## Appendix 11

## Solutions of Some Exercises

## A11.1 Chapter 1

## Exercise 1.4

In a dust universe with curvature and with a cosmological constant the Friedmann equation can be written in the form

$$
\begin{equation*}
\dot{a}^{2}=a^{2}\left[-K+\frac{C}{a}+\frac{1}{3} \Lambda a^{2}\right] \equiv G(a) . \tag{A11.1}
\end{equation*}
$$

Here

$$
C=\frac{8 \pi G}{3} \rho_{m} a^{3}=\Omega_{m} H_{0}^{2} a_{0}^{3}= \begin{cases}\frac{\Omega_{m}}{H_{0}\left|\Omega_{k}\right|^{3 / 2}}=\frac{2 q_{0}}{H_{0}\left|1-2 q_{0}\right|^{3 / 2}} & \text { if } \Omega_{k} \neq 0  \tag{A11.2}\\ \Omega_{m} H_{0}^{2} & \text { if } \Omega_{k}=0 \\ \text { and } a_{0}=1\end{cases}
$$

If the curvature is negative and $\Lambda>0, G$ is strictly positive and we find an expanding solution for all times. At late times, curvature becomes negligible and the universe expands like $a \propto 1 /|t| \propto \exp (\sqrt{\Lambda / 3} \tau)$. If $\Lambda<0$ the square bracket is decreasing and $G$ has a zero, $G\left(a_{c}\right)=0$. At this point expansion turns into contraction and the universe recollapes.

The case $K=0$ can be solved explicitly, leading to

$$
a^{3}(\tau)= \begin{cases}\frac{3 C}{2 \Lambda}(\cosh (\sqrt{3 \Lambda} t)-1) & \Lambda>0, a_{\min }=(3 C / 2 \Lambda)^{1 / 3}  \tag{A11.3}\\ \frac{-3 C}{2 \Lambda}(1-\cos (\sqrt{-3 \Lambda} t)) & \Lambda<0\end{cases}
$$

The qualitative behavior is like for $K<0$.
The case $K>0$ is most interesting. The function $G$ can be written as $G(a)=a P(a)$, where $P$ is a third-order polynomial that has one or three real roots. In the dashed region of Fig. A11.1, $P$ has one real root, but for a negative value of $a$. Hence the universe expands forever. In the upper left region, with a high cosmological constant, the scale factor has a minimum. Such a universe has no big bang but comes out of a previous contracting phase. It is called a bouncing solution. For a value of $\Omega_{m}>0.01$ one finds a maximum redshift $z_{\max }<4$ for a bouncing universe. Hence they cannot explain cosmological data like quasars and galaxies at a redshift of 6 or even the CMB. Solutions below the dashed region


Fig. A11.1 The kinematics of a universe with matter density parameter $\Omega_{m}$ and cosmological constant parameter $\Omega_{\Lambda}$. The universes with parameters above the dashed line are positively curved, those below negatively. The universes with values ( $\Omega_{m}, \Omega_{\Lambda}$ ) in the dashed region emerge from a big bang and expand forever. Those below emerge from a big bang and recollapse into a big crunch, and those above emerge from a collapsing universe; they have no big bang in the past.
emerge from a big bang but recollapse eventually, when either the negative cosmological constant or the positive curvature term render $G\left(a_{\max }\right)=0$.

## A11.2 Chapter 2

## Exercise 2.1

We want to show that

$$
\begin{align*}
L_{X} g=a^{2} & {\left[-2\left(\frac{\dot{a}}{a} T+\dot{T}\right) d t^{2}+2\left(\dot{L}_{i}-T_{, i}\right) d t d x^{i}\right.} \\
& \left.+\left(2 \frac{\dot{a}}{a} T \gamma_{i j}+L_{i \mid j}+L_{j \mid i}\right) d x^{i} d x^{j}\right] \tag{A11.4}
\end{align*}
$$

for $X=T \partial_{t}+L^{i} \partial_{i}$ and $g=a^{2}(t)\left[-d t^{2}+\gamma_{i j} d x^{i} d x^{j}\right]=a^{2}(t) S_{\mu \nu} d x^{\mu} d x^{\nu}$.
We use $L_{X} a^{2}=2 \dot{a} a T$ and $L_{X}\left(a^{2} S\right)=L_{X}\left(a^{2}\right) S+a^{2} L_{X} S$. Furthermore, we show in the text that follows that for an arbitrary metric $S$, we have

$$
\begin{equation*}
\left(L_{X} S\right)_{\mu \nu}=X_{\mu ; \nu}+X_{\nu ; \mu} \tag{A11.5}
\end{equation*}
$$

where here ; denotes the covariant derivative w.r.t. the metric $S$. For our metric $S$ all Christoffel symbols involving a " 0 " vanish, so that $X_{\nu ; 0}=X_{\nu, 0}$ and $X_{0 ; \nu}=X_{0, v}$. Furthermore $X_{i ; j}=X_{i \mid j}$, where ${ }_{\mid}$denotes the covariant derivative w.r.t. the three-dimensional metric $\gamma$. With this we obtain

$$
\begin{equation*}
L_{X} g=2 \frac{\dot{a}}{a} T a^{2} S+a^{2}\left(-2 \dot{T} d t^{2}-2\left(T_{, i}-\dot{L}_{i}\right) d t d x^{i}+\left(L_{i \mid j}+L_{j \mid i}\right) d x^{i} d x^{j}\right) \tag{A11.6}
\end{equation*}
$$

which agrees with Eq. (A11.4). It remains to show Eq. (A11.5). For this we use the general expression (A2.20). For a doubly covariant tensor field this gives

$$
\begin{aligned}
\left(L_{X} S\right)_{\alpha \beta} & =X^{\mu} S_{\alpha \beta},{ }_{\mu}+X^{\mu},{ }_{\alpha} S_{\mu \beta}+X^{\mu},{ }_{\beta} S_{\mu \alpha} \\
& =X_{v}\left(S^{\mu v} S_{\alpha \beta}, \mu+S^{\mu v}{ }_{\alpha} S_{\mu \beta}+S^{\mu \nu},{ }_{\beta} S_{\mu \alpha}\right)+X_{\alpha, \beta}+X_{\beta, \alpha} .
\end{aligned}
$$

For the last equals sign we simply inserted $X^{\mu}=X_{\nu} S^{\nu \mu}$. We now take the derivative of the identity $S^{\nu \mu} S_{\mu \beta}=\delta_{\beta}^{\nu}$ w.r.t. $\alpha$. This yields $S^{\mu \nu}{ }_{\alpha} S_{\mu \beta}=-S^{\mu \nu} S_{\mu \beta, \alpha}$. Correspondingly $S^{\mu \nu}{ }_{, \beta} S_{\mu \alpha}=-S^{\mu \nu} S_{\mu \alpha, \beta}$. Inserting this above and using the definition

$$
X_{\alpha ; \beta}=X_{\alpha, \beta}-\Gamma_{\alpha \beta}^{\mu} X_{\mu} \text { with } \quad \Gamma_{\mu \nu}^{\beta}=\frac{1}{2} S^{\beta \alpha}\left(S_{\mu \alpha, \nu}+S_{\nu \alpha, \mu}-S_{\mu v, \alpha}\right),
$$

we obtain (A11.5).

## Exercise 2.3

In synchronous gauge ( $A=B=0$ ) we have

$$
\begin{align*}
& \Psi=-k^{-1}(\mathcal{H} \sigma+\dot{\sigma}) \quad \text { and }  \tag{A11.7}\\
& V=v-\sigma \tag{A11.8}
\end{align*}
$$

For a pure dust universe Eq. (2.119) reduces to

$$
\begin{equation*}
\dot{V}+\mathcal{H} V=k \Psi \tag{A11.9}
\end{equation*}
$$

Inserting the expressions above this yields

$$
\begin{equation*}
\dot{v}+\mathcal{H} v=0 \tag{A11.10}
\end{equation*}
$$

which only has a decaying solution, $v \propto 1 / a$, that is, the only possible nondecaying solution is $v=0$.

## Exercise 2.4

We consider a perturbed FL universe containing two noninteracting fluids with energy densities $\rho_{\alpha}$ and pressure $P_{\alpha}$. The total energy density and pressure are $\rho=\rho_{1}+\rho_{2}$ and $P=P_{1}+P_{2}$. We first note that for both components the intrinsic entropy perturbation is given by

$$
\begin{equation*}
\Gamma_{\alpha}=\pi_{L}^{(\alpha)}-\frac{c_{\alpha}^{2}}{w_{\alpha}} \delta_{\alpha}=\frac{\delta P_{\alpha}}{P_{\alpha}}-c_{\alpha}^{2} \frac{\delta \rho_{\alpha}}{P_{\alpha}} \tag{A11.11}
\end{equation*}
$$

and the total sound speed is

$$
\begin{equation*}
c_{s}^{2}=\frac{\dot{P}_{1}+\dot{P}_{2}}{\dot{\rho}}=\frac{c_{1}^{2} \dot{\rho}_{1}+c_{2}^{2} \dot{\rho}_{2}}{\dot{\rho}}=\frac{\left(1+w_{1}\right) c_{1}^{2} \rho_{1}+\left(1+w_{2}\right) c_{2}^{2} \rho_{2}}{(1+w) \rho} . \tag{A11.12}
\end{equation*}
$$

For the second equality sign we have used that both components are separately conserved.
Defining now $R=\rho_{2} / \rho$, so that $\rho_{1} / \rho=1-R$, we can also write

$$
\begin{equation*}
(1+w) c_{s}^{2}=\left(1+w_{1}\right) c_{1}^{2}(1-R)+c_{2}^{2}\left(1+w_{2}\right) R . \tag{A11.13}
\end{equation*}
$$

Let us first assume $\Gamma_{\alpha}=0$, so that $\delta P_{\alpha}=c_{\alpha}^{2} \delta \rho_{\alpha}$. The total entropy perturbation is then given by $\Gamma=\Gamma_{\text {rel }}$ with

$$
\begin{equation*}
P \Gamma_{\mathrm{rel}}=c_{1}^{2} \delta \rho_{1}+c_{2}^{2} \delta \rho_{2}-c_{s}^{2}\left(\delta \rho_{1}+\delta \rho_{2}\right)=\left(c_{1}^{2}-c_{s}^{2}\right) \delta \rho_{1}+\left(c_{2}^{2}-c_{s}^{2}\right) \delta \rho_{2} \tag{A11.14}
\end{equation*}
$$

To express $\Gamma_{\text {rel }}$ in terms of gauge-invariant variables we now use

$$
\delta \rho_{\alpha}=\left[D_{g}^{(\alpha)}+\left(1+w_{\alpha}\right)\left(3 H_{L}+H_{T}\right)\right] \rho_{\alpha}
$$

Inserting this in Eq. (A11.14) yields

$$
\begin{align*}
w \Gamma_{\mathrm{rel}}= & \left(c_{1}^{2}-c_{s}^{2}\right)(1-R) D_{g}^{(1)}+\left(c_{2}^{2}-c_{s}^{2}\right) R D_{g}^{(2)}+\left(3 H_{L}+H_{T}\right) \\
& \times\left[\left(c_{1}^{2}-c_{s}^{2}\right)(1-R)\left(1+w_{1}\right)+\left(c_{2}^{2}-c_{s}^{2}\right) R\left(1+w_{2}\right)\right] . \tag{A11.15}
\end{align*}
$$

Using Eq. (A11.13) and

$$
1+w=\frac{\rho+P}{\rho}=\frac{\rho_{1}+P_{1}+\rho_{2}+P_{2}}{\rho}=\left(1+w_{1}\right)(1-R)+\left(1+w_{2}\right) R,
$$

we find that the square bracket above vanishes and $\Gamma_{\text {rel }}$ is gauge invariant, as it should be. In fact, with the relation (A11.13)

$$
[\quad]=c_{1}^{2}(1-R)\left(1+w_{1}\right)+c_{2}^{2} R\left(1+w_{2}\right)-c_{s}^{2}(1+w)=0
$$

Multiplying Eq. (A11.15) with $1+w$ and using Eq. (A11.13) to replace $c_{s}^{2}$ finally leads to

$$
\begin{equation*}
w(1+w) \Gamma_{\mathrm{rel}}=R(1-R)\left(c_{1}^{2}-c_{2}^{2}\right)\left[\left(1+w_{2}\right) D_{g}^{(1)}-\left(1+w_{1}\right) D_{g}^{(2)}\right] \tag{A11.16}
\end{equation*}
$$

From this equation we already conclude that $\Gamma_{\text {rel }}$ vanishes if both sound speeds are equal, $c_{1}^{2}=c_{2}^{2}$, or if one of the two components is largely subdominant, $R \simeq 0$ or $R \simeq 1$. If neither of these conditions is fulfilled, perturbations are adiabatic if

$$
\begin{equation*}
\left(1+w_{2}\right) D_{g}^{(1)}=\left(1+w_{1}\right) D_{g}^{(2)} \quad \text { (adiabaticity) } \tag{A11.17}
\end{equation*}
$$

To determine $\Gamma$ when $\Gamma_{\alpha} \neq 0$ we simply note that in this case $\delta P_{\alpha}=P_{\alpha} \Gamma_{\alpha}+c_{\alpha}^{2} \delta \rho_{\alpha}$ so that

$$
P \Gamma=P_{1} \Gamma_{1}+P_{2} \Gamma_{2}+P \Gamma_{\text {rel }} .
$$

Inserting our result for $\Gamma_{\text {rel }}$ we find

$$
\begin{equation*}
\Gamma=\frac{w_{1}}{w}(1-R) \Gamma_{1}+\frac{w_{2}}{w} R \Gamma_{2}+\Gamma_{\text {rel }} . \tag{A11.18}
\end{equation*}
$$

We now want to derive an evolution equation for $\Gamma_{\text {rel }}$ in the case where $\Gamma_{\alpha}=0$ and $w_{\alpha}=$ $c_{\alpha}^{2}=$ constant for both components. We use the conservation equation (2.115), which in this case reduces to

$$
\begin{equation*}
\dot{D}_{g}^{(\alpha)}=-k\left(1+w_{\alpha}\right) V_{\alpha} . \tag{A11.19}
\end{equation*}
$$

Defining

$$
f=\frac{R(1-R)}{w(1+w)}\left(c_{1}^{2}-c_{2}^{2}\right),
$$

the derivative of $\Gamma_{\text {rel }}$ can be written as

$$
\begin{equation*}
\dot{\Gamma}_{\text {rel }}=\frac{\dot{f}}{f} \Gamma_{\text {rel }}+k f\left(1+w_{1}\right)\left(1+w_{2}\right)\left[V_{2}-V_{1}\right] \tag{A11.20}
\end{equation*}
$$

This shows that even if perturbations of a two-component fluid are initially adiabatic, they develop a relative entropy perturbation if $V_{1} \neq V_{2}$. This is already clear from the adiabaticity condition (A11.17), which cannot be maintained if $V_{1} \neq V_{2}$ due to the time evolution of $D_{g}^{(\alpha)}$ given in Eq. (A11.19). Especially on sub-Hubble scales, where $V_{1}$ and $V_{2}$ evolve differently (we consider the nontrivial case $c_{1} \neq c_{2}$ ), adiabaticity between different components cannot be maintained. When talking about adiabatic perturbations, we therefore always refer to super-Hubble scales.

## A11.3 Chapter 3

## Exercise 3.1

We want to show that only exponential potentials allow for power law inflation, $a \propto t^{q}$ with some constant $q$, and we want to express $q$ in terms of the parameters of the potential. We assume a spatially flat FL universe, $K=0$.

For a spatially flat FL universe, the Friedmann equation and energy-momentum conservation (or the first and second Friedmann equations) imply

$$
\dot{\mathcal{H}}=-\frac{1+3 w}{2} \mathcal{H}^{2}
$$

Now if $a \propto t^{q}$ we have $\mathcal{H}=q / t$ and $\dot{\mathcal{H}}=-q / t^{2}$. Inserting this above gives

$$
q=\frac{2}{1+3 w} \quad \text { hence } \quad w=\frac{2-q}{3 q}=\text { constant }
$$

From this we also conclude that inflation, that is, $w<-1 / 3$, is obtained if and only if $q<0$. Hence for an expanding and inflating universe with an expansion law of the form $a \propto\left(t / t_{0}\right)^{q}$ we have to choose $t$ and $t_{0}$ negative in order for $a$ to increase with $t$. That is, conformal time is negative during inflation.

Furthermore, integrating $d \tau=a d t \propto t^{q} d t$ yields $\tau \propto t^{q+1}$; hence

$$
a \propto \tau^{p} \quad \text { with } \quad p=\frac{q}{q+1}=\frac{2}{3+3 w} .
$$

Since

$$
w=P / \rho=a^{2} P /\left(a^{2} \rho\right)=\frac{\frac{1}{2} \dot{\varphi}^{2}-a^{2} W}{\frac{1}{2} \dot{\varphi}^{2}+a^{2} W}=\text { constant }
$$

and

$$
a^{2} \rho=\frac{1}{2} \dot{\varphi}^{2}+a^{2} W=3 M_{P}^{2} \mathcal{H}^{2} \propto 1 / t^{2}
$$

it follows that both $\frac{1}{2} \dot{\varphi}^{2}$ and $a^{2} W$ are proportional to $1 / t^{2}$. More precisely,

$$
\begin{align*}
& \dot{\varphi}=\sqrt{a^{2}(\rho+P)}=\sqrt{3(1+w)} M_{P} \mathcal{H}=\sqrt{3(1+w)} q \frac{M_{P}}{t}  \tag{A11.21}\\
& \varphi=M_{P} \sqrt{2 q(1+q)} \log \left(t / t_{*}\right) \tag{A11.22}
\end{align*}
$$

where $t_{*}$ is an integration constant. But also $W=(\rho-P) / 2$ is a power law in $t$. This is possible only if $W \propto \exp \left(-\alpha \varphi / M_{P}\right) \propto t^{-\alpha \sqrt{2 q(1+q)}}$. To determine $\alpha$ we use that $a^{2} W \propto$ $1 / t^{2}$; hence

$$
\begin{equation*}
t^{2 q-\alpha \sqrt{2 q(1+q)}} \propto t^{-2} \tag{A11.23}
\end{equation*}
$$

which implies

$$
\begin{equation*}
\alpha^{2}=\frac{2(1+q)}{q} \quad \text { or } \quad q(\alpha)=\frac{2}{\alpha^{2}-2} . \tag{A11.24}
\end{equation*}
$$

Inserting this into the expressions for $w$ and $p$ we find

$$
\begin{equation*}
w(\alpha)=\frac{\alpha^{2}-3}{3}, \quad p(\alpha)=\frac{2}{\alpha^{2}} . \tag{A11.25}
\end{equation*}
$$

The universe is inflating when $w<-1 / 3$; hence $\alpha^{2}<2$. De Sitter inflation is obtained in the limit $\alpha \rightarrow 0$.

## A11.4 Chapter 4

## Exercise 4.4

We start with Eq. (4.137), which yields

$$
\begin{aligned}
& \frac{1}{4 \pi} \sum_{\ell}(2 \ell+1) C_{\ell}^{(V)} P_{\ell}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right) \\
& \quad=\sum_{\ell} \frac{(2 \ell+1)^{2}}{(2 \pi)^{3}} \int d^{3} k M_{\ell}^{(V)}(k) P_{\ell}(\mu) P_{\ell}\left(\mu^{\prime}\right)\left(\mathbf{n} \cdot \mathbf{n}^{\prime}-\mu \mu^{\prime}\right)
\end{aligned}
$$

For the last factor we made use of Eq. (4.138). Before we continue we now show Eq. (4.194). The addition theorem of spherical harmonics yields

$$
\begin{aligned}
& \int d \Omega_{\hat{\mathbf{k}}} P_{\ell}(\mu) P_{\ell^{\prime}}\left(\mu^{\prime}\right) \\
& \quad=\frac{(4 \pi)^{2}}{(2 \ell+1)\left(2 \ell^{\prime}+1\right)} \sum_{m m^{\prime}} \int d \Omega_{\hat{\mathbf{k}}} Y_{\ell m}(\hat{\mathbf{k}}) Y_{\ell m}^{*}(\mathbf{n}) Y_{\ell^{\prime} m^{\prime}}^{*}(\hat{\mathbf{k}}) Y_{\ell^{\prime} m^{\prime}}\left(\mathbf{n}^{\prime}\right) .
\end{aligned}
$$

Using the orthogonality of spherical harmonics, this implies

$$
\begin{aligned}
\int d \Omega_{\hat{\mathbf{k}}} P_{\ell}(\mu) P_{\ell^{\prime}}\left(\mu^{\prime}\right) & =\delta_{\ell \ell^{\prime}} \frac{(4 \pi)^{2}}{(2 \ell+1)^{2}} \sum_{m} Y_{\ell m}^{*}(\mathbf{n}) Y_{\ell m}\left(\mathbf{n}^{\prime}\right) \\
& =\frac{4 \pi}{2 \ell+1} P_{\ell}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)
\end{aligned}
$$

For the last equals sign we have again applied the addition theorem. With the help of the recursion relation

$$
\mu P_{\ell}(\mu)=\frac{\ell+1}{2 \ell+1} P_{\ell+1}(\mu)+\frac{\ell}{2 \ell+1} P_{\ell-1}(\mu),
$$

we can now perform the angular integration,

$$
\begin{aligned}
\int & d^{3} k M_{\ell}^{(V)}(k) P_{\ell}(\mu) P_{\ell}\left(\mu^{\prime}\right)\left(\mathbf{n} \cdot \mathbf{n}^{\prime}-\mu \mu^{\prime}\right) \\
= & 4 \pi \int d k k^{2} M_{\ell}^{(V)}(k)\left[\frac{1}{2 \ell+1}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right) P_{\ell}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)\right. \\
& \left.-\frac{(\ell+1)^{2}}{(2 \ell+1)^{2}(2 \ell+3)} P_{\ell+1}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)-\frac{\ell^{2}}{(2 \ell+1)^{2}(2 \ell-1)} P_{\ell-1}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)\right] \\
= & \frac{4 \pi}{(2 \ell+1)^{2}} \int d k k^{2} M_{\ell}^{(V)}(k)\left[\frac{(\ell+1)(\ell+2)}{2 \ell+3} P_{\ell+1}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)+\frac{\ell(\ell-1)}{2 \ell-1} P_{\ell-1}\left(\mathbf{n} \cdot \mathbf{n}^{\prime}\right)\right] .
\end{aligned}
$$

Identifying the coefficient of $P_{\ell}$ finally results in

$$
\begin{equation*}
C_{\ell}=\frac{2 \ell(\ell+1)}{\pi(2 \ell+1)^{2}} \int d k k^{2}\left[M_{\ell+1}^{(V)}(k)+M_{\ell-1}^{(V)}(k)\right] \tag{A11.26}
\end{equation*}
$$

## A11.5 Chapter 5

## Exercise 5.3

We consider the following parameterization of a 2D tensor field:

$$
\begin{equation*}
T_{a b}=\alpha \delta_{a b}+\gamma \epsilon_{a b}+\left(\partial_{a} \partial_{b}-\frac{1}{2} \delta_{a b} \Delta\right) \varepsilon+\frac{1}{2}\left(\epsilon_{a c} \partial^{c} \partial_{b}+\epsilon_{b c} \partial^{c} \partial_{a}\right) \beta . \tag{A11.27}
\end{equation*}
$$

We want to show that every tensor field can be written in this form. Clearly, there are as many parameters on the right-hand side as there are components of $T_{a b}$, so this may work as a general parameterization. Note that in flat space raising and lower indices is done with $\delta_{a b}$, so it does not change anything.
(1) We first determine the parameters $\alpha$ to $\beta$ from $T_{a b}$. A straightforward calculation yields

$$
\begin{align*}
& \alpha=\frac{1}{2} \operatorname{trace} T=\frac{1}{2}\left(T_{11}+T_{22}\right)  \tag{A11.28}\\
& \gamma=\frac{1}{2} \epsilon^{a b} T_{a b}=\frac{1}{2}\left(T_{12}-T_{21}\right)  \tag{A11.29}\\
& \varepsilon=2 \Delta^{-2}\left(\partial^{a} \partial^{b} T_{a b}\right)-\Delta^{-1}\left(T_{11}+T_{22}\right)  \tag{A11.30}\\
& \beta=2 \Delta^{-2}\left(\epsilon_{a c} \partial^{c} \partial_{b} T^{a b}\right)-\Delta^{-1}\left(T_{12}-T_{21}\right) \tag{A11.31}
\end{align*}
$$

These equations have unique solutions $\alpha, \gamma, \varepsilon, \beta$ [we assume that our functions decay at infinity, e.g., that they are in $\left.L^{2}\left(\mathbb{R}^{2}\right)\right]$. Inversely we obtain

$$
\begin{align*}
& T_{11}=\alpha+\frac{1}{2}\left(\partial_{1}^{2}-\partial_{2}^{2}\right) \varepsilon+\partial_{1} \partial_{2} \beta  \tag{A11.32}\\
& T_{12}=\gamma+\partial_{1} \partial_{2} \varepsilon+\frac{1}{2}\left(\partial_{2}^{2}-\partial_{1}^{2}\right) \beta  \tag{A11.33}\\
& T_{22}=\alpha-\frac{1}{2}\left(\partial_{1}^{2}-\partial_{2}^{2}\right) \varepsilon-\partial_{1} \partial_{2} \beta  \tag{A11.34}\\
& T_{21}=-\gamma+\partial_{1} \partial_{2} \varepsilon+\frac{1}{2}\left(\partial_{2}^{2}-\partial_{1}^{2}\right) \beta \tag{A11.35}
\end{align*}
$$

Inserting the expressions for $\alpha, \gamma, \varepsilon, \beta$ shows that our identities are consistent.
(2) As $\epsilon_{a b}$ changes sign under parity and $\delta_{a b}$ as well as $\partial_{a} \partial_{b}$ do not, we find that for $T_{a b}$ to be a normal 2-tensor that does not change sign under parity, we must request that $\alpha$ and $\varepsilon$ are even under parity while $\gamma$ and $\beta$ change sign under parity. In other words, $\alpha$ and $\varepsilon$ are scalars while $\gamma$ and $\beta$ are pseudo-scalars.
(3) The polarization from Thomson scattering is a symmetric and traceless tensor; hence $\alpha=\gamma=0$ and it is of the form

$$
\begin{equation*}
\mathcal{P}_{a b}=\left(\partial_{a} \partial_{b}-\frac{1}{2} \delta_{a b} \Delta\right) \varepsilon+\frac{1}{2}\left(\epsilon_{a c} \partial^{c} \partial_{b}+\epsilon_{b c} \partial^{c} \partial_{a}\right) \beta \tag{A11.36}
\end{equation*}
$$

Using Eqs. (5.24) we have

$$
\begin{equation*}
\mathcal{E}=\partial_{a} \partial_{b} \mathcal{P}_{a b}-\epsilon_{c d} \epsilon_{a b} \partial_{c} \partial_{a} \mathcal{P}_{b d}=2 \partial_{a} \partial_{b} \mathcal{P}_{a b}=\Delta^{2} \varepsilon \tag{A11.37}
\end{equation*}
$$

For the second equality we used that in two dimensions $\epsilon_{c d} \epsilon_{a b}=\delta_{c a} \delta_{d b}-\delta_{d a} \delta_{c b}$ and that $P_{a b}$ is traceless. Using also the fact that $P_{a b}$ is symmetric, we find, inserting (5.25) for $\mathcal{B}$,

$$
\begin{equation*}
\mathcal{B}=-2 \epsilon_{b c} \partial_{a} \partial_{b} \mathcal{P}_{a c}=\Delta^{2} \beta \tag{A11.38}
\end{equation*}
$$

Therefore, the decomposition (A11.36) is entirely equivalent to the decomposition of the polarization into $\mathcal{E}$ and $\mathcal{B}$ modes.

## A11.6 Chapter 6

## Exercise 6.1

Because of statistical homogeneity, a 3-point function depends only on the differences $\mathbf{r}_{i j}=\mathbf{x}_{i}-\mathbf{x}_{j}$,

$$
\begin{equation*}
\left\langle X\left(\mathbf{x}_{1}\right) X\left(\mathbf{x}_{3}\right) X\left(\mathbf{x}_{3}\right)\right\rangle=\xi_{3}\left(\mathbf{r}_{12}, \mathbf{r}_{32}\right) \tag{A11.39}
\end{equation*}
$$

Here we use that $\mathbf{r}_{13}=\mathbf{r}_{12}-\mathbf{r}_{32}$ is not an independent variable. Fourier transforming this expression we obtain

$$
\begin{align*}
& \int d^{3} x_{1} d^{3} x_{2} d^{3} x_{3} e^{i\left(\mathbf{k}_{1} \mathbf{x}_{1}+\mathbf{k}_{2} \mathbf{x}_{2}+\mathbf{k}_{3} \mathbf{x}_{3}\right)}\left\langle X\left(\mathbf{x}_{1}\right) X\left(\mathbf{x}_{3}\right) X\left(\mathbf{x}_{3}\right)\right\rangle \\
& \quad=\int d^{3} r_{12} d^{3} x_{2} d^{3} r_{32} e^{i\left(\mathbf{k}_{1} \mathbf{r}_{12}+\mathbf{k}_{3} \mathbf{r}_{32}+\mathbf{x}_{2}\left(\mathbf{k}_{1}+\mathbf{k}_{2}+\mathbf{k}_{3}\right)\right)} \xi_{3}\left(\mathbf{r}_{12}, \mathbf{r}_{32}\right) \\
& \quad=(2 \pi)^{3} \delta\left(\mathbf{k}_{1}+\mathbf{k}_{2}+\mathbf{k}_{3}\right) \int d^{3} r_{12} d^{3} r_{32} e^{i\left(\mathbf{k}_{1} \mathbf{r}_{12}+\mathbf{k}_{3} \mathbf{r}_{32}\right)} \xi_{3}\left(\mathbf{r}_{12}, \mathbf{r}_{32}\right) \\
&  \tag{A11.40}\\
& =(2 \pi)^{3} \delta\left(\mathbf{k}_{1}+\mathbf{k}_{2}+\mathbf{k}_{3}\right) B\left(\mathbf{k}_{1}, \mathbf{k}_{3}\right) .
\end{align*}
$$

Since the first line of this equation as well as the Dirac delta are symmetric in $\mathbf{k}_{1}, \mathbf{k}_{2}$, and $\mathbf{k}_{3}$, this is also true for $B\left(\mathbf{k}_{1}, \mathbf{k}_{3}\right)$. Here we have suppressed the variable $\mathbf{k}_{2}=-\left(\mathbf{k}_{1}+\right.$ $\mathbf{k}_{3}$ ). We now want to show that $B$ depends only on the moduli $k_{i}=\left|\mathbf{k}_{i}\right|$. For this we use that the cosine of the angle between $\mathbf{k}_{1}$ and $\mathbf{k}_{3}$ is given by

$$
\begin{equation*}
\mu \equiv \frac{\mathbf{k}_{1} \cdot \mathbf{k}_{3}}{k_{1} k_{3}}=\frac{k_{2}^{2}-k_{1}^{2}-k_{3}^{2}}{2 k_{1} k_{3}} \tag{A11.41}
\end{equation*}
$$

Therefore if we can show that $B$ depends only on $k_{1}, k_{3}$, and $\mu$ we are done. For this, without loss of generality, we choose the $z$-direction as the direction of $\mathbf{k}_{1}$ and denote by $\mu_{i j}=\cos \theta_{i j}$, where $\theta_{i j}$ is the polar angle of $\mathbf{r}_{i j}$. We also use that

$$
\begin{equation*}
\mathbf{k}_{3}=k_{3}\left(\mu \mathbf{e}_{z}+\sqrt{1-\mu^{2}} \mathbf{e}_{\perp}\right) \tag{A11.42}
\end{equation*}
$$

and, again without loss of generality, we identify the direction $\mathbf{e}_{\perp}$ that is normal to $\mathbf{e}_{z}$ with the $x$-direction, so that $\mathbf{r}_{i j} \mathbf{e}_{\perp}=r_{i j} \cos \varphi_{i j}$. With these choices of coordinate directions we have

$$
\begin{equation*}
\mathbf{k}_{1} \mathbf{r}_{12}=\mu_{12} r_{12} k_{1} \text { and } \quad \mathbf{k}_{3} \mathbf{r}_{32}=\left(\mu \mu_{32}+\sqrt{1-\mu^{2}} \sqrt{1-\mu_{32}^{2}}\right) r_{32} k_{3} \tag{A11.43}
\end{equation*}
$$

Note also that due to statistical isotropy apart from $r_{12}$ and $r_{32}, \xi_{3}$ depends only on the cosine of the angle between $\mathbf{r}_{12}$ and $\mathbf{r}_{32}$, which is given by

$$
\begin{equation*}
v=\mu_{12} \mu_{32}+\sqrt{\left(1-\mu_{12}^{2}\right)\left(1-\mu_{32}^{2}\right)} \cos \left(\varphi_{12}-\varphi_{32}\right) \tag{A11.44}
\end{equation*}
$$

Using spherical coordinates we can transform $\varphi_{12} \rightarrow \varphi_{12}-\varphi_{32} \equiv \varphi$. With this the integral (A11.40) becomes

$$
\begin{align*}
B\left(k_{1}, k_{3}, \mu\right)= & \int r_{12}^{2} d r_{12} d \mu_{12} d \varphi r_{32}^{2} d r_{32} d \mu_{32} d \varphi_{32} \xi_{3}\left(r_{12}, r_{32}, \nu\left(\mu_{12}, \mu_{32}, \varphi\right)\right) \\
& \times \exp \left[i\left(r_{12} k_{1} \mu_{12}+r_{32} k_{3}\left(\mu \mu_{32}+\sqrt{1-\mu^{2}} \sqrt{1-\mu_{32}^{2}} \cos \varphi_{32}\right)\right)\right] \tag{A11.45}
\end{align*}
$$

Finally, one can perform the integration over $\varphi_{32}$, which yields

$$
\begin{align*}
B\left(k_{1}, k_{3}, \mu\right)= & 2 \pi \int r_{12}^{2} d r_{12} d \mu_{12} d \varphi r_{32}^{2} d r_{32} d \mu_{32} \xi_{3}\left(r_{12}, r_{32}, \nu\left(\mu_{12}, \mu_{32}, \varphi\right)\right) \\
& \times J_{0}\left(r_{32} k_{3} \sqrt{1-\mu^{2}} \sqrt{1-\mu_{32}^{2}}\right) \exp \left[i\left(r_{12} k_{1} \mu_{12}+r_{32} k_{3} \mu_{32}\right)\right] \tag{A11.46}
\end{align*}
$$

Here $J_{0}$ denotes the Bessel function of order 0 and $\mu$ can be written as a function of the $k_{i}$ via Eq. (A11.41). In principle one could also convert the integral over $\varphi$ or the one over $\mu_{32}$ into an integral over $v$ but with awkward boundary conditions and with a not very illuminating result. In Eq. (A11.46) it is no longer evident that $B$ is symmetric under the exchange of the $k_{i}$. But we know that this must be true because of the original expression given on the first line of Eq. (A11.40).

## Exercise 6.2

The coefficient $\left\langle a_{\ell_{1} m_{1}} a_{\ell_{2} m_{2}} a_{\ell_{3} m_{3}}\right\rangle$ is obtained from the 3-point function by integrating with the corresponding spherical harmonics,

$$
\begin{equation*}
\left\langle a_{\ell_{1} m_{1}} a_{\ell_{2} m_{2}} a_{\ell_{3} m_{3}}\right\rangle=\int \xi_{3}\left(\mathbf{n}_{1}, \mathbf{n}_{2}, \mathbf{n}_{2}\right) Y_{\ell_{1} m_{1}}^{*}\left(\mathbf{n}_{1}\right) Y_{\ell_{2} m_{2}}^{*}\left(\mathbf{n}_{2}\right) Y_{\ell_{3} m_{3}}^{*}\left(\mathbf{n}_{3}\right) d \Omega_{1} d \Omega_{2} d \Omega_{3} . \tag{A11.47}
\end{equation*}
$$

On the other hand, we have expression (6.38) for $\xi_{3}$. Using the addition theorem of spherical harmonics,

$$
\begin{equation*}
P_{L}\left(\mu_{i j}\right)=\frac{4 \pi}{2 L+1} \sum_{M} Y_{L M}\left(\mathbf{n}_{i}\right) Y_{L M}^{*}\left(\mathbf{n}_{j}\right) \tag{A11.48}
\end{equation*}
$$

Eq. (6.38) leads to three integrals of the following form:

$$
\begin{equation*}
\int Y_{\ell_{i} m_{i}}^{*}\left(\mathbf{n}_{i}\right) Y_{L_{i} M_{i}}\left(\mathbf{n}_{i}\right) Y_{L_{[i-1]} M_{[i-1]}}^{*}\left(\mathbf{n}_{i}\right), \tag{A11.49}
\end{equation*}
$$

where $[i-1]=i-1$ for $i=2,3$ and $[i-1]=3$ for $i=1$. Using the triple integrals of spherical harmonics given in Appendix 4, Section A4.2.3, we find

$$
\begin{align*}
\left\langle a_{\ell_{1} m_{1}} a_{\ell_{2} m_{2}} a_{\ell_{3} m_{3}}\right\rangle= & (4 \pi)^{3 / 2} \sum_{L_{i}, M_{i}} \prod_{i=1}^{3} \sqrt{2 \ell_{i}+1}\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
0 & 0 & 0
\end{array}\right) \\
& \times\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
m_{i} & M_{i} & M_{[i-1]}
\end{array}\right) b_{L_{1} L_{2} L_{3}}^{(2)} . \tag{A11.50}
\end{align*}
$$

The factors $(-1)^{M_{i}}$ multiply together to 1 since $M_{1}+M_{2}+M_{3}=0$, as is easy to check. Now together with (A4.61) Eq. (6.40) implies

$$
\begin{align*}
& \sqrt{\frac{\prod_{i=1}^{3}\left(2 \ell_{i}+1\right)}{4 \pi}}\left(\begin{array}{ccc}
\ell_{1} & \ell_{2} & \ell_{3} \\
0 & 0 & 0
\end{array}\right) b_{\ell_{1} \ell_{2} \ell_{3}} \\
&= \sum_{m_{1} m_{2} m_{3}}\left\langle a_{\ell_{1} m_{1}} a_{\ell_{2} m_{2}} a_{\ell_{3} m_{3}}\right\rangle\left(\begin{array}{ccc}
\ell_{1} & \ell_{2} & \ell_{3} \\
m_{1} & m_{2} & m_{3}
\end{array}\right)  \tag{A11.51}\\
&=(4 \pi)^{3 / 2} \sum_{m_{1} m_{2} m_{3} ; L_{i} M_{i}}\left(\begin{array}{ccc}
\ell_{1} & \ell_{2} & \ell_{3} \\
m_{1} & m_{2} & m_{3}
\end{array}\right) \prod_{i=1}^{3} \sqrt{2 \ell_{i}+1} \\
& \times\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
0 & 0 & 0
\end{array}\right)\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
m_{i} & M_{i} & M_{[i-1]}
\end{array}\right) b_{L_{1} L_{2} L_{3}}^{(2)} . \tag{A11.52}
\end{align*}
$$

Deviding by the prefactor we find

$$
\begin{align*}
b_{\ell_{1} \ell_{2} \ell_{3}}= & \sum_{L_{i}} Q_{\ell_{1} \ell_{2} \ell_{3}}^{L_{1} L_{2} L_{3}} b_{L_{1} L_{2} L_{3}}^{(2)}, \text { where }  \tag{A11.53}\\
Q_{\ell_{1} \ell_{2} \ell_{3}}^{L_{1} L_{2} L_{3}}= & (4 \pi)^{2}\left(\begin{array}{ccc}
\ell_{1} & \ell_{2} & \ell_{3} \\
0 & 0 & 0
\end{array}\right)^{-1} \sum_{m_{i} ; M_{i}}\left(\begin{array}{ccc}
\ell_{1} & \ell_{2} & \ell_{3} \\
m_{1} & m_{2} & m_{3}
\end{array}\right) \prod_{i=1}^{3}\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
0 & 0 & 0
\end{array}\right) \\
& \times\left(\begin{array}{ccc}
\ell_{i} & L_{i} & L_{[i-1]} \\
m_{i} & M_{i} & M_{[i-1]}
\end{array}\right) . \tag{A11.54}
\end{align*}
$$

Here the sums over all $m_{i}$ and $M_{i}$ are always understood as sums from $-\ell_{1}$ to $\ell_{i}$ and $-L_{i}$ to $L_{i}$ respectively.

## Exercise 6.7

Let us first show that

$$
\begin{equation*}
V_{1}(\nu)=\int_{\partial K(\nu)} d s=\frac{1}{4} \int_{\mathbb{S}^{2}} d \Omega \delta(u(\mathbf{n})-v) \sqrt{(\nabla u)^{2}} \tag{A11.55}
\end{equation*}
$$

To show this let us consider a small neighborhood of a given point $\mathbf{n}_{0}$ on the curve $u(\mathbf{n})=\nu$. In this neighborhood we may parameterize this curve by some function $\mathbf{n}(t)$. (We assume that $v$ is not a local maximum; otherwise the curve $u(\mathbf{n})=v$ shrinks to a point.) The length of a part of our curve is then given by the integral of $\sqrt{\dot{\mathbf{n}}^{2}} d t$. Choosing local coordinates on the sphere along the curve and orthogonal to it we find for the small part of the curve that we parameterize as $\mathbf{n}(t)$

$$
\begin{equation*}
L=\int \sqrt{(\dot{\mathbf{n}})^{2}} d t=\int \delta(u(\mathbf{n})-v) \sqrt{(\dot{\mathbf{n}})^{2}} d \Omega \tag{A11.56}
\end{equation*}
$$

By construction $u(\mathbf{n}(t))=v$ and therefore

$$
\begin{equation*}
\frac{d u(\mathbf{n}(t))}{d t}=\nabla u(\mathbf{n}(t)) \cdot \dot{\mathbf{n}}=0 \tag{A11.57}
\end{equation*}
$$

Since we are in two dimensions, this implies that

$$
\begin{equation*}
\dot{n}_{i}=\alpha \epsilon_{i j} \nabla_{j} u(\mathbf{n}(t)), \text { or equivalently } \epsilon_{k i} \dot{n}_{i}=-\alpha \nabla_{k} u(\mathbf{n}(t)) \tag{A11.58}
\end{equation*}
$$

The proportionality factor depends on our parameterization and we can choose it to be unity. Equation (A11.58) then implies that $(\dot{\mathbf{n}})^{2}=(\nabla u)^{2}$, which leads to (A11.55).

We now also want to show that

$$
\begin{equation*}
V_{2}(\nu)=\int_{\partial K(v)} \kappa(s) d s=\frac{1}{2 \pi} \int_{\mathbb{S}^{2}} d \Omega \delta(u(\mathbf{n})-v) \frac{\sum_{i j=1}^{2}(-1)^{j+i+1} \nabla_{i} u \nabla_{j} u \nabla_{i} \nabla_{j} u}{(\nabla u)^{2}} \tag{A11.59}
\end{equation*}
$$

Using $d s=\delta(u(\mathbf{n})-v) \sqrt{(\nabla u)^{2}} d \Omega$ we simply need to show that on the curve $u(\mathbf{n})=v$ the geodesic curvature is given by

$$
\begin{equation*}
\kappa(\mathbf{n})=\frac{\sum_{i j=1}^{2}(-1)^{j+i+1} \nabla_{i} u \nabla_{j} u \nabla_{i} \nabla_{j} u}{(\nabla u)^{3 / 2}} \tag{A11.60}
\end{equation*}
$$

To show this we now derive $u(\mathbf{n}(t))=v$ a second time, leading to

$$
\begin{array}{r}
\nabla_{i} \nabla_{j} u(\mathbf{n}(t)) \dot{n}_{i} \dot{n}_{j}+\nabla_{j} u(\mathbf{n}(t)) \ddot{n}_{j}=0, \\
\nabla_{i} \nabla_{j} u(\mathbf{n}(t)) \dot{n}_{i} \dot{n}_{j}=-\dot{n}_{i} \ddot{n}_{j} \epsilon_{i j} . \tag{A11.62}
\end{array}
$$

For the second equality we made use of Eq. (A11.58) (with $\alpha=1$ ). Now the general expression for the geodesic curvature of an arbitrary line can be found in a generic geometry book; it is

$$
\begin{equation*}
\kappa(\mathbf{n}(t))=\frac{\dot{n}_{i} \ddot{n}_{j} \epsilon_{i j}}{(\dot{\mathbf{n}})^{3 / 2}} . \tag{A11.63}
\end{equation*}
$$

Inserting $\dot{\mathbf{n}}$ and $\dot{n}_{i} \ddot{n}_{j} \epsilon_{i j}$ from Eqs. (A11.58) and (A11.62) we find Eq. (A11.60).

## A11.7 Chapter 7

## Exercise 7.1

We consider a mass $M$ positioned at $\mathbf{x}=0$ with gravitational potential $\Psi=G M / r$. To first order in $\Psi$ the corresponding metric is given by

$$
d s^{2}=-(1+2 \Psi) d t^{2}+(1-2 \Psi) d \mathbf{x}^{2}
$$

We want to determine the deflection of a photon in this metric. Angles are invariant under conformal transformations of the geometry. We may therefore calculate the deflection in the conformally related metric $d \tilde{s}^{2}=(1+2 \Psi) d s^{2}$. To first order in $\Psi$ we have

$$
d \tilde{s}^{2}=-(1+4 \Psi) d t^{2}+d \mathbf{x}^{2}
$$



Fig. A11.2 A photon passing the mass $M$ in direction $\mathbf{n}$ with impact parameter $d$.

We consider a photon along the unperturbed path $\mathbf{x}(s)=d \mathbf{e}+s \mathbf{n}$. The spatial unit vector $\mathbf{n}$ is the direction of motion of the photon and $\mathbf{e}$ is a spatial unit vector normal to $\mathbf{n}$. Hence $d$ is the impact parameter, that is, the closest distance of the photon from the mass $M$ at $\mathbf{x}=0$; see Fig. A11.2. The unperturbed photon velocity is given by $\left(n^{\mu}\right)=(1, \mathbf{n})$. Since $\Psi$ is spherically symmetric, angular momentum is conserved and also the perturbed motion will be in the plane $(\mathbf{e}, \mathbf{n})$. We define the perturbed velocity by

$$
\left(n^{\mu}+\delta n^{\mu}\right)=\left(1+\delta n^{0}, \mathbf{n}+\delta \mathbf{n}\right) .
$$

As it lies in the plane $(\mathbf{e}, \mathbf{n})$, the spatial part of $\delta n^{\mu}$ is of the form $\delta \mathbf{n}=\varphi \mathbf{e}+\alpha \mathbf{n}$, where $\varphi$ is the deflection angle and $\alpha$ is related to the gravitational redshift. The Christoffel symbols are of first order in $\Psi$, so that the first-order equation of motion for the photon trajectory gives

$$
\delta \dot{n}^{\mu}+\tilde{\Gamma}_{00}^{\mu}+2 \tilde{\Gamma}_{0 j}^{\mu} n^{j}+\tilde{\Gamma}_{i j}^{\mu} n^{i} n^{j}=0 .
$$

For the metric $d \tilde{s}^{2}$ the only nonvanishing Christoffel symbols are

$$
\tilde{\Gamma}_{0 i}^{0}=\tilde{\Gamma}_{i 0}^{0}=\tilde{\Gamma}_{00}^{i}=2 \partial_{i} \Psi .
$$

For the deflection angle we therefore obtain

$$
\dot{\varphi}=(\delta \dot{\mathbf{n}} \cdot \mathbf{e})=-2 \mathbf{e} \cdot \nabla \Psi=2 M G \frac{d}{\left(d^{2}+s^{2}\right)^{3 / 2}}
$$

Integrating this from $s=-\infty$ to $s=\infty$ yields

$$
\begin{equation*}
\varphi=\frac{4 M G}{d} \tag{A11.64}
\end{equation*}
$$

## A11.8 Chapter 8

## Exercise 8.1

We want to show the following theorem:
Theorem: Let $\xi(\mathbf{r})$ be a correlation function that depends on the orientation of $\mathbf{r}$ only via its scalar product with one fixed given direction $\mathbf{n}$ (e.g., the line of sight). Denoting the corresponding direction cosine by $\mu$ and expanding $\xi$ in Legendre polynomials, we have

$$
\begin{equation*}
\xi(\mathbf{r})=\sum_{n} \xi_{n}(r) L_{n}(\mu), \quad \mu=\hat{\mathbf{r}} \cdot \mathbf{n} . \tag{A11.65}
\end{equation*}
$$

In this situation the Fourier transform of $\xi$, the power spectrum, is of the form

$$
\begin{align*}
& P(\mathbf{k})=\sum_{n} p_{n}(k) L_{n}(v), \quad v=\hat{\mathbf{k}} \cdot \mathbf{n} \quad \text { where }  \tag{A11.66}\\
& p_{n}(k)=4 \pi i^{n} \int_{0}^{\infty} d r r^{2} j_{n}(k r) \xi_{n}(r), \quad \text { and }  \tag{A11.67}\\
& \xi_{n}(r)=\frac{(-i)^{n}}{2 \pi^{2}} \int_{0}^{\infty} d k k^{2} j_{n}(k r) p_{n}(k) . \tag{A11.68}
\end{align*}
$$

Proof The Fourier transform of $\xi$ is defined as

$$
\begin{equation*}
P(\mathbf{k})=\int d^{3} r e^{i \mathbf{r} \cdot \mathbf{k}} \xi(\mathbf{r}) \tag{A11.69}
\end{equation*}
$$

We use that

$$
e^{i \mathbf{r} \cdot \mathbf{k}}=\sum_{\ell} i^{\ell}(2 \ell+1) j_{\ell}(k r) L_{\ell}(\hat{\mathbf{k}} \cdot \hat{\mathbf{r}})
$$

where $L_{\ell}$ is the Legendre polynomial of degree $\ell$. Hence

$$
L_{\ell}(\hat{\mathbf{k}} \cdot \hat{\mathbf{r}})=\frac{4 \pi}{2 \ell+1} \sum_{m=-\ell}^{\ell} Y_{\ell m}(\hat{\mathbf{k}}) Y_{\ell m}^{*}(\hat{\mathbf{r}})=\frac{4 \pi}{2 \ell+1} \sum_{m=-\ell}^{\ell} Y_{\ell m}(\hat{\mathbf{r}}) Y_{\ell m}^{*}(\hat{\mathbf{k}}) ;
$$

$Y_{\ell m}$ are the spherical harmonics. Inserting these identities in (A11.69) using the ansatz (A11.65) for the correlation function, we obtain

$$
\begin{equation*}
P(\mathbf{k})=\sum_{\ell m} \sum_{n m^{\prime}} \frac{(4 \pi)^{2} i^{\ell}}{2 \ell+1} \int d^{3} r \xi_{n}(r) j_{\ell}(k r) Y_{\ell m}(\hat{\mathbf{k}}) Y_{\ell m}^{*}(\hat{\mathbf{r}}) Y_{n m^{\prime}}(\hat{\mathbf{r}}) Y_{n m^{\prime}}^{*}(\mathbf{n}) \tag{A11.70}
\end{equation*}
$$

Using the orthogonality relation of spherical harmonics, the integration over directions gives

$$
\begin{equation*}
P(\mathbf{k})=4 \pi \sum_{n} i^{n} \int_{0}^{\infty} d r r^{2} \xi_{n}(r) j_{n}(k r) L_{n}(\nu) \tag{A11.71}
\end{equation*}
$$

Identification of the expansion coefficients yields (A11.67). Equation (A11.68) is obtained in the same way using the inverse Fourier transform,

$$
\xi(\mathbf{r})=\frac{1}{(2 \pi)^{3}} \int d^{3} k e^{-i \mathbf{k} \cdot \mathbf{r}} P(\mathbf{k})
$$

Clearly, if $\xi(\mathbf{r})=\langle\Delta(\mathbf{x}) \Delta(\mathbf{x}+\mathbf{r})\rangle$ is independent of $\mathbf{x}$ ( $\Delta$ is statistically homogeneous), $\xi$ does not depend on the sign of $\mathbf{r}$ and in the sum above only $\xi_{n}$ with even $n$ 's can contribute so that $P(\mathbf{k})$ is real.

## A11.9 Chapter 9

## Exercise 9.9.2

We parameterize the initial conditions by

$$
C_{i j}=\left\langle X_{i}(\mathbf{k}) X_{j}^{*}\left(\mathbf{k}^{\prime}\right)\right\rangle=A_{i j}\left(k / H_{0}\right)^{n_{i j}} \delta\left(\mathbf{k}-\mathbf{k}^{\prime}\right)
$$

Clearly, for $C_{i j}$ to be positive semidefinite for all values of $k$, the matrix $A_{i j}=C_{i j}\left(k=H_{0}\right)$ has to be positive semidefinite. Let us now consider $i \neq j$ with $A_{i j} \neq 0$. If neither $n_{i i} \leq$ $n_{i j} \leq n_{j j}$ nor $n_{j j} \leq n_{i j} \leq n_{i i}$ is true, $n_{i j}$ is either the largest or the smallest of these three spectral indices. Let us first assume it to be the smallest. To show that $C_{i j}$ is not positive semidefinite, we have to find a vector $V$ so that $C_{m n} V^{m} V^{n}<0$. If $A_{i j}>0$, we choose $V^{i}=-V^{j}=1$, and if $A_{i j}<0$, we choose $V^{i}=V^{j}=1$, so that $A_{i j} V^{i} V^{j}=-\left|A_{i j}\right|$ (no sum!). Since $n_{i j}$ is smaller than $n_{i i}$ and $n_{j j}$ we can choose $k$ to be sufficiently small so that $\left|A_{i j}\right|\left(k / H_{0}\right)^{n_{i j}} \gg\left|A_{i i}\right|\left(k / H_{0}\right)^{n_{i i}}$ and $\left|A_{i j}\right|\left(k / H_{0}\right)^{n_{i j}} \gg\left|A_{j j}\right|\left(k / H_{0}\right)^{n_{j j}}$. Setting all other components of $V$ to 0 we obtain for such values of $k$

$$
\sum_{m n} V^{m} V^{n} C_{m n}(k)=-\left|A_{i j}\right|\left(k / H_{0}\right)^{n_{i j}}+A_{i i}\left(k / H_{0}\right)^{n_{i i}}+A_{j j}\left(k / H_{0}\right)^{n_{j j}}<0 .
$$

If $n_{i j}$ is larger than $n_{i i}$ and $n_{j j}$ we just have to choose $k$ sufficiently large.

## A11.10 Chapter 10

## Exercise 10.3

We want to compute the integral

$$
\begin{equation*}
J_{B E}\left(x_{c}\right)=\int_{0}^{1} \frac{d x}{x} \frac{e^{x} \exp \left[-2 x_{c} / x\right]}{e^{x}-1} \tag{A11.72}
\end{equation*}
$$

for small values of $x_{c}$; more precisely, $0<x_{c} \ll 1$. We want to show that

$$
\begin{equation*}
J_{B E}\left(x_{c}\right)=\frac{1}{2 x_{c}}-\frac{1}{2} \log \left(x_{c}\right)+\text { higher order }, \tag{A11.73}
\end{equation*}
$$

where "higher order" denotes terms of order unity and terms that vanish for $x_{c} \rightarrow 0$. To compute the integral (A11.72) we use the series expansion

$$
\begin{equation*}
\frac{t e^{t}}{e^{t}-1}=\sum_{m=0} B_{m} \frac{t^{m}}{m!} \tag{A11.74}
\end{equation*}
$$

Here $B_{m}$ are the Bernoulli numbers (see Abramowitz and Stegun, 1970), given by ${ }^{1}$

$$
\begin{align*}
B_{0} & =1, \quad B_{1}=1 / 2, \quad B_{2}=1 / 6, \quad B_{3}=0, \quad B_{4}=-1 / 30,  \tag{A11.75}\\
B_{2 n+1} & =0, \quad B_{2 n}=-(-1)^{n} \frac{2(2 n)!}{(2 \pi)^{2 n}} \zeta(2 n), \quad \text { for } n>1 . \tag{A11.76}
\end{align*}
$$

Here $\zeta$ denotes the Riemann zeta-function. With this

$$
\begin{align*}
J_{B E}\left(x_{c}\right) & =\sum_{m=0} B_{m} \frac{I_{m}\left(x_{c}\right)}{m!} \quad \text { where }  \tag{A11.77}\\
I_{m}\left(x_{c}\right) & =\int_{0}^{1} d x x^{m-2} \exp \left[-2 x_{c} / x\right] \tag{A11.78}
\end{align*}
$$

With the variable transform $y=1 / x$ we obtain

$$
\begin{equation*}
I_{m}\left(x_{c}\right)=\int_{1}^{\infty} d y y^{-m} \exp \left[-2 x_{c} y\right]=E_{m}\left(2 x_{c}\right) \tag{A11.79}
\end{equation*}
$$

where $E_{m}$ denotes the well-known exponential integral function of order $m . E_{0}$ is elementary and yields the first part of our result, $I_{0}=e^{-2 x_{c}} /\left(2 x_{c}\right) \simeq 1 /\left(2 x_{c}\right)$. The exponential integral of order 1 has the asymptotic behavior $E_{1}\left(2 x_{c}\right)=\operatorname{Ei}\left(2 x_{c}\right) \simeq-\log \left(2 x_{c}\right)-$ $\gamma+\mathcal{O}\left(x_{c}\right)$, where $\gamma \simeq 0.577$ is the Euler-Mascheroni constant. Then, as $E_{m}^{\prime}\left(2 x_{c}\right)=$ $-E_{m-1}\left(2 x_{c}\right)$ it follows that the exponential integral of order $m \geq 1$ behaves as $E_{m}\left(2 x_{c}\right) \simeq$ $\left(2 x_{c}\right)^{m-1} \log \left(2 x_{c}\right)+$ const. for small $x_{c} \ll 1$. This proves Eq. (A11.73).

As a final remark let me mention that such integrals are often estimated using a saddle point approximation. While the behavior $J_{B C} \propto x_{c}^{-1}$ is recovered by this method also here, the prefactor is wrong. One can actually show that in this case the saddle point approximation obtains corrections that scale like $x_{c}^{-1}$ at every order and is therefore useless.

[^0]
[^0]:    ${ }^{1}$ One often finds $B_{1}=-1 / 2$. This depends on the definition of $B_{n}$ as $B_{n}=b_{n}(0)$ or $B_{n}=b_{n}(1)$, where $b_{n}(x)$ are the Bernoulli polynomials; see Abramowitz and Stegun (1970). Here we use the second identification, which gives $B_{1}=1 / 2$.

