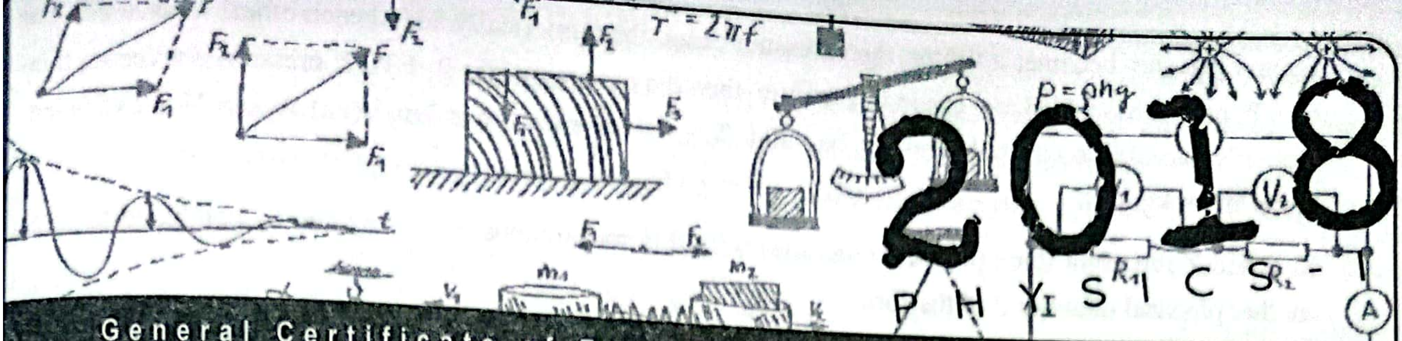


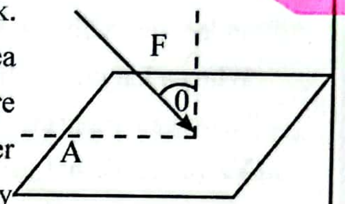
General Certificate of Education (Adv. Level) Examination

1. Unit of pressure of.

(1) $kg\,ms^{-2}$ (2) $kg\,m^2s^{-2}$ (3) $kg\,m^{-1}s^{-2}$ (4) $kg\,m^2s^{-3}$ (5) $kg\,m^{-2}s^{-2}A^{-1}$

Unit and Dimentions

Does pressure is a vector or a scalar? This is a question that most of the students ask. Pressure $P = F/A$; F = force and A = area. F force is a vector. If we consider the area as a scalar quantity, then there is a fair argument that questions how can pressure be a scalar quantity? As I mentioned in the review of 2017, it is wrong to consider the area as a scalar quantity. Even the division of two vectors is not interpreted by mathematicians. I think that pressure should be interpreted as the division of force that perpendicularly acts on a certain area by that area. However, a force that acts along an area cannot create a pressure on an area.

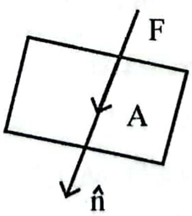


If there is an inclined force F that acts on the area of A, then the pressure is only felt by the area from the perpendicular component of F to the area. That means $P = F\cos\theta/A$

From the component of $F\sin\theta$ there is no pressure on the area. $F\sin\theta$ acts along the surface. If we interpret as below, then we can get over from this question.

Pressure = The force that acts perpendicularly to a certain surface/ the area of that surface

The figure has shown a perpendicular force F that acts on a particular surface A. Now $P = F/A$; As we need to consider the area as a vector . is vector· $\vec{A} = A\hat{n}$. $\vec{A}$ is the numerical value of the area. is the unit vector that acts along the direction that F acts perpendicularly to the area. The children who do Mathematics know these things very well. As there is no interpretation of division of two vectors, I will transform $P = F/A$ into a multiplication. Then $PA = F$



Now A and F vector are acting to one side. They are acting along the same direction. They are two vectors that act towards the same direction. If so, then definitely P cannot be a vector. P should be a scalar. If we write the expression considering the vectors, then $P(A\hat{n}) = (F\hat{n})$

From this you can see that pressure P should be a scalar. When a vector is multiplied by a scalar, the direction of the vector does not change.

Even if pressure becomes a vector, then pressure cannot be numerically added each other. When a gas that is in P_1 pressure is added to a gas in P_2 pressure, then the net pressure is $P_1 + P_2$. If pressure is a vector, then Dolton's partial pressure will also not be valid. The unit of pressure is Nm^{-2} (Pa). When N is expanded, $kgms^{-2}m^{-2} = kgm^{-1}s^{-2}$

02. X, F and Z represent three physical quantities with different dimensions. They can be combined to form another physical quantity P of the form,

$$P -- AX + BY + CZ$$

Which of the following expressions has different dimensions from the rest?

(1) AX (2) $AX - CZ$ (3) $\frac{(AX)(CZ)}{BY}$ (4) $\frac{(BY)^2}{P}$ (5) $(BY)(CZ)$

Unit and Dimentions

Consider a physical equation which is given by P = X + Y + Z. To be a valid equation, then the dimensions of X, Y, Z and P should be same. Consider that three lengths [L] are being added. When the lengths are added, you will get another length. The dimensions of (X-Y) or (X-Z) also have the dimensions of length. When a length is subtracted by another, you will get another length. The dimensions of XZ/Y or Y^2/X also have the dimensions of length. $[L][L]/[L] = [L]$ and $[L]^2/[L] = [L]$. But the dimensions of the multiples of YZ or XY do not have the dimensions of length. $[L][L] = [L]^2$. Therefore, the dimensions of the multiple of two quantities will not be equal to the dimension of each quantity. All you have consider is this point.

When X, Y and Z are multiplied by constants with dimension, the argument is valid as before. The dimensions of X, Y and Z are different from each other and they are being multiplied by three quantities of a, b and c respectively. If aX, bY and cZ can be added together, then the dimensions of aX, bY and cZ quantities should be same. Otherwise, they cannot be added together. The above fair argument can be applied even it is like this way. The addition, subtraction of two quantities, multiply two and divide from one, square a quantity and divide by one, all have the same dimension. But when two quantities are multiplied, you will get the square of the dimension. Look at this question.

P = X + Y + Z represents a valid physical equation. Out of the following, which one has the different dimensions?

1) X 2) X-Z 3) XZ/Y 4) Y^2/P 5) YZ

The answer is YZ.

03. Which of the following statements is not true?

(1) LASER light consists of transverse waves.

(2) Gamma rays are transverse waves.

(3) Primary waves (P-waves) travelling through the crust of the Earth are longitudinal waves.

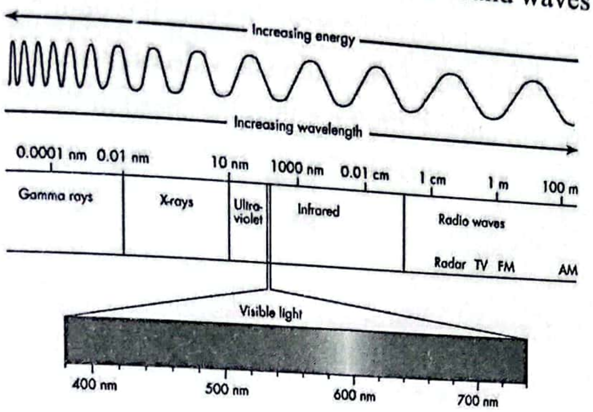
(4) Ultrasound waves are longitudinal waves.

(5) FM waves are longitudinal waves.

Electromagnetic Waves 03

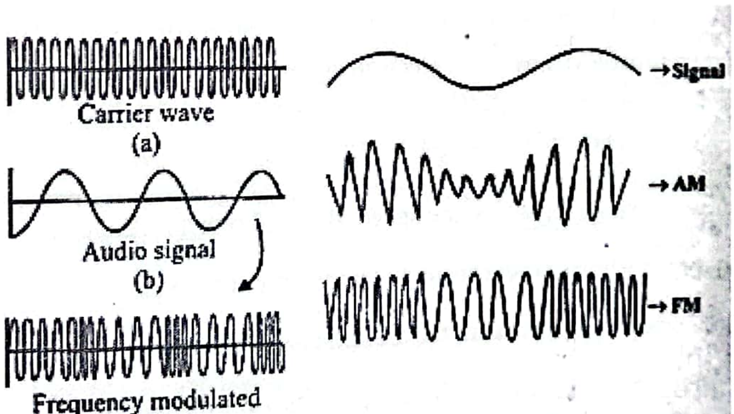
We know that electromagnetic waves are transverse waves. From radio waves to Gamma waves, the electromganetci spectrum as shown below is made totally from transverse waves.

Even laser light is characteristically changed light. Sound waves are longitudinal. Ultra-sound waves are sound waves which have high frequencies exceeding our hearing limit. But these are also sound waves. Therefore, ultra-sound waves are alos longitudinal waves. Actually, we do not hear transverse waves. When a string is vibrated in a guitar, we hear it according to the sound waves created in the air.



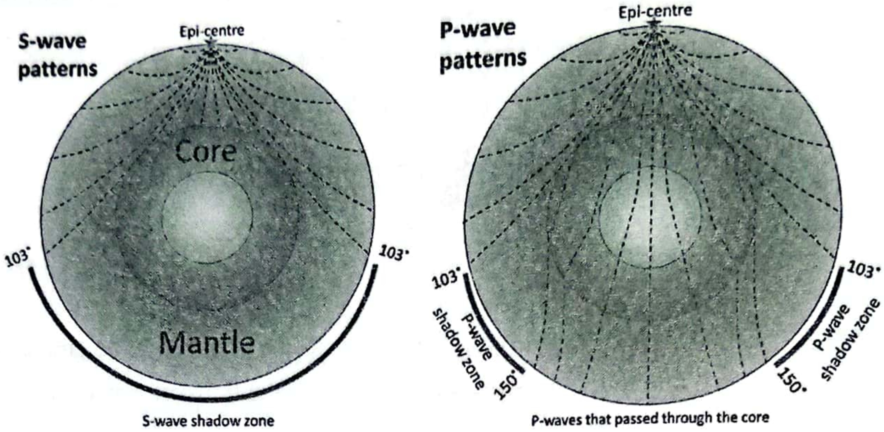
If our ear gets sensitive to transverse waves, then the structure of it should be changed more than this way. The ear drum is vibrated due to the pressure variations from the transverse waves.

FM waves are type of a radio wave. What are FM waves (Frequency Modulated waves)? Even this is not in the syllabus, I will present it in a simple way for general understanding. Audible waves (voice), music cannot be just broadcast. The audible waves should be mixed with a radio wave with a higher frequency. If we need to go a journey that is far away, then we will get into the vehicle. The vehicle takes us. Later, once we arrived the destination, we get off from the vehicle. Wave broadcasting is also like this way.



The wave consisting the audio and video information with low frequency (this is called the base wave) and the radio wave with higher frequency (this is known as the carrier wave) are shown in the figure. The necessary information is mixed and taken far away from this carrier wave. There are two commonly used methods to superposition these two waves. One method is, varying the amplitude of the carrier wave according to the variation of the audio and video wave. This method is called amplitude modulation (AM). The other method is varying the frequency of the carrier wave without varying the amplitude. This is known as frequency modulation (FM). How the carrier waves are displayed is shown in the last figure. It has the information of the base wave.

If the carrier wave frequency is f_c you will see that, when the base wave goes to its maximum positive displacement, then the frequency of the modulated wave is higher ($>f_c$) whereas when the base wave goes to its maximum negative displacement, then the frequency of the modulated wave is lower ($<f_c$). When the displacement of the base wave is zero, then the frequency of the modulated wave is exactly equal to the frequency of the carrier frequency ($=f_c$). Even the base wave is not a beautiful sinusoidal wave as shown in the figure, due to its variations, the frequency of the corresponding carrier wave varies. Like we get out from the vehicle, when the FM wave is captured at the receiving end, it is de-modulated and the mixed based wave can be separated from the carrier wave. If the audio and video signals are not attached this way to a radio wave and broadcast, then we will not get information at the speed of light.



You have learnt back in 2015 that P waves created in an earthquake are longitudinal. There is an explanatory discussion in the review of 2015. P waves are longitudinal whereas S waves are transverse. P waves can propagate across both solids and liquids but S waves cannot propagate through liquids.

Even transverse waves can propagate across a liquid surface, they cannot propagate through a liquid. Many said that I have explained beautifully about how this is happened. Even if you forget about P waves, you clearly know as a fact that FM waves are electromagnetic waves and electromagnetic waves are transverse. It can be seen quickly that FM waves cannot be longitudinal. How many FM channels are there in Sri Lanka?

04. Consider the following stage regarding the speed of sound o in an ideal gas.

(A) υ is directly prosportional to the absolute temperature of the gas.

(B) υ is inversely prosportional onal to the molar mass of the gas.

(C) υ depends on the ratio of the molar heat capacities y for the gas.

Of the above statements,

(1) only A is true. (2) only C is true. (3) only A and B are true.
(4) only B and C are true. (5) all A, B and C are true.

Velocity of Sounds 03

We know that the speed of sound in an ideal gasv is $\sqrt{\frac{\gamma RT}{M}}$. Speed v is proportional to the square of the absolute temperature $v \propto \sqrt{T}$. Likewise,$v \propto 1/\sqrt{M}$. Speed v is inversely proportional to the square root of the molar mass of the gas. But we cannot say that the speed of sound is proportional to T. Likewise, we cannot say that v is inversely proportional to M. We can say that v is dependent upon T, M and γ. Based on T, M and γ, v changes. It is true. But when we talk of proportionality or inverse proportionality you really need to consider the parameters of the equation as it is.

05. **Which of the following statements made regarding optical instruments under normal adjustment is not true?**

(1) **In a simple microscope, the image of the object is virtual.**

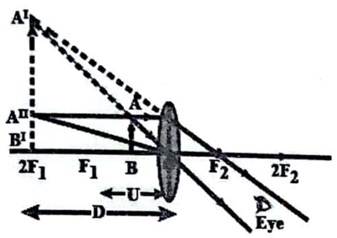
(2) **When reading small letters using a simple microscope, a short-sighted person has an advantage over a long-sighted person.**

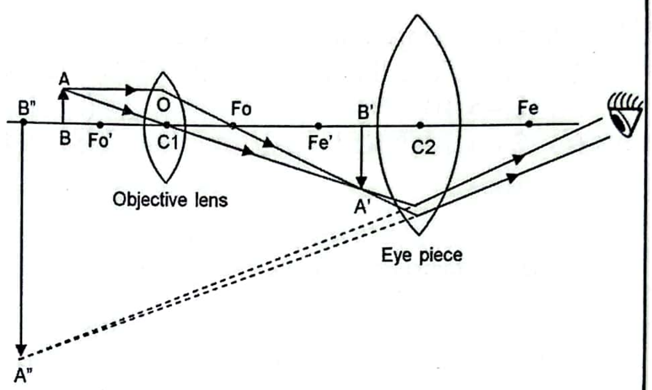
(3) **In a compound microscope, the eyepiece acts as a simple microscope.**

(4) **In a compound microscope, the final image is inverted.**

(5) **In an astronomical telescope, the object distance and the image distance are both considered to be very large**

There were many statements regarding simple microscope, combined microscope and astronomical telescope in the previous papers. The final image of a combined telescope is unreal and inverted. The final image of a simple microscope is unreal and non-inverted. In a microscope, the object is placed in between the optical centre of the lens and the focus.





The image of the objective in the combined telescope is placed in between the focus of the eyepiece and its optical centre. Therefore, the lens of the eyepiece is working as a simple microscope. Nearsightedness (myopia) means that you cannot see long distances. A simple microscope is used to enlarge something small nearby. Therefore, inability to see the long distances does not impact on to see nearby small objects by enlarging.

Farsightedness (hyperopia) means that you cannot see near distances. In such a person the least distance of distinct vision is greater than 25 cm. Therefore, when such a person is using a simple microscope, if s/he wants to view the enlarged non-inverted created image, then s/he has to make the image in his/her relevant least distance of distinct vision. The magnification of a simple microscope is given by (1 + D/f). As D is greater than 25 cm in a farsighted person, greater magnification can be obtained when such a person is using the simple microscope. In the usage of simple microscope, farsighted person has a certain advantage than a person of nearsightedness. But the reciprocal of this situation is not true. The objects that are viewed from the astronomical telescope are however at far away distances. Astronomical objects are viewed from it. When the astronomical telescope is **only in normal adjustment**, the image distance is very far way (at infinity).

06. In a certain thermodynamic process in which an ideal gas is used, the increase of the internal energy of the gas is equal to the heat supplied to the gas. This process is

(1) a cyclic process. (2) an adiabatic process. (3) a constant pressure process.
(4) a constant volume process. (5) an isothermal process

Thermodynamics 04

In an ideal gas, the internal energy change (ΔU) is equal to the exchanging heat amount (ΔQ). Under which process does this happen? To be like this, the work done by the gas or the work done on the gas (ΔW) should be zero.

$\Delta U = \Delta Q - \Delta W$; If $\Delta U = \Delta Q$, then $\Delta W = 0$. So, this should be an isochoric (constant volume) process. If the volume is constant, then $\Delta V = 0$. There is no work from PV.

If the process is adiabatic, then $\Delta Q = 0$. If it is a cyclic process, then $\Delta U = 0$. The internal energy of an ideal gas is dependent upon its temperature only. Therefore, in an isothermal process also $\Delta U = 0$. If the process is isobaric (constant pressure), ΔW can be calculated by $P\Delta V$.

7. When the temperature of a metal rod is increased by 100 °C, its fractional change in length is 2.4×10^{-5}. The linear expansivity of the material of the rod is

(1) 2.4×10^{-3} °C^{-1} (2) 2.4×10^{-4} °C^{-1} (3) 2.4×10^{-5} °C^{-1}
(4) 2.4×10^{-6} °C^{-1} (5) 2.4×10^{-7} °C^{-1}

Expansion of Solids 04

Linear expansivity = fractional change of length/ temperature difference. $l = l_0(1 + \alpha\theta) \rightarrow$

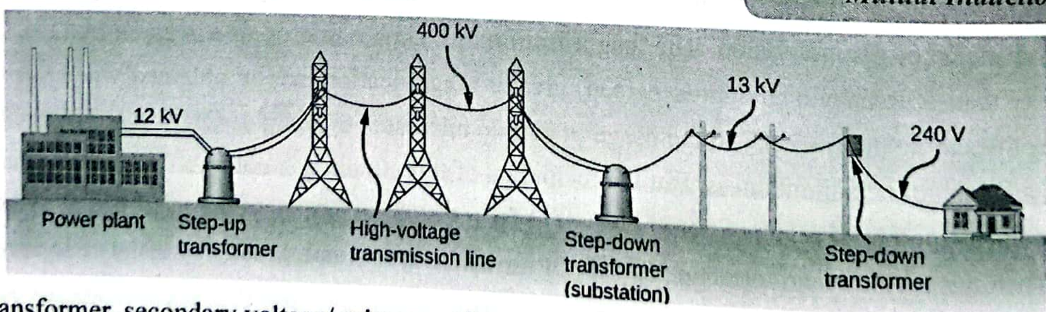
$$\alpha = \frac{(l - l_0)}{l_0\theta}$$

(Look at the 10th question of paper 1998) If the fractional change is 2.5 X 10^{-5} and the temperature difference is 100°C, then the linear expansivity is 2.5 X 10^{-7}°C^{-1}. Do it from your memory. Do not try to do rough work. Cannot you divide a number by 100? Only the power of 10 is changed.

08. A certain transformer has 900 turns in the primary coil and 30 turns in the secondary coil. When 240 V alternating voltage is applied across the primary coil, the voltage across the secondary coil is

(1) 0V (2) 8V (3) 12V
(4) 72V (5) 7.2kV

Mutual Induction 08



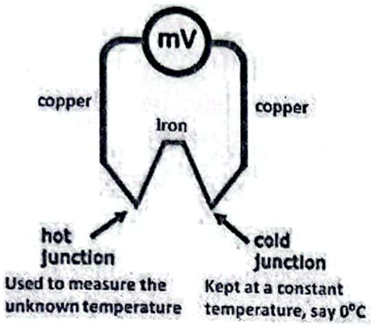
For a transformer, secondary voltage/ primary voltage = number of turns in the secondary/ number of turns in the primary. This relation has been checked many times. If the voltage of the fundamental power to our homes is 240 V and it should be reduced to 8V for the activity of a certain device, then the number of turns in the secondary should be 30 times reduced compared to the primary. Do these by your memory. Do you need rough work to divide 240 by 30?

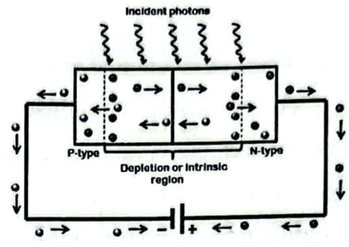
09. Which of the following is not a source of e.m.f.?

(1) Electrochemical cell (2) Photodiode (3) Rezoelectric crystal

(4) Thermocouple (5) Charged capacitor

A certain device is needed to provide an e. m. f to a closed circuit with a continuous current. Such a device is called an e. m. f source. Examples for e. m. f sources are batteries (cells), dynamos, solar cells, thermo-electric couple etc. All these devices use a particular energy (mechanical, chemical, thermal, optical) where it is converted into electric potential energy and provide it to the circuit where the device is connected.





PIN photodiode

A water fountain allows water to go up and allows the same water to fall down again. Same water is again risen. If the water is not spilled to the outside, then the water flow is conserved. As such a water fountain needs a water pump, an e. m. f source is needed to have a continuous current flow in an electric circuit. An e. m. f can be obtained from a charged capacitor. It is true. But it can be done for a smaller period of time. When the charge is discharged, it should be recharged. Again, the charges have to be pumped. Therefore, a charged capacitor is not considered as an e. m. f **source**. Here the word 'source' is very important. An e. m. f can be obtained from a capacitor but it is not an e. m. f source. It cannot continuously retain an electric potential across the circuit.

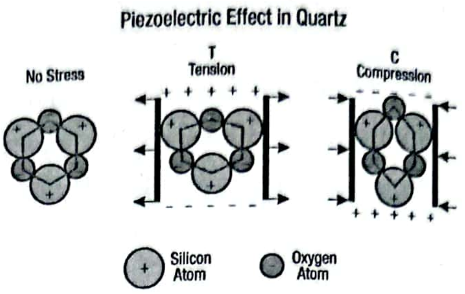
A well with a good water spring is a water source. It can keep water continuously to a certain extent. But a water storing tank cannot be considered as a water **source**. Once water is taken from it, you need to refill water from another source here.

In a chemical cell, chemical energy is turned into electrical energy. In a photo diode, the light energy is converted into electricity whereas in a thermo-electric couple, the thermal energy is converted into electrical energy and a piezoelectric crystal converts a pressure variation into electrical energy. But a charged capacitor cannot be expressed like this way. A charged capacitor has stored electric energy too. It is true that this energy should be taken from another source. But we tend to get electricity from the stored electric energy of the capacitor.

Even students may have heard about the other devices, you may not have heard about piezoelectric crystals. But all these devices are given as e. m. f sources in the teachers' guide. If you have not heard about piezoelectric crystals, then you will be tempted to pick that choice.

When a piezoelectric crystal is pressed or applied with a mechanical stress, then the generation of an electrical potential across its ends where the stress was applied is known as the piezoelectric effect. If it is expressed in a simple way, then it means that one surface of the crystal is charged as positive whereas the other surface is charged as negative and acts as a small cell. Quartz crystals behave in this manner. Even it is not needed to know how does it happen, in a simpler way it can be explained like this way. There is no net charge normally in a piezoelectric crystal as there is no such symmetrical charge distribution. Therefore, there is no net charge remained in the crystal surfaces. But when the crystal is pressed (subjected to a stress) or a tension is applied to it, the Silicon (Si^+) and Oxygen (O^-) atoms in quartz crystal lattice gets deformed. Then one surface of the crystal is charged as net positive whereas the opposite surface is charged as net negative. Now the crystal is like a small cell.

Such crystals are being used to convert the sound energy in microphones into electric energy. Even in olden days, the needle which travelled in the gramophone disks had a piezoelectric crystal piece. The corresponding electrical signals are given by the crystal when a mechanical pressure variation (going up and down) occurs while the tip of the needle is gone through the grooves.

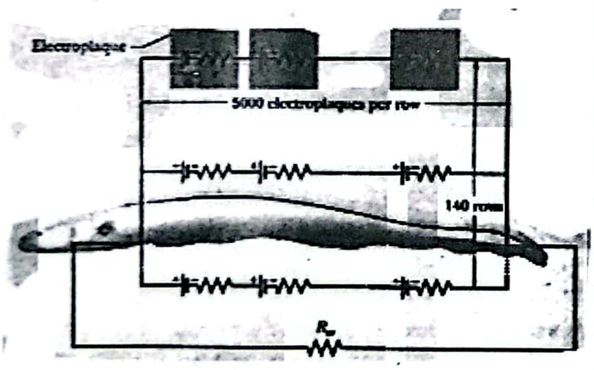


Likewise, when a mechanical energy, pressure variation or a movement is needed to convert into an electrical energy, such crystals are being used. Even in some cigarette lighters have piezoelectric crystals. When the switch is pressed, the crystal is pressed and the generated voltage creates a spark across the small hole.

Such crystals can be used in the inverse way too. When electric pulses are given across the crystal, the crystal lattice is subjected to rapid mechanical vibrations. Ultra sound waves are generated like this way. As the vibrational frequency of the crystals are high, we do not hear the generated sound waves from them. Even the time measurement of the watches that has quartz crystals are done from such vibrations of the crystal.

Some children have chosen photo diode as the photo diode can be used in two ways, I guess. Normally, a phot diode is acting as an e. m. f source when it is acting as a photovoltaic cell (solar cell).

A figure of an electric eel is shown here. Using a biological cell group called electro plaques, it generates an e. m. f. There are 5000 electro plaques each with the e. m. f of 0.15 V along with 140 such rows of electric boards.



For the reverse biased photo diode, some voltage has to be given from the outside. In such an occasion, the photo diode cannot be considered as an e. m. f source is my opinion. I can be wrong. Instead of the charged capacitor, if you consider a capacitor (without the word 'charged'), then the selection of the answer is easy. Then it can turn into a dumb question also. When it is mentioned as a charged capacitor, it can be considered as an e. m. f source as a current can be obtained even for a small time as it has been already charged. There are some hints in some books for this.

But it is considered as a source, when something can be generated from it. The sun is a light source. A glowing light is a light source. We do not supply light to the bulb. Radioactive nuclei are radioactive sources. Hearts are sources of love, kindness and affection. If the spring of the well is dried, then there is no water. But as long as there is a supply, the well is a water source. We do not fill water into the well and use that water for our usage.

10. **The logic circuit shown in figure (a) is equivalent**

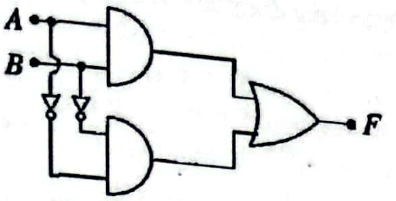
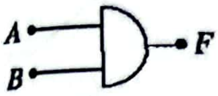
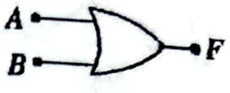


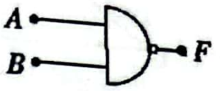
Figure (a)



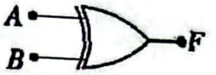
(1)



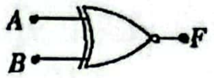
(2)



(3)



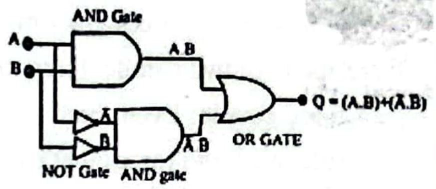
(4)



(5)

Logic Gates 09

An equivalent circuit of Ex-NOR gate is shown here. Ex-OR gate is completely devoted for OR gate. The output is high (1), when there is one only. In a normal OR gate, the output is high when both are there. Ex-NOR should be the inverse of Ex-OR gate. The output is 1 only when both are 0 (both are absent) and when both are 1 (both are present).

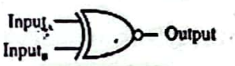


On other occasions, the output is low (0). If we say in another way, then the output of Ex-NOR (XNOR) gate is high when the logic levels are equal (0,0 or 1,1). Write the Boolean expression on the logic circuit. Even if we look at the circuit, then we can say that it cannot be an equivalent to AND, OR or NAND gate. Once you see , it is understood that it is a XNOR gate. It is good if both are bad. Even it is good if both are good. The family life cannot be existent when one is good and the other is bad.

Exclusive - OR gate

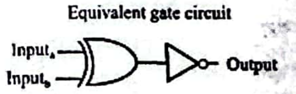
A	B	Output
0	0	0
0	1	1
1	0	1
1	1	0

Exclusive-NOR gate



A	B	Output
0	0	1
0	1	0
1	0	0
1	1	1

Equivalent gate circuit



11. **The accelerations due to gravity on the surfaces of a uniform spherical planet A of radius RA and a uniform spherical planet B of radius RB are equal. If the mass of A is twice the mass of B,**

(1) $R_A = \sqrt{2}R_B$ (2) $R_A = 2R_B$ (3) $R_A = \frac{R_B}{\sqrt{2}}$ (4) $R_A = \frac{R_B}{2}$ (5) $R_A = R_B$

Gravitational Force 05

The gravitational acceleration $g \propto \frac{M}{R^2}$ [mg =GMm/R^2]

$2M/R_A^2 = M/R_B^2 \rightarrow R_A = \sqrt{2}R_B$

12. A, B, C, D and E are five coplanar forces of equal magnitudes acting on a body as shown in figure (a). Which of the following diagrams best represents the direction of the resultant of these forces?

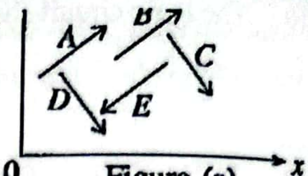
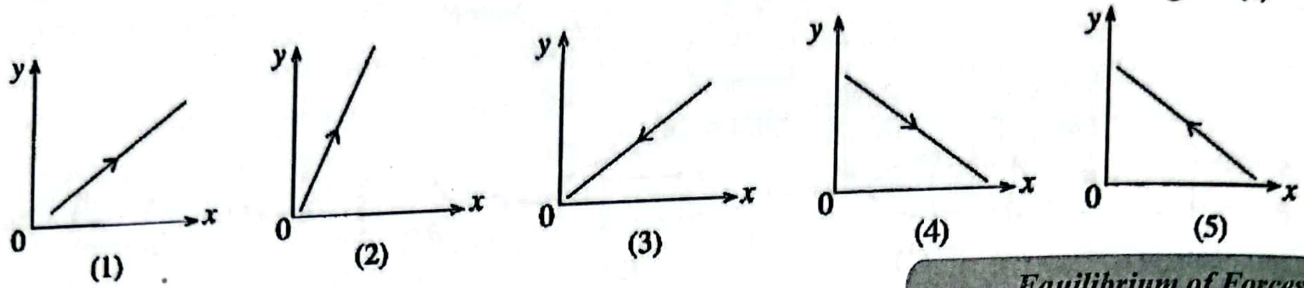
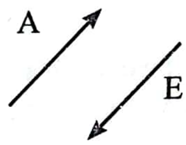


Figure (a)

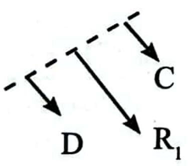


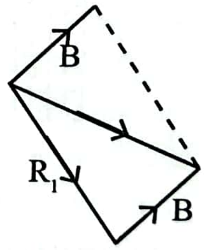
The direction of the resultant of a coplanar force system which is not parallel can be easily found by drawing the polygon of forces. This method can be applied even whether the forces meet or not at one point. The resultant of two parallel forces that are equal in magnitude but act in opposite direction is zero. But it creates a force couple. Even there is a net moment of force in the force couple, there is no resultant force. The resultant of two forces that act on the same direction acts in the middle of the action line parallel to those forces.

From A and E there is no resultant force. So, forget about these forces.



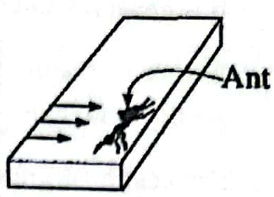
The resultant of C and D (R_1) is acting on the direction that is shown in the figure. The magnitude is double the amount of one force. Now the resultant of R_1 and B is acting to the direction that is shown in the figure. The method of force polygon can be used if needed. Now the complete resultant is acting on the line that is shown in three arrows. Even if you do not apply force polygon method, as the resultant of D and C is acting on the place where force B is applied, the final resultant can be decided from the parallelogram of forces method.





However, if someone has drawn the relevant direction to downwards  as one direction, then there is no other direction than this one. There is no use from other mango trees than this mango tree. Give merits.

13. An ant of mass 2×10^{-6} kg (2 milligrams), which is stationary at the edge of a horizontal smooth strip is removed in 0.2 s by blowing with mouth. The direction of blowing is horizontal as shown by the arrows in the figure. If the ant is thrown out in the direction of the blowing with a horizontal velocity of $0.5\ \mathrm{m\,s^{-1}}$, the average force exerted on the ant by the blow is



(1) 5×10^{-6} N (2) 1×10^{-5} N (3) 2×10^{-5} N (4) 1×10^{-3} N (5) 5×10^{-3} N

Newton's Law and Momentum **02**

A piece of Styrofoam of mass 2.0 X 10^{-6} kg which is kept on a horizontal table is blown away by wind in 0.2 s with a velocity of 0.5 $\mathrm{ms^{-1}}$. What is the average force that was applied on the piece by the wind?

Force = rate of change of momentum

$$P = \frac{2\times10^{-6}}{0.2} \times \frac{1}{2}\left(\frac{mv-0}{t}\right) = 5 \text{ X } 10^{-6}\ \text{N}$$

*Here o.5 $\mathrm{ms^{-1}}$ is written as ½ for the convenience of simplification.

14. A small object of mass m placed on the horizontal surface of a frozen pond is given a kick imparting an initial speed υ_0 along the horizontal direction. The object moves on the surface in a horizontal straight line without rotation. The coefficient of kinetic friction between the object and the suJace is ɥ. If the air resistance can be neglected, the distance that the object moves before coming to rest is,

(1) $\frac{v_0^2}{2\mu g}$ (2) $\frac{v_0^2}{\mu g}$ (3) $\frac{2v_0^2}{\mu g}$ (4) $\frac{v_0^2}{2g}$ (5) $\frac{2v_0^2}{g}$

Friction **02**

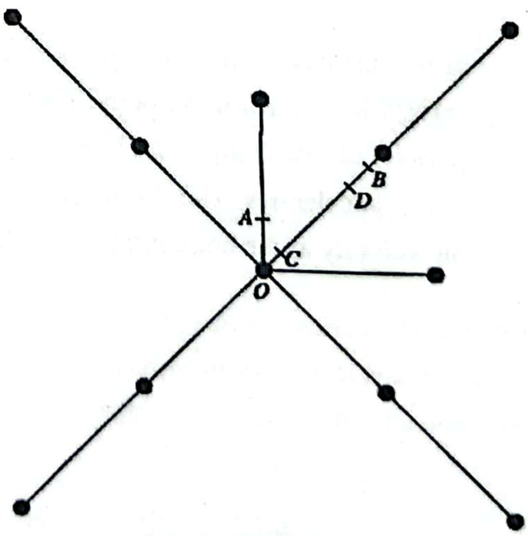
Velocity υ was given in the horizontal direction to an object of mass m which is kept on a horizontal surface with a static coefficient of friction μ. If the air resistance is negligible, then what is the distance that the object travels before it comes to a rest?

The kinetic energy loss = the work done against the friction

½ mv² = μmgs → s = υ²/2μg

This is only the part of Physics. There is ice and others to make the question beautiful.

15. A coplanar structure is made by connecting eleven identical spheres each of mass m using ten identical light rods as shown in the figure. The centre of gravity of the structure is most likely to be at the point



(1) O

(2) A

(3) B

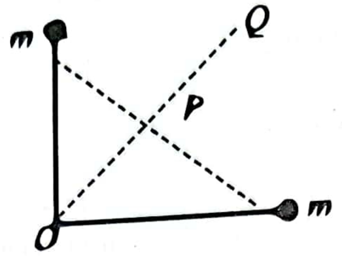
(4) C

(5) D

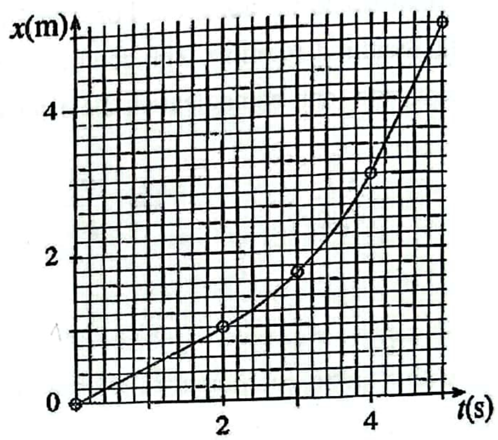
Centre of Gravity **02**

It is general knowledge to find the centre of gravity as the point O in the middle of five identical balls. If such a structure is rotated to any direction, then it is clear that, each extra structure will have the centre of gravity at O. Now look at the following arrangement.

The centre of gravity of shown two masses m is in the middle of the line joining those the two masses. That is at point P. The centre of gravity of all the other points are located at point O. Therefore, the centre of gravity of the total structure should lie on OQ line near to point O. There are many masses that are concentrated at point O. There is 2m in point P. Therefore, the centre of gravity of the combined structure should be closer to point O. There is one point that is closer to point O. Select that point by giving merits.

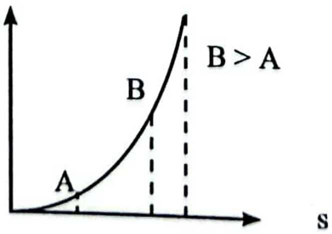


16. A block of mass 2 kg is pushed along a horizontal surface. The variation of the displacement z, of the block with time i, is shown in the figure. The values of the resultant force F acting on the block along the direction of motion during each of the time intervals 0 < i < 2, 2 < t < 4 and 4 < t < 5 do not change. Which of the following correctly represents the magnitude of F in each of the time intervals?



	F(N) (0 < t < 2)	F(N) (2 < t < 4)	F(N) (4 < t < 5)
(1)	0	0	0
(2)	0	1.5	0
(3)	0	2	0
(4)	1	0	0
(5)	2	1.5	1

If the displacement – time graph is a simple inclined line, then it indicates a uniform velocity. If an object travels in a uniform velocity, then the resultant force that acts on it is zero. Therefore, during the time periods of 0<t<2 and 4<t<5 in this graph, the resultant force on the object is zero. In the middle part, the object accelerates. The displacement – time graph is a curve. If you need to find the acceleration, then find the velocity at the beginning of acceleration and you can apply x =ut + ½ at2 for the object.



The displacement- time graph for an object which travels with a non- uniform velocity

The initial velocity means the gradient of the first straight line. It can be obtained from your memory. ½ = 0.5 ms-1. The distance travelled in 2 s (4 s -2 s) = 2 (3-1).

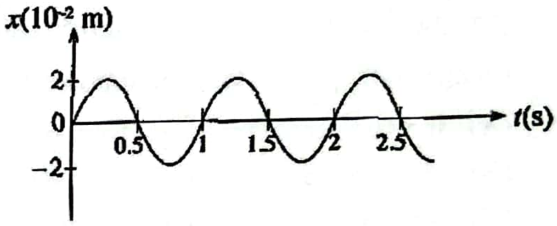
2 = 0.5 X 2 + ½ a X 4 →a =1/2 ms-2. Force (F =ma) F = 2 X ½ = 1 N. This answer is not there. Maybe it was taken using v = u +at.

The final velocity of the object = the gradient of the last straight line = (5-3)/ (5-4) = 2 ms^{-1}

2 = 0.5 + a X 2→a =1.5/2. When we apply F =ma, F = 2 X 1.5/2 =1.5 N. This answer is there. But the answers are different from the equations of x =ut + ½ at^2 and v = u +at. Therefore, it has been given as an all answer. It is easy to use x =ut + ½ at^2 as you do not need to find the final velocity. From the graph directly s and t can be taken. Even the third part of the motion could not have been given. The measured Physics from it is already evaluted from the first part. But if a has to be found by applying v = u +at is on your mind, then the third part is needed. Then the next section of uniform velocity should be drawn matching the curve.

Final velocity = v = u + at = 0.5 + ½ X 2 = 1.5 ms-1 (by applying v = u + at for the second part). If the final velocity has to be 1.5 ms-1, then x = 4.5 m at t = 5 s. Such mistakes happen from every one of us.

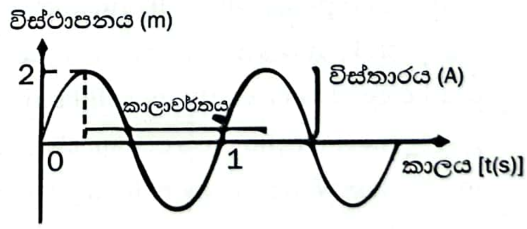
17. Figure shows a displacement (x) — time (t) curve for an object executing simple harmonic motion. For this motion, magnitudes of the period T, the frequency / the angular speed m, the maximum speed rgp and the maximum acceleration app are given by,



	T (s)	f (Hz)	ω (s^{-1})	$v_{max} \times 10^{-2}$ (m s^{-1})	$a_{max} \times 10^{-2}$ (m s^{-2})
(1)	0.5	2	4π	4	16
(2)	1	1	2π	4π	$8\pi^2$
(3)	1	2π	2	4π	8
(4)	1	1	2π	8π	$16\pi^2$
(5)	1	1	4π	8	16

Simple harmonic Motion **03**

The displacement – time graph of an object which undergoes simple harmonic motion is shown here. Period = 1 s and frequency f = 1/T = 1 Hz. Then, $\omega = 2\pi f = 2\pi$ $rads^{-1}$, $v_{max} = A\omega = 2 X 2\pi$ ms^{-1}, $a_{max} = \omega^2 A = 4\pi^2 X 2 = 8\pi^2$ ms^{-2}



The equation for the velocity of the simple harmonic motion, which is $v = \omega\sqrt{A^2 - x^2}$ is not in the syllabus as I know. When x =0, v gets maximum. Therefore, Vmax can be taken as $v_{max} = A\omega$.

18. An elephant at rest is observed by a person 1 km away from his location. The sound intensity of trumpet of the elephant heard by the person is 10^{-10} W m^{-2}. Assume that the sound comes from a point source. If the threshold of hearing of the person is 10^{-12} W m^{-2}, what is the maximum distance from which he can hear this trumpet?

(1) 1 km (2) 2 km (3) 4.5 km (4) 10 km (5) 20 km

Intensity of Sound **03**

When a man with 10^{-12} Wm^{-2} threhold intensity of hearing is away from a sound source by 1 km, the sound intensity that he hears is 10^{-10} Wm^{-2}. What is the maximum distance that he can hear this sound? A point sound sourcce will equally distribute sound to every direction. In such an instance, the sound intensity of a certain point is inversely proportional to the square of the distance. (Look at the 41st question of paper 2006). That means $I \propto \frac{1}{r^2}$. If there is no energy loss, then the sound energy also obeys the inverse sqaure law.

10^{-10} has to be reduced till 10^{-12}. The amount that should be reduced will be 10^{-2} or 1/100. If so, then the relevant distance should be 10 km. $1/10^2 = 1/100$. You can do this from your memory. If you write equations, then $10^{-10} \propto \frac{1}{1^2}, 10^{-12} \propto \frac{1}{r^2}$. When both relations are divided from each other, then $10^{-10}/10^{-12} = r^2$, $r^2 = 10^2$; r = 10 km.

You do not have to solve like this. You can do it from your memory as I told before. To make 10^{-10} into 10^{-12}, you need to reduce by 100. As I is related with $1/r^2$, then to get a change of 100, r should be 10 km. Even I is changed by a power of ten. Then the distance should be a power of ten. In the answers, there is 10 as a power of ten (apart from 1). Most of the time, such questions give the changes of I only in the powers of ten.

19. Two mercury-in-glass thermometers P and Q are to be constructed with P having a larger bulb of mercury than that of Q, and both calibrated in the range 0 °C — 100°C. Assume that walls of both bulbs have the same thickness. Consider the following statements.

Using capillary tubes with appropriate uniform bore radii, the two thermometers can be constructed to have

(A) the same capillary length between 0°C and 100 °C markings.

(B) the same response time for rapid changes in the measuring temperature.

(C) a higher sensitivity in thermometer P than the sensitivity of Q thermometer. Of the above statements,

(1) only A is true. (2) only B is true (3) only B and C are true.

(4) only A and C are true. (5) all A, B and C are true.

If the bulb of the thermometer is greater than the other, then the thermometer with the larger bulb has more mercury in it. Therefore, the volume expansion of mercury is greater for a certain temperature increment. This is general knowledge [$\Delta V = V_0 \gamma \Delta\theta$]. If the initial volumeis high, then the volume difference is also high (ΔV). Therefore, if the capillary radii of the two thermometers are same, then the mercury expansion is greater along the capillary in the thermometer with the bigger bulb. That means the sensitivity of the thermometer is higher. It shows a greater expansion to a particular temperature difference. Then the temperature readings can be done very easily by callibrating the marks with the required distance. Increased sensitivity means giving more reaction to a certain change. As written in previous reviews, these points are relevant to humans too. People who react more to tiny issues are very sensitive. If there is no reaction for a change, then they are insensitive to that change. My wife is sensitive to a small change of mine. She can feel even a small change of mine.

If the capillary diameters are equal, then the thermometer with the bigger bulb is more sensitive. This statement is true. Even if the capillary diameter is increased, then the increment of the mercury column length can be cancellled off for a certain temperature change. By increasing the capillary radius of the thermometer with the bigger bulb and by reducing capillary radius of the thermometer with the smaller bulb according to the requirement (or keeping on the same value while only increasing the other one), the same capillary distance can be designed to get on each thermometers between two temperature marks. Therefore, this statement is also true.

But if there is more mercury in the bulb, then more heat should be absorbed from the required material to measure the temperature. If there is small amount of mercury in the bulb, then the absorption happens very quickly. If there is more mercury, then the heat that should be absorbed is more as well as it takes some time. Therefore, these thermometers cannot be desigened to get same reaction time. Even if the measured temperature is rapidly changed, then you cannot however get the same reaction time. When there is more mercury, it consumes some time to absorb from the liquid and reach the continuous state of temeprature. If there is less mercury, then it quickly absorbes heat and get settled.

There cannot be any effect on the reaction time by the increase or the decrease of the capillary hole. First, mercury in the bulb should absorb heat to rise in the capillary. What will give an accurate temeprature value? It is the thermometer with the small bulb. The mercury in it has a less heat capacity. Therefore, it does not absorb more from the material on which it absorbs. So, it measures the temperature of the material which is much similar to the actual temeprature. If there is more mercury, then more heat is absorbed. Then the measured temeprature will be a lesser value than the actual value (Look at the 57th question of paper 2007, the 37th question of paper 1996). If there is no equal thickness to the walls of the bulbs, then that point also can affect the reaction time. If the walls of the bulb is thicker, then the heat will not transfer quickly towards mercury.

20. Water at 0°C is continuously fed into a fully insulated boiler fixed with an immersion heater at a constant rate of 1 x 1£i=2 kg s 1.The specific heat capacity and the specific latent heat of vaporization of water are 4.2 x lW J kg' °C*' and 2.25 x 106 I kg ' respectively. If the steam at 100 °C is to be produced at the same rate as that of supply of water, the power of the immersion heater should be

(1) 4.2 kW (2) 22.5 kW (3) 26.7 kW (4) 42.0 kW (5) 267.0 kW

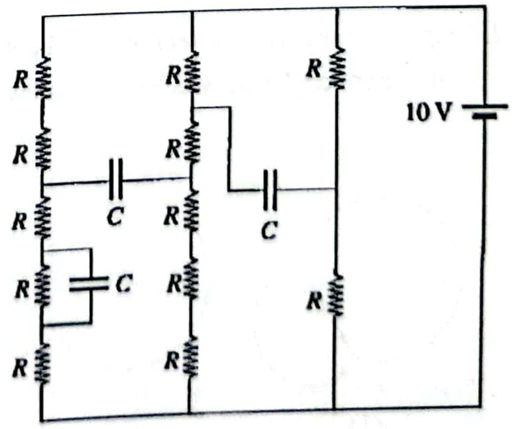
Calorimetry **04**

How much heat is needed to convert 10 g of 0°C water into 100°C of vapour?(the specific heat capacity of water = 4.2×10^3 $Jkg^{-1 0}C^{-1}$, The latent heat of evaporation in water = 2.25×10^6 Jkg^{-1})

The required heat = $(10 \times 10^{-3} \times 4.2 \times 10^3 \times 100) + (10 \times 10^{-3} \times 2.25 \times 10^6) = 4.2 \times 10^3 + 2.25 \times 10^4 = (4.2 + 22.5) \times 10^3 = 26.7$ kJ

10 g (10^{-2} kg) of 0°C water flows per second and with that same flow if you need to produce 10 g of vapour, then the above heat amount should be given per second. Here the heat loss to the environment has been neglected. The heat capacity of the container with water has been neglected. Otherwise, it has been assumed that the container is in 100°C.

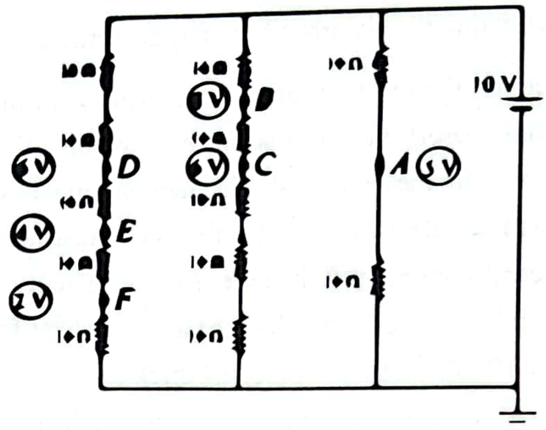
21. In the circuit shown, value of each capacitor is 1 pF. When the capacitors are fully charged, the total charge stored in capacitors is



(1) 2 μC (2) 4 μC (3) 5 μC
(4) 8 μC (5) 10 μC

Ohm's Law combination of Resistances **08**

The charge of a capacitor can be obtained by Q = CV. If you know the capacity of the capacitor, then to find Q, you need to know the potential differnece between the two plates. Consider the resistor network here. For convenience, we will earth the negative end. Think that we need to find the potential of A, B, C, D, E and F points. These values can be obtained from the memory. The value of the resistors are equal. To find the potential, these values are not needed (as they are equal).



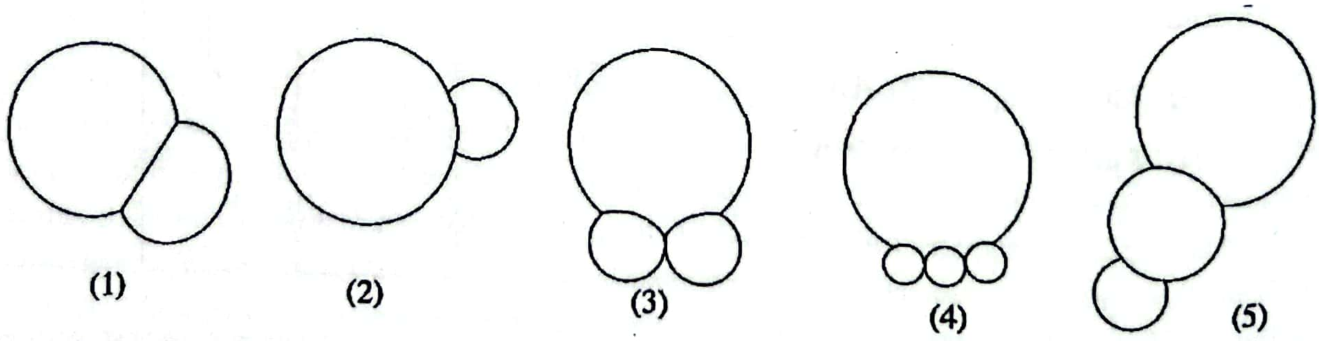
There are two resistor from the right side. Therefore, the potential of A is 5 V (10 V is divided equally). Mark 5V near to point A.

Next two branches have equal resistors of five in each. Therefore, 10 V should be divided equally by 5. The potential difference across each resistor is 2 V. The potential of point B is 8 V whereas it is 6 V in point C, 6 V in point D, 4 V in point E and 2 V in point F. Mark these values quickly near to the respective points. Now if 1μF capacitor is connected across point A and B, then the potential differnece across its plates will be 3 V (8-5). Therefore, accroding to Q = CV, then the stored charge is 3 μC.

There is no potential difference between the points of C and D as both have 6 V. Therefore, there will not be any stored charge if a capacitor is connected across C and D. The potential difference between E and F is 2V. Therefore, if a capacitance of 1μF is connected across point E and F, the stored charge will be 1 X 2 = 2 μC. Therefore, the total charge that is stored on the capacitors is 5 μF (3 + 2).

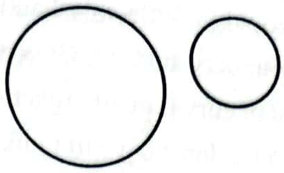
In such questions, always earth the negative end of the cell. Then you can quickly find the potentials of the needed points. If you mark them at the relevant points, then the work will be very convenient.

22. Figures show five clusters of soap bubbles in air, as drawn by a student. F centres of the bubbles in each cluster are coplanar, which of the following shows the cluster with physically possible correct shape?

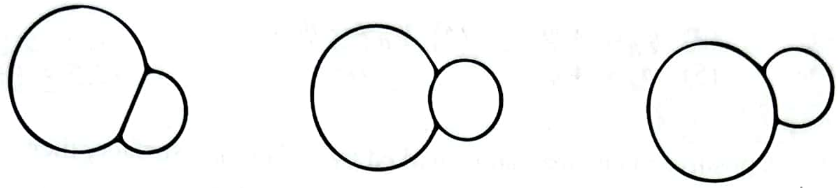


(1) (2) (3) (4) (5)

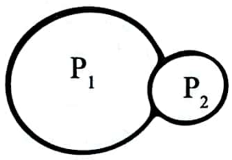
Two soap bubbles in the air with a greater and lesser radius has been shown here.



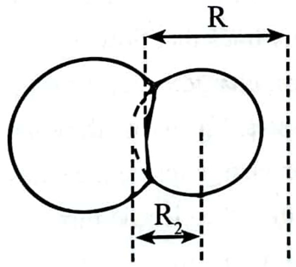
When these are merged, the curvature of the common surface for the boundary of both bubbles should definitely change. The common boundary surface cannot be a flat one. The radius of curvature is infinte in a flat surface. If the radius of curvature is infinite, then the pressure difference gets zero. According to these points, the following shapes can be removed.



Then what will be left out? You can quickly find the answer. Even there is no need, this can be shown by writing the equations. Consider that the bigger bubble radius as R_1, the smaller bubble radius as R_2 and π as the atmosphere pressure. Take the pressure inside the bigger bubble as P_1 and the pressure inside the smaller bubble as P_2. Now think for a moment that they will be like this way when the bubbles are merged.



$P_2 - \pi = 4T/R_2$ .... (1) $P_1 - \pi = 4T/R_1$ .... (2) As $R_1 > R_2$ then $P_2 > P_1$. Now if the intermediate boundary surface radius is R, then $P_2 - P_1 = 4T/R$ .... (3)

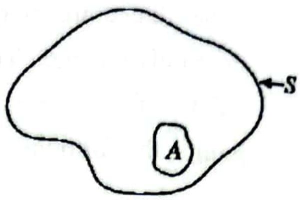


As P_1 is greater than π, by considering (1) and (3), we can directly decide that $R > R_2$. That means the common surface should be bit biased towards to the smaller bubble. Look at the figure.

When two equal bubbles with same radius are merged, it should be like this way. The pressure is equal on both bubbles. $P - P = 4T/\infty$.The boundary surface should be flat.

The above arguments are not needed. Look only at the boundary surface of the bigger and the smaller bubble. Those boundary surface should be little smashed towards the side of the smaller bubble. Find the figure that has a difference in the boundary surfaces (It is not flat). The radius of curvature of the boundary surface cannot be equal to the radius of curvature of bigger or smaller bubble. From this knowledge also you can find the correct figure quickly. You don't have to think anything else. Give merits.

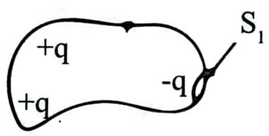
23. A Gaussian surface S is drawn enclosing a charge distribution of net positive charge as shown in the figure. If the electric flux through the portion of the surface marked as A is —9 (9 > 0), which of the following is true regarding the electric flux 9p through the rest of the Gaussian surface?



(1) $\psi_R = -\psi$ **(2)** $\psi_R = +\psi$ **(3)** $\psi_R < -\psi$
(4) $\psi_R < +\psi$ **(5)** $\psi_R > +\psi$

If the net charge is positive in an area that is marked by a Gaussian surface, then the flux is going outward of the surface across the Gaussian surface. That means the net flux is positive.

The electric flux is going outwards from the positive charges inside the Gaussian surface whereas the the electric flux is flown inside by the negative charges. But the net flux is positive. Consider the charge distribution as shown below.

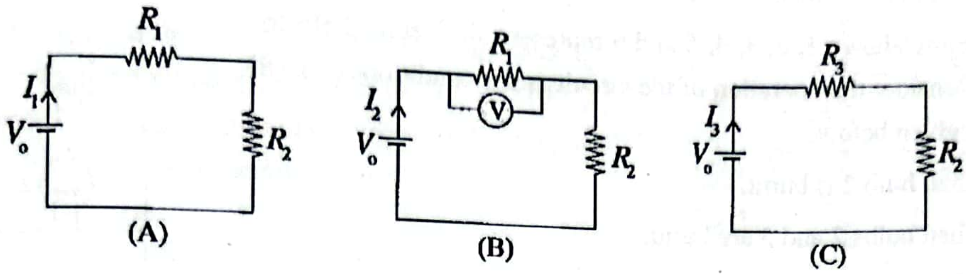


Near to the left side of the Gaussian surface, there are two +q cahrges and near to the right side there is a -q charge. We will consider a small S_1 area which is near to -q cahrge on the Gaussian surface.

The flux lines flow towards -q charge. As S_1 surface near to -q than +q charges, the flux lines that flow into S_1 can be greater than the flux lines that flow outside from S_1. The reason is that positive charges are far awa y from S_1 and the negative charge is closer to S_1. Therefore, the electric flux across S_1 can take a negative value. Let us think that as $-\varphi$. But across the complete Gaussian surface the net electric flux should be positive. Why? Because the net charge inside the Gaussian surface is positive (+q+q-q =+q). Therefore, if the net flux has to be positive, then the electric flux across the rest of the section other than S_1 should be greater than $+\varphi$. If the electric flux across the rest of the area is φ_1, then to have a positve net flux across the Gaussian surface $(\varphi_1 - \varphi) > 0$. $\varphi_1 > \varphi$. However, the value of φ_1 cannot be a zero. Then the total net flux will be negative. It cannot be less than φ also. Even $\varphi_1 = \varphi$ cannot be happened.

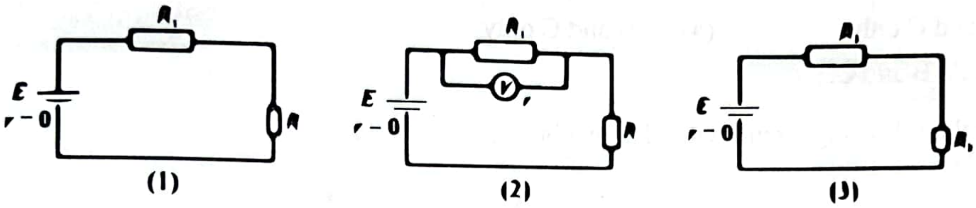
Then the net flux will be zero. Consider the positve charges as sprayers that injects water to the outside whereas consider the negative charges are as the water absorbing sinkers. A person who is closer to a sprayer can experience the of water injection more. Likewise, to a person who is closer to a sinker, may feel the water absorbing nature more. But if sprayers are there more than sinkers and when a cloth is spread to cover the sinkers and sprayers, then does not the cloth get more water by injection?

24. The three identical voltage sources in the circuits (A), (B) and (C), have a negligible intema resistance. In circuit (B), (V) represents a voltmeter having internal resistance r. If $R_3 = \dfrac{R_1 r}{R_1 + r}$,which of the following is true regarding I_1, I_2 and I_3 shown in the circuits?



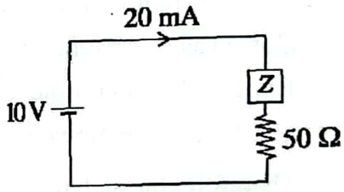
(1) $I_1 = I_2 = I_3$ (2) $I_1 > I_2 > I_3$ (3) $I_1 > I_2 = I_3$
(4) $I_2 = I_3 > I_1$ (5) $I_3 > I_2 > I_1$

Moving Coil Meters **08**



In all of the three circuits, the resistance of R_2 has not been changed. The e. m. f of the cell is also same. Therefore, the current in the circuits are being controlled by the rest of resistors/ or the values of the resistor arrangements. If R_3 has been given as R_1r/R_1+r, then it is clear that R_1r/R_1+r is the equivalent resistance of R_1 and r. R_1 and V are parallel to each other. R_3 is the resistance of the parallel arrangement of R_1 and r. Therefore, $I_2 = I_3$. You can decide this by without writing any equations. When R_1 and r is parallel, the equivalent resistance is less than R_1. Through out the history, it has been checked that when the resistors are parallel, the equivalent resistance is lesser than the value of each individual resistances. Therefore, $R_1 > R_3$. That means the current in circuit (1) is lesser than the other two. $I_3 = I_2 > I_1$

25. In the figure shown, Z represents a network consisting of resistors of unknown values. If the internal resistance of the voltage source is negligible, the power dissipated by the network is



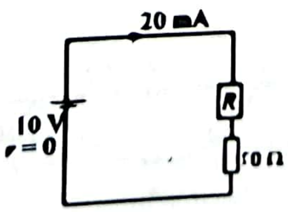
(1) 60 mW (2) 90 mW (3) 120 mW
(4) 150 mW (5) 180 mW

Heating Effects of Electric Current **08**

What is the generated power in the shown circuit by R? As there is no internal resistance in the cell, it does not generate heat. A shorter method is conserve the energy (power). If P is the generated power across R, then $10 \times 20 \times 10^{-3} = P + 50 \times (20 \times 10^{-3})^2$ [$Vi = P + i^2R$]

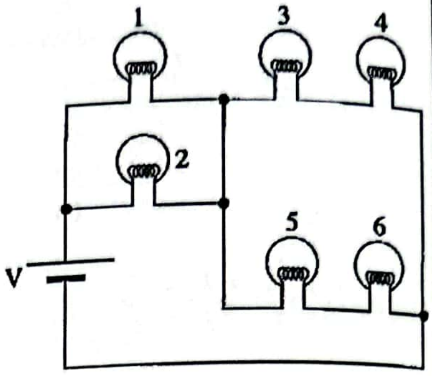
$0.2 = P + 50 \times 4 \times 10^{-4}$; $0.2 = P + 0.02$

$P = 0.18$ W $= 180$ mW



Otherwise, another shorter method is to find the potential difference across 50 Ω. It is $20 \times 10^{-3} \times 50 = 1$ V. Therefore, the voltage drop across R is 9 V (10-1). Now the power generation across R = (9 x 20) mW = 180 mW. I feel that this method is the easiest method. If needed, find R and then you can find I^2R.

26. In the figure shown 1, 2, 3, 4, 5 and 6 represent six identical electric bulbs. Consider the operation of the circuit under conditions (A), (B) and (C) given below.



(A) When bulb 2 is burnt.

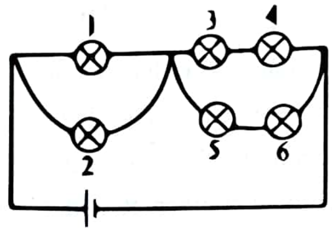
(B) When bulbs 2 and 5 are burnt.

(C) When none of the bulbs are burnt.

Unbumt bulbs in the circuit can be seen glowing at the same brightness in,

(1) B only. (2) C only.
(3) A and C only. (4) B and C only
(5) all A, B and C.

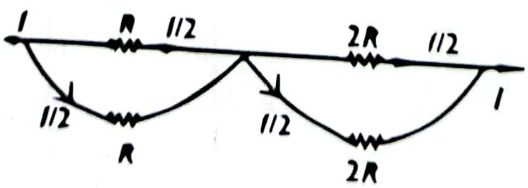
Consider the following circuit with identical bulbs.



If the first or the second bulb gets burnt (by burning the filament), then the unburnt bulb is glowed with more brightness compared to the other four. The current across the unburnt first or second bulb, then divides into two (equally). Therefore, all the bulbs will not be glowed with the same brightness. When fifth bulb is burnt, then the sixth will not just be glowed. Bulb 5 and 6 are connected in series. Then the 6^{th} bulb will not be glowed as its current flow is blocked but not because it is burnt. Same current will be flown across the bulbs 1, 3 and 4 by that time. Second bulb has been burnt. As 5^{th} bulb is burnt, 6^{th} bulb is not glowed. Even 1, 3 and 4 bulbs are glowed with same brightness, bulb 6 is not glowed even it is not burnt.

Therefore, the statement that 'All the unburnt bulbs are glowed with same brightness' is not proved. There can be a mistake in the judgement. As 1, 3 and 4 bulbs are lit in same brightness, it can be taken as true/ relevant as a decision for that situation. This is not a Physics decision but it is adapting to the phrase of 'unburnt bulbs of the circuit'. The filament of the 6^{th} bulb is not burnt but it is not glowed. It has become a victim of the sin of the burnt filament of bulb 5.

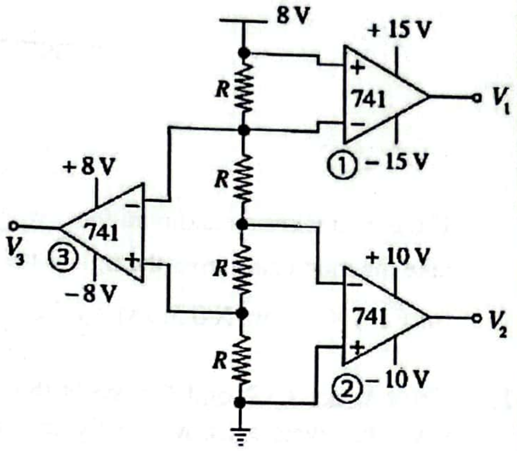
Even 1, 3 and 4 bulbs are glowed with same brightness, the unburnt 6 is not glowed. When one is out from the last pair, the other has to leave the ground even he is not out. When none of the bulbs are burnt, then all are glowed with the same brightness. The current that divides in between bulb 1 and 2 is also equally divided in between the two bulb pairs of 3, 4, 5 and 6.



The first branch has R and R. The second branch has 2R and 2R. But the current in both occasions is divided equally. There is no change in that. Whatever the difficulties are in both roads, how can you treat differently if the difficulties are same? But, due to the values of the difficulties, the total current flow is changed.

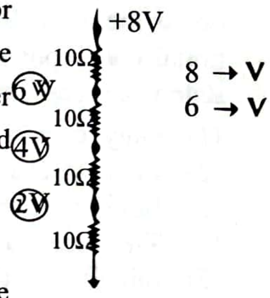
27. In the given circuit, the three 741 operational amplifiers I, Q2 and Q3 are operated by power supplies of z15 V, *10 V, and +8 V, respectively. The approximate values of the output voltages V, , Y2 and V3 are respectively given by,

(1) +2V, −4V, −4V
(2) +15 V, −10 V, −8 V
(3) +2V, +4V, −4V
(4) −15 V, +10 V, +8 V
(5) +15 V, +10 V, +8 V

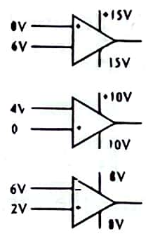


Transisters 09

According to the capacitor problem in 21, divide the potential across each resistor while marking the relevant values. Between 8 V and 6 V, as 8 V is connected to the non-inverting end and 6 V is connected to the inverting end, the operational amplifier is $V_+ > V_-$. So, the output is saturated for +15 V. The other ones can also be decided like this way.



There is no need of calculation. Mark the potentials quickly. Then look at the voltages for the + and – inputs. If $V_+ > V_-$, then the output is positively saturated. If $V_- > V_+$, then the output is negatively saturated. The difference between the input ends is in Volts. They are not in mV. Therefore, the outputs are saturated positively or negatively.

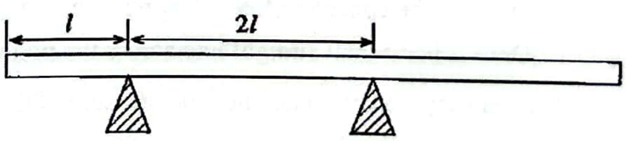


The output is +15 V as $V_+ > V_-$

The output is -10 V as $V_- > V_+$

The output is -8 V as $V_- > V_+$

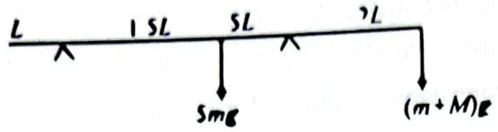
28. A uniform straight heavy plank of length 5i and mass 5m is kept horizontal on two supports separated by a distance 2/ as shown in the figure. A painter of mass m needs to walk along the entire length of the plank carrying his bucket of paint. What is the maximum mass of the bucket of paint that can be carried by the painter without toppling the plank?



(1) $\frac{15m}{2}$ (2) $\frac{13m}{2}$ (3) $\frac{5m}{4}$ (4) m (5) $\frac{m}{4}$

Equilibrium of Forces 02

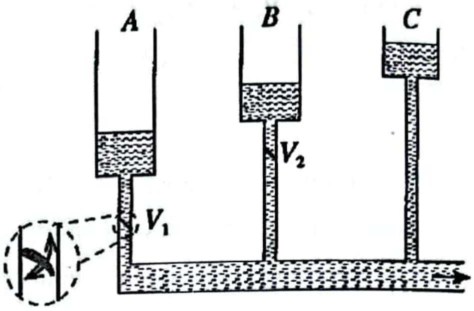
There will be a greater moment across the support when the man with the bucket comes to the right-hand corner. The distance from the right-hand support to the right-hand corner is 2L. Even the least distance to the centre of gravity from a support is from the right-hand support. Therefore, the highest tendency to roll is there when the man comes towards the right-hand corner. When the man goes to the right-hand corner, the clockwise moment across the right-hand support gets maximum.



If the man takes a maximum mass when he goes, then the rod is slightly risen from the left support. If we take the moments across the right-hand support, then

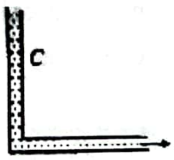
(m + M) X 2 = 5m X 0.5; 2M = 2.5 m – 2m = 0.5 m; M = 0.5 m/2 = ¼ m where M = the mass of the bucket.

29. Three tanks *A*, *B*, and *C* open at the top are initially filled with water to levels as shown in figure. They provide water at very slow speed to an outlet where static conditions could be applied. The two valves V_1 and V_2 allow water to flow only downwards when the pressure above the valve is greater than the pressure below the valve. When the system is put into operation with the initial conditions shown in the figure, which of the following statements best describes subsequent operation of the system?

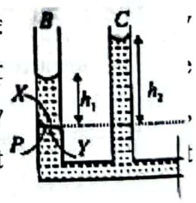


(1) Only *C* will contribute to the flow at the outlet.
(2) Initially, *C* starts to contribute to the flow at the outlet followed by *B* and then *A* in succession.
(3) Initially, *A* starts to contribute to the flow at the outlet followed by *B* and then *C* in succession.
(4) The three tanks will never contribute to the flow at the outlet simultaneously.
(5) Initially, all three tanks contribute to the flow at the outlet with major contribution from *C*.

When the vertical arm of a L shaped tube is filled with water, even a small child also can tell that the water is flowing from the outlet of the horizontal section. It is general knowledge. Therefore, definitely water should emit from C. There are no two words about it. According to simple intelligence, you can remove (3) and (4). Now let us consider the tube of B with the valve.

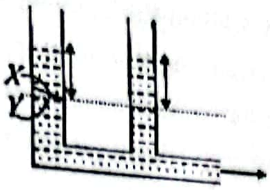


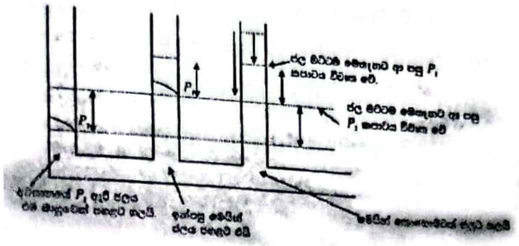
The valve of P is opened, only when the pressure above the valve is greater than the pre the valve. The pressure of X which is above the valve is (the atmospheric pressure + the p upper water level). As there is atmospheric pressure above any arm, it is common to every it should not be considered. So, the pressure of X = h1dg. How to find the pressure below t Y? Draw a horizontal straight line across the point Y.



If we can apply static state, then the pressure of the points on the same horizontal line of the same and freely existing liquid should be equal. If so, then the pressure of point Y = $h_2$dg (once you forget the atmospheric pressure). This logic cannot be applied to a horizontal straight line across X. The valve that is below the point X has blocked the water. But point Y and corresponding liquid points in the tube are at the same condition.

As h_2>h1, the pressure of point Y is greater than the pressure of point X. Therefore, initially the valve is closed. It will not open. But when water is removed from the arm with no valve automatically, its water level is gradually decreasing. When h_2 is less than h_1, the pressure of point X gets greater than the pressure of point Y. Then the valve will be opened and the downward water flow occurs.

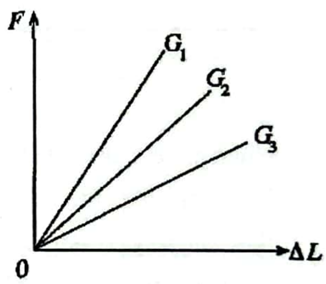




When the arm which has the valve starts to flow water can be depicted from this figure. After a moment form this situation, the pressure in point Y gets lesser than the pressure of point X. If there is another tube with a valve, then this same logic can be applied there.

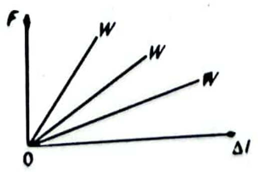
To apply this logic, the water should not flow. When the water flows, the static pressure of a certain point in water is not equal to hdg. You need to consider the dynamic pressure term of ½ dv^2 also (Bernoulli's principle). But if the speed of water flow is very slow, then this can be ignored and static state can be applied to it. To solve the problem, you need to assume this situation essentially.

30. In an experiment to find Young's modulus, three different wires W_1, W_2 and W_3 of the same material have been used and obtained three curves G_1, G_2 and G_3 respectively for the graph of applied tensile force F with extension ΔL as shown in the figure. Which of the following statements, made for the reason of obtaining different graphs, is true?



(1) The wire W_1 may have a larger length and a smaller area of cross-section than W_2.
(2) The wire W_1 may have the same length as W_2 but a smaller area of cross-section than W_2.
(3) The wire W_3 may have the same area of cross-section as W_1 but a length larger than W_1.
(4) The wire W_2 may have a smaller area of cross-section, but a larger length than W_3.
(5) The wire W_3 may have a larger value for the ratio, $\frac{\text{Area of cross-section}}{\text{Length}}$ than that of W_1.

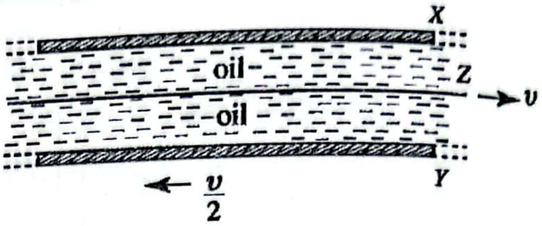
Till you find the correct one, you need to go across each statement. The graphs of the extension (Δl) for three wires of W_1, W_2, and W_3 that are made from the same material and tensile force (F) are shown here. Out of the following statements, which are true? E = F/A. L/ Δl; F = EA/L. Δl

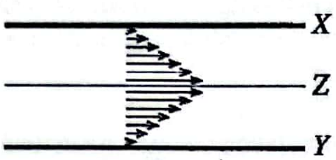


As the Young modulus E is same, all you have to consider the factor of A/L. Consider only the gradients of the three straight lines. W_1 cannot have a longer length and less cross- sectional area. Even if L is increased and A is decreased, the value of A/L will be reduced. Then the gradient should be reduced but the relevant gradient for W_1 is greater.

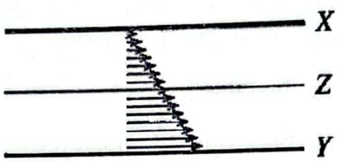
If there is same length with reduced A, then the gradient is reduced. Therefore, it is also wrong. W_3 has the minimum gradient. Therefore, if A gets equal and length gets increased, then you can get the minimum gradient. Now do not look at the other two statements. Once you find the correct one, why do you have to check on other things of others? However, A/L ratio should be highest in W_1 and least in W_3.

31. A thin flat plate Z is placed midway between two large horizontal plates X and F, and the space is filled with a viscous oil as shown in the figure. Now, consider a situation that the plate Z is pulled horizontally to the right with constant speed υ and plate F is pulled horizontally to the left with constant speed $\frac{\upsilon}{2}$ while keeping X stationary. The velocity vectors of thin oil layers between plates X and Y are best represented in,

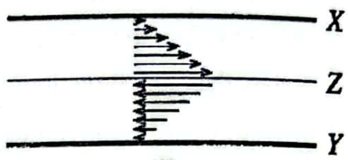




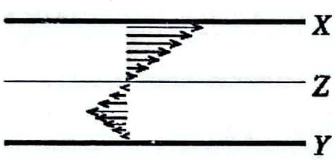
(1)



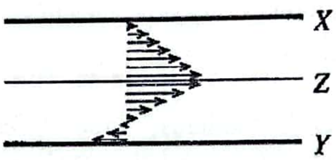
(2)



(3)

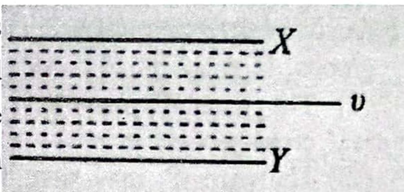


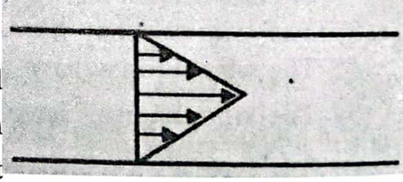
(4)



(5)

When plate Y is kept at rest with X and the middle plate is taken towards the right side, the velocity vectors of the oil layers are represented like this way. It is general knowledge. From this you can remove the first three figures. As plate Y is moving to the left side with a certain velocity, the oil layer just above that plate cannot be at rest. From that, (4) is removed. Only (5) is left out. Give merits.

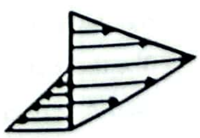




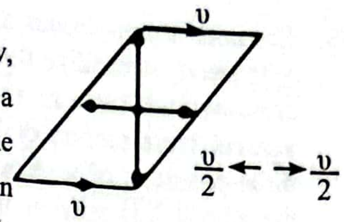
To get it logically, fist you need to assume that plate Y is at rest and think of the velocity vectors. Then you will get the above shown variation. Now keep the middle plate at rest and take plate Y to the left side. Then the velocity vectors of the oil layers in between the middle plate and plate Y is obtained like this way.

Next, superposition these two motions with each other. There is not impact to the middle velocity vectors of plate X and the middle plate. But the velocity vectors between the middle plate and plate Y gets modified (changed) due to the motion of the lower plate. The velocity of the oil layer that is nearest to plate Y is v/2.

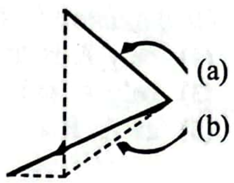
Due to the left side motion of plate Y, the magnitude of right-side velocities of the oil layers in between the middle plate and plate Y will be reduced by a certain amount. The middle layer is pulling the oil layers to the right. Plate Y pulls the oil layers to the left. Therefore, the symmetry of the velocity vectors which was around when there was only the pull of the middle



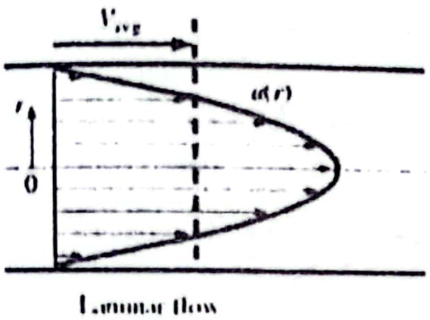
plate, is broken in the region between the middle plate and plate Y.
Even the velocity of the lower oil layer that touches the middle plate is v,
the velocities of the layers below are lower than before and at a layer in a
certain place will be at rest. The velocities of the lower layers are towards the
left side. Think that plate Y is moved to the left with a velocity of v. Then
the velocity vectors are looked like this way. If so, the layer that is at rest is
situated in the very middle of the plate Y and the middle plate. But as the plate Y is moving to left in v/2 not
v, the layer which is at rest is situated below the very middle.

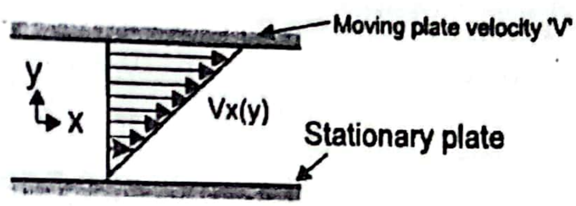


If you just think that, as plate Y is going to left, then the fluid layer that touches
with the plate cannot be at rest. That layer should be pulled towards left- side.
There is only one answer which is drawn like that way. As the plate Y is moved
towards left, the place where the modified profile of velocity vectors lie is shown
in (a) and (b) shows the line in which where the velocity vector profiles will lie
if plate Y is not moved (It is symmetric with the upper section).



As these plates are rectangular or square, the velocity profile is located like this way. If it is a circular tube, then the vector profiles get parabolic.





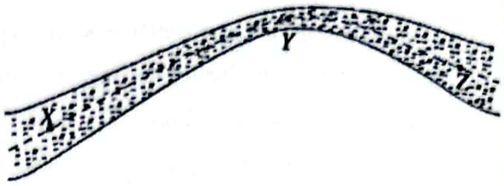
32. Radioactive element pX transforms to stable 2§ Pb after emitting eight o particles and six particles in successive decays. The numbers of protons and neutrons in the element X respectively are

(1) 92, 130 (2) 92, 146 (3) 92, 238
(4) 104, 148 (5) 146, 92

Radiactivity **11**

When an α particle is emitted, the value of Z will be reduced by 2. If 8 α particles are emitted, then Z is value is reduced by 16 (8 x 2). When a β particle is emitted, what happens is the neutron in the nucleus is turning into a proton. $n \rightarrow p + \beta^{-1} + \bar{\nu}_e$. Therefore, when a β particle is emitted, the number of protons will be increased by one. If so, when 6 β particles are emitted, then the number of protons will be increased by 6. According to this, it should be Z -16 +6 =82. Z =92. When an α particle is emitted, the number of neutrons will be reduced by 2. But in a β^- emission, the number of neutrons will be reduced by 1. According to this, it should be N-16-6 =124 (206-82 =124), N =146. If not, when Z is found you can find N after finding A also. When an α particle is emitted, the value of A is reduced by 4. The reduced number for 8 α particles will be 32. In a β^- emission, the value of A is unchanged. Therefore A -32 =206, A = 238. As Z = 92, N = 238-92 =146. X is Uranium 238. But you do not have to know this.

33. Consider a non-viscous and incompressible fluid moving with steady streamline flow through a tube of non-uniform cross-sectional area in a vertical plane. Figure shows the vertical cross-section of the tube. X, Y and Z represent three positions of a streamline. Area of cross-section of the tube at X is same as that at Z. Consider the following inequalities for the relative magnitudes of the kinetic energies per unit volume (KE_X, KE_Y, KE_Z) potential energies per unit volume (PE_X, PE_Y, PE_Z) and the fluid pressures (P_X, P_Y, P_Z) at the positions X, Y and Z respectively.



(A) $KE_Z < KE_X < KE_Y$ (B) $PE_X < PE_Z < PE_Y$ (C) $P_Y < P_Z < P_X$

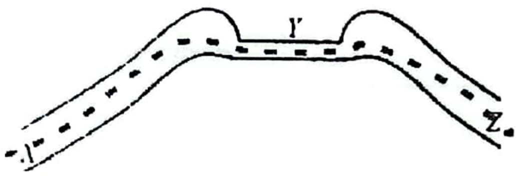
Of the above inequalities,

(1) only A is true.
(2) only B is true.
(3) only A and B are true.
(4) only B and C are true,
(5) all A, B and C are true.

Hydrodynamics **02**

A fluid with characteristics where Bernoulli's equation can be applied flows according to the figure in a tube which is positioned in a vertical plane. The inlet (X) and the outlet (Z) of the tube has equal cross-sectional areas. If the kinetic energies per a unit volume in the fluid at the places of X, Y and Z are (K_X, K_Y, K_Z), potential energies (V_X, V_Y, V_Z) and the fluid pressure (P_X, P_Y, P_Z) respectively, then what is true out of the following?

(A) $K_Z < K_X < K_Y$ (B) $V_X < V_Z < V_Y$ (C) $P_Y < P_Z < P_X$



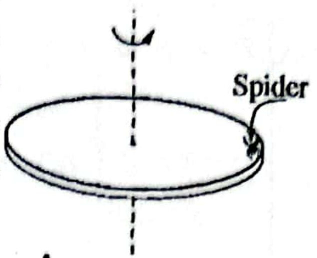
As the cross-sectional area of the tube which the fluid enters and departs is same, $K_X = K_Z$. According to the equation of continuity $A_X v_X = A_Z v_Z$, the real true statement is $K_Y > K_X = K_Z$. From this, the inequality of (A) is just removed. Point Y is situated above the points of X and Z. Point Z is placed above the point X. Therefore, at a glance we can say that the potential energy of Y is greater and the potential energy of Z is greater than of X but it can be seen that it should be less than the potential energy of Y. Therefore, $V_X < V_Z < V_Y$ is true. As the tube is in a vertical plane, this is true. If it was in a horizontal plane, then $V_X = V_Z = V_Y$.

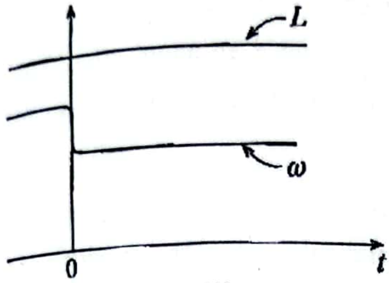
According to Bernoulli's equation, for the liquid $K + V + P =$ constant. At point Y, K_Y is greater as the tube is narrow and it is with less cross-sectional area on that place. Therefore, the speed of the fluid is greater. At point Y, V_Y is also high as it is the highest place. Therefore, P_Y should be a minimum.

As $K_X + V_X + P_X = K_Y + V_Y + P_Y = K_Z + V_Z + P_Z$, if K_Y and V_Y are greater and the total of each three quantities should be equal, then P_Y should be a minimum. K_X and K_Z are equal but as $V_X < V_Z$, then $P_Z < P_X$. Therefore, inequality of (C) is also true.

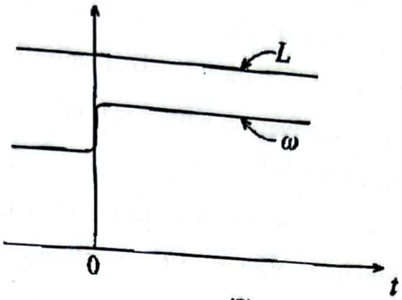
You can decide the true or false nature of inequalities (A) and (B) at once. You need to think a bit for (C). At Y, the kinetic energy as well as the potential energy is also high. Therefore, the pressure should be a minimum. If needed, then you can take the horizontal position across X as zero potential energy level. So, the potential energy of point X is zero. At X and Z, even the kinetic energies are equal, the potential energy of point Z is relatively higher than point X. Therefore, definitely $P_X > P_Z$. When the potential energy and the kinetic energy are greater, the pressure energy gets a minimum.

34. A disc freely rotates without friction at a certain angular speed, about a fixed vertical axis normal to the plane of the disc and passing through its centre. At time i = 0, a spider vertically lowers itself with negligible speed onto the rim of the rotating disc as shown in figure, and becomes rest. Variation of the magnitudes of the angular momentum (L) and the angular speed (m) of the disc only with time (i) is best represented by,

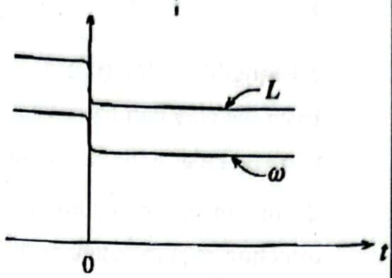




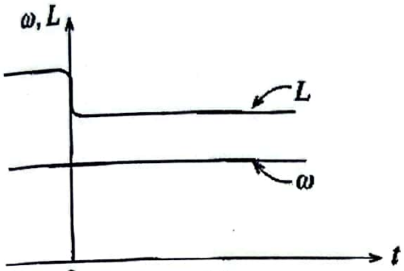
(1)



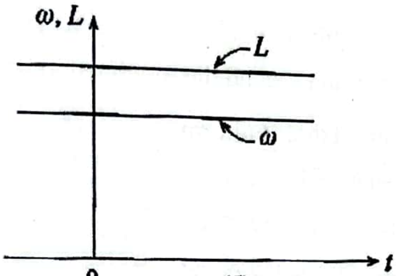
(2)



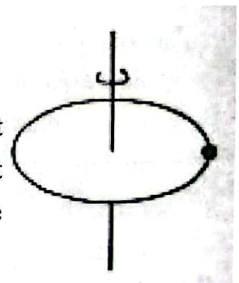
(3)



(4)



(5)



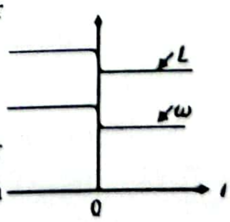
A disk is being freely rotated around an axis that is shown in the figure. When time t =0, a small clay ball is fallen with a negligible velocity to the edge of the disk and get attached. How the angular momentum (L) of the disk and the angular velocity of the system (ω) are varied with time?

When the clay ball is fallen and stuck, the moment of inertia of the system around the rotational axis gets increased by little. Before t =0, only the disk was rotated. Now at t=0, a visitor material has fallen into the disk. Even the clay is fallen away from the rotational axis.

Therefore, when the lay ball fell down, the moment of inertia around the rotational axis gets increased. According to the law of conservation of angular momentum, when the moment of inertia is increased, then the angular velocity should be decreased. If you need to write an equation, then $I\,\omega = (I + mr^2)\,\omega'$ where $\omega' < \omega$.

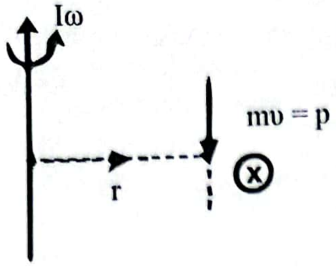
I = the moment of inertia of the disk; ω = the previous angular velocity of the disk; m = mass of the clay ball; r = the radius of the disk; ω' = the new angular velocity of the system

You should not be confused with the angular momentum only of the disk. The total angular momentum of the system is not changed. But as ω is reduced, if you consider only the disk, then its angular momentum is reduced. The previous angular momentum of the disk $L_1 = I\,\omega$ whereas the later angular momentum of the disk $L_2 = I\,\omega'$



As $\omega' < \omega$, $L_2 < L_1$. Therefore, the correct variation is this one. Both get reduced.

From a mistake, you may choose the other one that is shown below. If the angular momentum of the system was asked, then this variation is correct. However, when the clay ball fell, as the angular velocity should be decreased, you need to choose

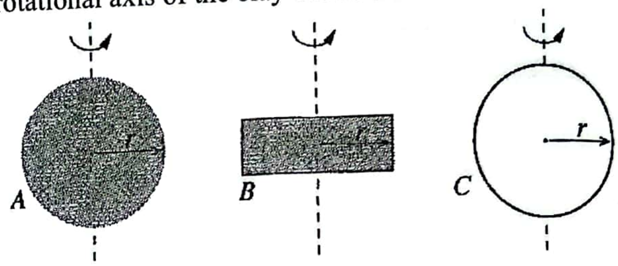


either one of these two variations. As the total angular momentum is always in the head and it should be conserved, you may tend to pick the variation which has a constant L. If so, it is a waste. It happens not just due to ignorance. It is because you did not consider the bold letters. **Always pay your attention to the words that are bold.**

Here the system is (the disk + clay ball). Therefore, the angular momentum is conserved in (the disk + clay ball) system. As the clay ball is falling at a negligible velocity, you can consider that the clay ball initially did not have an angular momentum. Even the clay ball has a considerable vertical velocity, the direction of the angular momentum from that is perpendicular to the rotational axis of the disk.

By the linear momentum mv, the angular momentum around the rotational axis is acting into the paper. Its direction is perpendicular to Iω direction. You need to use the right-hand rule to find the direction of the angular momentum around the rotational axis by the linear momentum of the clay ball. Keep the right thumb perpendicularly to the other fingers and when the fingers should be rotated from the direction of r vector to the direction of p vector, then the thumb is directed towards the paper. As the angle between r and p is 90^0, the magnitude of the angular momentum around the rotational axis of the clay ball is rmυ

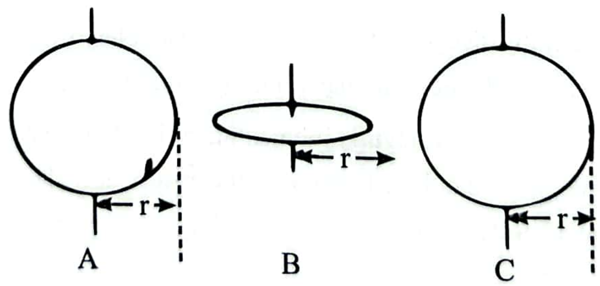
35. Figure shows vertical cross-sections of three uniform bodies A, B and C of identical masses.ñ is a solid sphere of radius r. C is a hollow sphere of radius r and having thin walls. The spheres can be rotated about vertical axes passing through respective centres. B is a disc of radius r which can be rotated about an axis normal to the plane of the disc passing through its centre. All figures are drawn to same scale. If KE_A, KE_B and KE_C respectively are the rotational kinetic energies to be provided for the bodies A, B and C to attain equal angular speeds, which of the following expressions is true?



(1) $KE_A < KE_B < KE_C$
(2) $KE_C < KE_A < KE_B$
(3) $KE_C < KE_B < KE_A$
(4) $KE_A < KE_C < KE_B$
(5) $KE_A = KE_B = KE_C$

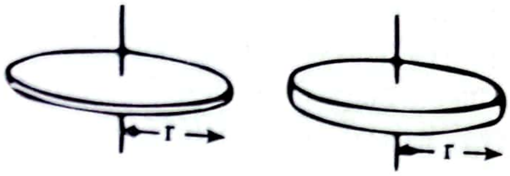
Work Power and Energy 02

A is a solid sphere with a radius r. B is a disk with a radius r. C is a hollow sphere with thin walls and radius r. The masses of them are equal. If the inertia of these three across an axis of the centre are I_A, I_B and I_C as shown, then what is true amongst the following inequalities?

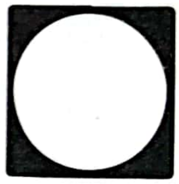


It is a known truth that the highest inertia is there for the hollow sphere. This has been checked before too. The total mass of the hollow sphere is concentrated at the wall. There is nothing in the middle. Therefore, the mass distribution from the rotational axis is far away in C. In other two, the mass is distributed from the axis. They are there nearby and far away. But all of C is at far away. Therefore, definitely I_C should be the highest. $I_A < I_B < I_C$. If the biggest inequality is there in one answer, then quickly select that. What else to do? Give merits.

To have such intelligence and investigation is not a bad non-scientific way. These are Physics. Is not choosing the correct thing quickly is a life skill? Out of objects A and B, how can we find a logic to decide which one has the highest inertia?



Consider a thin disk with radius r and a thick disk with radius r. If the masses of these disks are equal, then the moment of inertia of both disks across the shown axis is the same. The moment of inertia is depending upon the mass distribution away from the rotational axis (far away from the axis). If we look from mathematics, then the moment of inertia of a disk of mass M and radius r across the perpendicular axis towards the disk surface is $\frac{1}{2} Mr^2$. Here it does not have the thickness of the disk. Even we can consider the disk as a tall, thick cylinder too. Now if the middle of the cylinder is dug and then think that the solid sphere of radius r is put in.



When you consider the solid sphere and the solid cylinder, apart from the common volume for both, there is a coloured material part which belongs to the solid cylinder as shown in the figure. They are being concentrated on the four corners. What can we decide from this is that the moment of inertia of solid cylinder with radius r should be little bit higher than the moment of inertia of the solid sphere with radius r. The flesh of the solid cylinder has come out of the solid sphere a bit.

The equations of the moment of inertia are unknown by the students. Therefore, this can be solved from the logic than from mathematics. If you need mathematical equations, then $I = 0.4\ mr^2$ in the solid sphere; $I = 0.5\ mr^2$ in the solid disk and $I = 0.67\ mr^2$ in the hollow sphere.

If the rotational kinetic energies are asked to give equal angular velocities for these objects, then it indirectly asks about the moment of inertias. As the rotational kinetic energy $= \frac{1}{2} I\omega^2$, if ω is constant, then the rotational kinetic energy is directly proportional to the moment of inertia.

36. **A whistle that is used to train a dog produces a frequency of 22 kHz which is more than the thereshold of hearing of the human. The trainer of the dog wants to make sure that the whistle is working. While the trainer stands by the side of a long straight road, he asks a friend to blow this whistle from a car moving along the same road. The required speed of the car and the direction of motion of the car for the trainer to hear the whistle at his threshold of hearing of 20 kHz is (Speed of sound in air is 340 m s')**

(1) **31 m s^{-1}, away from the trainer.**
(2) **32 m s^{-1}, away from the trainer.**
(3) **34 ms^{-1}, away from the trainer.**
(4) **32 ms^{-1}, towards the trainer.**
(5) **34 m s^{-1}, towards the trainer.**

Doppler effect **03**

A vehicle is travelling in a straight road by using a horn of 22 kHz of frequency. What should be the speed and its direction of the vehicle that it should have if an observer who stays at rest near the road heard as 20 kHz? (The sound speed of the air is 340 ms^{-1})

This is a very simple question. When the dogs and elephants are added to the question, it gets more beautiful. But unfortunately, the many students who write the exam do not see that during the exam. The attractive nature of the question is seen by the people who study the questions after the exam. The children who do the exam see such questions as questions with more words. 22 kHz should be reduced till 20 KHz. If so, the vehicle should travel away from the observer. This is general knowledge. The observer is at rest. The speed of the observer v_o =0; According to Doppler's effect, $20 = (340 \times 22)/(340 + v_s)$; $340 + v_s = 34 \times 11$; v_s = 374 -340 = 34 ms^{-1}. The vehicle should travel at a speed of 34 ms^{-1} away from the observer. As 22 should be reduced to 20, the denominator should have $(v + v_s)$.

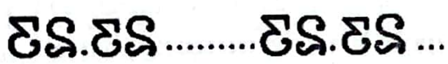
37. Number 23 is written on a piece of paper placed on the flat horizontal surface of a table. A thin convex lens is held just above the number and then it is slowly taken vertically upward while looking at the image of the number through it, and keeping its optical ads vertical. Which of the following best represents the variation of the size and shape of the image as the lens is gradually moved up from the number 23?

(1) 

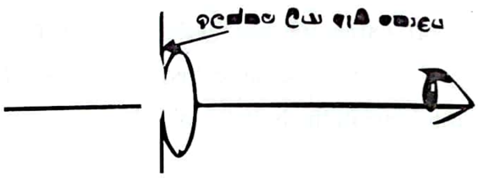
(2) 

(3) 

(4) 

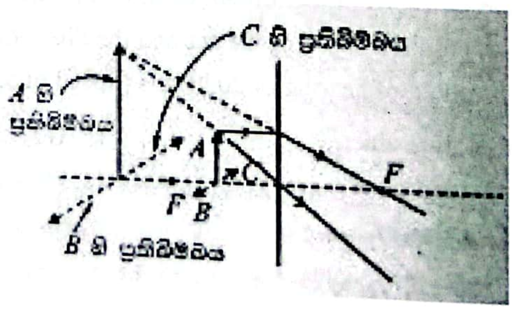
(5) 

According to the given figure, a white paper is kept perpendicularly to the optical axis of a convex lens. To the middle of the paper, a number is written and the image of it observed from the lens.



How will the image be seen when the paper is taken away from the lens along the optical axis?

When deciding the images from a lens, we most of the time draw a thin spike where the thickness of it is not considered. To find an answer to this question, we need to think of a broader object. A number is not a single line. It has a shape and a width. To find the exact image pattern I suggest the following method.



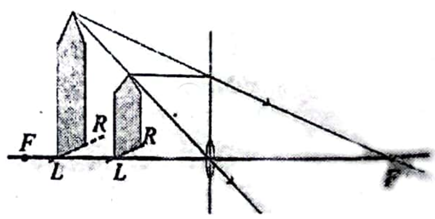
Think of three objects which are placed on the optical axis. One is kept vertically and perpendicularly to the optical axis (A) and the other two are (B and C) perpendicular to the optical axis but in a horizontal plane.

First, keep these objects in the middle of the optical centre and the focus of the lens. Look at the figure.

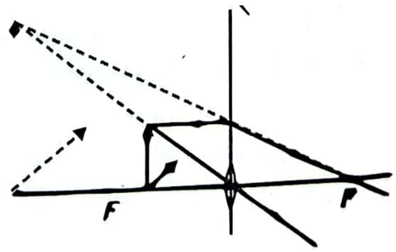
The images of the three objects A, B and C are shown in the figure. All the objects are seen in a non-inverted magnified way. Therefore, when the objects are placed in between the optical centre and the focus, the image is magnified and they are not inverted. Even their sides are not changed. Sides unchanged means when the object is viewed from front, the image that is on the left side is viewed from the left-hand side and the image that is on the right side is viewed from the right-hand side. If simply said, then the right and left sides are not changed in the object as well as in the image. If it is technically said, then there is no lateral inversion of the image.

Next, according to the given figure consider a thick object which is placed in between the optical centre and the focus. The image of the object is seen in a non-inverted and magnified way. But compared to the object the sides of the image are not changed. The left (L) of the object is the left of the image. Likewise, the right (R) of the object is the right of the image. An unreal, bigger and a non-inverted image is created.

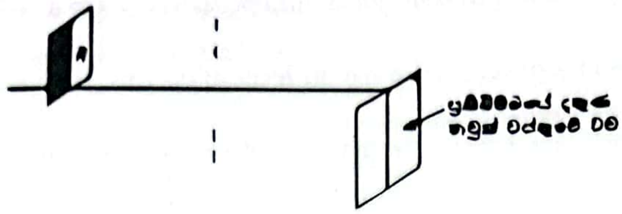
When you design the image of a thick object, we can consider two thinner objects that are perpendicular (horizontal and vertical) to each other.

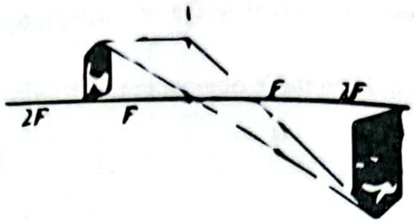


Now let us keep the thick object in between the focal length (F) and 2F. We know that a bigger inverted image will be created. Apart from that, another change has occurred. That is the change of sides. Compared to the object, the left image of the object is changed to the right whereas the right image of the object is changed to the left. The object is subjected to lateral inversion if it is said more technically. When the object is kept in between the optical centre and the focus, there is no lateral inversion occurring in the image.

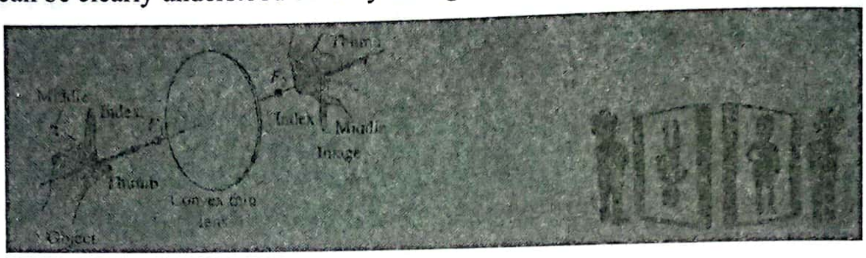


When the object is placed in between F and 2F symmetrically to the optical axis, this lateral inversion can be visible more clearly.





If the left of the object is taken as 2 and the right of the object is taken as 3, then 2and 3 will be seen in a magnified way. Both numbers are inverted and the sides of 2 and 3 are changed here along the way. The lateral inversion can be clearly understood even by the figures which were obtained from Internet.



Keep the thumb, index finger and the middle finger of a hand perpendicularly to each other and look at the way that their images are seen when these fingers are kept away from the focus of a convex lens (look at the figure). From this also this effect is more clearly understood.

Complete the following table for a convex lens:

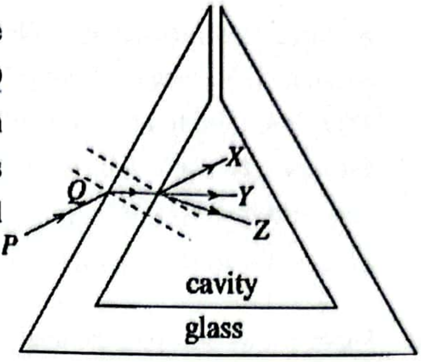
Object Position	Image Position	Nature of image
< F	Same side as the object	Virtual, erect , magnified
F	At infinity	No image is formed
>F and 2F	>2F	Real , inverted, magnified
2F	2F	Real , inverted, same size
>2F	>F and 2F	Real , inverted, diminished
At infinity	on focal plane	Real , inverted, dimisished

If you find a difficulty in finding the answer, then follow the removal of choices by looking each of them. Actually, you should do this. First, when the number is in between the optic centre and the focus, we know without a doubt that the number should be seen in a magnified way without being inverted. According to this, (4) and (5) choices can be removed. When the number is kept in between the optical centre and the focus and go away from the optic centre, the magnification should be increased. According to that, only (1), (2) or (3) are correct. 23 cannot be visible as 32.

Next, when the number is in between F and 2F, the image should be seen in a magnified way. When it is away from 2F, the image should be seen smaller. It cannot be visible in the same size continuously. According to that, (3) is removed. So, only (1) and (2) is left out. In (2), there is no upward inversion (to the vertical direction) in 2 and 3. Therefore, only (1) is left. Even if you do not think of lateral inversion, only in (1) the numbers of 2 and 3 are being inverted. When the number is gone away from the focus (F), only in (1) the numbers are drawn in an inverted way (the lower pieces of 2 and 3 have gone up and the upper parts have come down). Therefore, definitely the answer should be (1).

If the judgement of the answer is hard, then remove each choice/s by using small logic for such questions. Then the correct answer can be found without difficulty most of the time.

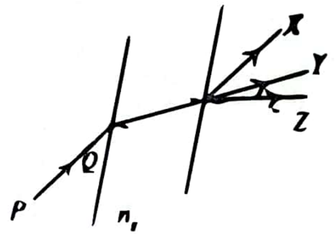
38. The hollow glass prism with thick walls as shown in figure, is made of a material of refractive index yg. A ray of monochromatic light PQ travelling in air is incident on the glass surface as shown in figure. In order to make the emergent ray travel along X, Y and Z directions respectively, the cavity of the hollow prism should be separately filled with transparent fluids of refractive index y, where



(1) $\mu < \mu_g$, $\mu = \mu_g$, and $\mu > \mu_g$ respectively.
(2) $\mu > \mu_g$, $\mu < \mu_g$, and $\mu = 1$ respectively.
(3) $\mu = 1$, $\mu = \mu_g$, and $\mu < \mu_g$ respectively.
(4) $\mu = 1$, $\mu < \mu_g$, and $\mu > \mu_g$ respectively.
(5) $\mu = \mu_g$, $\mu = 1$, and $\mu = \mu_g$ respectively.

Refraction of Prism **03**

A monochromatic PQ light is incident on a glass plate with the refractive index of n_g which has parallel sides. Which inequality should be satisfied by the emitted medium refractive index of n when the emitted ray is need to be travelled in X, Y and Z directions respectively?



The ray of X is going away from the perpendicular which is drawn to second parallel surface. Therefore, n< n_g. The ray of Y is not refracted and it goes without a deflection. So, n =n_g. The ray of Z is n towards the drawn perpendicular line after the refraction in the second surface. Therefore, n> n_g. It is a very simple question. n< n_g, n =n_g and n> n_g. The ray of X is drawn parallel to PQ ray. Therefore, to emit like that way, one can argue that the emitted medium should be the air. If you think like that, then you can take n=1 (air). But there is no choice as n =1, n =n_g and n> n_g. If there was a choice like that, then it is also correct.

39. Biscuits of a freshly opened packet of biscuits were placed inside a container, and it was tightly closed with a lid, so that air cannot enter or leave the container. The initial relative humidity inside the container was also found to be 80'7 . After a few days, it was found that the relative humidity inside the container, had decreased to 309a and the mass of the biscuits had increased by m. If the temperature inside the container remained constant throughout, the mass of water vapour initially present inside the container was,

(1) $\frac{5m}{8}$ (2) $\frac{11m}{8}$ (3) $\frac{8m}{5}$ (4) $\frac{5m}{3}$ (5) $\frac{8m}{3}$

Hygrometry **04**

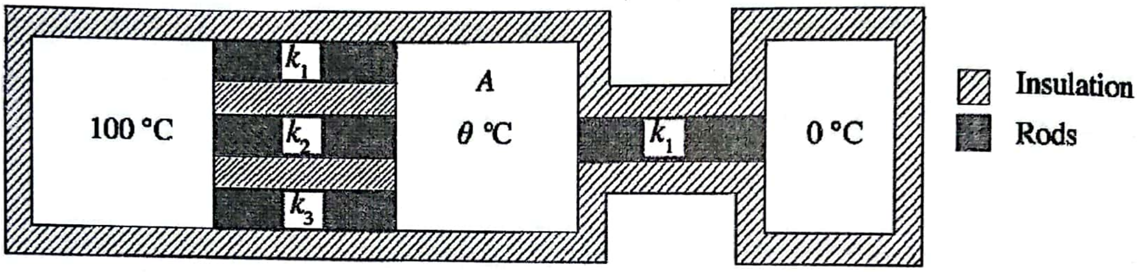
Some of dry cream crackers were put inside a container with the initial relative humidity is 80% and the lid of the container was closed. After some days, the relative humidity of the container was reduced up to 30% and the mass of the cream crackers were increased by the amount of m. If there was no temperature change inside the container, then what was the initial mass of the water vapour? It is a simple question. The fundamental interpretation of the relative humidity is,

$$\frac{\text{the mass of water vapour in a certain volume}}{\text{the mass of water vapour that is needed to saturate that volume}} \times 100\ \%$$

As there is no change in temperature, the amount of water vapour mass that is needed to saturate the container is unchanged. Therefore, if M is the initial mass of water vapour in the container, then $80 \propto M.$ There is no need to write the percentage. As initially dry cream crackers are put inside, their mass increment has happened due to the absorption of water vapour. Now the mass of water vapour in the container = M =m. Therefore $30 \propto M - m.$ Now divide the two relations with each other.

M - m/ M = 30/80; 1- m/M = 3/8 → m/M = 1- 3/8 =5/8 → M = 8m/5

40. Figure shows how four thermally insulated heat conducting rods of equal lengths and equal cross- sectional areas are connected between two heat reservoirs maintained at temperatures 100 °C and 0 °C. A is a thermally insulated heat reservoir which always remains at constant temperature *0*. The thermal conductivities k_1, k_2 and k_3 of the rods are 10, 30 and 50 W m^{-1} K^{-1} respectively. At the steady state, the temperature 8 of the reservoir A is



(1) 90 °C **(2) 85 °C** **(3) 80 °C** **(4) 75 °C** **(5) 65 °C**

Conductivity **04**

A group of properly insulated rods with thermal conductivity $k_1, k_2, k_3, \ldots$ with equal lengths and cross- sectional areas are kept parallelly. If the ends of the rods are kept at constant temperature (in the continuous state), the group of rods can be displaced by a single rod with a net thermal conductivity of $k_1+k_2+k_3+\ldots$ This can be shown very easily. If the rate of heat flow across the rods are $Q_1, Q_2, Q_3\ldots$ then $Q_1 \propto k_1$ $[Q_1 = \frac{k_1 A\,(\theta_2-\theta_1)}{L}]$. You do not have to write : A, L and $(\theta_2 - \theta_1)$. . Likewise, $Q_2 \propto k_2;\ Q_3 \propto k_3$.

That means the total heat flow rate $Q = Q_1+Q_2+Q_3+\ldots.$ $k_1+k_2+k_3+\ldots$

There is no doubt that you have learnt this. Now consider this instance.

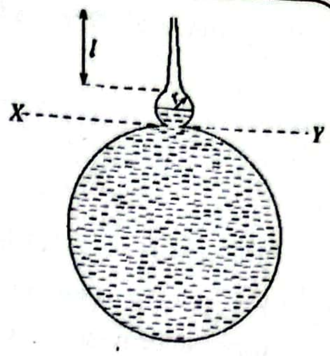
X X X X X   X X X X X

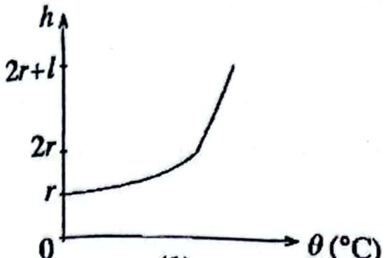
100°C $k_1 + k_2 + k_3$ θ k_1 0°C

X X X X X   X X X X X

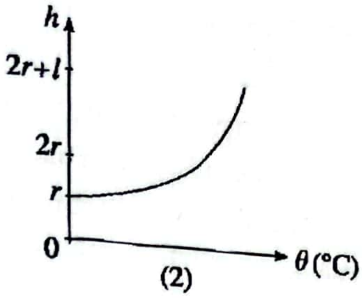
The heat flow rate of the net heat conductivity of $k_1+ k_2+ k_3$ should flow from the rod of k_1 (as they are in series). So, it should be $(k_1+ k_2+ k_3)\,(100 - \theta) = k_1\theta\,(\theta - 0)$. When the values are substituted, $90(100 - \theta) = 10\,\theta$; $9 \times 100 = 10\,\theta \rightarrow \theta = 90$°C

41. A glass bottle of special shape, a vertical cross-section of which is shown in figure consists of a large cavity, a small spherical cavity of radius r and a narrow tapered tube of length l. The total volume of the large cavity and the half of the volume of the small cavity are initially filled with water at 0 °C as shown. If the expansion of the bottle is negligible, the variation of height (h) of the water surface measured from the level XP, with the temperature of the water (θ) is best represented by

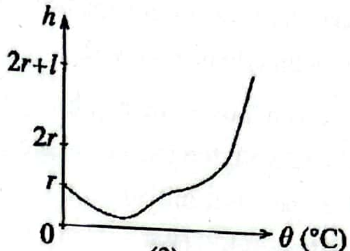




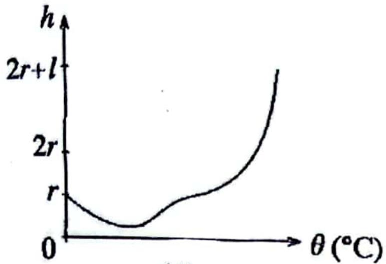
(1)



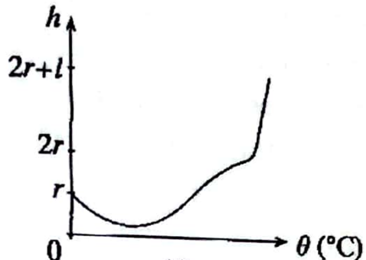
(2)



(3)

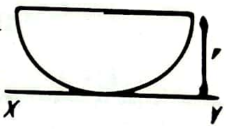


(4)



(5)

Here you need to consider the anomalous expansion of water. From 0°C to 4°C when the temperature of water is increased, the volume of the water is decreased. Therefore, from 0°C-4°C, the height of the water level should be subjected to a decrement. At 0°C, the height of water from XY (h) is r. Therefore, from 0°C-4°C, when the temperature is increased, h should be reduced. From this (1) and (2) can be removed. Next, when the temperature is increased, the water is expanded.

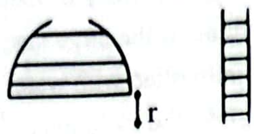


Again, when the water rises till r, even if h is increased, the rate of increment of h should be reduced. Why? Even the volume expansion of water that happens at a certain temperature difference is at a same value, the water filling cross-sectional area is not a constant. When the water is filled from XY to the diameter of the hole (that means till h=r), the water is filled to a cross-sectional area that is gradually increasing. When same volumes are spread over a greater cross- sectional area, even if h is gradually increased, the rate of h increment is reduced. Therefore, the variation of h should happen like this way.

When h=r is passed, the water is gradually filled to a gradually reducing area. Therefore, the rate of increment of h should gradually increase. So, the variation of h should be like this way.

When h =2r is passed, the water is filled to a narrow tube which gradually reduces the radius. Inside of it, the height of the water level should rapidly increase. As the cross-sectional area of the tube is continuously reduced, the height that a certain constant water volume is held should continuously increase. The variation of h cannot be linear. If it has to be linear, then the internal radius (cross-section) of the tube should be a constant. If the tube is like this way, then for each change of volume, the increment of h is same. If so, the variation of h is like this way.

But as the tube is getting narrower, the variation of h should be like this way.



These facts are satisfied by the graph of (4). If you remember the anomalous expansion of water, then you will quickly remove (1) and (2). The final parts of (3) and (5) (after h = 2r) are straight lines. From those, they can be removed. Only (4) is left. Actually, you do not have to look at the middle parts. If you forget about the anomalous expansion of water, then you will pick (2). The last part of (1) is linear. Therefore, it can definitely be removed.

Once you look at the graphs, at a glance you may think that, "my God! How can the value of height (h) be reduced with the increasing temperature?". It can happen. You may think that (3), (4) and (5) choices are false ones. But initially if 0°C water is clicked for you, then remove (1) and (2) quickly. After h =2r, the variation of h cannot be linear. It should be a curve. Instantly you can get the correct variation.

42. In the circuit shown in figure (a), the box with broken lines contains a resistor network. The 2V battery has a negligible internal resistance. An ideal voltmeter connected across ab gives a reading of IV. When the voltmeter is replaced by an ideal ammeter it indicates 2 mA. The resistor network inside the box with broken lines is

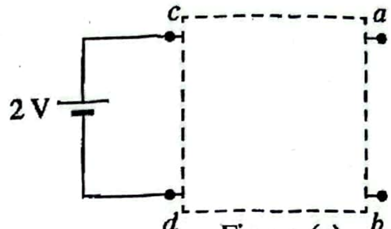
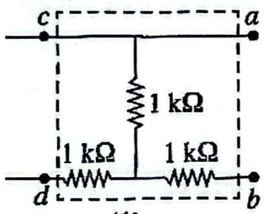
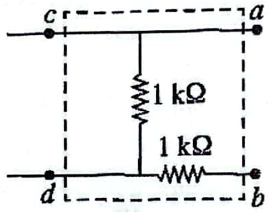


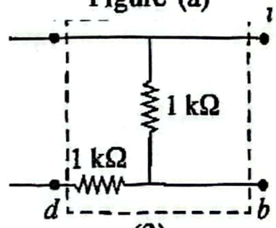
Figure (a)



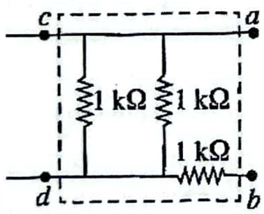
(1)



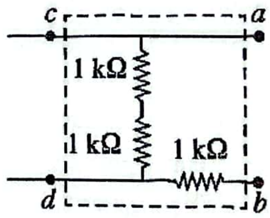
(2)



(3)

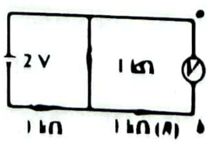


(4)



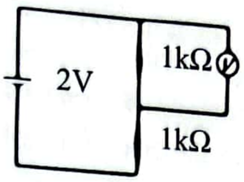
(5)

The easiest way to get the answer is by considering each resistor network in the answers and check which one satisfies the given data. According to the given facts, do not try find the matching resistor network by calculation. Find which data are fitting to the given networks.



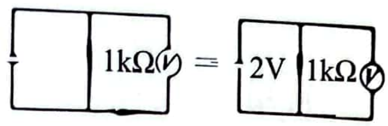
First, connect the voltmeter. Even separate circuits are drawn for explanation, when the problem is untied think that there are voltmeters in each network across ab. The voltmeter is ideal. Therefore, when a resistor of 1 kΩ is connected, there is no usage from that resistor. As the voltmeter is ideal, there is no current flow across it. So, there is no current across 1 kΩ (R).

Therefore, 2 V is divided equally amongst the rest of 1 kΩ. As there is no current flow across R, you can consider that the rest of two 1 kΩ are in series. If you think in another way, then the above circuit is like this way. Therefore, the reading of the voltmeter is 1 V. 2V is equally divided among two 1 kΩ. So, this circuit is correct according to the voltmeter reading.

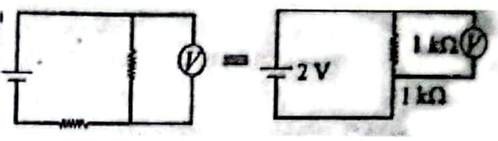


As the voltmeter is ideal, there is no current across any resistor which connects to it in series. The net is infinity if 1 kΩ is added to infinity (bigger resistor). Apply this logic into every network. For an infinite love, there is no use in adding other people's love.

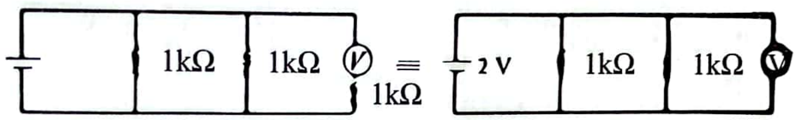
Here the voltmeter reading is 2 V. As there is a negligible resistor for the battery, the voltmeter reads the e. m. f of the battery. This circuit can be removed.



Now, let us go for the third circuit. Here the voltmeter reading is 1V. Therefore, for the time being it can be considered as correct.

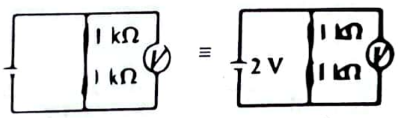


Let us loot at the fourth circuit.

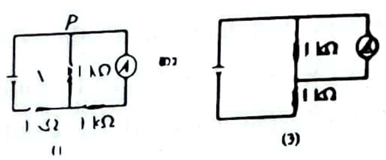


Here the voltmeter reading is 2 V. Remove it. The voltmeter is directly connected to the battery.

Let us look at the fifth circuit.



Here also the voltmeter is directly connected across the battery. The reading is 2V. Remove it. According to the voltmeter reading you can remove three out of five. What are left will be only (1) and (3). Now connect the ideal ammeter.

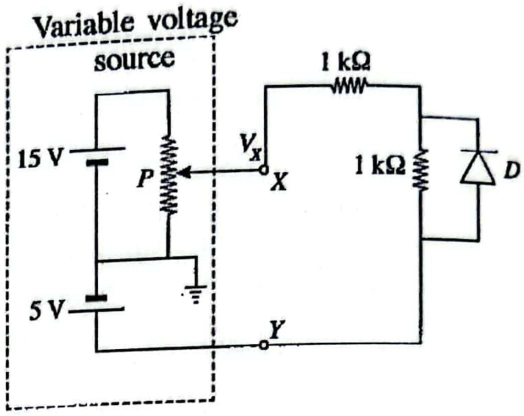


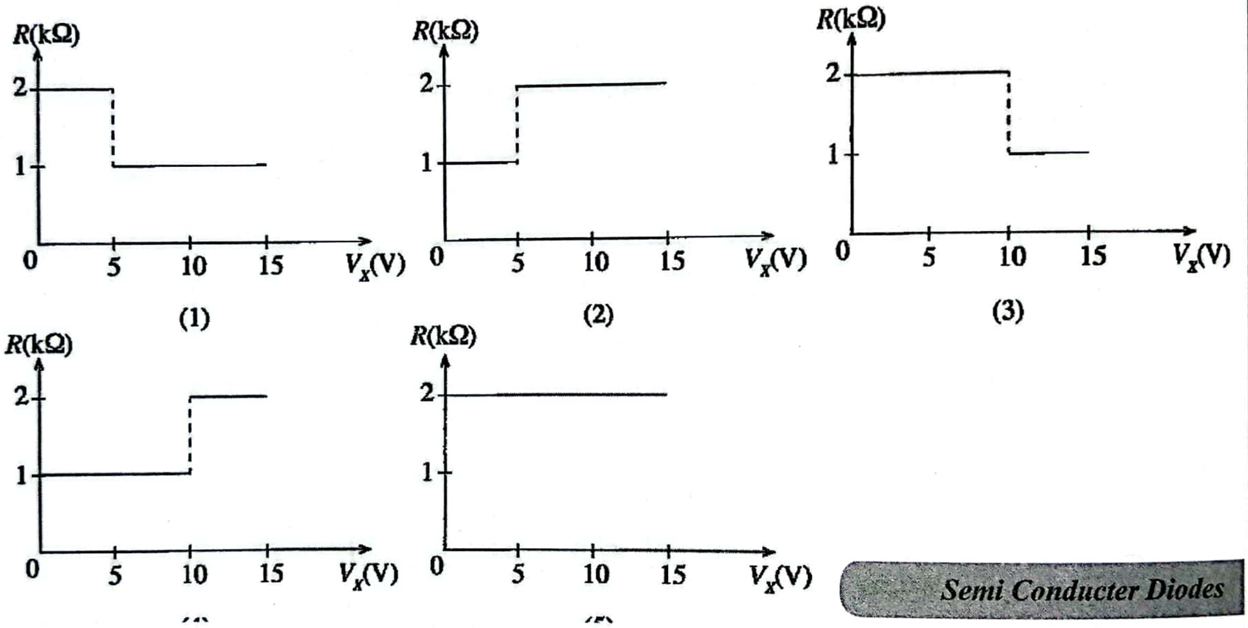
As the ammeter is ideal, there is no internal resistor. Therefore, there is no current across the resistor of 1 kΩ which is parallel with the ammeter. Which fool goes on a road with resistance when there is a road with no resistance?

So, the current of the 3rd circuit goes across the ammeter and goes only across the below resistor of 1 kΩ. Therefore, the current of the circuit (ammeter reading) = 2/ (1 X 10^3) = 2 mA. The correct network is (3). Even it is unnecessary to find the ammeter reading of first circuit, it can be clearly seen that it is not 2 mA. At point P, the current is equally divided into two aprts.

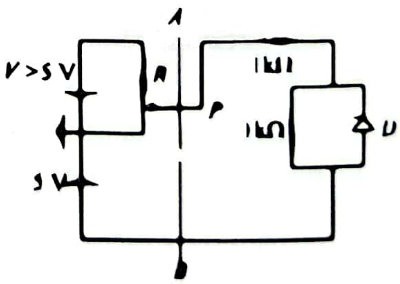
Even this much was written for explanation, in such problems you need to take one data and take it across the choices and clear them. As the voltmeter data is given at the beginning, first consider that fact. As the voltmeter is ideal, there is no current flow across the resistors that connected in series. From this logic, you can remove (2), (4) and (5) choices directly. In those three networks, the voltmeter reading should be 2 V when the resistor of 1 kΩ is removed that is in series with the voltmeter. Next, when the ideal ammeter is connected, to have the ammeter reading as 2 mA, you can undoubtedly decide that, 3rd network should be definitely correct. If the current is needed as 2 mA, then the current should flow across only one of 1 kΩ resistor. According to that also you can remove (1).

43. In the circuit shown, X and F represent the terminals of a variable voltage source situated in the box with broken lines. P is a variable resistor. D is an ideal diode. As the value of the voltage Vg at point X is gradually increased from 0 to 15 V, which of the following graphs correctly represents the variation of the overall resistance R of the section of the circuit to the right of

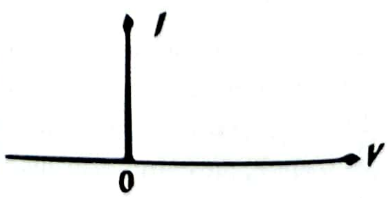




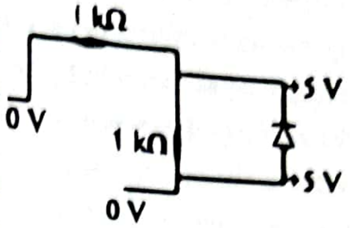
Even it is seen as a complex question, there is a simple logic. In the shown figure, R is a resistor. D is an ideal diode. When the voltage of point P is increased from 0 V to 15 V, how does the net resistance (R) is changed on the circuit section that is right side of AB? The only thing that can happen to the right side is the current flow/block across the diode.



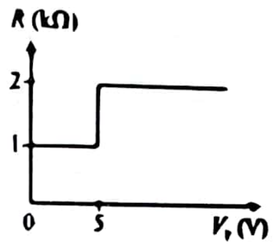
Therefore, you need to decide whether the diode is forward biased or reverse biased only. Even if the voltage of point P is kept at any value, by changing the value of the variable resistor, then the anode voltage of the diode is +5 V. The I-V curve of an ideal diode is shown here. When the diode is in forward biased, its resistance is zero. R =ΔV/ΔI; Even if

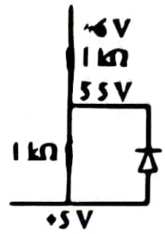


As the diode is forward biased, there is a current flow across the diode. The internal resistance of the diode is zero. Therefore, there is no current across the resistor of 1 kΩ that is connected parallel with the diode. Therefore, there is no use from that resistor. So, the other 1 kΩ only contributes to the entire resistance of this circuit part. As the diode is ideal, there is no voltage drop of 0.7 V. So, when the diode is forward biased (in conducting mode), there is no potential difference across it. This situation prevails until $V_P = 5$ V. When $V_P > 5$ V, the diode is reverse biased. There is no current across the diode. For a $V_P = 6$ V value, the voltage values of each point are as follows.



Now, the current is flown across the resistors only. Top +6 V; Bottom +5 V; in the middle 5.5 V. If we consider the diode, the voltage of the anode is +5 V whereas the cathode is +5.5 V. So, the diode is reverse biased now. The current flows across the two resistors of 1 kΩ which are connected in series. The current flows downwards. It is not an issue for the total resistance. The correct variation is like this way.





If the diode is not ideal, then its resistance is in Ω range at the conduction state (1Ω -25 Ω). When it is in reverse biased state, the resistance of the diode is in MΩ range. If the diode is Silicon, then at the forward biased state, the cathode is + 4.3 V if the anode is +5 V. There is a voltage drop of 0.7 V across the diode.

If such problems are judged as complex/hard problems, then you will be mentally down. The right side of AB has two resistors and a diode. If you understand that the net resistance of that part is dependent upon the action of the diode, then it is easy to obtain the variation. There is a simple logic. For whatever the value of V_P, as the anode voltage of the diode is in +5 V, it is indirectly felt that whatever the change should occur in +5 V. When the diode is conducting, there is no current across the connected 1 kΩ resistor. The equivalent resistance of the circuit part is 1 kΩ (the rest of the resistor). When the diode is not conducting, the current flows across the resistor only. The equivalent resistance is 2 kΩ.

44. When a long capillary tube of uniform bore radius is dipped vertically in a beaker of water of density dp, the water column in the capillary tube rises to a height ɦq as shown in figure (a). Now, an oil of density d0 (< d p) is poured onto the surface of the water in the beaker slowly without disturbing the water as shown in figure (b). Assume that the oil and water are immiscible liquids. The variation of heightɦ of the water column inside the capillary tube, measured from the water surface, with the height h of the oil layer is best represented in

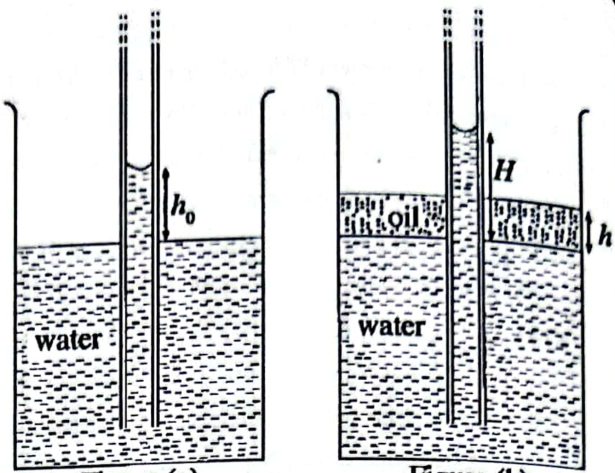
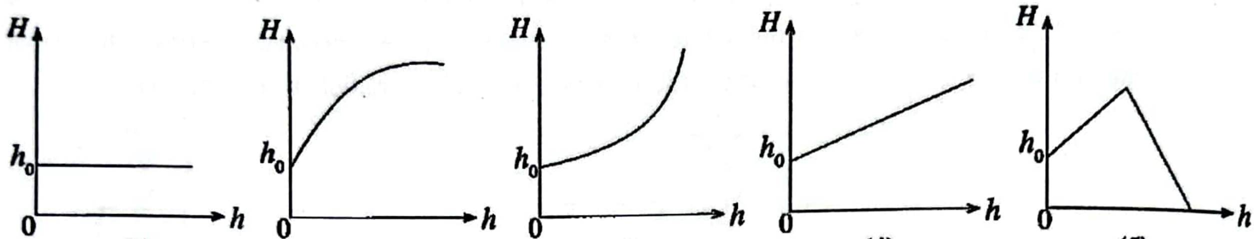


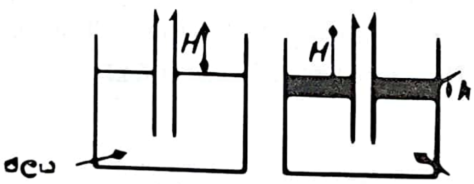
Figure (a) Figure (b)



(1) (2) (3) (4) (5)

As shown in the figure, a uniform capillary tube is sunk in a water container. If an oil layer that is not mixed with water is put on the water, how will H change with h?

If the contact angle between water and glass is zero, then we know that the upward surface tension force is $2\pi r\gamma$ in the water meniscus inside the capillary tube. This will not change once there is oil insertion. What happens from oil insertion is that there is an excess created pressure on the water meniscus. Initially, it had the atmospheric pressure on top of the water surface. But when oil is applied, it creates an extra pressure/ pull. Therefore, it can be said even from general knowledge that when h is increased H also should be increased. When it is pulled from the outside, then it will rise from the middle.



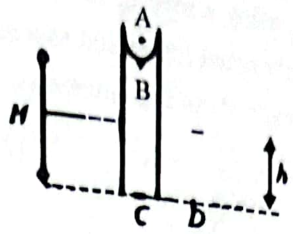
(1) (2) oil

Water

As the excess pressure from oil is hd_0g (d_0 = the density of oil), we just think that the variation of h and H should be linear. There cannot be squared terms or square root terms in it. You quickly can get a relation too. $P_A = \pi$ (atmospheric pressure).

$\pi - P_B = \frac{4\gamma}{r}$ .... (1) $P_B + Hd_wg = P_C = P_D = \pi + hd_og$.... (2)

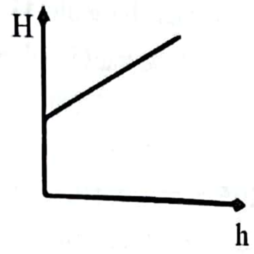
(1) + (2) $Hd_wg = hd_og + \frac{2\gamma}{r}$



So, the graph of h versus H should be a straight line with a positive intercept and a positive gradient. If h =0, then (there is no oil) . $Hd_wg = \frac{2\gamma}{r}$.

This is the equation that we are familiar with.

If the liquid that is newly put is mixed with water, then that mixture will go into the capillary tube and the surface tension of the mixture will be changed. The correct variation is this one.

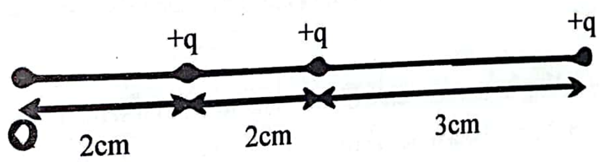


45. Charges in an isolated distribution of three +q point charges are located at distances 2cm, 3 cm and 6 cm from a point 0. Another charge can be brought from infinity to the point 0 without doing any work once a point charge of — q is placed at a distance r from the point O. The value of r is

(1) lcm (2) 2cm (3) 3cm (4) 4cm (5) 5cm

Electrostatic Potential 06

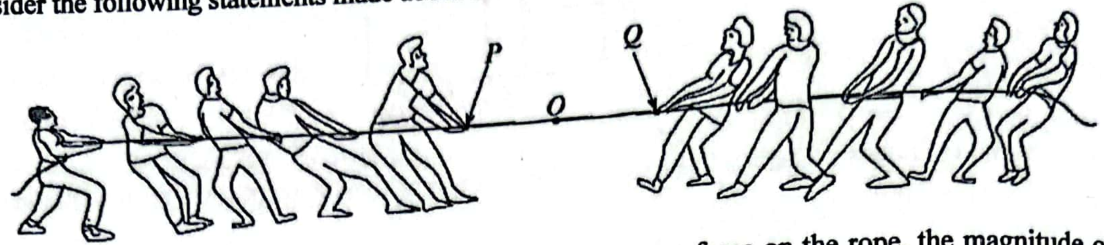
According to the given figure, three point-charges (each of +q) are kept from the point O.



What is the distance (r) from O that a point charge of -q should be kept, if another charge from the infinity is needed to be brought to the point O without doing a work against the electric forces? It is simple even it is question number 45. If we consider that the electric potential at the infinity as zero and there is no work done when moving a charge to O, then the electric potential of point O should be zero. Then there is no potential difference between the infinity and point O. It is also zero. If the potential difference is zero, then the net work that should be done against the electric forces is zero.

Electric potential is proportional to q/r. Therefore, q/2 + q/3 + q/6 - q/r =0; 1/r = (3 +2 +1)/6 → r = 1cm (You do not have to write q if needed)

46. Two teams start to play tug-of-war using a rope of uniform strength on a hard flat horizontal surface as shown in the figure. Both teams apply equal forces and as a result, the point 0 on the rope does not move. Consider the following statements made about this situation



(A) If each of the members of the two groups applies the same force on the rope, the magnitude of the tension throughout the rope is the same.

(B) If the magnitude of tension on the rope exceeds its breaking tension, the rope will break only at a point between P and @.

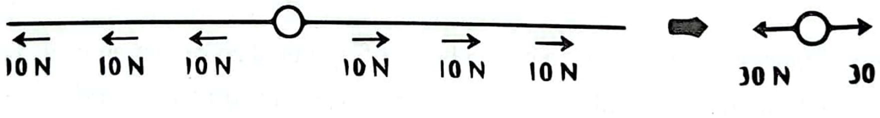
(C) The magnitude of the maximum force that can be applied by an individual on the rope depends on the coefficient of static friction between feet of the individual and the surface.

Of the above statements,

(1) only A is true. (2) only B is true. (3) only A and B are true.
(4) only B and C are true. (5) all A, B and C are true.

It is just clear that if each one of the participants applies same force on the rope, the tension throughout the rope will not be same in a competition of tug of war. For convenience, think that there is ring in the middle. Let us think that three people on the left apply 10 N each on the rope.

Then is not the net force acting to the left- side on the ring 30 N? If a same force is applied on the right-side, then the resultant force on the ring becomes zero.

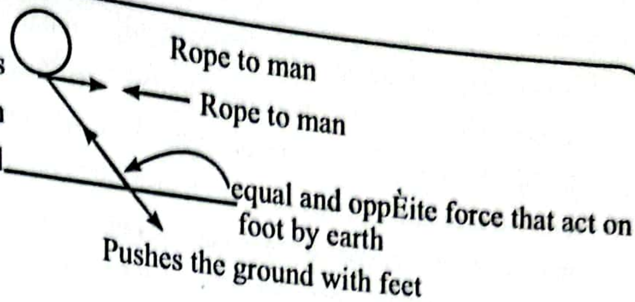


From this even you can understand from general knowledge that the tension of the rope cannot be taken as one value and the tension is maximum in the middle of the rope (near the ring). This is equivalent to a string which has weights at different places. According to the figure, if the tension parts of the string is T_1, T_2 and T_3, then it is easy to understand that $T_1 = 10$ N, $T_2 = 20$ N and $T_3 = 30$ N.

As the string is vertical, if there is a weight, then it should also be considered. Here I have neglected the weight of the string. If a part of the string is cut and consider its equilibrium, then $T_2 = 10 + T_1$ but $T_1 = 10$ N, therefore, $T_2 = 20$ N.

the last person by the end of the rope.

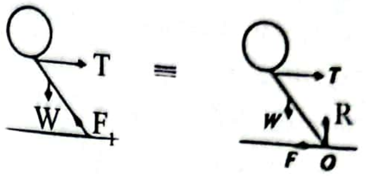
If you only consider the man, then the forces acting on himself is shown here. The force from the floor to the feet of the man has been resolved into horizontally and vertically.



R = perpendicular reaction

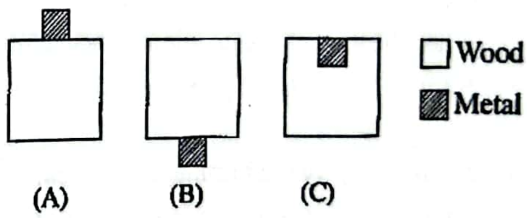
F = frictional force

As I have always mentioned, the force of F_1 only acts on the feet of the man. When it is resolved, the vertical force is called as the perpendicular reaction and the horizontal force is called as the frictional force. If the forces are in equilibrium on man, then R =W (this is most of the time true) and T =F. The maximum value that F can take is F = μ R. It can be applied for the limiting (maximum) frictional force only. Therefore, it is true that the maximum applied force is dependent upon μ. Other than that, the applied force is dependent upon the frictional force is also true. There is an important fact that is indicated from this. On a smooth floor, the tug of war cannot be done even how many strong men you have as they tend to slide. To increase the frictional force, you can wear necessary slippers. Some people dig small holes on the floor and keep the feet in them and pull the rope. Then the magnitude of F_1 can be increased. But it is against the rules of the competition.



If it is mentioned that the party who can strongly apply thrust on the floor can win from tug of war, then I do not see anything wrong in that statement. Even the friction of the rope is also important. If the rope is smooth, then the hands slid away. In addition, if the moment from the weight around the point O is greater than the moment created from T, then the person falls down. Most of the time, this happens when you win.

47. Figure shows three objects (A), (B) and (C) which are made using three uniform wooden cubes of identical dimensions made out of the same material, and three identical uniform metal cubes. In (A) and (B) the metal cubes are glued onto the top and the bottom of the wooden cubes respectively. In (C), the metal cube is (A) embedded in the wooden cube, as shown in the figure.



The three objects (A), (B) and (C) are now slowly lowered without changing their orientation and made to float vertically in a pool of water. If the depths to which the wooden cubes are immersed in water are Mfg, HB and ftp respectively, which of the following relationsNps is true?

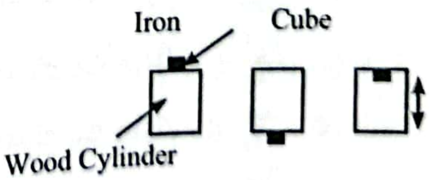
(1) $H_A > H_B > H_C$ (2) $H_A = H_B > H_C$

(3) $H_A = H_B = H_C$ (4) $H_C > H_B > H_A$

(5) $H_A > H_C > H_B$

Hydrostatics 02

As shown form the figure, three uniform and identical wooden cylinders are floating in water. Three identical iron cubes are there on the top, at the bottom and inside of it according to the figure. If the sunk depths of the wooden cylinders in water are h_1, h_2 and h_3 respectively, then what is the relation between these three quantities?



It can be just understood that the combined weight of the first and second cylinder are equal. But in the second cylinder, the iron cube is inside the water. In the first cylinder, the iron cube is not inside the cylinder.

Therefore, an extra upthrust is given by the cube in the second cylinder. So, the second cylinder is not sunk very much compared to the first cylinder. Therefore, $h_1>h_2$. From this relation, other relations of $h_1= h_2> h_3$, $h_1= h_2= h_3$ and $h_3> h_2> h_1$ are removed. What is left will be only $h_1> h_2> h_3$ and $h_1> h_3> h_2$. Even if you pick a blind choice, then all you have to choose either one of these. If you think of the first and third cylinders, then the difference between them is that the cube is inside the third cylinder. For that, a certain wooden volume has to be removed. Therefore, the weight of the first cylinder should be greater than the weight of the third cylinder by a certain amount. Due to both cubes, there is no created upthrust. They are not sunk in water. Therefore, it should be $h_1> h_3$.

Now we need consider cylinder (2) and (3). The weight of (2) is higher than (3). But by considering only weight we cannot take as $h_2> h_3$. Why? Due to the iron cube, there is an upthrust in (2). The iron in (3) is not creating an upthrust. Therefore, the sunken depth can be decided by comparing the upthrust on the iron cube by (2) and the weight of the wood in (3). If V is the volume of the iron cube, then the upthrust ↑= Vd_1g (d_1 = the density of water). The wooden volume that should be removed when the iron cube is inserted is also V. But the weight of the removed wood = Vd_2g (d_2 = the density of wood)

If we think that if the density of water is greater than the density of wood, then $d_1>d_2$. So, $Vd_1g> Vd_2g$

That means the upthrust on the iron cube is greater than the weight of the removed wood. Therefore, $h_3> h_2$.

Now $h_1> h_2$, $h_1> h_3$ and $h_3> h_2$. So, the correct inequality is $h_1> h_3> h_2$. You need to consider that the density of wood is less than the density of water. This is normally true. If you do not consider as $d_1> d_2$, you cannot get the answer.

Do not try to go and write equations. Always try to get the answer by logical arguments. Even it is not needed, you can write the below equations for (2) and (3). W = the mass of the wooden cylinder, m= the mass of the iron cube and A = the cross-sectional area of the cylinder.

For (2), $(W + m)\, g = Vd_1g + Ah_2d_1g$

For (3), $(W - Vd_2 + m)\, g = Ah_3d_1g$

When the second equation is subtracted from the first, $Vd_2= Vd_1+ Ad_1(h_2- h_3)$

$V (d_2 - d_1) = Ad_1(h_2- h_3)$ → If $d_1> d_2$, then it should be $h_3> h_2$. If $d_2> d_1$, then $h_2> h_3$. If so, the correct one is $h_1> h_2> h_3$.

48. An infinitely long thin straight wire held perpendicular to the plane of the paper at point 0 carries a current I into the paper as shown in the figure. Nine other similar infinitely long wires, parallel to the above wire, and held on the circumference of a circle of radius r with centre at point 0, each carries a current I into the paper. Except for wires A and B, the angular separation between any two consecutive wires is 30° as shown. The magnitude and the direction of the magnetic force per unit length on the wire held at the centre d due to other wires are, (Take $\cos 30° = \frac{\sqrt{3}}{2}$.)

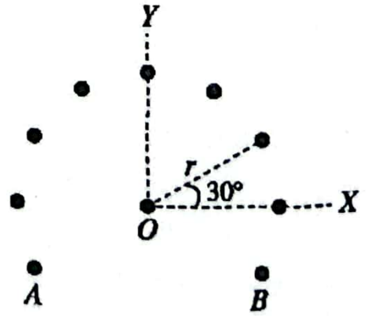
(1) $\frac{\mu_0 I^2}{2\pi r}(1+\sqrt{3})$ in the direction of *YO*.

(2) $\frac{\mu_0 I^2}{2\pi r}(1+\sqrt{3})$ in the direction of *OY*.

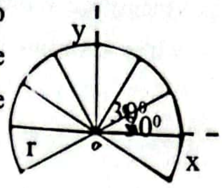
(3) $\frac{\mu_0 I^2}{\pi r}(1+\sqrt{3})$ in the direction of *OY*.

(4) $\frac{\mu_0 I^2}{2r}(1+\sqrt{3})$ in the direction of *OX*.

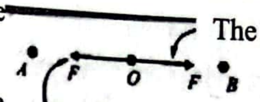
(5) $\frac{3\mu_0 I^2}{2\pi r}$ in the direction of *YO*.



The figure shows ten thin straight wires with infinite length which carry a current I into the paper. Nine wires are fixed at radius of r circle's circumference and the other wire goes across the centre of the circle O. Due to nine wires, what is the magnitude and the direction of the magnetic force per unit length in the wire across O?



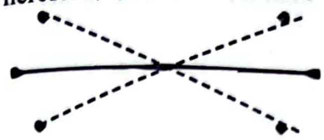
The magnetic force is attractive between two wires which has current to the same direction. Even there are many wires in this arrangement, you do not need to consider all of them. Relative to O, the currents that are in the opposite direction with each other cancel the forces acting on the wire across O. Because these forces are acting on equal and opposite directions.



The Force from B to O

The Force from A to O

Therefore, you can remove the wires which carry opposite currents to each other. What is left will be the upper three wires that are not in the opposite direction.

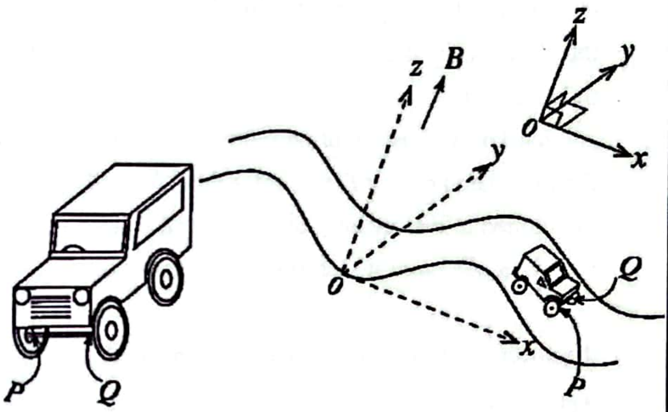


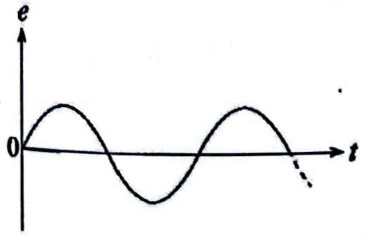
Due to the horizontal components of the inclined forces, they are being cancelled off. Therefore, the resultant force on the wire at O = F + 2F cos 30° ↑ OY direction; The force per unit length of the wire at O = IB (IIB, l =1)

$B = \frac{\mu_0 I}{2\pi r}$

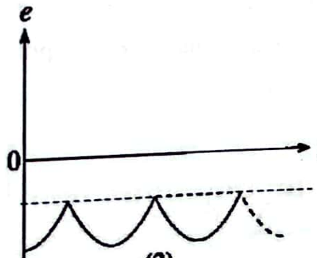
Therefore, net force = $\frac{\mu_0 I^2}{2\pi r}(1 + 2\cos 30^0) == \frac{\mu_0 I^2}{2\pi r}(1 + \sqrt{3})$

49. A toy car shown in figure (a) having an isolated metal axle *PQ*, travels with a constant speed *v* along a sinusoidal path whose vertical cross-section is in *zx* plane as shown in figure (b). At time *t* = 0, the axle *PQ* coincides with the *y* axis. If a uniform magnetic field of flux density *B* exists throughout the region in the +*z* direction and normal to the *xy* plane, the variation of the induced e.m.f.(*e*) at end *P* of the axle with respect to end *Q*, with time (*t*) is best represented by, (Neglect the effect of the Earth magnetic field.)

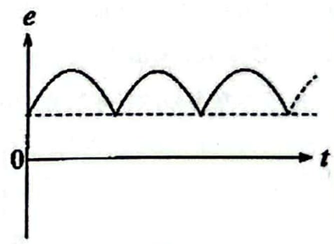




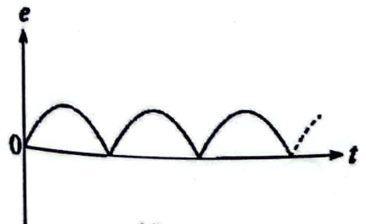
(1)



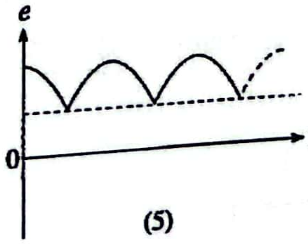
(2)



(3)

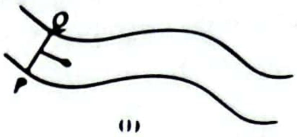


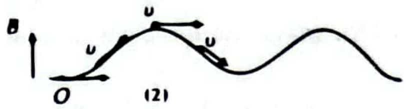
(4)



(5)

According to the figure (1), a thin metal rod of PQ is travelling in a uniform v speed in a wide wavy road with a sinusoidal shape. Figure (2) shows a vertical cross- section of the system.

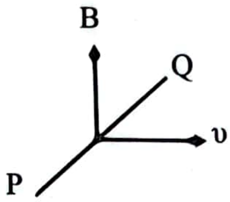

(1)



(2)

If there is a uniform magnetic field with flux density of B vertically upwards throughout the region and when t =0, the rod starts its motion from point O, then which graph shows the variation of the induced e. m f of P end relative to Q end with time (t)?

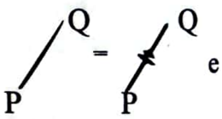
If cars or rotating swings are connected, then the question becomes more beautiful. But Physics part is there to find the variation of induced e. m. f on a metal rod that moves in a sinusoidal path when a vertical uniform magnetic field is applied.



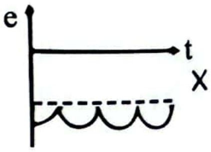
First, consider the moment where v and B are perpendicular to each other (at position O). The generated e. m. f is vlB (I = length of the rod). Keep the right-hand thumb perpendicular to the other fingers and rotate the fingers from the direction of v to B. The direction that the right-hand thumb indicates the direction of the induced e. m. f.

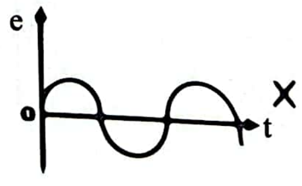
As you know, I use right-hand rule to decide all the directions according to magnetic fields. You can use the right-hand rule of Fleming for this. According to this the induced e. m f is acting from Q to P side. As I mentioned in the reviews, think of a cell instead of a rod. Then the sign of the induced e. m. f can be easily found on one point with respect to another point.

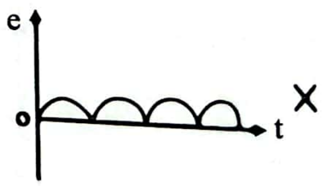
Accordingly, P end is positive relative to Q end. So, remove the variation which has drawn as negative.



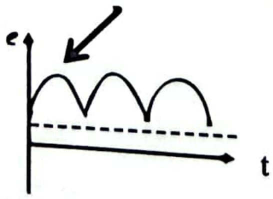
When the rod is at point O, the induced e. m. f should take a maximum value. Why? Because v and B are perpendicular to each other (vlB). At point O, e cannot be zero. From this, these two variations are removed.

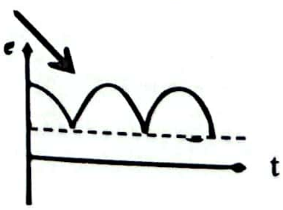






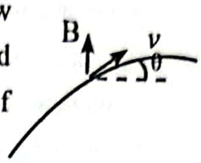
Initially, e takes a minimum value here. Remove this also. Then this will only be left.





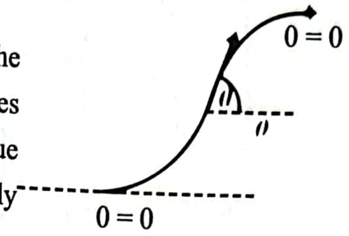
I wonder if it is not good to find the answer in such an easy way! As I get the blame, we will do the work well.

Consider a fair moment where the rod is kept in the sinusoidal path. In such a place, how do you find the induced e. m. f on the rod? We will resolve v into horizontal ($v \cos \theta$) and vertical ($v \sin \theta$) direction. As $v \sin \theta$ is in the direction of B, there is no induced e. m. f from it. Therefore,

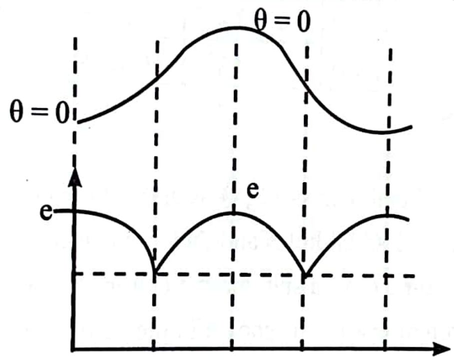


$e = lB\, v \cos \theta$. If you study the variation of θ ($\cos \theta$) when the rod is moving along the path, then you are done.

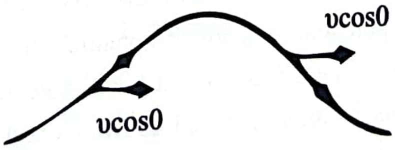
Can you see that the value of θ increases when you move from the bottom to the middle in the sinusoidal path and the value of θ decreases when you go from the middle to the peak of the mountain? The value of $\theta = 0$ at the bottom and at the peak. From zero, θ should be gradually increased and then it should be gradually decreased.



If so, then the value of $\cos \theta$ should gradually decrease (as θ increases), and after the middle of the path the value of $\cos \theta$ should increase (as θ decreases). When $\theta = 0$ at the bottom, $\cos \theta = 1$ (maximum e); Again, at the top of the path as $\theta = 0$, then $\cos 0 = 1$ (maximum e); The minimum value of e is in the middle of the path (θ maximum $\rightarrow \cos \theta$ minimum). Now the work is done. Look at the following figure.



The value of e will never be zero. It varies between the maximum and minimum. The direction of e also cannot be changed. Even the rod climbs the mountain or descends, the direction of $v \cos \theta$ is not changed. The direction gets reversed in $v \sin \theta$. But there is no induced e. m. f from that component. Do not get scared once you see such questions. First, find the direction of the induced e. m. f. Then you can find whether one end of the rod is positive or negative relative to the other end. Next, consider the moments where you can have the maximum and minimum instances of the induced e. m. f. Then without redesigning the whole shape of the variation, you can judge the correct variation. The value of θ may not be a zero at the bottom of the path in a rod of a vehicle.



50. A, B, C and D represent vertical cross sections of four identical parallel rectangular metal plates placed normal to the plane of the paper. Each of the plates B, C, and D has a small hole at its centre. The three plates are arranged so that their holes are coaxially placed as shown in figure (a). Plate A is grounded and the entire system is in a vacuum. A stationary electron is created at time t – 0, at the position 0 on the axis through the holes as shown. Which of the voltages VB, Vp and VD should be applied to the plates to obtain the velocity (o) – time (i) curve shown in figure (b) for the electron? (Assume that the given voltages are suitable for practical use and that the edge effects and gravitational effects can be neglected.)

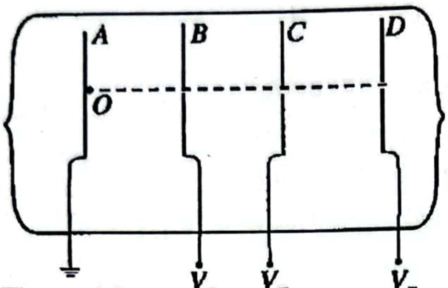


Figure (a)

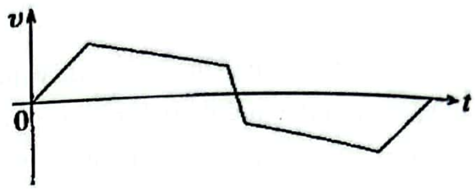
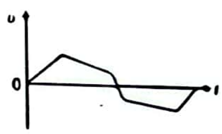


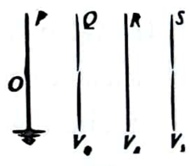
Figure (b)

	V_B	V_C	V_D
(1)	-3 kV	$+2.6$ kV	0 V
(2)	$+2.5$ kV	-2.6 kV	$+3$ kV
(3)	$+2.5$ kV	$+2.4$ kV	$+200$ V
(4)	$+3$ kV	$+2.6$ kV	-2.8 kV
(5)	$+3$ kV	$+3.2$ kV	-2.2 kV

06 *Electrostraic Potential*

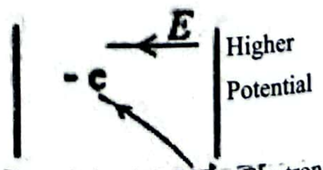

(2)

As shown in figure (1), four metal plates of P, Q, R and S have been kept in a vacuum. The middle of Q, R and S have small holes and they are placed in the same axis. The plate of P is earthed and at time t =0, a stationary electron is generated from point O. If the motion of the electron afterwards is shown in the velocity (v)- time (t) graph in figure (2), then what VQ, VR and VS voltages should be applied to the plates?


(1)

First, look at the v-t curve. Initially, the electron is accelerated and then it is subjected to a certain deceleration. Next, it undergoes a rapid deceleration and the velocity quickly becomes a zero where it turns back again (the velocity gets negative). Also, v-t curve is also symmetric around the time axis.

As there is a negative charge to the electron, if it has to be subjected to acceleration, then it should move from a lower potential to a higher potential. If you think in another way, then the electron should go along the direction which is opposite to the direction of the electric field. If it is a positive charge, then to accelerate it should go towards the direction of the electric field. Likewise, if an electron has to be decelerated, then it should move from a higher potential to a lower potential. If we summarize, then for acceleration it should go from a lower potential to a higher potential and for a deceleration it should go from a higher potential to a lower potential. Keep these facts in mind and go across each answer sets. Some answers can be removed.



1) V_Q = -3 kV, V_R = +2.6 kV and V_S = 0 V. VQ takes a negative value. V_Q<0 → from a higher potential to a lower potential→

It does not accelerate. Do not look at V_R and V_S values. Remove this choice. However, as the electrons are emitted from a zero potential, if the electron is accelerated when the eletron is travelling from P to Q, then you can just decide that V_Q potential should be positive.

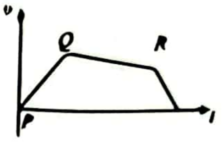
2) V_Q = +2.5 kV, V_R = -2.6 kV and V_S = +3 kV. V_Q >0 (accelerated), $V_Q > V_R$ (decelerated) and $V_R <$ V_S (it is not decelerated). Remove the answer (2).

3) V_Q = +2.5 kV, V_R = +2.4 kV and V_S = + 200 V. V_Q >0 (accelerated), $V_Q > V_R$ (decelerated) and $V_R > V_S$ (decelerated). The answer seems correct. There is a big chancw that this choice can be selected. As it is the 50th question, let us look at the other answers. Eventhough it is seen as correct, there is an upset issue here. We will look at that later.

4) V_Q = +3 kV, V_R = +2.6 kV and V_S = -2.8 kV. V_Q >0 (accelerated), $V_Q > V_R$ (decelerated) and $V_R >$ V_S (decelerated). The answer also seems correct.

5) V_Q = +3 kV, V_R = +3.2 kV and V_S = 2.2 kV. V_Q >0 (accelerated), $V_Q < V_R$ (it is not decelerated). Remove it. (from a lower potential to a higher potential→ It should accelerate) From these arguments, seems that both (3) and (4) are correct. When we look into these answers, there is not much difference in the values that are given for V_Q and V_R but V_S in (3) is +200 V (not kV). But V_S in (4) is -2.8 kV.

When you take a look at the v-t curve, it is clear that the electron is subjected to a rapid deceleration as it is removed from plate R. Therefore, V_R-V_S value (voltage difference) should be a greater value.



In (3), $V_R - V_S$ = (2.4 – 0.2) kV = 2.2 kV (200 V = 0.2 kV)

In (4), $V_R - V_S$ = [+ 2.6 – (-2.8)] kV = 5.4 kV

The difference of V_R-V_S value in (4) is greater than (3). Therefore, the most suitable choice is (4) not (3). You may not be tempted to think like this way. Both (1) and (2) are wrong. As (3) seems ok, you may not look at the choices of (4) and (5). However, 98% of our children blindly pick the answer for such questions. This question can be analyzed more. When looking at v-t curve, the electron should turn back before it tries to come to plate S (in any place in between R and S plates) or near to the S plate. If it goes away from plate S, then there is no chance for a turn back. But according to v-t graph, the electron has turned back again. If so, then in between → any place of R and S, the electron has quickly come to a rest and turn back.

Here you can do an investigation regarding the conservation of energy. The electron gets the energy only in between the plates of P and Q. As it decelerates afterwards, the kinetic energy is lost. Now let us go to the choice (3). V_Q = 2.5 kV, the increment of the kinetic energy of the electron when it goes from plate P to Q = 2.5q kJ (q = the charge of the electron numerically)

The kinetic energy loss when it goes from Q to R = q (2.5-2.4) = 0.1q kJ

If we think that the electron is going near to plate S, then the kinetic energy loss when it goes from R to S = q (2.4- 0.2) = 2.2q kJ

Therefore, the total kinetic energy loss when the electron goes from Q to S = (0.1q + 2.2q) kJ = 2.3q kJ

This is less than the gained kinetic energy of 2.5q kJ (2.3< 2.5).

It indicates that, there is a certain kinetic energy for the electron when it comes near to the hole of plate S (2.5q kJ -2.3q kJ = 0.2q kJ). If so, then the electron cannot turn back. It goes away from the hole of S. Afterwards, it goes in a uniform speed.

Now, let us look at (4).

As before, the gained kinetic energy of the electron when it goes from plate P to Q = 3q kJ

The kinetic energy loss when it goes from Q to R = (3- 2.6) q kJ = 0.4 q kJ

If we think that the electron is going to the near to plate S, then the kinetic energy loss when it goes from R to S = 2.6q(-2.8q) kJ = 5.4 q kJ

Therefore, the total kinetic energy loss when the electron goes from Q to S = (0.4q + 5.4q) kJ = 5.8q kJ

This is less than the gained kinetic energy of 3q kJ. According to the conservation of energy, this cannot happen. What can we decide from this? The electron gets its kinetic energy zero, before it comes to plate S. As the kinetic energy loss from Q to R is 0.4 q, then the kinetic energy when the electron is departing from the hole of R = (3q -0.4q) kJ = 2.6q kJ.

This kinetic energy is lost when the electron comes in between R and S plate. If it goes from R to S, then there is a loss of 5.4q kJ. But when the electron goes away from the hole of R, the kinetic energy it has which is 2.6q is less than the half of 5.4q. Therefore, definitely the electron is turning back in between R and S. Therefore, the correct answer is (4) not (3).

The path of the electron can be shown like this way.

However, this argument in conservation of energy is not remembered when doing the question as that logic is long. Therefore, this is the short argument that I suggest.

(1) As mentioned initially, the electron is accelerated from P to Q. From Q to R as well as from R to S, the electron is decelerated. If so, then it should be $V_Q > 0$; $V_Q > V_R$ and $V_R > V_S$. Then what is left after the argument are +2.5 kV, 2.4 kV, +200 V and +3 kV, +2.6 kV, -2.8 kV.

(2) To find the correct value out of the two choices, you can use this simple logic. The deceleration after R is greater than the deceleration from Q to R. The velocity is quickly reduced and gets zero. Therefore, the difference of V_R-V_S should be with a greater value.

For (3), $V_R - V_S$ = 2.2 kV and for (4), $V_R - V_S$ = 5.4 kV.

If we look at the v-t curve, then you can clearly see that the electron turns back. It can be also seen that the turning back happens when the electron passes the plate R. According to the conservation of energy as mentioned before, the gained kinetic energy of the electron should be completely lost.

The kinetic energy is increased when the electron travels from the earthed plate to the plate Q. This kinetic energy value is proportional to 2.5 according to (3).

When it is going from Q to R, the kinetic energy is reduced. The reduced amount is proportional to (2.5- 2.4) 0.1. Even the electron travels from R to S, the kinetic energy loss from R to S is proportional to (2.4 -0.2) 2.2.

As 2.5>0.1+2.2, the electron is not stopped. It goes out from the hole of plate S. If so, then the electron does not obey the given v-t curve. In (4), 3< 0.4+ 5.4. The electron is stopped in between R and S. It happens due to the deceleration due to the electric force created opposite to the direction of motion. The velocity quickly gets zero and get accelerated backwards. When coming backwards, the electron is accelerated from the middle of R and S to R and then from R to Q (to the negative direction of velocity). From Q to P the electron is decelerated and the velocity gets zero when it reaches the plate P. If there is no energy loss, then this process happens continuously.

Always you need to use the removal of answers by looking each of them when solving such questions. You cannot get the correct answer by doing calculations separately. You need to understand the logic and pick the correct answer by going across each of the given choices.