction gratings are used one of the two ways, either as reflection grating or as transmission grating. A reflection grating consists of a series of fine parallel grooves on a flat metallic surface. This grating was made by Fraunhofer in 1820. The transmission grating consists of series of parallel rulings made on flat glass plate.

A good quality of grating requires large number of slits about 15,000 per each. Another requirement for a good quality of grating is that the lines should be as equally spaced as possible consequently the pitch of screw must be constant. The distance between any two consecutive lines is 'd'. If width of each slit is 'a' then the combined width of a ruling ($a+d$) is called grating element.

If there are 15,000 lines per inch on the grating surface, the spacing between lines is

$$d = \frac{2.54}{15000}$$

$$d = 1.693\times10^{-4}\,cm$$

Grating Spectrum:

The positions of the principal maxima are given by

$$(a+d)\sin\theta_n = n\lambda \text{ where } n=0,\ 1,\ 2...$$

where ($a+d$) is the grating element, n is the order of maxima, λ is the wavelength of incident light

The relation is called as grating equation.

The angle of diffraction depends upon the wavelength λ. The corresponding spectrum is called grating spectrum.

When the number of lines on the grating are large, the maxima appears sharp and the bright lines parallel to the ruling of the grating and are termed as spectral lines. The principal maxima occurs at $\theta=0$ and is irrespective of the wavelength λ.

The first order maxima obtained for $n=1$ then $(a+d)\ \sin\theta_1 = \lambda$

$$\frac{r^2}{R}=(2n-1)\frac{\lambda}{2}$$

$$r^2=\frac{(2n-1)\lambda R}{2}$$

If D is the diameter of the ring, $r=\frac{D}{2}$

$$\frac{D^2}{4}=\frac{(2n-1)\lambda R}{2}$$

$$D^2=2\lambda R(2n-1)$$

$$D=\sqrt{2\lambda R(2n-1)}$$

$$D_n\propto\sqrt{(2n-1)}$$

Therefore the diameter of the bright ring is proportional to the square root of the odd natural numbers.

For dark rings, $2\times\frac{r^2}{2R}=n\lambda$

$$\frac{r^2}{R}=n\lambda$$

$$r^2=n\lambda R$$

If D is the diameter of the ring, $r=\frac{D}{2}$

$$\frac{D^2}{4}=n\lambda R$$

$$D^2=4n\lambda R$$

$$D=2\sqrt{n\lambda R}$$

$$D_n\propto\sqrt{n}$$

Therefore, the diameter of the dark ring is proportional to the square root of natural numbers.

A. 6(b) 1.The sources of the waves must be coherent, which means they emit identical waves with a constant phase difference.
2.The waves should be monochromatic - they should be of a single wavelength.

Let LOL' be the lens placed on a glass plate AB.

Let R be the radius of curvature of lens and r be the radius of Newton's ring corresponding to the constant film thickness t.

The rings are observed in the reflected light, an additional path $\lambda/2$ is introduced.

The effective path difference between the rays

$$\delta = 2\mu t\cos r + \frac{\lambda}{2} \quad \ldots\ldots (1)$$

For air film $\mu=1$ and for normal incidence r = 0

$$\delta = 2t + \frac{\lambda}{2} \quad \ldots\ldots (2)$$

At the point of contact $t = 0$, $\delta = \frac{\lambda}{2}$, this is the condition for minimum intensity. Hence the central spot is dark.

The condition for bright ring is

$$\delta = 2t + \frac{\lambda}{2} = n\lambda$$

$$\Rightarrow 2t = (2n-1)\frac{\lambda}{2} \quad \ldots\ldots. (3)$$ Where n=0, 1, 2, 3 ....

The condition for dark ring is

$$\delta = 2t + \frac{\lambda}{2} = (2n+1)\frac{\lambda}{2}$$

$$\Rightarrow 2t = n\lambda \quad \ldots\ldots. (4)$$ Where n=0, 1, 2, 3 ....

Let us consider the curved surface of the lens as an arc of a circle whose center is at C.

$$NP \times NQ = NO \times NO'$$

$$r \times r = t \times (2R - t)$$

$$r^2 = 2Rt - t^2$$

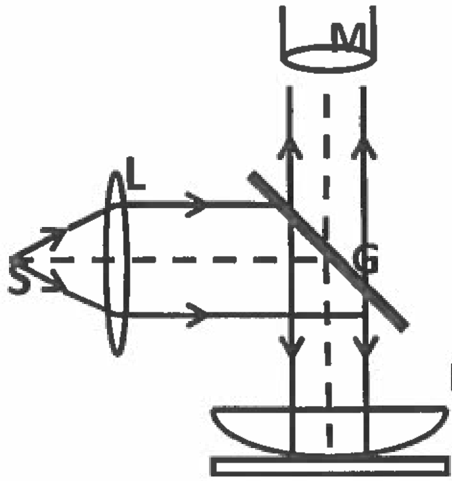
$$r^2 \approx 2Rt$$ (Since t is small t^2 is very small)

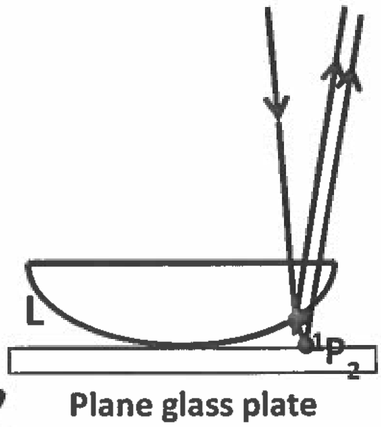
$$r^2 = 2Rt$$

$$t = \frac{r^2}{2R} \quad \ldots\ldots (5)$$

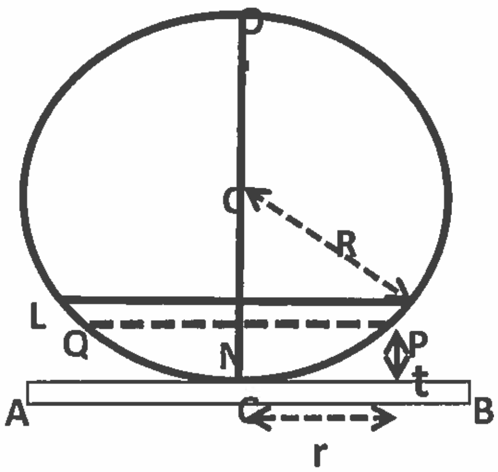
For bright rings, $2 \times \frac{r^2}{2R} = (2n-1)\frac{\lambda}{2}$

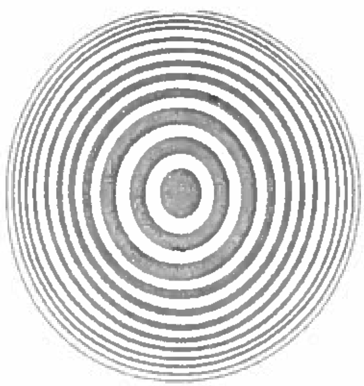
surface of the lens L and remaining is transmitted which is reflected back from the plane surface of the glass plate P. These two reflected rays (P_1 and P_2) are interfering and produce an interference pattern in the form of bright and dark circular rings. These rings can be viewed in a microscope M focused on the film.





Theory:





(1)

1.Superposition is due to two separate wave fronts originating from two coherent sources.	1.Superposition is due to secondary wavelets originating from different parts of the same wave front.
2.The fringe width may be may not be equal.	2.The fringe width of various fringes is never equal.

(2) Total Internal Reflection:

Optical fiber works on the principle of total internal reflection. When light traveling in an optically dense medium hits a boundary at a steep angle (larger than the critical angle for the boundary), the light is completely reflected. This is called total internal reflection.

(3) 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.

(4) Lighter the particle, greater the wavelength associated with it.

Smaller the velocity of the particle, longer the wavelength associated with it.

(5) Materials having small energy band gap (=1eV) are known as semiconductors.
Materials having very high energy band gap (=6eV) are known as insulators.

A. 6(a) When a plano-convex lens of long focal length with its convex surface is placed on a plane glass plate. At the point of contact where the lens touches the glass plate the thickness of the air film is zero and when moved gradually towards the edge of the lens, the thickness of the air film is increases. If a monochromatic light is allowed to fall normally and the film is viewed in reflected light, alternate dark and bright concentric circular rings are observed around the point of contact. This phenomenon was first observed by Newton, the rings are called Newton's rings.

***Experimental arrangements*:**

A plano-convex lens L of large radius of curvature and is placed on a plane glass plate. The light from monochromatic source is incident on a glass plate, which is placed at an angle of 45° with vertical. The glass plate reflects normally a part of incident light towards the air film enclosed by the lens L and the glass plate P. A part of the incident light is reflected by the curved

12(a) Concept of deBroglie waves 2M
Properties & expression for wavelength 8M

(b) $\lambda = 0.21 nm$ $v = ?$ 1M

$$\lambda = \frac{h}{mv} = \frac{6.625 \times 10^{-34}}{9.1 \times 10^{-31} \times v} = 0.21$$

$$v = \frac{6.625 \times 10^{-34}}{9.1 \times 10^{-31} \times 0.21 \times 10^{-9}} = 0.5232 \times 10^{6} \ m/s$$ 1M

(OR)

(13(a) Any 4 differences between classical free electron theory and quantum free electron theory 8M
(4X2 = 8M)

(b) Fermi-Dirac distribution function 2M
Temperature dependence curve 2M

14(a) Brief note on energy band formation 3M

(b) Introduction and potential curve 3M
Kronig-Penny equation (qualitative only) 3M
Energy bands from graph 3M

(OR)

15(a) Distinction between conductors, insulators and semiconductors 6M
Basic definitions of Intrinsic & Extrinsic semiconductors 2M

(b) Any four applications of Hall effect 4M

8 (a) List out various applications of LASERS (Scientific, Industrial and commercial) 8 M

(b) Any two differences between Spontaneous & Stimulated Emission each (2M) 4 M.

(OR)

9 (a) Classification based on Modes 4 M

Classification based on refractive index profile 6 M

(b) Numerical Aperture = 0.02

Cladding refractive index $n_2 = 1.59$

Core refractive index $n_1 = ?$ 1 M

$$NA = \sqrt{n_1^2 - n_2^2} \quad (\text{Assume } n_0 = 1)$$

$$n_1^2 - n_2^2 = (NA)^2$$

$$n_1^2 = (NA)^2 + n_2^2 = 0.0004 + 2.5281$$

$$n_1^2 = 2.5285 \Rightarrow n_1 = \sqrt{2.5285} = 1.590125$$ 1 M

10 (a) Classification, Properties and examples. (Diamagnetic, Paramagnetic & Ferromagnetic) 10 M

(b) $\chi_p = 3.7 \times 10^{-3}$, $\mu_r = ?$ 1 M

$\mu_r = 1 + \chi_p = 1 + 3.7 \times 10^{-3} = 1.0037$ 1 M.

(OR)

11 (a) Diagram with brief description 2 M

Each component of Internal field (E_0, E_1, E_2 and E_3 each 2M.) 4x2 = 8 M

No derivations – only qualitative.

(b) Basic definitions of dielectric const. 1 M

and Electric dipole 1 M

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20BSX33: Applied Physics Scheme of valuation

Part-A

(1.) Any two differences between Interference & diffraction (Each 1M) 2M.

(2) Brief note on Total Internal Reflection (TIR) 2M.

(3) Any two applications of dielectric materials (each 1M) 2M.

(4) Any two properties of matter waves (each 1M) 2M.

(5) Definition of semiconductor and insulator (each 1M) 2M.

Part-B.

6(a) principle and diagrams 3M

condition for bright ring 3M

condition for dark ring 3M

(b) Conditions for getting interference 3M
(Amplitude, frequency, phase)

(OR)

7(a) Principle and construction 4M

Grating Equation 2M

Grating spectrum 2M

(b) $N = \frac{15000}{2.54} = 5905$ lines/cm, $\lambda = 600\,nm = 600\times10^{-9}\,m = 6000\,Å$ 2M

$n \leq \frac{1}{N\lambda}$ 1M

$n \leq \frac{1}{5905\times6000\times10^{-8}}$

$n_{max} = 3$ 1M

12 (a)	Explain de-Broglie's concept of matter waves. Derive an expression for the de-Broglie wavelength	10M	20BSX33.4	L2
12 (b)	Calculate the velocity of an electron having wavelength of 0.21 nm	2M	20BSX33.4	L2
	OR			
13 (a)	Distinguish between the classical free electron theory and quantum free electron theory of metals	8M	20BSX33.4	L2
13 (b)	Illustrate the effect of temperature on the Fermi – Dirac distribution function	4M	20BSX33.4	L2
14 (a)	Discuss the origin of energy band formation in solids	3M	20BSX33.5	L2
14 (b)	Discuss Kronig-Penny model. Extend the conclusions drawn from the graph	9M	20BSX33.5	L2
	OR			
15 (a)	Distinguish between conductors, insulators and semi conductors, intrinsic and extrinsic semiconductors	3M	20BSX33.5	L2
15 (b)	Write the applications of Hall Effect	4M	20BSX33.5	L2

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Semester End Examination, Sept/Oct, 2021

Degree	B. Tech. (U. G.)	Program	ECE			Academic Year	2020 - 2021
Course Code	20BSX33	Test Duration	3 Hrs.	Max. Marks	70	Semester	II
Course	APPLIED PHYSICS						

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	Differentiate phenomena's of interference and diffraction exhibited by light	20BSX33.1	L2
2	What is the principle behind propagation of light signal through an optical fiber?	20BSX33.2	L1
3	Mention any two applications of dielectric materials	20BSX33.3	L1
4	Write any two properties of matter waves	20BSX33.4	L1
5	Define semiconductor and insulator	20BSX33.5	L1

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	In Newton's ring experiments derive conditions for dark and bright rings	9M	20BSX33.1	L2
6 (b)	What are the conditions to get interference?	3M	20BSX33.1	L2
	OR			
7 (a)	Explain the theory of Fraunhofer diffraction due to diffraction grating. Discuss its construction.	8M	20BSX33.1	L2
7 (b)	Find the highest order that can be seen with a grating having 15000 lines per inch. The wavelength of light used is 600 nm	4M	20BSX33.1	L2
8 (a)	Write the applications of lasers	8M	20BSX33.2	L2
8 (b)	Write any two differences between spontaneous and stimulated emissions	4M	20BSX33.2	L2
	OR			
9 (a)	Explain the classification of fibers	10M	20BSX33.2	L2
9 (b)	An optical fiber has a numerical aperture of 0.02 and a cladding refractive index of 1.59. Determine the value of refractive index of core	2M	20BSX33.2	L2
10 (a)	Distinguish between diamagnetic, paramagnetic and ferromagnetic materials. Explain their behavior with the help of examples	10M	20BSX33.3	L2
10 (b)	A paramagnetic material has the susceptibility of 3.7×10^{-3}, calculate the relative permeability	2M	20BSX33.3	L2
	OR			
11 (a)	Describe Lorentz method to calculate the internal field of a dielectric material	10M	20BSX33.3	L2
11 (b)	Define the terms: (a) Dielectric constant (b) Electric dipole	2M	20BSX33.3	L2

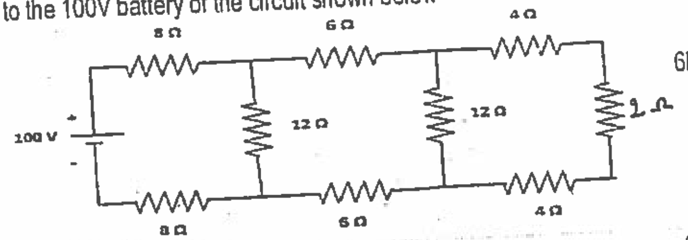
Semester End Examination, Sept./Oct., 2021

Degree	B. Tech. (U. G.)	**Program**	CE/ME/CSE/CSM/CSD	**Academic Year**	2020 - 2021
Course Code	20ESX05	**Test Duration**	3 Hrs. Max. Marks 70	**Semester**	II
Course	**Basic Electrical and Electronics Engineering**				

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

No.	Questions (1 through 5)	Learning Outcome (s)	DoK
1	State Kirchhoff's voltage and current laws	20ESX05.1	L1
2	Draw the Torque-Speed characteristic curve of the DC Shunt motor	20ESX05.2	L2
3	Write the expression for the starting torque of the 3φ induction motor $T_s = \frac{K E_2^2 R_2}{\sqrt{R_2^2 + X_2^2}}$	20ESX05.3	L2
4	Define minimum regulation and maximum regulation of 1φ transformer	20ESX05.4	L1
5	List any four applications of Op-Amps	20ESX05.5	L2

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

No.	Questions (6 through 15)	Marks	Learning Outcome (s)	DoK
6 (a)	Calculate (i) equivalent resistance across the terminal of the supply ii) total current supplied by the source, iii) power delivered to the 100V battery of the circuit shown below 	6M	20ESX05.1	L3
6 (b)	Briefly discuss various network elements	6M	20ESX05.1	L1
	OR			
7 (a)	The impedance of the series circuit is $Z_1 = (4+j6)$ ohms and $Z_2 = (12-j8)$ ohms. If the applied voltage is 220 V, find (i) current and power factor of each branch (ii) overall current (iii) power consumed by each impedance	8M	20ESX05.1	L3
7 (b)	Explain the phasor relation for series RL and RC elements	4M	20ESX05.1	L1
8	Explain the construction and principle of operation of the DC motor with neat diagrams	12M	20ESX05.2	L2
	OR			
9 (a)	Explain the speed control techniques of the DC motor	6M	20ESX05.2	L2
9 (b)	What is the need for the starter in the DC motor? With a neat diagram, explain the operation of the 3-point starter	6M	20ESX05.2	L2
10	Explain the procedure to determine the voltage regulation of the 3Φ alternator using the synchronous impedance method in detail with supportive diagrams	12M	20ESX05.3	L2
	OR			
11 (a)	Explain the operation of the 3Φ induction motor	6M	20ESX05.3	L2
11 (b)	With a neat diagram, explain the speed-torque characteristics of the 3Φ induction motor	6M	20ESX05.3	L2

12 (a)	Explain the operation of the single-phase transformer	6M	20ESX05.4	L2
12 (b)	Derive the EMF equation of the transformer	6M	20ESX05.4	L2
	OR			
13	Explain the OC and SC test of a single-phase transformer with neat sketches	12M	20ESX05.4	L2
14	With a neat schematic diagram and waveform, explain the operation of the bridge diode rectifier and write the average output voltage and current equations	12M	20ESX05.5	L2
	OR			
15(a)	Explain the characteristics of the Op-Amp	6M	20ESX05.5	L2
15(b)	Explain the following: (i) Non-inverting Amplifier (ii) Differentiator	3M 3M	20ESX05.5	L2

Network elements

(i) Active elements: The elements which are capable of delivering the energy from one element to other

Ex: voltage source, current source

(ii) passive elements which only receives power

PART-A

1. Definition of KCL - 1M
Definition of KVL - 1M

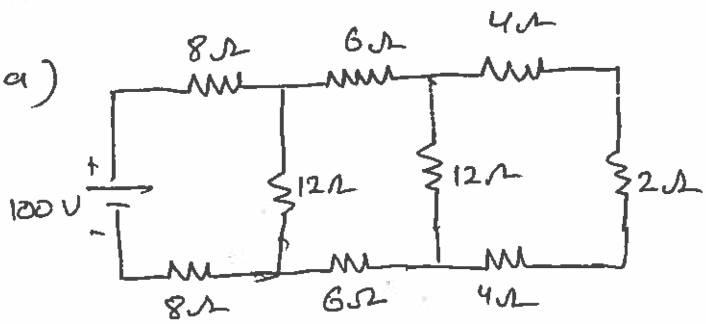
2. Speed - torque graph - 2M

3. Definition of formulate of starting torque - 1M
final equation - 1M

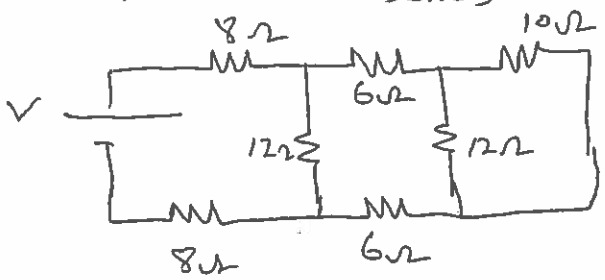
4. Definition of regulation - 1M
min & max regulation of equation - 1M

5. 4. Applications - 2M

6 a)



4Ω, 2Ω, 4Ω are series



10Ω, 12Ω ∥el

$$\frac{10\times12}{10+12} = \frac{120}{22} = 5.45\Omega$$

5.45Ω, 6Ω, 6Ω series

$$5.45+6+6 = 17.45\Omega$$

17.45Ω, 12Ω ∥el

$$\frac{17.45\times12}{17.45+12} = \frac{209.4}{29.45} = 7.11\Omega$$

7.11Ω, 8Ω, 8Ω are series.

$$7.11+8+8 = 23.11\Omega$$

$\boxed{R_{eq} = 23.11\Omega}$

$V = 100V$ $I_T = \frac{V}{R} = \frac{100}{23.11}$

$\boxed{I_T = 4.32A}$

$P = VI = 100\times4.327$

$\boxed{P = 432.71W}$

7(b) Phasor relation for series RL & RC elements

Circuit diagrams of RL & RC - 2M

Equations & its derivation - 2M

Phasor diagrams of RL & RC - 2M

6(b) :- Types of elements

8 elements — 3M for naming each
[Active, Passive, unilateral, bilateral, linear,
non-linear, lumped distributed.

for explaining each element [3m]

8) Construction and principle of operation of DC motor

Construction diagram – 4m
Principle of operation – 4m
Theory and explanation of parts – 4m

9.a) Speed control techniques

1) Armature or flux control – 3m
2) field control – 3m

9b) 3-point starter:-

Need of 3-point starter – 2m

Diagram – 2m

construction and operation – 2m

10)

10) Definition of voltage regulation - 2m

formula and explanation - 2

Types to find voltage regulation - 1m

open circuit & short circuit Characteristics - 5m

final formula - 2m.

11a) operation of 3ϕ Im

- Diagram - 2m
- Principle of operation - - 2m
- Explanation - 2m

b) speed Torque of 3-ϕ I.M.

Diagram - 3m

explanation - 3m.

15(a) characteristics of OP-AMP

Definition & Diagram - 4M

Explanation & theory - 4M

final equation & wave forms } - 4M

15(b)

(i) Non-inverting amplifier

Definition & Diagram - 2m

Explanation - 1m

(ii) Differentiator

Definition & Diagram - 2m

Explanation - 1m

1 to 5 - 2M

8, 10, 13, 14 - 12M

7 → 8, 4 M

HOD-EEE 8/10/21.

12 (a) operation of Single phase transformer

principle of operation - 2M

Diagram - 2M

Explanation part - 2M

12 (b) EMF Equation of transformer

Definition - 2M

Derivation - 3M

final equation - 1M.

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13 OC & SC TEST on 1 phase transformer

Need to Conduct OC & SC test - 2M

Required diagrams - 4M

Process to conduct exp - 4M

final equations - 2M

14. Bridge diode Rectifier

Definition - 2M

Diagram & waveforms - 5M

Explanation - 3M

Voltage & Current eqn - 2M

12

PART A

Sl.No Solutions

1. Kirchhoffs Current Law

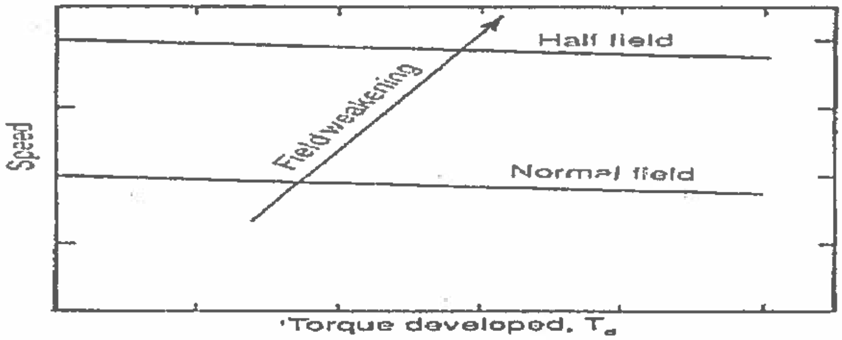
Here, the three currents entering the node, I1, I2, I3 are all positive in value and the two currents leaving the node, I4 and I5 are negative in value. Then this means we can also rewrite the equation as;

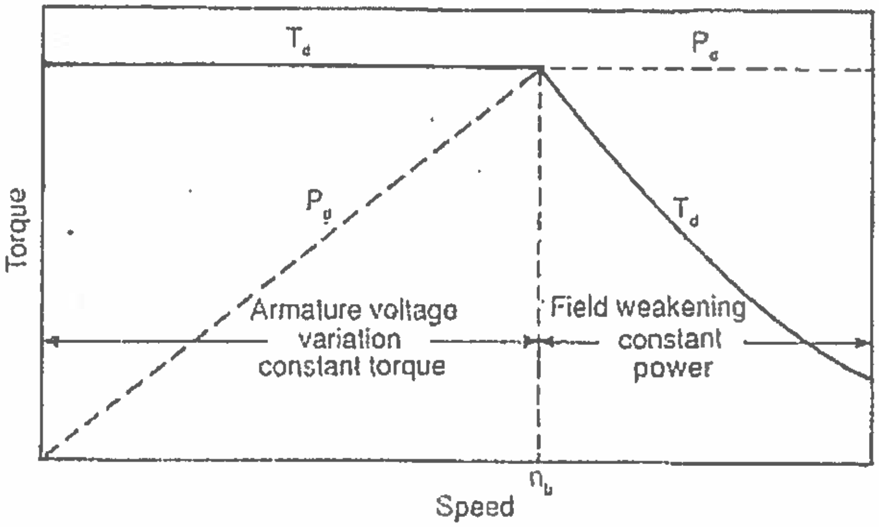
$$I1 + I2 + I3 - I4 - I5 = 0$$

Kirchhoffs Second Law – The Voltage Law, (KVL)

Kirchhoffs Voltage Law or KVL, states that "in any closed loop network, the total voltage around the loop is equal to the sum of all the voltage drops within the same loop" which is also equal to zero. In other words the algebraic sum of all voltages within the loop must be equal to zero. This idea by Kirchhoff is known as the Conservation of Energy.

2. Speed torque characteristics:





3. Expression for stating torque in induction motor

$$T = KsE_2^2 \frac{R_2}{\sqrt{R_2^2 + (sX_2)^2}}$$

$$This\ constant\ K = \frac{3}{2\pi n_s}$$

4. Voltage regulation:

Voltage regulation is a measure of change in the voltage magnitude between the sending and receiving end of a component. It is commonly used in power engineering to describe the percentage voltage difference between no load and full load voltages distribution lines, transmission lines, and transformers.

$$Voltage\ regulation(\%) = \frac{E_2 - V_2}{V_2} \times 100\%$$

Voltage regulation of transformer at lagging power factor,

$$Voltage\ regulation\ (\%) = \frac{E_2 - V_2}{V_2} \times 100(\%)$$

$$= \frac{I_2R_2 \cos\theta_2 + I_2X_2 \sin\theta_2}{V_2} \times 100(\%)$$

Voltage regulation of transformer at leading power factor,

$$Voltage\ regulation\ (\%) = \frac{E_2 - V_2}{V_2} \times 100(\%)$$

$$= \frac{I_2R_2 \cos\theta_2 - I_2X_2 \sin\theta_2}{V_2} \times 100(\%)$$

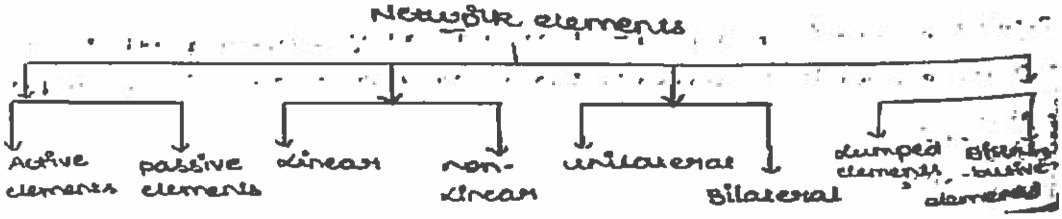
5. Applications

voltage follower, selective inversion circuit, a current-to-voltage converter, active rectifier, integrator

PART B

Sl.No Solutions

6b

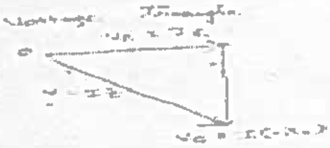


Sl.No Solutions

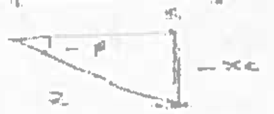
7b

Series combination of RC circuit

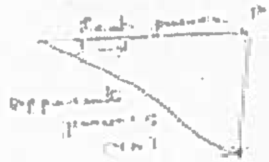
Voltage Triangle



Impedance Triangle



Power Triangle



Real power

Reactive power

Apparent power

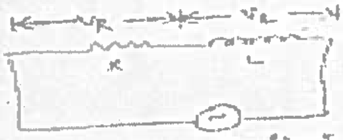
Instantaneous Power

Average Power

Case: Combination of RL circuit

Consider the circuit with pure resistor R and pure inductive coil of L henry

In series circuit current i_t flows through R & L.



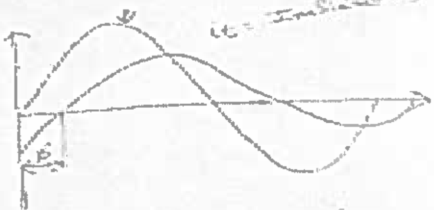
$v_t = V_m \sin \omega t$

But different v will be passing through it

So $v_t = v_R + v_L$

$v_t = V_m \sin \omega t$

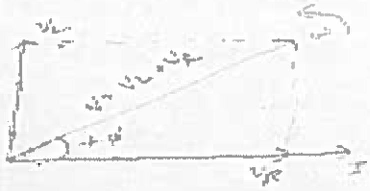
$i_t = I_m \sin(\omega t - \phi)$



For pure resistor 'R', V & I are both in phase

So V_R is drawn superimposed onto current vector

For pure inductor current I lags V_L by $90°$



If supply voltage

$v = V_m \sin \omega t$

$i = I_m \sin(\omega t - \phi)$

$I_m = \frac{V_m}{Z}$

voltage triangle

$V_R = IR$

$V_L = IX_L$

$V = \sqrt{V_R^2 + V_L^2}$

$= \sqrt{(IR)^2 + (IX_L)^2}$

$= I\left(\sqrt{R^2 + X_L^2}\right)$

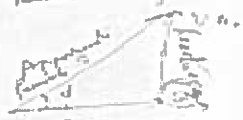
$= IZ$

Impedance triangle

$Z = \sqrt{R^2 + X_L^2}$

$\phi = \tan^{-1}\frac{X_L}{R}$

Power triangle



Real power, P

$P = VI\cos\phi = I^2R$

$Q = VI\sin\phi = I^2X_L$

$S = VI = I^2Z$

Power factor: $\cos\phi = R/Z$

Instantaneous Power

$P_i = v \times i$

$= V_m \sin\omega t \times I_m \sin(\omega t - \phi)$

multiply & divide by 2

$= \frac{2 V_m \sin\omega t \times I_m \sin(\omega t - \phi)}{2}$

$= \frac{V_m I_m}{2}\left(\cos\phi - \cos(2\omega t - \phi)\right)$

average Power

$P_{ave} = \frac{1}{2\pi}\int_0^{2\pi} \frac{V_m I_m}{2}\left[\cos\phi - \cos(2\omega t - \phi)\right] d\omega t$

$= \frac{V_m I_m}{2\pi}\int_0^{2\pi} \frac{1}{2}\left[\cos\phi - \cos(2\omega t - \phi)\right] d\omega t$

$= \frac{V_m I_m}{4\pi}\left[\cos\phi(\omega t)_0^{2\pi} - \left(\frac{\sin(2\omega t - \phi)}{2}\right)_0^{2\pi}\right]$

$= \frac{V_m I_m}{4\pi}\left[2\pi\cos\phi\right]$

$- \frac{V_m I_m}{8\pi}\left[\sin(4\pi - \phi) - \sin(-\phi)\right]$

$= \frac{V_m I_m}{2}\left[\cos\phi\right] - \frac{V_m I_m}{8\pi}\left[-\sin\phi + \sin\phi\right] = \frac{V_m I_m}{2}\cos\phi$

$= \frac{V_m}{\sqrt{2}}\frac{I_m}{\sqrt{2}}\cos\phi = VI\cos\phi$

8 Principle and Operation of DC motor

PRINCIPLE OF OPERATION

When a current-carrying conductor is placed in a magnetic field, it experiences a torque and has a tendency to move.

In other words, when a magnetic field and an electric field interact, a mechanical force is produced. The DC motor or direct current motor works on that principle. This is known as motoring action

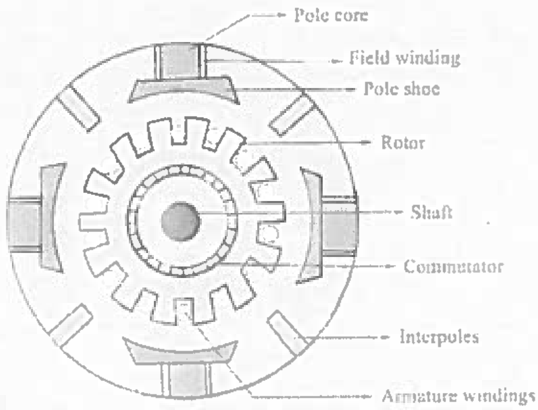
The direction of rotation of this motor is given by Fleming's left hand rule, which states that if the index finger, middle finger, and thumb of your left hand are extended mutually perpendicular to each other and if the index finger represents the direction of the magnetic field, middle finger indicates the direction of the current, then the thumb represents the direction in which force is experienced by the shaft of the DC motor.

Structurally and construction wise a direct current motor is exactly similar to a DC generator, but electrically it is just the opposite.

The DC machine consists of two parts: One part is rotating, called rotor and the other part is stationary, called stator.

The major components of a DC machine are:

- Magnetic frame or yoke
- Pole core and pole shoe
- Field coil or winding
- Armature core and winding
- Commutator
- Brushes
- Bearings and shaft



1. Magnetic Frame or yoke

It is the stationary part of the machine in the shape of hollow cylinder. Poles are fixed at the inner periphery of the yoke.

This operational amplifier circuit performs the mathematical operation of Differentiation, that is it "produces a voltage output which is directly proportional to the input voltage's rate-of-change with respect to time". In other words the faster or larger the change to the input voltage signal, the greater the input current, the greater will be the output voltage change in response, becoming more of a "spike" in shape.

As with the integrator circuit, we have a resistor and capacitor forming an RC Network across the operational amplifier and the reactance (Xc) of the capacitor plays a major role in the performance of a Op-amp Differentiator.
The input signal to the differentiator is applied to the capacitor. The capacitor blocks any DC content so there is no current flow to the amplifier summing point, X resulting in zero output voltage. The capacitor only allows AC type input voltage changes to pass through and whose frequency is dependant on the rate of change of the input signal.

At low frequencies the reactance of the capacitor is "High" resulting in a low gain (Rf/Xc) and low output voltage from the op-amp. At higher frequencies the reactance of the capacitor is much lower resulting in a higher gain and higher output voltage from the differentiator amplifier.

However, at high frequencies an op-amp differentiator circuit becomes unstable and will start to oscillate. This is due mainly to the first-order effect, which determines the frequency response of the op-amp circuit causing a second-order response which, at high frequencies gives an output voltage far higher than what would be expected. To avoid this the high frequency gain of the circuit needs to be reduced by adding an additional small value capacitor across the feedback resistor Rf.

$$V_1 = \frac{R_2}{R_2 + R_F} \times V_{OUT}$$

Ideal Summing Point: $V_1 = V_{IN}$

Voltage Gain, $A_{(V)}$ is equal to: $\dfrac{V_{OUT}}{V_{IN}}$

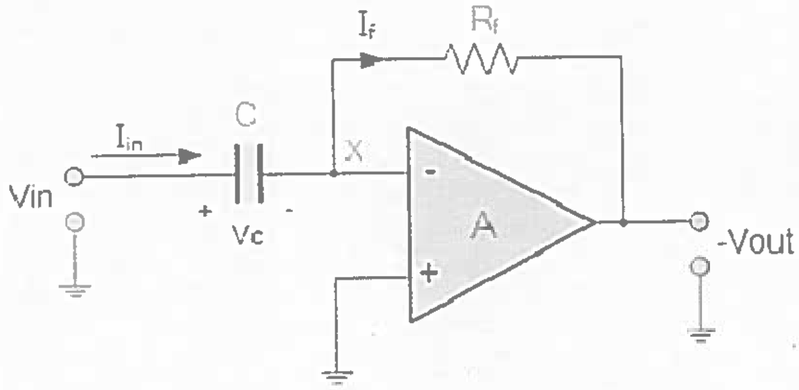
$$\text{Then, } A_{(V)} = \frac{V_{OUT}}{V_{IN}} = \frac{R_2 + R_F}{R_2}$$

$$\text{Transpose to give: } A_{(V)} = \frac{V_{OUT}}{V_{IN}} = 1 + \frac{R_F}{R_2}$$

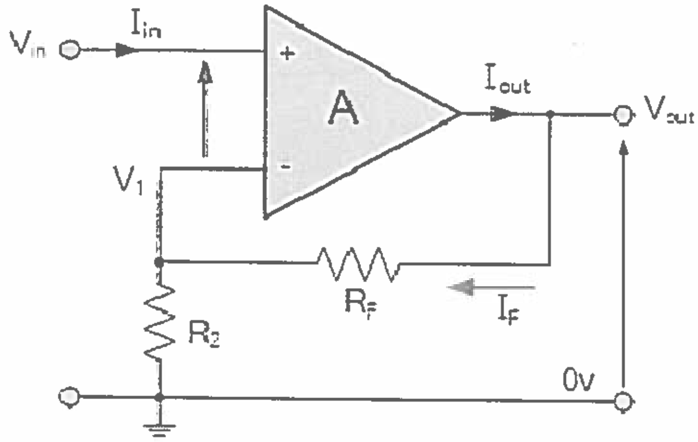
Then the closed loop voltage gain of a **Non-inverting Operational Amplifier** will be given as:

$$A_{(V)} = 1 + \frac{R_F}{R_2}$$

Differentiator:



Here, the position of the capacitor and resistor have been reversed and now the reactance, XC is connected to the input terminal of the inverting amplifier while the resistor, Rf forms the negative feedback element across the operational amplifier as normal.



In this configuration, the input voltage signal, (VIN) is applied directly to the non-inverting (+) input terminal which means that the output gain of the amplifier becomes "Positive" in value in contrast to the "Inverting Amplifier" circuit we saw in the last tutorial whose output gain is negative in value. The result of this is that the output signal is "in-phase" with the input signal.

Feedback control of the non-inverting operational amplifier is achieved by applying a small part of the output voltage signal back to the inverting (–) input terminal via a Rf – R2 voltage divider network, again producing negative feedback. This closed-loop configuration produces a non-inverting amplifier circuit with very good stability, a very high input impedance, Rin approaching infinity, as no current flows into the positive input terminal, (ideal conditions) and a low outpuIn the previous Inverting Amplifier tutorial, we said that for an ideal op-amp "No current flows into the input terminal" of the amplifier and that "V1 always equals V2". This was because the junction of the input and feedback signal (V1) are at the same potential.

In other words the junction is a "virtual earth" summing point. Because of this virtual earth node the resistors, Rf and R2 form a simple potential divider network across the non-inverting amplifier with the voltage gain of the circuit being determined by the ratios of R2 and Rf as shown below.t impedance, Rout as shown below.
Then using the formula to calculate the output voltage of a potential divider network, we can calculate the closed-loop voltage gain (Av) of the **Non-inverting Amplifier** as follows:

impedance of 10-20 kΩ. An ideal op amp behaves like a perfect voltage source delivering current without any internal losses. The internal resistance reduce the voltage available to the load.

Bandwidth(BW)

An ideal op amp has an infinite bandwidth that is it can amplify any signal from DC to the highest AC frequencies without any losses. So therefore, an ideal op amp is said to have infinite frequency response. In real op amps, the bandwidth is generally limited. The limit depends on the gain bandwidth (GB) product. GB is defined as the frequency where the amplifier gain becomes unity.

Offset Voltage(V_{io})

The offset voltage of an ideal op amp is zero, which means that the output voltage will be zero if the difference between the inverting and non-inverting terminal is zero. If both the terminals are grounded, the output voltage will be zero. But real op amps have an offset voltage.

Common Mode Rejection Ratio(CMRR)

Common mode refers to the situation when the same voltage is applied to both the inverting and non-inverting terminal of the op amp. The common mode rejection refers to the ability of the op amp to reject the common mode signal. Now we are in a position to understand the term common mode rejection ratio.

The common mode rejection ratio refers to the measure of the ability of the op amp to reject the common mode signal. Mathematically it is defined as

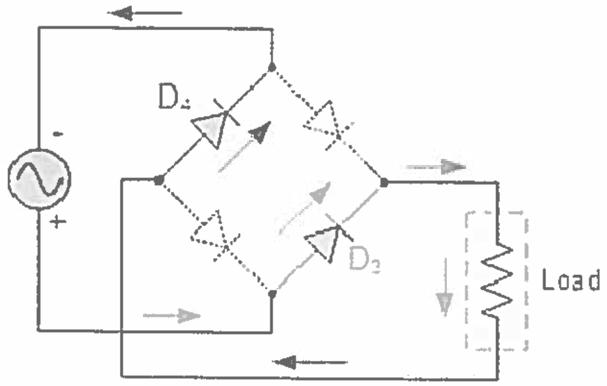
Where, A_D is the differential gain of the op amp, ∞ for an ideal op amp. A_{CM} refers to the common mode gain of the op-amp.

The CMRR of an ideal op amp is ∞. That means it is able to reject all common mode signal. Also from the formula, we can see the A_D is infinite for an ideal op amp and A_{CM} is zero. Therefore the CMRR of an ideal op-amp is infinite. Therefore it will reject any signal which is common to both.

However, real omp have finite CMRR, and does not reject all common mode signals.

15b Non inverting amplifier:

The Negative Half-cycle



As the current flowing through the load is unidirectional, so the voltage developed across the load is also unidirectional the same as for the previous two diode full-wave rectifier, therefore the average DC voltage across the load is $0.637V_{max}$.

15a Operational amplifier or op amps as they are usually referred are linear devices that can give ideal DC amplification. They are fundamentally voltage amplifying devices used with external feedback components like resistors or capacitors. An op amp is a three terminal device, with one terminal called the inverting input, other the non-inverting input and the last one is the output. Below is a diagram of a typical op amp:

As you can see from the diagram, op amp has three terminals for input and output and 2 for power supply.

Before we understand the operation of an op amp, we must learn about the op amp characteristics of an op amp. We will explain them one by one here:

Open Loop Voltage Gain(A)

The open loop voltage gain without any feedback for an ideal op amp is infinite. But typical values of open loop voltage gain for a real op amp ranges from 20,000 to 2, 00,000. Let the input voltage be V_{in}. Let A be the open loop voltage gain. Then the output voltage is $V_{out} = AV_{in}$. The value of a typically is in the range specified above but for an ideal op amp, it is infinite.

Input Impedance(Z_{in})

Input Impedance is defined as the input voltage by the input current. The input impedance of an ideal op amp is infinite. That is there no current flowing in the input circuit. However, an ideal op amp has certain current flowing in the input circuit of the magnitude of few pico-amps to a few milli-amps.

Output Impedance (Z_{out})

Output impedance is defined as the ratio of the output voltage to the input current. The output impedance of an ideal op amp is zero, however, real op amps have an output

$$Then,\ Z_e = \frac{V_{sc}}{I_L}$$

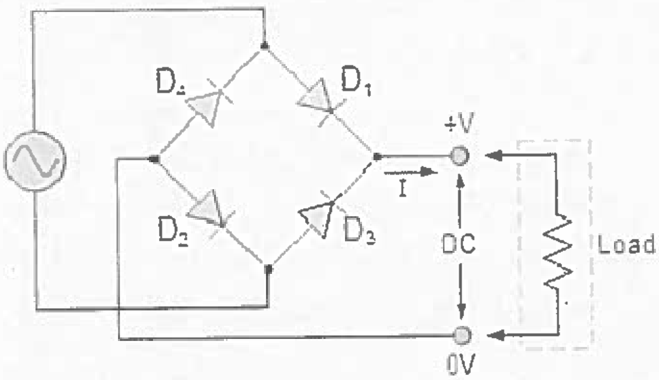
Therefore, if equivalent reactance of transformer is X_e.

$$Then,\ X_e^2 = Z_e^2 - R_e^2$$

These values are referred to the HV side of the transformer as the test is conducted on the HV side of the transformer. These values could easily be converted to the LV side by dividing these values with the square of transformation ratio.

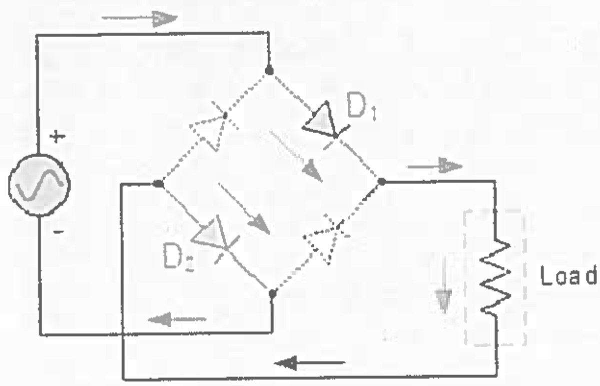
Hence the **short-circuit test of a transformer** is used to determine copper losses in the transformer at full load. It is also used to obtain the parameters to approximate the equivalent circuit of a transformer.

14 **The Diode Bridge Rectifier**



The four diodes labelled D_1 to D_4 are arranged in "series pairs" with only two diodes conducting current during each half cycle. During the positive half cycle of the supply, diodes D1 and D2 conduct in series while diodes D3 and D4 are reverse biased and the current flows through the load as shown below.

The Positive Half-cycle



During the negative half cycle of the supply, diodes D3 and D4 conduct in series, but diodes D1 and D2 switch "OFF" as they are now reverse biased. The current flowing through the load is the same direction as before.

$$Then, \left(\frac{1}{X_m}\right)^2 = \left(\frac{1}{Z_m}\right)^2 - \left(\frac{1}{R_m}\right)^2$$

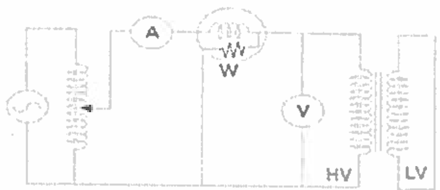
These values are referred to the LV side of the transformer due to the tests being conducted on the LV side of transformer. These values could easily be referred to HV side by multiplying these values with square of transformation ratio.

Therefore it is seen that the **open circuit test on transformer** is used to determine core losses in transformer and parameters of the shunt branch of the equivalent circuit of the transformer.

Short Circuit Test on Transformer

The connection diagram for the short circuit test on the **transformer** is shown in the figure below. A voltmeter, wattmeter, and an ammeter are connected in HV side of the transformer as shown. A low voltage of around 5-10% is applied to that HV side with the help of a variac (i.e. a variable ratio auto transformer). We short-circuit the LV side of the transformer. Now with the help of variac applied voltage is slowly increased until the wattmeter, and an ammeter gives reading equal to the rated current of the HV side.

After reaching the rated current of the HV side, we record all the three instrument readings (Voltmeter, Ammeter and Watt-meter readings). The ammeter reading gives the primary equivalent of full load current IL. As the voltage applied for full load current in a short circuit test on the transformer is quite small compared to the rated primary voltage of the transformer, the core losses in the transformer can be taken as negligible here.



Short Circuit Test on Transformer

Let's say, voltmeter reading is V_{sc}. The watt-meter reading indicates the input power during the test. As we have short-circuited the transformer, there is no output; hence the input power here consists of copper losses in the transformer. Since the applied voltage V_{sc} is short circuit voltage in the transformer and hence it is quite small compared to the rated voltage, so, we can neglect the core loss due to the small applied voltage. Hence the wattmeter reading can be taken as equal to copper losses in the transformer. Let us consider wattmeter reading is P_{sc}.

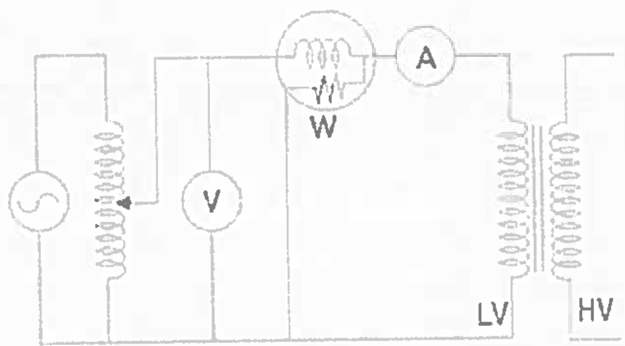
$$P_{sc} = R_e I_L^2$$

Where, R_e is equivalent resistance of transformer.

If, Z_e is equivalent impedance of transformer.

voltmeter, wattmeter, and an ammeter are connected in LV side of the transformer as shown. The voltage at rated frequency is applied to that LV side with the help of a variac of variable ratio auto transformer.

The HV side of the transformer is kept open. Now with the help of variac, applied voltage gets slowly increased until the voltmeter gives reading equal to the rated voltage of the LV side. After reaching rated LV side voltage, we record all the three instruments reading (Voltmeter, Ammeter and Wattmeter readings).



Open Circuit Test on Transformer

The ammeter reading gives the no load current I_e. As no load current I_e is quite small compared to rated current of the transformer, the voltage drops due to this current that can be taken as negligible.

Since voltmeter reading V_1 can be considered equal to the secondary induced voltage of the transformer, wattmeter reading indicates the input power during the test. As the transformer is open circuited, there is no output, hence the input power here consists of core losses in transformer and copper loss in transformer during no load condition. But as said earlier, the no-load current in the transformer is quite small compared to the full load current so, we can neglect the copper loss due to the no-load current. Hence, can take the wattmeter reading as equal to the core losses in the transformer.

Let us consider wattmeter reading is P_o.

Where, R_m is shunt branch resistance of transformer.

If, Z_m is shunt branch impedance of transformer.

$$Then,\ Z_m = \frac{V_1}{I_e}$$

Therefore, if shunt branch reactance of transformer is X_m,

$E_1 = 4.44 f N_1 \Phi_m$ .... eq 1

Similarly, RMS induced emf in secondary winding (E_2) can be given as

$E_2 = 4.44 f N_2 \Phi_m$. .... eq 2

from the above equations 1 and 2,

$$\frac{E_1}{N_1} = \frac{E_2}{N_2} = 4.44 f \Phi m$$

This is called the **emf equation of transformer**, which shows, emf / number of turns is same for both primary and secondary winding.

For an ideal transformer on no load, $E_1 = V_1$ and $E_2 = V_2$.
where, V_1 = supply voltage of primary winding
V_2 = terminal voltage of secondary winding

Voltage Transformation Ratio (K)
As derived above,

$$\frac{E_1}{N_1} = \frac{E_2}{N_2} = K$$

Where, K = constant
This constant K is known as **voltage transformation ratio.**

- If $N_2 > N_1$, i.e. $K > 1$, then the transformer is called step-up transformer.
- If $N_2 < N_1$, i.e. $K < 1$, then the transformer is called step-down transformer.

13 Open and short circuit tests are performed on a transformer to determine the:

1. Equivalent circuit of transformer
2. Voltage regulation of transformer
3. Efficiency of transformer

The power required for **open circuit tests and short circuit tests on a transformer** is equal to the power loss occurring in the transformer.

Open Circuit Test on Transformer

The connection diagram for **open circuit test on transformer** is shown in the figure. A of the transformer.

EMF Equation Of The Transformer

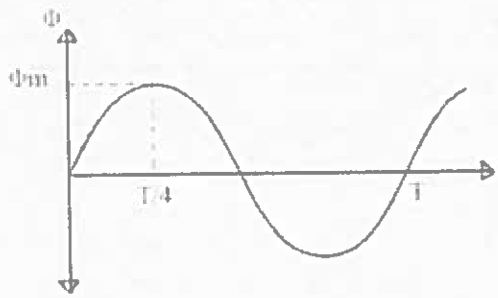
Let,

N_1 = Number of turns in primary winding

N_2 = Number of turns in secondary winding

Φ_m = Maximum flux in the core (in Wb) = (B_m x A)

f = frequency of the AC supply (in Hz)



As, shown in the fig., the flux rises sinusoidally to its maximum value Φ_m from 0. It reaches to the maximum value in one quarter of the cycle i.e in T/4 sec (where, T is time period of the sin wave of the supply = 1/f).

Therefore,

average rate of change of flux = $\Phi_m / (T/4) = \Phi_m / (1/4f)$

Therefore,

average rate of change of flux = 4f Φ_m .... (Wb/s).

Now,

Induced emf per turn = rate of change of flux per turn

Therefore, average emf per turn = 4f Φ_m ....(Volts).

Now, we know, Form factor = RMS value / average value

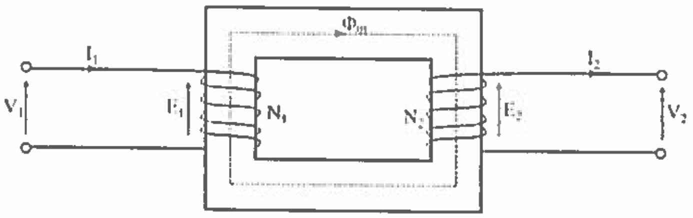
Therefore, RMS value of emf per turn = Form factor X average emf per turn.

As, the flux Φ varies sinusoidally, form factor of a sine wave is 1.11

Therefore, RMS value of emf per turn = 1.11 x 4f Φ_m = 4.44f Φ_m.

RMS value of induced emf in whole primary winding (E_1) = RMS value of emf per turn X Number of turns in primary winding rating) develop their maximum torque at a speed about 98% of synchronous speed.

12a The working of the transformer is based on the principle of mutual inductance between two coils wound on the same magnetic core.



When an alternating voltage (V1) is applied to the primary winding, an alternating magnetic flux (Φm) sets up in the core and links with the secondary winding, i.e. the magnetic flux links both the windings of the transformer magnetically. This magnetic flux induces EMF E1 in the primary winding and E2 in the secondary winding according to Faraday's law of electromagnetic induction.

According to Lenz' law,

SecondaryEMF,E2=−N2dϕmdt...(2)

Therefore,

E2/E1=N2/N1...(3)

From the above equations, it is clear that the induced EMFs in the primary and secondary windings depends upon the number of turn of the winding.

If $N1 > N2$, then $E1 > E2$ i.e. the primary EMF is greater than the secondary EMF, the transformer is called as step-down transformer.

If $N2 > N1$, then $E2 > E1$ i.e. the primary EMF is less than the secondary EMF, the transformer is called as step-up transformer.

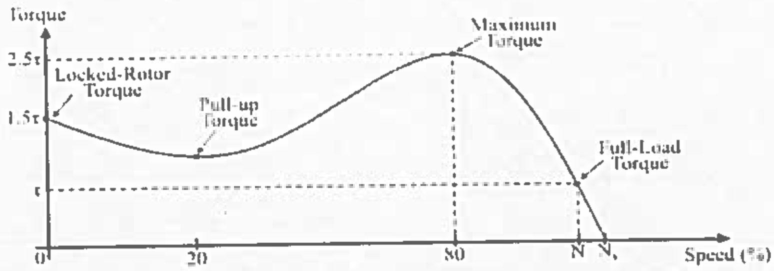
If a load is connected across the terminals of the secondary winding, the secondary EMF causes a current I2 to flow through the load. In this way, a transformer transfers AC power from one circuit to another circuit with a change in voltage level without any electrical connection between both the circuits i.e. the power from input circuit to output circuit transfers magnetically. During this transfer of electrical power, the frequency does not change.

12b In a transformer, source of alternating current is applied to the primary winding. Due to this, the current in the primary winding (called as magnetizing current) produces alternating flux in the core of transformer. This alternating flux gets linked with the secondary winding, and because of the phenomenon of mutual induction an emf gets induced in the secondary winding. Magnitude of this induced emf can be found by using the following **EMF equation**

- The RMF passes through air gap and cuts the rotor conductors, which are stationary at start. Due to relative motion between RMF and the stationary rotor, an EMF is induced in the rotor conductors. Since the rotor circuit is short-circuited, a current starts flowing in the rotor conductors.
- Now, the current carrying rotor conductors are in a magnetic field created by the stator. As a result of this, mechanical force acts on the rotor conductors. The sum of mechanical forces on all the rotor conductors produces a torque which tries to move the rotor in the same direction as the RMF.
- Hence, the induction motor starts to rotate. From, the above discussion, it can be seen that the three phase induction motor is self-starting motor.
- The three induction motor accelerates till the speed reached to a speed just below the synchronous speed.

11b The torque-speed characteristics of a 3-phase induction motor is defined as the curve plotted between torque developed and rotational speed of the motor. It gives the information about variation in the motor torque with the change in its speed.

As the torque of three-phase induction depends upon its speed but the relationship between them cannot be expressed by a simple equation. Therefore, we use the torque-speed curve to express the relationship between them.

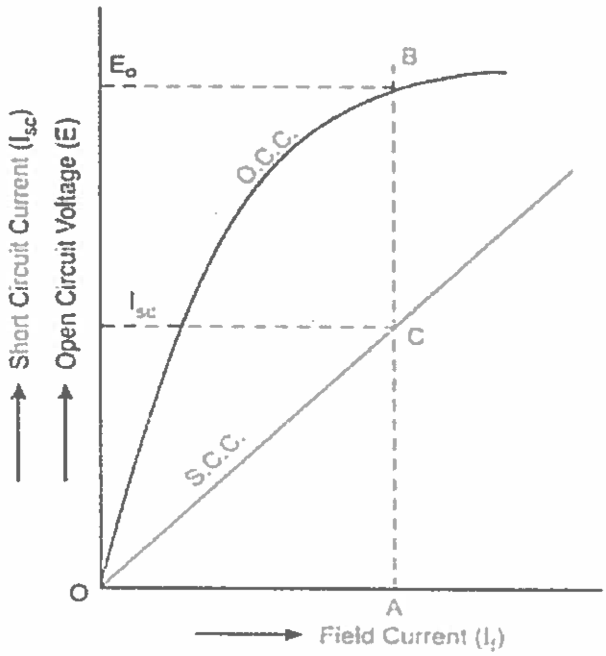


If the full-load torque is τ, then the starting torque or locked rotor torque is 1.5 times of τ and the maximum torque (also known as breakdown torque) is 2.5 times of τ.

The full load speed of the motor is N. If the mechanical load on the shaft is increased, the motor speed will decrease until the electromagnetic torque (or motor torque) is again equal to the load torque. As soon as the two torques are equal, the motor will run at a constant speed but lower than the previous speed. Although, if the torque exceeds the breakdown torque (2.5τ), the will suddenly stop.

The torque-speed characteristics of a three-phase induction motor is a straight line between the no-load and full-load operating points. The slope of the curve line depends upon the resistance of the rotor circuit i.e. the higher the rotor circuit resistance, the sharper the slope of the curve.

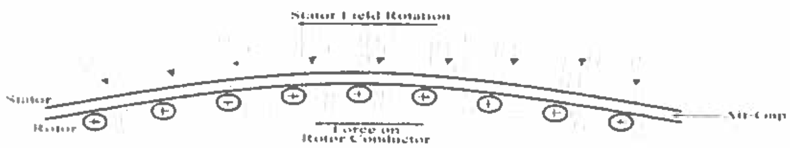
The small three-phase induction motors (below 10 kW rating) develop their maximum torque at a speed about 80% of synchronous speed whereas large motors (more than 1000 kW



To determine synchronous impedance of the alternator, from figure 4, let OA be the extension or field current (If). For this field current OA, the open circuit voltage is AB (Eo) and for the same field current the short circuit current is AC (Isc). When the alternator is short circuited terminal voltage is zero. Therefore, at short circuit, whole of the induced voltage (Eo) is being utilised for circulating the short circuit current (Isc) through the synchronous impedance (Zs). Zs = Open Circuit Voltage / Short Circuit Current Zs = Eo / Isc (at the same field current)

11a A three phase induction motor has a stator and a rotor. The stator carries a 3-phase winding called as stator winding while the rotor carries a short circuited winding called as rotor winding. The stator winding is fed from 3-phase supply and the rotor winding derives its voltage and power from the stator winding through electromagnetic induction. Therefore, the working principle of a 3-phase induction motor is fundamentally based on electromagnetic induction.

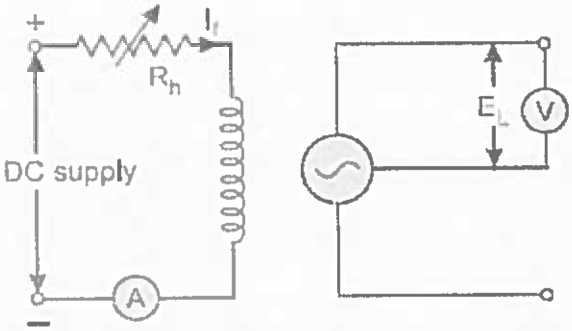
Consider a portion of a three phase induction motor (see the figure). Therefore, the working of a three phase induction motor can be explained as follows –



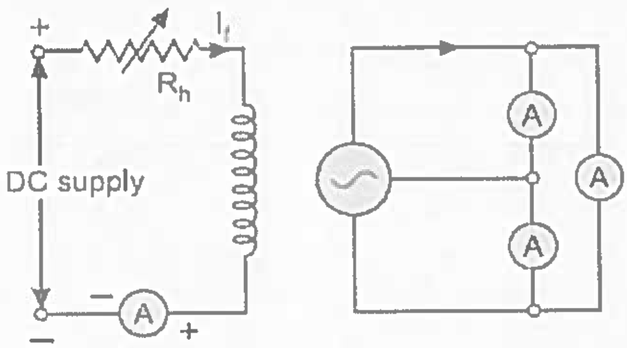
- When the stator winding is connected to a balanced three phase supply, a rotating magnetic field (RMF) is setup which rotates around the stator at synchronous speed (N_s). Where,

$$NS=120f/P$$

a rheostat. A voltmeter is connected across the terminals of the alternator to measure open circuit voltage (Eo) and an ammeter is connected in the field circuit to measure field current (If) as shown in Figure 2. The field current (excitation current) is gradually varied (increased in steps) and the voltage across the terminals of the alternator (Eo) is recorded for every change in the field current (If). A graph is plotted taking (If) along abscissa and (Eo) along the ordinate called open circuit characteristics (O.C.C.). The O.C.C. curve so obtained is shown in Figure 4. The curve rises steeply and then flattened due to saturation of the magnetic circuit.



Short circuit test: To perform short circuit test, the terminals of the alternator are short circuited by a thick strip or an ammeter as shown in Figure 3. And its rotor is rotated by the prime-mover at synchronous speed. The field current If is gradually increased and the short circuit current (Isc) is recorded for every change in the field current (If) with the help of ammeter connected across the alternator terminals. A graph is plotted taking (If) along abscissa and (Isc) along with ordinate called short circuit characteristics (S.C.C.).



speed. Finally, when the starter handle is in 'RUN' position, the entire starting resistance is eliminated, and the motor runs with normal speed

This is because back emf is developed consequently with speed to counter the supply voltage and reduce the armature current

So the external electrical resistance is not required anymore and is removed for optimum operation. The handle is moved manually from OFF to the RUN position with the development of speed. Now the obvious question is once the handle is taken to the RUN position how it is supposed to stay there, as long as the motor is running.

The supply to the field winding is derived through no voltage coil. So when field current flows, the NVC is magnetized.

Now when the handle is in the 'RUN' position, a soft iron piece is connected to the handle and gets attracted by the magnetic force produced by NVC, because of flow of current through it. The NVC is designed in such a way that it holds the handle in 'RUN' position against the force of the spring as long as supply is given to the motor.

Thus NVC holds the handle in the 'RUN' position and hence also called hold on coil.

Now when there is any kind of supply failure, the current flow through NVC is affected and it immediately loses its magnetic property and is unable to keep the soft iron piece on the handle, attracted

At this point under the action of the spring force, the handle comes back to OFF position, opening the circuit and thus switching off the motor. So due to the combination of NVC and the spring, the starter handle always comes back to OFF position whenever there is any supply problem. Thus it also acts as a protective device safeguarding the motor from any kind of abnormality.

10 The voltage regulation of an alternator is the difference between the noload voltage and the full-load voltage exprssed in percent of full load voltage or it is the rise in terminal voltage when a given load is removed, while the excitation and speed of alternator remaining constant, .i.e. Percentage Regulation $=(V_{no\ load} - V_{full\ load}) / V_{full\ load} * 100\% = (E_o - V_t) / V_t * 100\%$ Where E_o is no load terminal voltage, V_t is the full load rated voltage. The main causes for terminal voltage drop is illustated below: ➢ The resistance of armature winding. ➢ The leakage reactance of armature winding. ➢ Effect of armature reaction (this is the most predominant factor).

we should first determine the synchronous impedance (synchronous reactance and armature resistance) of the alternator. This requires an open-circuit and short circuit tests are to be performed on the alternator. By using these parameters, the regulation of the alternator can be determined at any load. 3.1.1. Determination of Synchronous Impedance: A- Open circuit test: To perform open circuit test, the terminals of the alternator are kept open and is rotated by the primemover at synchronous speed. A DC supply is given to the field winding through shown in the figure

The contact points of these sections are called studs and are shown separately as OFF, 1, 2, 3, 4, 5, RUN

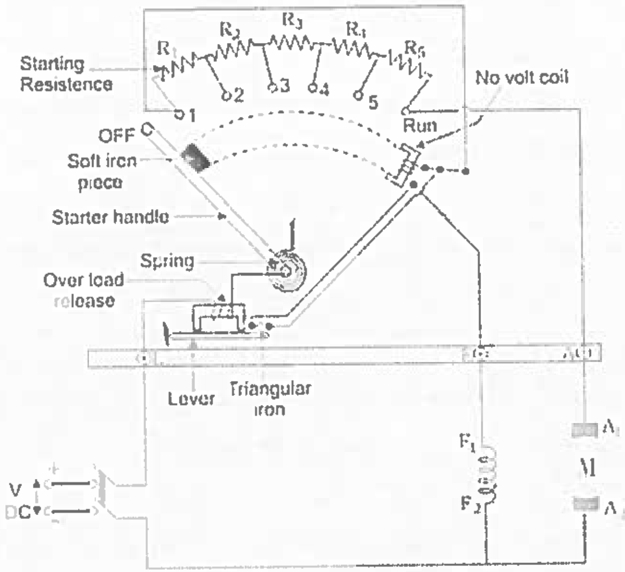
Other than that there are three main points, referred to as

'L' Line terminal (Connected to positive of supply)

'A' Armature terminal (Connected to the armature winding)

'F' Field terminal (Connected to the field winding)

And from there it gets the name 3 point starter. Now the construction of 3 point starter in further details reveals that the point 'L' is connected to an electromagnet called overload release (OLR) as shown in the figure



Three Point Starter

The other end of OLR is connected to the lower end of conducting lever of starter handle where spring is also attached with it, and the starter handle also contains a soft iron piece housed on it

This handle is free to move to the other side RUN against the force of the spring. This spring brings back the handle to its original OFF position under the influence of its own force

Another parallel path is derived from the stud '1', given to another electromagnet called No Volt Coil (NVC) which is further connected to terminal 'F.' The starting resistance at starting is entirely in series with the armature. The OLR and NVC act as the two protecting devices of the starter

(iii) Working of 3 point Starter

To start with the handle is in the OFF position when the supply to the DC motor is switched on. Then handle is slowly moved against the spring force to make contact with stud No. 1. At this point, field winding of the shunt or the compound motor gets supply through the parallel path provided to starting the resistance, through No Voltage Coil.

While entire starting resistance comes in series with the armature. The high starting armature current thus gets limited as the current equation at this stage becomes:

$$I_a = \frac{E}{(R_a + R_{st})}$$

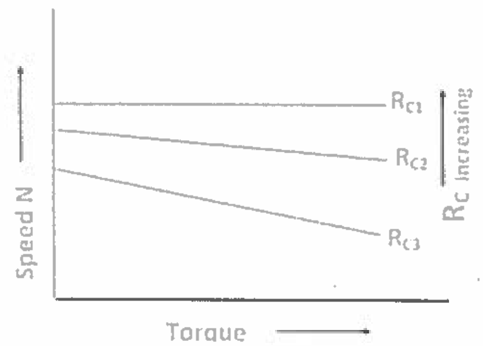
As the handle is moved further, it goes on making contact with studs 2, 3, 4, etc., thus gradually cutting off the series resistance from the armature circuit as the motor gathers equation

$$I_{sh} = \frac{V}{R_{sh} + R_C}$$

The connection of RC in the field reduces the field current, and hence the flux is also reduced. This reduction in flux increases the speed, and thus, the motor runs at a speed higher than the normal speed.

Therefore, this method is used to give motor speed above normal or to correct the fall of speed because of the load

The fig., shows the speed torque characteristics.



9b Three point starter

(i) Need of 3 point starter:

The main task of a motor starter is to start as well as stop the motor to which it is allied. Starters are particularly designed provide an overload protection for the motor. The starter gives the supply to the motor manually or automatically as well as protects the motor from the faults or overload. Based on the type of motor, the motor starters are available in different sizes with different ratings

A 3 point starter is a device that helps in the starting and running of a DC shunt motor or compound wound DC motor

it's due to the presence of back emf (Eb), which plays a critical role in governing the operation of the motor. The back emf develops as the motor armature starts to rotate in presence of the magnetic field, by generating action and counters the supply voltage. Hence the back emf at the starting of the motor is zero, but it develops gradually as the motor gathers speed

The general motor emf equation is

$E = E_b + I_a . R_a$

At the time of starting Eb = 0

Therefore,

$$\therefore I_a = \frac{E}{R_a}$$

from above equation, current will be dangerously high at starting So to limit the starting current to acceptably low value we need a 3 point starter

(ii) Construction of 3 point Starter

Construction wise a starter is a variable resistance, integrated into the number of sections as the mechanical power from or to the machine. All the rotating parts including the armature core, commutator, cooling parts and mounted and keyed to the shaft.

9a Speed control techniques of DC motor

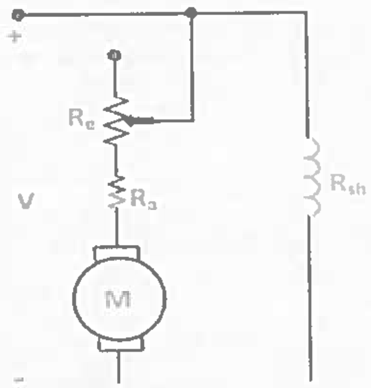
Back emf Eb of a DC motor is nothing but the induced emf in armature conductors due to rotation of the armature in magnetic field. Thus, the magnitude of Eb can be given by EMF equation of a DC generator.

but, for a DC motor A, P and Z are constants

Therefore, N ∝ K Eb/Ø (where, K=constant)

This shows the speed of a dc motor is directly proportional to the back emf and inversely proportional to the flux per pole.

I. **ARMATURE CONTROL METHODS:**

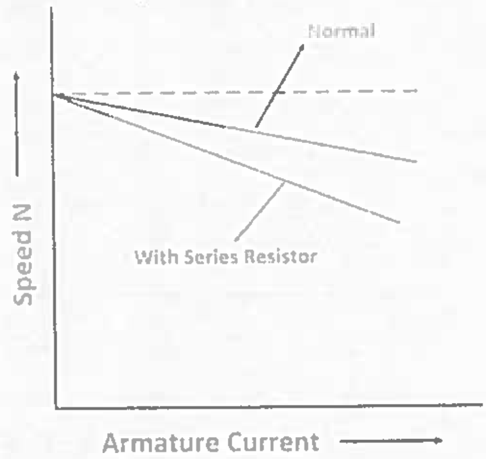


Speed of a dc motor is directly proportional to the back emf Eb;

Eb = V - IaRa.

That means, when supply voltage V and the armature resistance Ra are kept constant, then the speed is directly proportional to armature current Ia.

Speed and armature current characteristics are shown here.



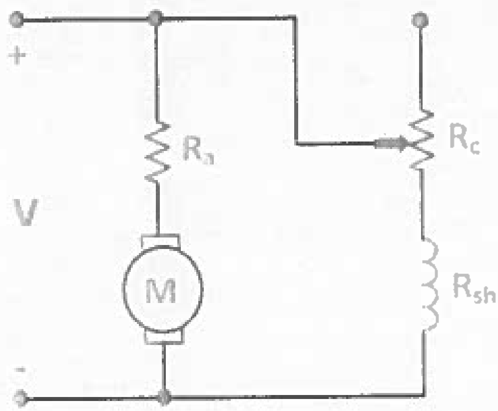
Thus, if we add resistance in series with the armature,

Ia decreases and, hence, the speed also decreases.

Greater the resistance in series with the armature,

greater the decrease in speed.

II. **FIELD FLUX CONTROL METHOD**



Flux is produced by the field current. Thus, the speed control by this method is achieved by control of the field current.

The shunt field current is given by the

It acts as the outer cover or frame for the entire machine and serves two main purposes: It is used to carry the magnetic flux produced by the poles. It acts as mechanical support for the machine.

Yoke is usually made of cast iron for small machine, because of its cheapness. But for large machines, it is made of cast steel or rolled steel, due to its high permeability.

The lifting eye, feet and the terminal box are welded to the frame afterwards.

2. Pole core and pole shoe

The field pole consists of pole cores, pole shoes and field winding. The poles are made of thin laminated sheets, to avoid heating and eddy current loss.

Pole cores are the projecting rectangular parts, which produce magnetic flux needed for the generator, when it is excited by the field winding. It is fitted to the yoke or frame by means of bolts and nuts or rivets.

The pole shoes are located at the end of pole core. The purpose of providing pole shoe in the poles is to make the magnetic field uniform on the surface of the armature.

Since the poles project inwards they are called as salient poles. Each pole as a pole shoe having a curved surface.

Following are the main function of the poles

it acts as a mechanical support to the field coil.

they reduce the reluctance of the magnetic path.

they guide and spread out the flux in the air gap

3. Field coil or winding

Field coil is made up of copper. They are mounted on the pole core and carry the dc current. The field coils are connected in such a way that adjacent poles have opposite polarity.

When the coils carry dc current, the pole core become an electromagnet and produces the magnetic flux. The magnetic flux passes through the pole core, the air gap, the armature and the yoke.

The number of poles in a DC Generator depends on the speed of the machine and the output for which the machine is designed.

There are several field constructions are adopted according to the type of excitation. In shunt field, more number of turns with small cross sectional are used, in series field only a few turns of large cross sectional area are used and in compound field, both shunt and series field winding are used.

4. Armature core and winding

In the construction of DC generator, armature core is designed as the rotating part and is built in cylindrical or drum shape with slots on its outer periphery. The purpose of armature is to house the winding and to rotate the conductors in the uniform magnetic field. It is mounted on the shaft.

It is build up of steel lamination which are insulated by each other by thin paper or thin coating of varnish as insulation. The thickness of each lamination is about 0.5 mm. These lamination will reduce the eddy current loss. If silicon sheet is used for armature core, the hysteresis loss will also reduce.

Due to losses, heat will be developed in the armature. To dissipate this heat, a fan is provided at one end of armature. Ventilating ducts (air holes) are also provided in the armature for the purpose of cooling. The width of the ventilating ducts varies from 5 to 10 mm.

The armature winding or coil is placed on slots available on the armature's outer periphery. The ends of the coils are joined with commutator segments. Insulated higher conductivity copper wire is used for making the coils. There are two types of winding.

lap winding – Lap winding is used for high current, low voltage generators.

Wave winding – Wave winding is used for high voltage, low current generators.

5. Commutator

The commutator provides the electrical connection between the rotating armature coil and the stationary external circuit. It is essentially a cylindrical structure and is built up of wedge shaped copper segments insulated from each other by mica sheets and mounted on the shaft of the machine.

The commutator is a mechanical rectifier which converts the alternating emf generator in the armature winding into direct voltage across the brushes. The ends of the armature coil or winding are connected to commutator segments.

Great care is to be taken while building the commutator because even slight eccentricity will cause the brushes to bounce, which can cause high sparking.

6. Brushes

The function of brush is to collect the current from the commutator and supply it to the external load circuit. The brushes are manufactured in a variety of compositions to suit the commutation requirements. It is made of carbon, graphite metal graphite or copper and is rectangular in shape.

The brushes are placed in the brush holders which is mounted on rocker arm. The brushes are arranged in rocker arm in such a way that, it touches the commutator.

Brush pressure is adjusted by means of adjustable springs. If the bush pressure is high, the friction produces heating of the commutator and the brushes. If the pressure is too weak, the imperfect contact with the commutator may produce spark.

7. Bearings and Shaft

For construction of smaller DC generator, ball bearings are used at both the ends of the shaft but for larger machines, roller bearings are used at the driving end and ball bearings are used at the non driving end of the machine.

The shaft is made up of mild steel having maximum breaking strength. It is used to transfer