12th CBSE (2013-14)
SUBJECT : PHYSICS
SOLUTION(Code No. 55/3)

1. Anticlock wise(according to lenz's law)

2. Metal A

3. One ampare is that value of steady current, which on flowing in each of the two parallel infinitely long conductors of negligible cross-section placed in vacuum at a distance of 1m from each other, produces between them a force of 2×10^{-7} newton per metre of their length.

4. Converging lens since refractive index of sourrounding is greater then refractive index of lens.

5. ELOF of resultant electric field can never intersect with each other. Because at the intersect point there are two direction of electric field which is not possible.

6. Microwaves

7. Neutrino hypothesis(Explanation)

8. **(i)** AC generators are strong and do not require much attention. The absence of commutator in AC generator avoids sparkings and increases the efficiency.

(ii) The AC voltage can be easily varied with the help of a transformer which is a device for changing alternating voltages. AC voltage can be easily stepped up or down as per requirement.

9. Displacement current is that current which comes into existence. In addition to the conduction current, whenever the electric field and hence the electrci flux chagnes with time.
To maintain the dimensional consistency, the displacement current is given the form :

$$I_d = \varepsilon_0 \frac{d\phi_E}{dt}$$

Where ϕ_E = electric filed × area = EA, is the electric flux across the loop
$\therefore$ Total current across the closed loop

$$= I_c + I_d = I_c + \varepsilon_0 \frac{d\phi_E}{dt}$$

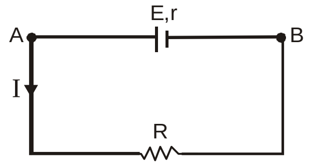
Hence the modified form of the Ampere's law is

$$\oint \vec{B}.\vec{d\ell} = \mu_0\left[I_c + \varepsilon_0 \frac{d\phi_E}{dt}\right]$$

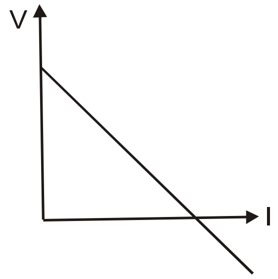
10. $I = neAV_d$

$$V_d = \frac{I}{neA} = \frac{2.7}{9\times10^{28}\times1.6\times10^{-19}\times2.5\times19^{-7}} = 7.5 \times 10^{-4} \text{ m/s}$$

11.



V = E –Ir



Interception on v-axis gives emf and slop will give internal reistance.

12. $U_1 = \frac{1}{2}CV^2$

$V' = \frac{CV + C \times 0}{C + C}$

$V' = \frac{V}{2}$

now, $U_2 = \frac{1}{2}(C + C)\left(\frac{V^2}{4}\right)$

$\frac{U_2}{U_1} = \frac{1}{2}$

13. Let the total energy of the electron be E. It is the sum of kinetic energy and potential energy.

E = kinetic energy + potential energy

$E = \frac{1}{2}mv^2 - \frac{KZe^2}{r}$

We know $mv^2 = kze^2/r$

so

$E = \frac{KZe^2}{2r} - \frac{KZe^2}{r} = -\frac{KZe^2}{2r}$

Putting the value of r, $r = \frac{n^2h^2}{4\pi^2 mKe^2}$

$E = -\frac{KZe^2}{2} \times \frac{4\pi^2 mKZe^2}{n^2h^2} = -\frac{2\pi^2 Z^2 K^2 me^4}{n^2h^2}$

For hydrogen atom, Z = 1

So, $E = -\frac{2\pi^2 K^2 me^4}{n^2h^2}$

Putting the values of π, K, m, e and h.

$E = -\frac{2 \times (3.14)^2 \times (9 \times 10^9)^2 \times (9.1 \times 10^{-31}) \times (1.6 \times 10^{-19})^4}{(6.625 \times 10^{-34})^2}$

$= -\ 21.79 \times 10^{-19}$ J per atom

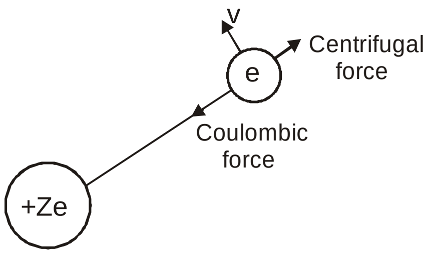
= –13.6 eV per atom (1 J = 6.2419 × 10^{18} eV)

Here negative sign show that electron required 13.6 eV to eject from hydrogen atom.

“OR”

Consider an electron of mass ‘m’ and charge ‘e’ revolving around the nucleus of charge ‘Ze’ (Z = atomic number). Let ‘v’ be the velocity of the revolving electron and ‘r’ the radius of the orbit. The electrostatic force of attraction between the nucleus and electron (applying Coulomb’s law)

$$= \frac{KZe \times e}{r^2} = \frac{KZe^2}{r^2}$$



Here K is a constant. It is equal to $\frac{1}{4\pi\varepsilon_0}$, ε_0 being absolute permittivity of medium. In SI units , the numerical value of $\frac{1}{4\pi\varepsilon_0}$ is equal to 9 × 10^9 Nm^2/c^2.

So, $\frac{KZe^2}{r^2} = \frac{mv^2}{r}$ or $v^2 = \frac{KZe^2}{mr}$

$$v^2 = \frac{1}{4\pi\varepsilon_0} \times \frac{Ze^2}{mr}$$

According to one of the postulates,

Angular momentum = mvr = n $\frac{h}{2\pi}$

or $v = \frac{nh}{2\pi mr}$

Putting the value of ‘v’

$$\frac{n^2h^2}{4\pi^2m^2r^2} = \frac{KZe^2}{mr} \text{ or } \frac{n^2h^2}{4\pi^2mr} = KZe^2$$

or $r = \frac{n^2h^2}{4\pi^2mKZe^2}$

<u>Bohr’s radius :</u>

For hydrogen atom, Z = 1; so $r = \frac{n^2h^2}{4\pi^2mKe^2}$

Now putting the values of h, π, m, e and K.

$$r = \frac{n^2 \times (6.625 \times 10^{-34})^2}{4 \times (3.14)^2 \times (9.1 \times 10^{-31}) \times (9 \times 10^9) \times (1.6 \times 10^{-19})^2}$$

$= 0.529 \times n^2 \times 10^{-10}$ m $= 0.529 \times n^2$ Å

$= 0.529 \times 10^{-8} \times n^2$ cm

Thus, radius of 1st orbit

$= 0.529 \times 10^{-8} \times 1^2 = 0.529 \times 10^{-8}$ cm $= 0.529 \times 10^{-10}$ m $= 0.529$ Å

Radius of 2nd orbit

$= 0.529 \times 10^{-8} \times 2^2 = 2.11 \times 10^{-8}$ cm $= 2.11 \times 10^{-10}$ m $= 2.11$ Å

Radius of 3rd orbit

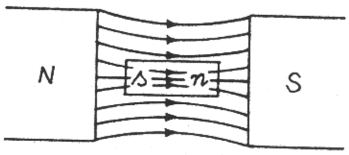
$= 0.529 \times 10^{-8} \times 3^2 = 4.76 \times 10^{-8}$ cm $= 4.76 \times 10^{-10}$ m $= 4.76$ Å

and so on

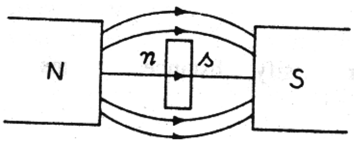
$\therefore r_n = r_1 \times n^2$ for hydrogen atom

and $r_n = 0.529 \times \frac{n^2}{Z}$ Å for hydrogen like species.

14. (a) When a rod of paramagnetic material is suspended freely between two magnetic poles, then its axis becomes parallel to the magnetic field The poles produced at the ends of the rod are opposite to the nearer magnetic poles.

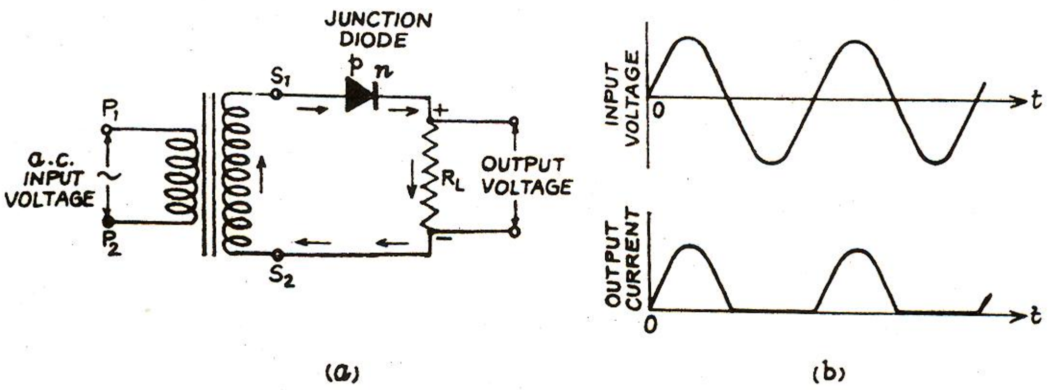


(b) When a rod of diamagnetic material is suspended freely between two magnetic poles, then its axis becomes perpendicular to the magnetic field.



15. **p-n Junction Diode as Half·wave Rectifier:** The half-wave rectifier circuit is shown in Fig. (a) and the input and output wave forms in Fig. (b). The a.c. input voltage is applied across the primary P_1P_2 of a transformer. S_1S_2 is the secondary coil of the same transformer. S_1 is connected to the p -type crystal of the junction diode and S_2 is connected to the n -type crystal through a load resistance R_L.

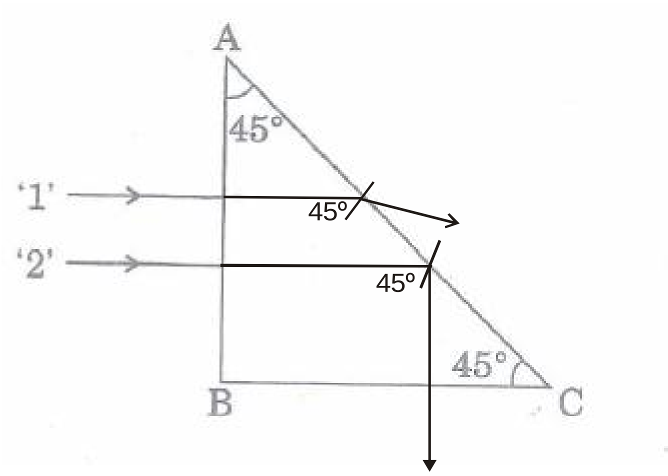
During the first half-cycle of the a.c. input, when the terminal S_1 of the secondary is suppose positive and S_2 is negative, the junction diode is forward-biased. Hence it conducts and current flows through the load R_L in the direction shown by arrows. The current produces across the load an output voltage of the same shape as the half-cycle of the input voltage. During the second half-cycle of the a.c. input, the terminal S_1 is negative and S_2 is positive. The diode is now reverse-biased. Hence there is almost zero current and zero output voltage across R_L. The process is repeated. Thus, the output current is unidirectional, but intermittent and pulsating, as shown in lower part of Fig. (b).



Since the output- current corresponds to one half of the input voltage wave, the other half being missing, the process is called half-wave rectification.

The purpose of the transformer is to supply the necessary voltage to the rectifier. If direct current at high voltage is to be obtained from the rectifier, as is necessary for power supply, then a step-up transformer is used, as shown in Fig. (a). In many solid-state equipments, however, direct current of low voltage is required. In that case, a step-down transformer is used in the rectifier.

16.



i = 45º

sin 45º = $\frac{1}{\mu} = 0.707$ ....(i)

$\sin i_{C_1} = \frac{1}{1.38} = 0.7246$ ....(ii)

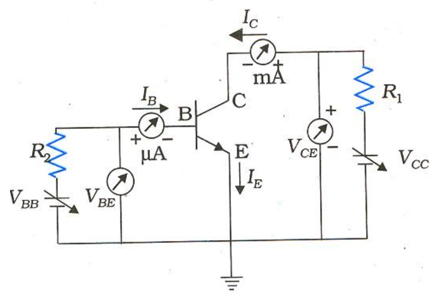
$\sin i_{C_2} = \frac{1}{1.52} = 0.637$ ....(iii)

So, ray1 will be refracted

and ray 2 will be totally reflected

17. **Common Emitter (CE) :** The transistor is most widely used in the CE configuration.

When a transistor is used in CE configuration, the input is between the base and the emitter and the output is between the collector and the emitter. The variation of the base current I_B with the base–emitter voltage V_{BE} is called the input characteristic. The output characteristics are controlled by the input characteristics. This implies that the collector current changes with the base current.



In CE configuration it used as an amplifier.

18. a demodulator, to separate the low frequency audio signal from the modulated signal.

19. **For lens :**

$$\frac{1}{v}=\frac{1}{f}+\frac{1}{u}$$

So, $\frac{1}{v}=\frac{1}{20}+\frac{1}{-40}$

V = 40 cm

Now for mirror :

u = 40– 15 = 25 cm

f = 10 cm

So, $\frac{1}{v}=\frac{1}{f}-\frac{1}{u}=\frac{1}{10}-\frac{1}{25}$

v = $\frac{50}{3}$ cm

The virtual image formed behind mirror will act as an object for the lens & again an image will be formed on front side of mirror (behind the lens)

<u>Considering virtual image (formed behind mirror) as an object for the lens :</u>

$u=-\left(15+\frac{50}{3}\right)$ cm

$=-\frac{95}{3}$ cm = –31.66 cm

$\because$ f = 20 cm

$\therefore \frac{1}{v}=\frac{1}{f}+\frac{1}{u}$

$=\frac{1}{20}-\frac{3}{95}=\frac{95.60}{1900}$

or v = 54 .28 cm

So the final image will be formed at 54.28 cm far from lens, on side of first object.

20. $\lambda = \frac{12.27}{\sqrt{50x10^3}}$ Å

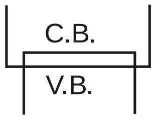
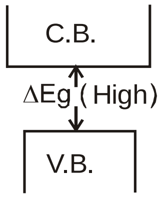
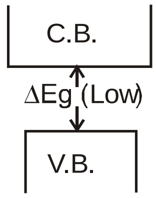
So, $\lambda = 5.5 \times 10^{-12}$ m

Resolving power of electron microscope = $\frac{1.22}{\lambda}$ Å = $\frac{1.22}{5600}$ Å

Resolving power of microscope = $\frac{1.22}{0.55 \times 10^{-12}}$ m

$$\frac{R_e}{R_0} = \frac{1.22 \times 5600}{0.55 \times 1.22} = 10181$$

21.

	Properties	**Conductor**	**Insulator**	**Semicoductor**
(i)	Band Structure			
(ii)	Energy gape E_0	zero	grater then(3eV)	Ge $\rightarrow$ 0.7 eV Si $\rightarrow$ 1.1 eV
(iii)	Position of C.B.& V.B.	C.B. & V.B are completely filled or C.B. is half empty	V.B. is completely filled & C.B. is completely empty	V.B. is litile empty & CB is little filled

22. **(i) Analog communication :** The signal which modulates the carrier signal for transmission is analog or representative of the original message or information to be transmitted. Note that the carrier signal may be sinusoidal or in the form of pulses. Only the modulating signal has to be the analog of the information.

(ii) Digital communication: In this case, the original message or information signal is first converted into discrete amplitude levels and then coded into a corresponding sequence of binary symbols 0 and 1. Subsequently, a suitable modulation method is used.

Amplitude Modulation : When a modulating AF wave superimposed on a high frequency carrier wave in a manner that the frequency of modulated wave is same as that of the carrier wave, but its amplitude is made proportional to the instantaneous amplitude of the audio frequency modulating voltage, the process is called amplitude modulation (AM).

Let the instantaneous carrier voltage (e_c) and modulating voltage (e_m) be represented by

$e_c = E_c \sin \omega_c t$ ...(1)

$e_m = E_m \sin \omega_m t$ ...(2)

Thus, in amplitude modulation, amplitude A of modulated wave is made proportional to the instantaneous modulating voltage e_m

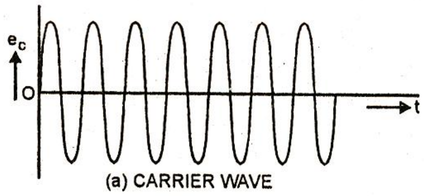
i.e. $A = E_c + ke_m$ ...(3)

where k is a constant of proportionality.

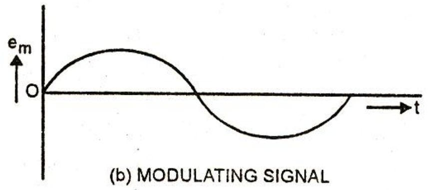
In amplitude modulation, the proportionality constant k is made equal to unity. Therefore, maximum positive amplitude of AM wave is given by

$A = E_c + e_m = E_c + E_m \sin \omega_m t$ ...(4)

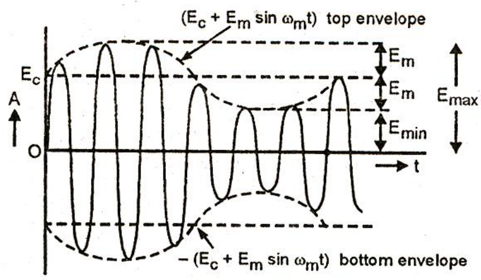
It is called top envelope.
The maximum negative amplitude of AM wave is given by
$– A = – Ec – e_m$
$= – (E_c + E_m \sin \omega_m t)$ ...(5)
This is called bottom envelope.
The modulated wave extends between these two limiting envelopes, and its frequency is equal to the unmodulated carrier frequency. Figure (a) shows the variation of voltage of carrier wave with time. Figure (b) shows one cycle of modulating sine wave and figure (c) shows amplitude modulated wave for this cycle.



(a) CARRIER WAVE



(b) MODULATING SIGNAL



(c) AMPLITUDE MODULATED WAVE

In amplitude modulation, the degree of modulation is defined by a term, called modulation index or modulation factor or depth of modulation represented by m_a . It is equal to the ratio of amplitude of modulating signal to the amplitude of carrier wave i.e.

$$m_a = \frac{E_m}{E_c} = \frac{E_{max} - E_{min}}{E_{max} + E_{min}}.$$

23. **(a)** They acts as connecting point with no resistance
(b) By obtaining balance point in the middle of bridge wire, percentage error in resistance can be minimized.
(c) magnanin, Constantan

"OR"

$$I = \frac{V}{\frac{\frac{R_0}{2} \times R}{\left(\frac{R_0}{2} + R\right)} + \frac{R_0}{2}}$$

$$I = \frac{V}{\frac{R \times R_0}{R_0 + 2R} + \frac{R_0}{2}}$$

Now, potential drop across R

$$V_R = I\left(\frac{\frac{R_0 \times R}{2}}{\frac{R_0}{2} + R}\right)$$

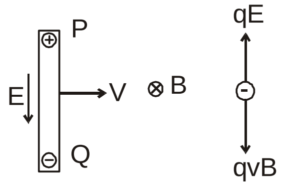
On solving $V_R = \frac{2VR}{4R + R_0}$

24. **(a)** She is curious, loving and caring about her family members.

(b) CO-60 is used in the treament of cancer it damege the cancer cell by γ–radiation

25. **(a)** If a rod is moving with velocity v in a magnetic field B ,as shown , the free electrons in a rod will experience a magnetic force in downward direction and hence free electrons will accumulate at the lower end and there will be a deficiency of free electrons and hence a surplus of positive charge at the upper end. These charges at the ends will produce an electric field in downward direction which will exert an upward force on electron. If the rod has been moving for quite some time enough charges will accumulate at the ends so that the two forces qE and qvB will balance each other. Thus E = v B.

$V_P - V_Q = V B \ \ell$



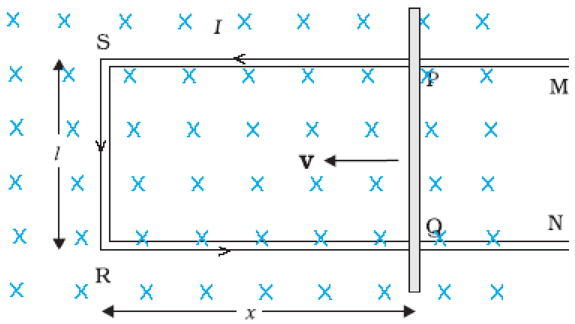
(b) Let us consider a straight conductor moving in a uniform and timeindependent magnetic field. Figure shows a rectangular conductor PQRS in which the conductor PQ is free to move. The rod PQ is moved towards the left with a constant velocity, V as shown in the figure. Assume that there is no loss of energy due to friction. PQRS forms a closed circuit enclosing an area that changes as PQ moves. It is placed in a uniform magnetic field B which is perpendicular to the plane of this system. If the length RQ = x and RS = ℓ , the magnetic flux Φ_B enclosed by the loop PQRS will be

$\Phi_B = B\ell x$

Since x is changing with time, the rate of flux Φ_B will induce an emf given by:

$$\varepsilon = \frac{-d\phi_B}{dt} = -\frac{d}{dt}(B\ell x) \qquad .....(i)$$

$$-B\ell\frac{dx}{dt} = B\ell v$$



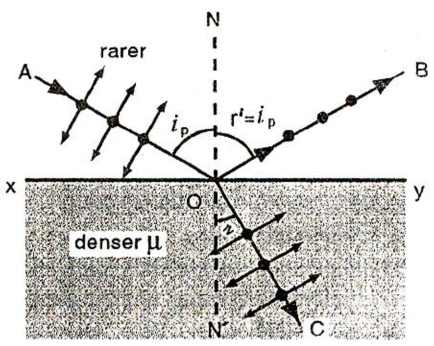
where we have used dx/dt = – v which is the speed of the conductor PQ. The induced emf $B\ell v$ is called motional emf motional emf. Thus, we are able to produce induced emf by moving a conductor instead of varying the magnetic field, that is, by changing the magnetic flux enclosed by the circuit.

It is also possible to explain the motional emf expression in Equation by invoking the Lorentz force acting on the free charge carriers of conductor PQ. Consider any arbitrary charge, q in the conductor PQ. When the rod moves with speed, v the charge will also be moving with speed, v in the magnetic field, B . The Lorentz force on this charge is qvB in magnitude, and its direction is towards Q. All charges experience the same force, in magnitude and direction, irrespective of their position in the rod PQ.

The work done in moving the charge from P to Q is,
$W = qvB\ell$
Since emf is the work done per unit charge,

$$\varepsilon = \frac{W}{q} = B\ell v$$

26. **(a)** When unpolarised light is reflected from a surface, the reflected light may be completely polarised, partially polarised or unpolarised. This would depend on the angle of incidence.
If angle of incidence is 0° or 90°, the reflected beam remains unpolarised. For angles of incidence between 0° and 90°, the reflected beam is polarised to varying degree.
The angle of incidence at which the reflected light is completely plane polarised is called polarising angle or Brewster's angle. It is represented by i_p. The value of i_p depends on the wavelength of light used. Therefore, complete polarisation is possible only for monochromatic light. In figure, unpolarised light is incident along AO at $\angle i_p$ on the interface XY separating air from a medium of refractive index μ. The light reflected along OB is completely plane polarised. The light refracted along OC continues to be unpolarised.



(b) $I = I_0 \cos^2\theta$
$I = I_0^2 \cos^2 60^\circ$

$$= I_0 \left(\frac{1}{2}\right)^2$$

So intensity from P_3 is $\frac{I_0}{4}$

and from P_2

$$I = I_0 \cos^2\theta = \frac{I_0}{4} \cos^2 30^\circ$$

$$= \frac{I_0}{4} \times \left(\frac{\sqrt{3}}{2}\right)^2 = \frac{3I_0}{16}$$

27. Consider an electrical device which may be a source, a capacitor, a resistor, an inductor or any combination of these. Let the potential difference be $v = V_A - V_B = V_m \sin\omega t$. Let the current through it be $i = I \sin(\omega t + \phi)$. Instantaneous power P consumed by the device $= v\,i = (V_m \sin \omega t)\,(I_m \sin(\omega t + \phi))$

$$\text{Average power consumed in a cycle} = \frac{\int_0^{2\pi/\omega} P\,dt}{\frac{2\pi}{\omega}} = \frac{1}{2} V_m I_m \cos\phi$$

$$= \frac{V_m}{\sqrt{2}} \cdot \frac{I_m}{\sqrt{2}} \cdot \cos\phi = V_{rms}\, I_{rms} \cos\phi.$$

Here cos ϕ is called **power factor**.

If a pure resistor is connected in the ac circuit then

$\phi = 0$, $\cos\phi = 1$

$\therefore \quad P_{av} = \frac{E_0 I_0}{2} = \frac{E_0^2}{2R} = E_{rms} I_{rms}$

Thus the power loss is maximum and electrical energy is converted in the form of heat.

If a pure inductor or a capacitor are connected in the ac circuit, then

$\phi = \pm 90^\circ$, $\cos\phi = 0$

$\therefore \quad P_{av} = 0$ (minimum)

Thus there is no loss of power.

28. **(a)** $\beta = \frac{\lambda D}{d}$

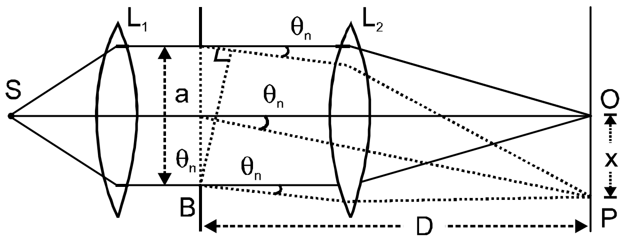
(Complete derivation reference NCERT-XII-Physics volume-2, page no. 363-64)

(b) $\frac{I_{min}}{I_{max}} = \frac{9}{25} = \frac{(d_1 - d_2)^2}{(d_1 + d_2)^2}$

$= \frac{d_1}{d_2} = \frac{|3-5|}{|3+5|} = \frac{1}{4}$

"OR"

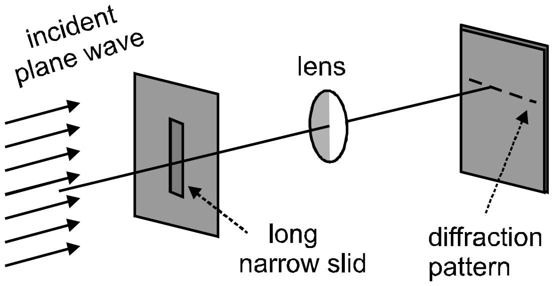
(a) Ab is single slit of width a, Plane wavefront is incident on a slit AB. Secondary wavelets coming from every part of AB reach the axial point P in same phase forming the central maxima. The intensity of central maxima is maximum in this diffrection. Where θ_n represents direction of n[th] minima Path difference BB' =a sin $\theta_{n\backslash}$



for n[th] minima a sin $\theta_n = n\lambda$ $\quad\therefore\quad$ $\sin\theta_n \approx \theta_n = \frac{n\lambda}{a}$ (if θ_n is small)

When path difference between the secondary wavelets coming from A and B is $n\lambda$ or $2n\left[\frac{\lambda}{2}\right]$

or even multiple of $\frac{\lambda}{2}$ then minima occurs



For minima $a \sin \theta_n = 2n\left[\frac{\lambda}{2}\right]$

When path difference between the secondary wavelets coming from A and B is $(2n + 1)\frac{\lambda}{2}$

or odd multiple of $\frac{\lambda}{2}$ then maxima occurs

For maxima $a \sin \theta_n = (2n + 1)\frac{\lambda}{2}$ where n = 1, 2, 3 ......

$n = 1 \rightarrow$ first maxima and $n = 2 \rightarrow$ second maxima.

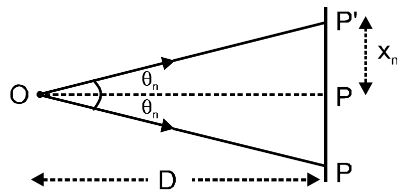
In alternate order minima and maxima occurs on both sides of central maxima.

For nth minima

If distance of nth minima from central maxima = x_n

distance of slit from screen = D, width of silt = a

Path difference $\delta = a \sin\theta_n = \frac{2n\lambda}{2} \Rightarrow \sin\theta_n = \frac{n\lambda}{a}$



In Δ POP' $\tan \theta_n = \frac{x_n}{D}$ If θ_n is small $\Rightarrow \sin \theta_n \approx \tan\theta_n \approx \theta_n$

$X_n = \frac{n\lambda D}{a} \Rightarrow \theta_n = \frac{x_n}{D} = \frac{n\lambda}{a}$ First minima occurs both sides on central maxima.

For first minima $x = \frac{D\lambda}{a}$ and $\theta = \frac{x}{D} = \frac{\lambda}{a}$

Linear width of central maxima $w_x = 2x \Rightarrow w_x = \frac{2d\lambda}{a}$

Angular width of central maxima $w_\theta = 2\theta = \frac{2\lambda}{a}$

(b) $d\sin\theta = \frac{3\lambda}{2}$

$d\frac{x_1}{D} = \frac{3\lambda}{2}$

$x_1 = \frac{3\lambda_1 D}{2}$ ....(i)

$x_2 = \frac{3\lambda_2 D}{2}$ ....(ii)

$x_2 - x_1$

$\frac{3\times 1.5}{2\times 2\times 10^{-6}}(596-590)10^{-9}$

$= 6.75 \times 10^{-3}$ m

29. **(a)**

$qvB = mv^2/R$

$v = \frac{qBR}{m}$

So, time period $T = \frac{2\pi R}{V}$

$= \frac{2\pi R\times m}{qBR}$

$T = \frac{1}{\nu_c} = \frac{2\pi m}{qB}$

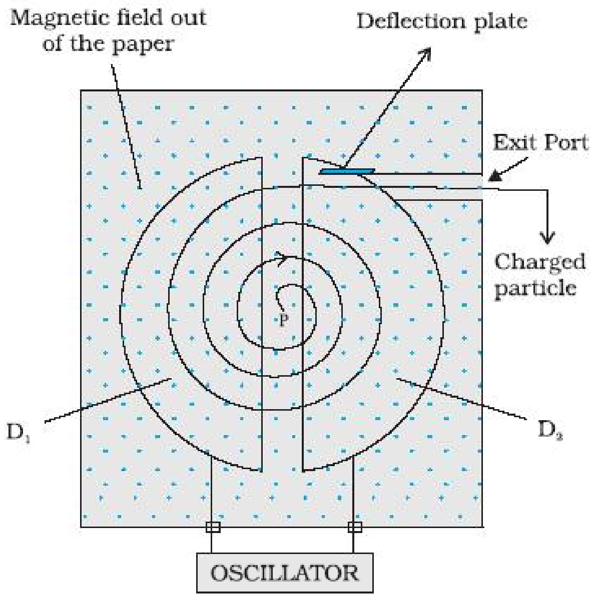
or $\nu_c = \frac{qB}{2\pi m}$

(b) Cyclotron : The cyclotron uses both electric and magnetic fields in combination to increase the energy of charged particles. As the fields are perpendicular to each other they are called *crossed fields*. Cyclotron uses the fact that the frequency of revolution of the charged particle in a magnetic field is independent of its energy. The particles move most of the time inside two semicircular disc-like metal containers, D_1 and D_2, which are called *dees* as they look like the letter D. Figure shows a schematic view of the cyclotron. Inside the metal boxes the particle is shielded and is not acted on by the electric field. The magnetic field, however, acts on the particle and makes it go round in a circular path inside a dee. Every time the particle moves from one dee to another it is acted upon by the electric field. The sign of the electric field is changed alternately in tune with the circular motion of the particle. This ensures that the particle is always accelerated by the electric field. Each time the acceleration increases the energy of the particle.

The whole assembly is evacuated to minimise collisions between the ions and the air molecules. A high frequency alternating voltage is applied to the dees. In the sketch shown in figure, positive ions or positively charged particles (e.g., protons) are released at the centre P. They move in a semi-circular path in one of the dees and arrive in the gap between the dees in a time interval *T*/2;

The phase of the supply is adjusted so that when the positive ions arrive at the edge of D_1, D_2 is at a lower potential and the ions are accelerated across the gap. Inside the dees the particles travel in a region free of the electric field. The increase in their kinetic energy is *qV* each time they cross from one dee to another (*V* refers to the voltage across the dees at that time). From Eq. , it is clear that the radius of their path goes on increasing each time their kinetic energy increases. The ions are repeatedly accelerated across the dees until they have the required energy to have a radius approximately that of the dees. They are then deflected by a magnetic field and leave the system via an exit slit.

$$v = \frac{qBR}{m}$$



where *R* is the radius of the trajectory at exit, and equals the radius of a dee.
Hence, the kinetic energy of the ions is,

$$\frac{1}{2}mv^2 = \frac{q^2B^2R^2}{2m}$$

The operation of the cyclotron is based on the fact that the time for one revolution of an ion is independent of its speed or radius of its orbit. The cyclotron is used to bombard nuclei with energetic particles, so accelerated by it, and study the resulting nuclear reactions. It is also used to implant ions into solids and modify their properties or even synthesise new materials. It is used in hospitals to produce radioactive substances which can be used in diagnosis and treatment.

"OR"

(a) draw the labelled diagram principle and working of moving coil galvanometer.
(b)(i) Soft iron core is inserted due to which eddy current are produced due to which the vibrations are stoped
(ii) deflection per unit current is known as current sensitivity of galvanometer. It is given by the following formula :
NBA/C so it does not depend on potential.

30. **VAN DE GRAFF GENERATOR :** This is a machine that can build up high voltages of the order of a few million volts. The resulting large electric fields are used to accelerate charged particles (electrons, protons, ions) to high energies needed for experiments to probe the small scale structure of matter.
Principle :
Suppose we have a large spherical conducting shell of radius R, on which we place a charge Q. This charge spreads itself uniformly all over the sphere. The field outside the sphere is just that of a point charge Q at the centre while the field inside the sphere vanishes. So the potential outside is that of a point charge and inside it is constant (at the radius R) . We thus have–

Potential inside conducting spherical shell of radius R carrying charge Q–

$$= \frac{1}{4\pi\varepsilon_0}\frac{Q}{R}$$

Let us suppose that in some way we introduce a small sphere of radius r, carrying some charge q, into the large one, and place it at the centre. The potential due to this new charge clearly has the following values at the radii indicated:
Potential due to small sphere of radius r carrying charge q at surface of small sphere is,

$$= \frac{1}{4\pi\varepsilon_0}\frac{q}{r} \text{ at surface of small sphere}$$

Potential due to small sphere of radius r carrying charge q at large shell of radius R is,

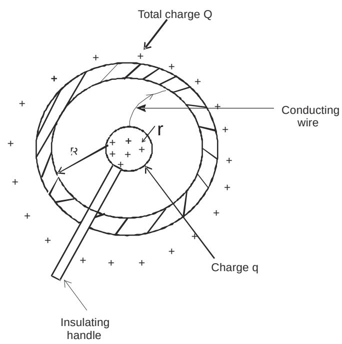
$$= \frac{1}{4\pi\varepsilon_0}\frac{q}{R} \text{ at large shell of radius R}$$

Taking both charges q and Q into account we have for the total potential V and the potential difference.
Potential on the surface (at R)

$$V(R) = \frac{1}{4\pi\varepsilon_0}\left(\frac{Q}{R}+\frac{q}{R}\right)$$

Potential on the surface r

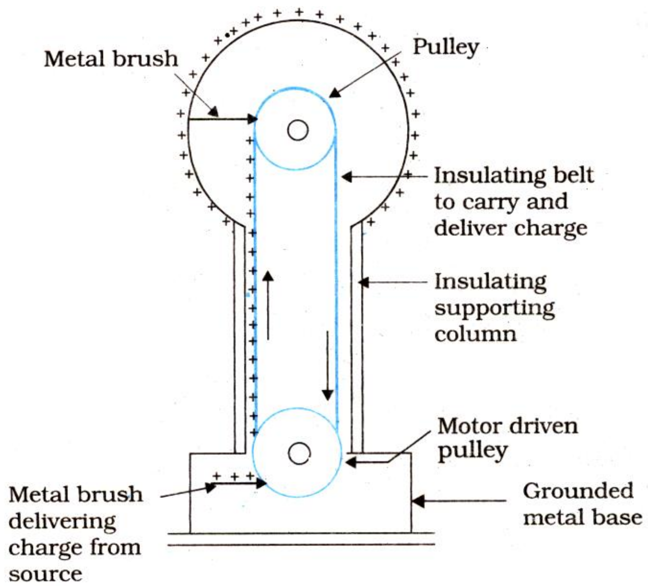
$$V(r) = \frac{1}{4\pi\varepsilon_0} = \left(\frac{Q}{R}+\frac{q}{r}\right)$$



$$\text{Potential difference} = V(r) - V(R) = \frac{q}{4\pi\varepsilon_0}\left(\frac{Q}{R}+\frac{q}{r}\right)$$

$$V(r) - V(R) = \frac{q}{4\pi\varepsilon_0}\left(\frac{1}{r}-\frac{1}{R}\right)$$

Assume now that q is positive. We see that, independent of the amount of charge Q that may have accumulated on the larger sphere and even if it is positive, the inner sphere is always at a higher potential and the difference V(r) – V(R) is positive. The potential due to Q is constant upto radius R and so cancels out in the difference
This means that if we now connect the smaller and larger sphere by a wire, the charge q on the former will immediately flow onto the matter, even though the charge Q may be quite large. The natural tendency is for positive charge to move from higher to lower potential. Thus, provided we are somehow able to
introduce the small charged sphere into the larger one, we can in this way keep piling up larger and larger amount of charge on the latter.



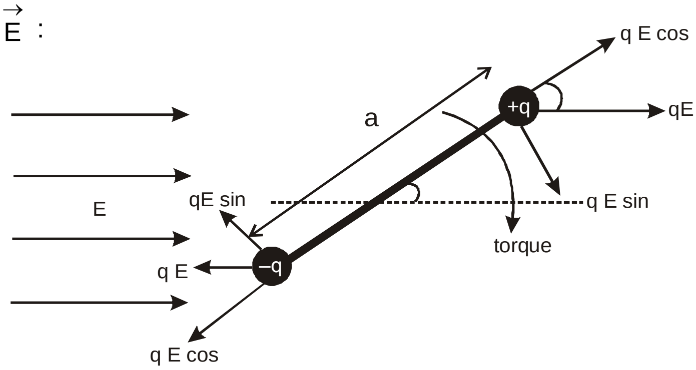
Construction and working : It is a machine capable of building up potential difference of a few million volts, and fields close to the breakdown field of air which is about 3×10^6 V/m. A schematic diagram of the van de Graff generator is given in figure. A large spherical conducting shell (of few metres radius) is supported at a height several meters above the ground on an insulating column. A long narrow endless belt insulating material like rubber or silk, is wound around two pulleys- one at ground level, one at the centre of the shell. This belt is kept continuously moving by a motor driving the lower pulley. It continuously carries positive charge, sprayed on to it by a brush at ground level, to the top. There it transfers its positive charge to another conducting brush connected to the large shell. Thus positive charge is transferred to the shell, where it spreads out uniformly on the outer surface. In this way voltage differences of as much as 6 or 8 million volts (with respect to ground) can be built up.

Uses : The high potential difference set up in a Vande Graaff generator is used to accelerate charged particles like protons, deutorns, α- particles, etc, to hight energies of about 10 MeV, needed for experiments to probe the small scale structure of matter.

Limitation : Potenetial on the outer matellic sphere will be not greater then 10^6 Volt. If it will be greater charge disperssion will be started through ionisation of atmospheric air.

"OR"

(a) If the dipole is placed at θ angle from $\vec{E}$:



In this case $F_{net} = 0$ but

Net torque $\tau = (qEa\sin\theta)$

Here $qa = P \Rightarrow \tau = PE\sin\theta$

in vector form $\vec{\tau} = \vec{P} \times \vec{E}$

(b) (i) $\phi_1 = \dfrac{2Q}{\varepsilon_0}$

$\phi_2 = \dfrac{6Q}{\varepsilon_0}$

$\dfrac{\phi_1}{\phi_2} = \dfrac{1}{3}$

(ii) Dielectric substance is filled inside sphere S_1, so there is no change in electric flux through S_1.