# Discrete one-dimensional quasi-periodic Schrödinger operators with pure point spectrum 

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## 1. Introduction

The discrete quasi-periodic Schrödinger operator in one dimension

$$
\mathcal{L}_{\theta}: l^{2}(\mathbf{Z}) \rightarrow l^{2}(\mathbf{Z})
$$

is defined by

$$
\left(\mathcal{L}_{\theta} u\right)_{n}=-\varepsilon\left(u_{n+1}+u_{n-1}\right)+E(\theta+n \omega) u_{n}
$$

where $\omega$ is a real number and where $E$ is a smooth function on the torus $\mathbf{T}=\mathbf{R} /(2 \pi \mathbf{Z})$, belonging to the Gevrey class

$$
\sup _{\theta \in \mathbf{T}}\left|\partial^{\nu} E(\theta)\right|<C(\nu!)^{2} K^{\nu} \quad \forall \nu \geqslant 0 .
$$

We shall assume that $\omega$ satisfies a Diophantine condition

$$
\|k \omega\|:=\inf _{n \in \mathbf{Z}}|k \omega-n 2 \pi| \geqslant \frac{\varkappa}{|k|^{\tau}} \quad \forall k \in \mathbf{Z} \backslash\{0\}
$$

for some constants $\varkappa>0, \tau>1$, and that $E$ satisfies a generic transversality condition

$$
\begin{cases}\max _{0 \leqslant \nu \leqslant s}\left|\partial_{x}^{\nu}(E(\theta+x)-E(\theta))\right| \geqslant \xi>0 & \forall \theta \forall x \\ \max _{0 \leqslant \nu \leqslant s}\left|\partial_{\theta}^{\nu}(E(\theta+x)-E(\theta))\right| \geqslant \xi\|x\| & \forall \theta \forall x\end{cases}
$$

Under these two assumptions we prove the following theorem.
Theorem. Assume that $E$ and $\omega$ are as above. Then there exists a constant $\varepsilon_{0}=$ $\varepsilon_{0}(C, K, \xi, s, \varkappa, \tau)$ such that if $|\varepsilon|<\varepsilon_{0}$, then $\mathcal{L}_{\theta}$ is pure point with a complete set of eigenvectors in $l^{2}(\mathbf{Z})$ for a.e. $\theta$. Moreover, the measure of the set $[\inf E, \sup E] \backslash \sigma\left(\mathcal{L}_{\theta}\right)$ goes to 0 as $\varepsilon \rightarrow 0$.

This result generalizes previous work by Fröhlich-Spencer-Wittver and by Sinai who considered the case when $E$ has only two critical points both of which are non-degenerate (see [7], [14], [8])-such functions clearly fulfill the generic condition.

It is easy to see that the generic condition is fulfilled for any analytic function which has no period shorter than $2 \pi$. And if it has a shorter period we can scale the period to become $2 \pi$ and then apply the theorem. In particular the theorem applies to any real-analytic function which is not constant.

It will also be clear (from Proposition 8) that the same result remains true if we replace in $\mathcal{L}_{\theta}$ the constant difference operator $u \rightarrow \varepsilon\left(u_{n+1}+u_{n-1}\right)$ by any other constant finite symmetric difference operator $u \rightarrow \varepsilon\left(\sum_{i=-N}^{N} a_{i} u_{n+i}\right)$. Therefore this result also generalizes the work [2].

Idea of proof. The formulation we choose for the problem is to consider the operator as a symmetric infinite-dimensional matrix that depends on the parameter $\theta$,

$$
D(\theta)+\varepsilon F(\theta)=\left(\begin{array}{ccccc}
\ddots & & & & 0 \\
& E(\theta-\omega) & -\varepsilon & & \\
& -\varepsilon & E(\theta) & -\varepsilon & \\
& & -\varepsilon & E(\theta+\omega) & \\
0 & & & & \ddots
\end{array}\right),
$$

and that is close to a diagonal matrix. This matrix satisfies the shift condition

$$
(D+\varepsilon F)_{m+k, n+k}(\theta)=(D+\varepsilon F)_{m, n}(\theta+k \omega)
$$

a property that permits us to control for example the whole spectrum of the matrix for a fixed but arbitrary parameter value $\theta_{0}$ by controlling the dependence of a single eigenvalue as a function of $\theta$.

Our goal is to conjugate the matrix $(D+\varepsilon F)(\theta)$ to a diagonal matrix $D_{\infty}(\theta, \varepsilon)$ by an orthogonal matrix made up of a complete set of orthonormal eigenvectors. The way to construct such a matrix is by an iterative procedure

$$
U_{j}^{*} \cdot \ldots \cdot U_{1}^{*} \cdot(D+\varepsilon F) \cdot U_{1} \cdot \ldots \cdot U_{j}=D_{j+1}+F_{j+1}
$$

that conjugates $D+\varepsilon F$ closer and closer to a diagonal matrix $D_{j}=\operatorname{diag}\left(E^{j}(\theta+k \omega)\right)$. During this iteration we will have to deal with almost multiple eigenvalues of the main part $D_{j}$-i.e. eigenvalues that are so close so that they prevent a purely perturbative construction of the transformation $U_{j}$ close to the identity. These "almost multiplicities" are of finite order and give rise to certain finite-dimensional symmetric submatrices of
$D_{j}+F_{j}$ that will be diagonalized in a non-perturbative manner. The "almost multiplicities" of $D_{1}=D$ are of order

$$
\leqslant 2^{s+4} \frac{C((s+1)!)^{2} K^{s+1}}{\xi}
$$

but in the following steps this order will grow in a controlled way and we will have to handle "almost multiplicities" of higher and higher order. This is a distinct feature from the approach of [7], [14], where the "almost multiplicities" remain bounded (by two) throughout the iteration. The "almost multiplicities" will be described at the $j$ th step by a decomposition $\bigcup_{i} \Lambda_{i}^{j}=\mathbf{Z}$.

In carrying out this procedure it will be essential that the order of the "almost multiplicities" do not increase too fast. Our means to control this will be through a non-degeneracy condition on the spectrum of $D_{j}$, which at the first step is simply the condition on $E(\theta+k \omega)$ described above-smoothness and transversality. Such a simple condition on the eigenvalues will not persist unchanged under perturbations, and we will have to consider in what way the non-degeneracy persists during the iteration. At the $j$ th step this will be through a condition on $E^{j}(\theta+k \omega)$ that involves all the "almost multiplicities" of the $k$ th eigenvalues that have been considered at the previous steps. These will be described by subsets $\Omega_{i}^{j} \supset \Lambda_{i}^{j}$. Another factor that will help us to control the "almost multiplicities" is the exponential decay of the perturbation off the diagonal.

The iterative construction will be carried out completely for all $\theta$. In order to establish pure point spectrum one is left with the problem of proving that the infinite sequence $U_{1} \cdot \ldots \cdot U_{j}$ of orthogonal matrices constructed will indeed converge to an orthogonal transformation. This turns out to be the case only for a.e. $\theta$. Our approach provides a way to analyze also the remaining $\theta$ 's. (Confer in this respect [4] where a similar approach is made for another perturbation problem.)

A technical problem concerns the smoothness. If we insist on always having the main part $D_{j}$ in diagonal form it will not be possible to control the smoothness of (the non-perturbative part of) the transformation $U_{j}$, hence the smoothness of $F_{j+1}$. The smoothness of the perturbation can be controlled if we give up the diagonal form of $D_{j}$ and only require it to be in block-diagonal form with finite-dimensional symmetric blocks-such a matrix can of course be diagonalized. However, such a decomposition into blocks would be somewhat arbitrary and would therefore destroy the shift condition on $D_{j}$ and $F_{j}$. We prefer to preserve both smoothness and shift and we will therefore have to work with a more general normal form which is not block diagonal but which can be block diagonalized-this requires a more complex formulation which will have to include the dimensions of the blocks, their location in $l^{2}(\mathbf{Z})$ and their smoothness. We describe the normal form in $\S 4$ and we give some more explanations of the idea there.

The use of Gevrey functions is also a technical point. We could work with piecewise analytic functions but would then have to count the number of pieces. This is avoided by Gevrey classes since they contain functions with compact support. There is of course nothing particular with the Gevrey class we have chosen above-we could just as well have chosen another.

Content of the paper. The paper is organized in the following way. In $\S 2$ we present some results on finite-dimensional symmetric matrices and in $\S 3$ we study the transversality condition-the proof of the lemmas in these two sections will be of no relevance for the proof of the theorem, only the statements will be of importance. In §4 we describe the normal form and a small divisor result related to it-Lemma 5 and Corollary 6. In $\S 5$ we prove the inductive lemma-Lemma 7 -common for all KAM techniques and in $\S 6$ we finally prove Proposition 8 and, as a simple consequence, the Theorem. In $\S 7$ we discuss some other spectral properties of the operator. Some basic results on functions in the Gevrey class, on parameter dependence of roots of polynomials, and on parameter-dependent Gram-Schmidt orthogonalization are given in an appendix.

Notations. For any smooth function $f$ defined on $I$ we use the norm

$$
|f|_{\mathcal{C}^{\nu}}:=\frac{1}{(\nu!)^{2}} \max _{\theta \leqslant k \leqslant \nu} \sup _{\theta \in I}\left|\partial^{\nu} f(\theta)\right| .
$$

If $f=\left(f_{1}, \ldots, f_{d}\right)$ then we let $\left|\partial^{\nu} f(\theta)\right|$ be the Euclidean norm

$$
\sqrt{\left(\partial^{\nu} f_{1}(\theta)\right)^{2}+\ldots+\left(\partial^{\nu} f_{d}(\theta)\right)^{2}}
$$

and if $f$ is a matrix then we let $\left|\partial^{\nu} f(\theta)\right|$ denote the operator norm.
The distance $\|\cdot\|$ on the circle $\mathbf{T}=\mathbf{R} / 2 \pi \mathbf{Z}$ is defined above.
If $A$ is a subset of $\mathbf{Z}$ and $n \in \mathbf{Z}$ then

$$
\operatorname{diam}(A):=\sup _{k, l \in A}|k-l| \quad \text { and } \quad \operatorname{dist}(n, A):=\inf _{k \in A}|n-k| .
$$

We shall consider positive functions $\gamma(z, x)$ defined for $z$ in some open set $U \subset \mathbf{R}^{n}$ and for $x \geqslant 1$ such that for some $j$

$$
\gamma(z, x) \leqslant \underbrace{\exp \circ \ldots \circ \exp }_{j}(A(z) x), \quad z \in U, x \geqslant 1 .
$$

We say that $\gamma$ increases (at most) superexponentially in $x$ and that $1 / \gamma$ decays (at most) superexponentially in $x$.

If $X$ is a subset of $\mathbf{R}$ we denote by $|X|$ the Lebesgue measure of $X$.
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## 2. Block diagonalization

Let $D$ be a $(d \times d)$-matrix, $2 \leqslant d$. We let $D_{m, n}$ denote the component in the $m$ th row and the $n$th column. For a subset $\Lambda \subset\{1, \ldots, d\}$ we define

$$
\begin{aligned}
D_{\Lambda} & = \begin{cases}D_{m, n} & \text { if } m, n \in \Lambda \\
\delta_{m}^{n} & \text { otherwise }\end{cases} \\
\mathbf{R}^{\Lambda} & =\left\{x \in \mathbf{R}^{d}: x_{i}=0 \text { if } i \notin \Lambda\right\}
\end{aligned}
$$

Then

$$
D_{\Lambda}: \mathbf{R}^{\Lambda}+\mathbf{R}^{\Lambda^{\perp}} \rightarrow \mathbf{R}^{\Lambda}+\mathbf{R}^{\Lambda^{\perp}}, \quad \Lambda^{\perp}=\{1, \ldots, d\} \backslash \Lambda
$$

acting as $\mathbf{R}^{\Lambda} \hookrightarrow \mathbf{R}^{d} \xrightarrow{D} \mathbf{R}^{d} \xrightarrow{\perp \text {-proj }} \mathbf{R}^{\Lambda}$ on the first component and as the identity on the second component. When there is no risk for confusion we will use $D_{\Lambda}$ also to denote its first component.

Let now $D(\theta)$ be symmetric and smoothly parametrized by $\theta$ on an interval $I$ and let

$$
\begin{equation*}
|D|_{\mathcal{C}^{\nu}}<C_{1} K_{1}^{\nu} \quad \forall \nu \geqslant 0 \tag{1}
\end{equation*}
$$

for some $C_{1}$ and $K_{1} \geqslant 1$. The dependence in the parameter $\theta$ of the eigenvectors, and more generally of the invariant subspaces, is related to the spacing between the eigenvalues. This is described in the following lemma which gives estimates on "block diagonalization" of $D$.

Lemma 1. For all $0<\delta<1$, there is a smooth orthogonal matrix $Q(\theta)$ such that for all $\theta$ in $I, \widetilde{D}(\theta)=Q^{*}(\theta) D(\theta) Q(\theta)$ is a product of commuting blocks $\prod_{i=1}^{p} \widetilde{D}(\theta)_{\Lambda_{i}}$ which are such that

$$
\left\{\begin{array}{l}
E_{m}(\theta) \text { and } E_{n}(\theta) \text { are eigenvalues of the same block } D(\theta)_{\Lambda_{i}} \\
\quad \Rightarrow\left|E_{m}(\theta)-E_{n}(\theta)\right|<4 d \delta, \\
E_{m}(\theta) \text { and } E_{n}(\theta) \text { are eigenvalues of different blocks } D(\theta)_{\Lambda_{i}} \text { and } D(\theta)_{\Lambda_{j}} \\
\quad \Rightarrow\left|E_{m}(\theta)-E_{n}(\theta)\right| \geqslant \delta,
\end{array}\right.
$$

and such that

$$
\begin{align*}
|Q|_{\mathcal{C}^{\nu}} & \leqslant K_{2}^{\nu} \quad \forall \nu \geqslant 0 \\
K_{2} & =\gamma_{1}\left(C_{1} / \delta\right)^{d(d+1)} K_{1} \tag{2}
\end{align*}
$$

where $\gamma_{1}=\gamma_{1}\left(\gamma_{0}, d\right)$ increases superexponentially in $d$, and $\gamma_{0}$ is defined in Lemma A1. The block decomposition $\bigcup \Lambda_{i}=\{1, \ldots, d\}$ depends on $\theta$ in a piecewise constant way. Moreover, the block decomposition is constant on intervals of length $\delta / C_{1} K_{1}$.

Proof. Choose a family of continuous eigenvalues $E_{1}(\theta), \ldots, E_{d}(\theta)$. Fix a $\theta$, say $\theta=0$, and decompose the eigenvalues into groups

$$
\left\{\begin{array}{l}
E_{1}(\theta), \ldots, E_{k_{1}}(\theta) \\
E_{k_{1}+1}(\theta), \ldots, E_{k_{2}}(\theta) \\
\vdots \\
E_{k_{p-1}+1}(\theta), \ldots, E_{d}(\theta)
\end{array}\right.
$$

such that any eigenvalue of one group is separated from any eigenvalue of any other group by at least $2 \delta$ and any two eigenvalues of one group are separated by not more than $2 d \delta$. It follows that on a whole symmetric neighborhood $I^{\prime} \ni 0$ of length $2 \delta^{\prime}<\delta / C_{1} K_{1}$, any eigenvalue of one group is separated from any eigenvalue of any other group by at least $\delta$, and any two eigenvalues of one group are separated by not more than $(2 d+1) \delta$. We now restrict the discussion to the smaller interval $I^{\prime}$.

By Lemma A2, the characteristic polynomial $P(\lambda, \theta)$ of $D(\theta)$ satisfies

$$
|P(\lambda, \cdot)|_{\mathcal{C}^{\nu}}<\gamma_{2} C_{1}^{d} K_{1}^{\nu} \quad \forall \nu \geqslant 0
$$

in $|\lambda|<R=C_{1}+1$; all $\gamma$-constants will depend on $\gamma_{0}$ and increase superexponentially in $d$. We can write $P(\lambda, \theta)=P_{1}(\lambda, \theta) \cdot \breve{P}_{1}(\lambda, \theta)$, where

$$
\breve{P}_{1}(\lambda, \theta)=\prod_{i>k_{1}}\left(\lambda-E_{i}(\theta)\right)=\sum_{j=0}^{d-k_{1}} e_{d-k_{1}-j} \lambda^{j}
$$

Since all eigenvalues lie in $\left|E_{i}\right|<R-1$ we get by Lemma A4

$$
\left|e_{j}\right|_{\mathcal{C}^{\nu}}<\left(\gamma_{3} R C_{1}^{d} \frac{1}{\delta^{d-1}}\right)^{j}\left(\gamma_{3} C_{1}^{d} \frac{1}{\delta^{d}} K_{1}\right)^{\nu}
$$

and the corresponding result holds of course for the other groups of eigenvalues.
Let $q^{1}, \ldots, q^{d}$ be a positively oriented orthonormal basis of eigenvectors for $D(0)$ and define

$$
v^{m}(\theta)= \begin{cases}\widetilde{P}_{1}(D(\theta), \theta) q^{m} / \breve{P}_{1}\left(E_{m}(0), 0\right), & m \leqslant k_{1} \\ \vdots & \\ \breve{P}_{p}(D(\theta), \theta) q^{m} / \breve{P}_{p}\left(E_{m}(0), 0\right), & k_{p-1}<m \leqslant d\end{cases}
$$

Then $v^{1}, \ldots, v^{d}$ will be an orthonormal basis for $\theta=0$ and will span $p$ orthogonal subspaces for all $\theta \in I^{\prime}$. Moreover,

$$
\left|v^{m}\right|_{\mathcal{C}^{\nu}}<\gamma_{4}\left(\frac{C_{1}}{\delta}\right)^{d^{2}}\left(\gamma_{4}\left(\frac{C_{1}}{\delta}\right)^{d} K_{1}\right)^{\nu}
$$

It follows that

$$
\left|v^{m}(\theta)-v^{m}(0)\right| \leqslant \gamma_{4}\left(\frac{C_{1}}{\delta}\right)^{d(d+1)} K_{1}|\theta|<\frac{1}{4 d^{2} 3^{d+5}}
$$

for all $\theta \in I^{\prime}$ if only

$$
\delta^{\prime}<\frac{\delta^{d(d+1)}}{\gamma_{5} C_{1}^{d(d+1)} K_{1}}
$$

Moreover, $v^{1}, \ldots, v^{d}$ will be a basis also on $I^{\prime}$-indeed on this interval we get

$$
\begin{gathered}
\left|\left\langle v^{m}, v^{n}\right\rangle\right|_{\mathcal{C}^{0}}<\frac{1}{4 d^{2} 3^{d+4}}, \quad m \neq n \\
1-\frac{1}{4 d^{2} 3^{d+4}}<\left|v^{m}\right|_{\mathcal{C}^{0}}<1+\frac{1}{4 d^{2} 3^{d+4}}
\end{gathered}
$$

Now we do Gram-Schmidt orthogonalization on this basis, obtaining an orthonormal basis $\hat{v}^{1}, \ldots, \hat{v}^{d}$ on $I^{\prime}$. By Lemma A3,

$$
Q=\left(\hat{v}^{1}, \ldots, \hat{v}^{d}\right)
$$

will verify the lemma with the estimate

$$
\left(\gamma_{6}\left(\frac{C_{1}}{\delta}\right)^{d(d+1)} K_{1}\right)^{\nu}=\left(K_{1}^{\prime}\right)^{\nu}
$$

Suppose now that we have made two such constructions of orthogonal matrices $Q_{1}, Q_{2}$ on two intervals $I_{1}^{\prime}$ and $I_{2}^{\prime}$ that intersect each other on an interval $\hat{I}$ of length $\delta^{\prime}$, let us say $\left.I_{1}^{\prime}=\right]-\delta^{\prime}, \delta^{\prime}\left[\right.$ and $\left.I_{2}^{\prime}=\right] 0,2 \delta^{\prime}\left[\right.$. To $Q_{1}$ and $Q_{2}$ are associated two decompositions $\bigcup \Lambda_{i}^{1}=\bigcup \Lambda_{i}^{2}=\{1, \ldots, d\}$. Let us first assume that the decomposition of $Q_{1}$ is finer than that of $Q_{2}$, so that $Q_{1}$ also has the block decomposition $\bigcup \Lambda_{i}^{2}$. We want to interpolate $Q_{1}$ on $\theta<\frac{1}{3} \delta^{\prime}$ with $Q_{2}$ on $\theta>\frac{2}{3} \delta^{\prime}$ by a matrix $Q$ such that $Q^{*} D Q$ is block diagonal with block decomposition $\bigcup \Lambda_{i}^{2}$. For this we consider $P=Q_{1}^{*} Q_{2}$. It is a block-diagonal matrix with block decomposition $\bigcup \Lambda_{i}^{2}$ and we need to interpolate it with the identity matrix by a block matrix $B$ with the same blocks as $P$. Then we can take $Q=Q_{1} B$.

Write $P(\theta)=P\left(\delta^{\prime}\right)(I+\widehat{P}(\theta))$. Then

$$
|\widehat{P}|_{\mathcal{C}^{\nu}}<4 \delta^{\prime} K_{1}^{\prime}\left(\gamma_{7} K_{1}^{\prime}\right)^{\nu}
$$

Let now $e^{S_{o}}=P\left(\delta^{\prime}\right)$ and $e^{S}=I+\widehat{P}$, and let $\widetilde{S}_{o}=\phi S_{o}, \widetilde{S}=\phi S$, where $\phi$ is a $\mathcal{C}^{\infty}$-cut-off function which is 0 for $\theta<\frac{1}{3} \delta^{\prime}$ and 1 for $\theta>\frac{2}{3} \delta^{\prime}$. Now $S$ can be expressed as a power series in $\widehat{P}$ each term of which can be estimated by Lemma A2, $\phi$ is estimated by Lemma A1 and the estimate of $\widetilde{S}$ is again obtained by Lemma A2. This gives
if $\delta^{\prime} \gamma_{8} K_{1}^{\prime}<1$ —the same estimate holds for $e^{\tilde{S}_{o}}$ with the only difference that the factor 2 is replaced by another constant. $e^{\widetilde{S}_{o}} e^{\widetilde{S}}$ is the matrix $B$.

If $Q_{1}^{\prime}$ and $Q_{2}^{\prime}$ are defined by decompositions of which neither one are finer than the other, then we can construct on $\left[0, \delta^{\prime}\right]$ a finer decomposition and an orthogonal matrix $Q_{3}$ associated to this decomposition. Using it we can now interpolate $Q_{1}$ and $Q_{2}$ via $Q_{3}$. We leave the details.

Let $F$ be a symmetric $(d \times d)$-matrix, smoothly parametrized by $\theta$, and assume that

$$
\begin{equation*}
|F|_{\mathcal{C}^{\nu}}<C_{3} K_{3}^{\nu} \quad \forall \nu \geqslant 0 \tag{3}
\end{equation*}
$$

Lemma 2. For all $0<\delta<1$ there are a smooth symmetric matrix $G$ and a smooth anti-symmetric matrix $X$ on $I$ such that

$$
\left\langle q(\theta), G(\theta) q^{\prime}(\theta)\right\rangle=0 \quad \text { if }\left|E(\theta)-E^{\prime}(\theta)\right| \geqslant 4 d \delta
$$

for any two eigenvalues $E(\theta), E^{\prime}(\theta)$ and corresponding eigenvectors $q(\theta), q^{\prime}(\theta)$, and

$$
[X, D]=F-G
$$

and such that

$$
\begin{array}{r}
\delta|X|_{\mathcal{C}^{\nu}}+|G|_{\mathcal{C}^{\nu}}<\gamma_{2} C_{3} K_{4}^{\nu} \quad \forall \nu \geqslant 0 \\
K_{4}=\max \left(\gamma_{2}\left(C_{1} / \delta\right)^{d(d+1)+1} K_{1}, K_{3}\right) \tag{4}
\end{array}
$$

where $\gamma_{2}\left(\gamma_{0}, d\right)$ increases superexponentially in d. Moreover, if $F$ has compact support in $I$, then $G$ and $X$ also.

Proof. Fix an interval $I^{\prime}$ of length $\delta^{\prime}=K_{2}^{-1}$, defined in Lemma 1, and let $Q$ be an orthogonal matrix on $I^{\prime}$ with constant decomposition $\bigcup \Lambda_{i}$ as in Lemma 1. Let $\widetilde{F}=Q^{*} F Q, \widetilde{D}=Q^{*} D Q$ and define

$$
\widetilde{G}_{m, n}= \begin{cases}\widetilde{F}_{m, n} & \text { if } m, n \text { belongs to the same } \Lambda_{i} \\ 0 & \text { otherwise }\end{cases}
$$

$\widetilde{F}$ and $\widetilde{D}$ are estimated by Lemma A2, and it follows that

$$
|\widetilde{G}|_{\mathcal{C}^{\nu}}<\gamma_{3} C_{3} \max \left(K_{2}, K_{3}\right)^{\nu}
$$

Let

$$
\begin{gathered}
\widetilde{X}_{\Lambda_{i}, \Lambda_{j}} \widetilde{D}_{\Lambda_{j}}-\widetilde{D}_{\Lambda_{i}} \widetilde{X}_{\Lambda_{i}, \Lambda_{j}}=(\widetilde{F}-\widetilde{G})_{\Lambda_{i}, \Lambda_{j}} \\
\tilde{X}_{m, n}=0 \quad \text { if } m, n \text { belongs to the same } \Lambda_{i}
\end{gathered}
$$

Since the eigenvalues of $\widetilde{D}_{\Lambda_{i}}$ and $\widetilde{D}_{\Lambda_{j}}$ are more than $\delta$ apart we get

$$
|\widetilde{X}|_{\mathcal{C}^{0}} \leqslant \frac{\gamma_{4}}{\delta}|\widetilde{F}-\widetilde{G}|_{\mathcal{C}^{0}}
$$

Differentiating the equality gives

$$
\begin{aligned}
& \partial^{\nu} \widetilde{X}_{\Lambda_{i}, \Lambda_{j}} \widetilde{D}_{\Lambda_{j}}-\widetilde{D}_{\Lambda_{i}} \partial^{\nu} \widetilde{X}_{\Lambda_{i}, \Lambda_{j}} \\
&=\partial^{\nu}(\widetilde{F}-\widetilde{G})_{\Lambda_{i}, \Lambda_{j}}-\sum_{i=1}^{\nu}\binom{\nu}{i}\left(\partial^{\nu-i} \widetilde{X}_{\Lambda_{i}, \Lambda_{j}} \partial^{i} \widetilde{D}_{\Lambda_{j}}-\partial^{i} \widetilde{D}_{\Lambda_{i}} \partial^{\nu-i} \widetilde{X}_{\Lambda_{i}, \Lambda_{j}}\right)
\end{aligned}
$$

from which the estimate

$$
|\widetilde{X}|_{\mathcal{C}^{\nu}}<\gamma_{5} \frac{1}{\delta} C_{3} \max \left(\gamma_{5} \frac{C_{1}}{\delta} K_{2}, K_{3}\right)^{\nu}
$$

follows by induction.
Choose now a partition of unity $\left\{\phi_{j}\right\}$ with $\operatorname{supp}\left(\phi_{j}\right) \subset I_{j}^{\prime},\left|I_{j}^{\prime}\right|<\delta^{\prime}$. On each $I_{j}^{\prime}$ we have $X_{j}$ and $G_{j}$ such that

$$
\begin{aligned}
{\left[X_{j}(\theta), D(\theta)\right] } & =F(\theta)-G_{j}(\theta) \\
\left\langle q(\theta), G_{j}(\theta) q^{\prime}(\theta)\right\rangle & =0 \quad \text { if }\left|E(\theta)-E^{\prime}(\theta)\right| \geqslant 4 d \delta
\end{aligned}
$$

with the above estimates. Now we only need to set $X=\sum \phi_{j} X_{j}$ and $G=\sum \phi_{j} G_{j}$, and we obtain the estimates from Lemmas A1 and A2.

## 3. Transversality

In this section we shall analyze a kind of transversality condition which should be fulfilled by the spectrum of our matrices. Such a condition was considered by Pyartli in relation to Diophantine approximation [12]. ( ${ }^{1}$ )

Let $I$ be an interval and let $u$ be a smooth function on $I$ such that

$$
\begin{gather*}
|u|_{\mathcal{C}^{\nu}<C_{5} K_{5}^{\nu} \quad \forall \nu \leqslant m+1,}^{\max _{0 \leqslant \nu \leqslant m}\left|\frac{1}{(\nu!)^{2} K_{5}^{\nu}} \partial^{\nu} u(\theta)\right| \geqslant \beta \quad \forall \theta \in I .} .
\end{gather*}
$$

${ }^{(1)}$ A similar condition has been used by N. Nekhoroshev under the name of steepness and by R. Krikorian under the name of Pyartli transversality.

Lemma 3. For all $\varepsilon>0$, there is a disjoint union of intervals $\bigcup_{i \in J} I_{i}$ such that

$$
\begin{aligned}
\# J & \leqslant 2^{m}\left[\frac{2 C_{5} K_{5}(m+1)^{2}}{\beta}|I|+1\right], \\
\max _{i \in J}\left|I_{i}\right| & \leqslant \frac{2}{K_{5}}\left(\frac{2 \varepsilon}{\beta}\right)^{1 / m}, \\
|u(\theta)| & \geqslant \varepsilon \quad \forall \theta \in I \backslash \bigcup I_{i} .
\end{aligned}
$$

Notice that the estimate of $\# J$ is independent of $\varepsilon$.
Proof. Assume for simplicity that $0 \in I$ and $\left|\partial^{m} u(0)\right| \geqslant \beta(m!)^{2} K_{5}^{m}$. Then there is an interval $\tilde{I}$ of length

$$
l=\frac{\beta}{2 C_{5} K_{5}(m+1)^{2}}
$$

such that $\left|\partial^{m} u(\theta)\right| \geqslant \frac{1}{2} \beta(m!)^{2} K_{5}^{m}$ on $\tilde{I}$.
Consider now $\partial^{m-1} u$ on $\tilde{I}$. There exists an interval $I_{1}$ such that

$$
\begin{aligned}
\left|\partial^{m-1} u(\theta)\right| & \geqslant \frac{1}{2} \beta\left(\frac{2 \varepsilon}{\beta}\right)^{1 / m}((m-1)!)^{2} K_{5}^{m-1} \quad \forall \theta \in \tilde{I} \backslash I_{1} \\
\left|I_{1}\right| & <\frac{1}{K_{5}}\left(\frac{2 \varepsilon}{\beta}\right)^{1 / m} .
\end{aligned}
$$

Consider now $\partial^{m-2} u$ on $\tilde{I} \backslash I_{1}$. There exist two intervals $I_{2}, I_{3}$ such that

$$
\begin{aligned}
\left|\partial^{m-2} u(\theta)\right| & \geqslant \frac{1}{2} \beta\left(\frac{2 \varepsilon}{\beta}\right)^{2 / m}((m-2)!)^{2} K_{5}^{m-2} \quad \forall \theta \in \tilde{I} \backslash I_{1} \cup I_{2} \cup I_{3} \\
\left|I_{i}\right| & <\frac{1}{K_{5}}\left(\frac{2 \varepsilon}{\beta}\right)^{1 / m},
\end{aligned}
$$

etc.
Hence we obtain in $\tilde{I}, 2^{m}-1$ (possibly void) intervals $I_{i}$ such that

$$
\begin{gathered}
|u(\theta)| \geqslant \varepsilon \quad \forall \theta \in \tilde{I} \backslash \bigcup I_{i} \\
\left|I_{i}\right|<\frac{2}{K_{5}}\left(\frac{2 \varepsilon}{\beta}\right)^{1 / m}
\end{gathered}
$$

On the whole interval $I$ we get at most

$$
2^{m} \times\left[\frac{2 C_{5} K_{5}(m+1)^{2}}{\beta}|I|+1\right]
$$

many such intervals.

Let now $u_{j}$ be a sequence of smooth functions defined on an open interval $I$ and satisfying

$$
\begin{gather*}
\left|u_{j}\right| \mathcal{C}^{\nu}<C_{6} K_{6}^{\nu} \quad \forall \nu \leqslant N=m_{1}+\ldots+m_{J} \\
\max _{0 \leqslant \nu \leqslant m_{j}}\left|\frac{1}{(\nu!)^{2} K_{6}^{\nu}} \partial^{\nu} u_{j}(\theta)\right| \geqslant \beta \quad \forall \theta \in I \tag{6}
\end{gather*}
$$

with $C_{6} \geqslant 1$.
LEMMA 4. If $v=u_{1} \cdot \ldots \cdot u_{J}$ then

$$
\begin{gathered}
|v|_{\mathcal{C}^{\nu}<4^{J-1} C_{6}^{J} K_{6}^{\nu} \quad \forall \nu \leqslant N=m_{1}+\ldots+m_{J},}^{\max _{0 \leqslant \nu \leqslant N}\left|\frac{1}{(\nu!)^{2} K_{6}^{\nu}} \partial^{\nu} v(\theta)\right| \geqslant\left(\frac{1}{e}\right)^{J^{2 N}}\left(\frac{\beta}{C_{6}^{2}}\right)^{J^{N+1}} \quad \forall \theta \in I .} .
\end{gathered}
$$

Proof. The first part follows from Lemma A2 in the appendix so we concentrate on the second estimate. Fix a $\theta$, say $\theta=0$. We can assume without restriction that

$$
\frac{1}{K_{6}^{m_{i}}\left(m_{i}!\right)^{2}}\left|\partial^{m_{i}} u_{i}(0)\right| \geqslant \beta \quad \forall i
$$

Then

$$
\begin{aligned}
& \frac{1}{K_{6}^{N}(N!)^{2}} \partial^{N}\left(u_{1} \cdot \ldots \cdot u_{J}\right)(0)= \\
& \quad \sum_{\iota_{1}+\ldots+\iota_{J}=N} \frac{\iota_{1}!\cdot \ldots \cdot \iota_{J}!}{N!}\left[\frac{1}{K_{6}^{\iota_{1}\left(\iota_{1}!\right)^{2}}} \partial^{\iota_{1}} u_{1}(0) \cdot \ldots \cdot \frac{1}{K_{6}^{\iota_{J}\left(\iota_{J}!\right)^{2}}} \partial^{\iota_{J}} u_{J}(0)\right]
\end{aligned}
$$

Consider now a vector $\left(\iota_{1}, \ldots, \iota_{J}\right)$ in $\mathbf{Z}_{+}^{J}$ such that

$$
\iota_{1}+\ldots+\iota_{J}=N \quad \text { and } \quad\left|m_{1}-\iota_{1}\right|+\ldots+\left|m_{J}-\iota_{J}\right|=2 l .
$$

Then

$$
\min \left(m_{1}, \iota_{1}\right)+\ldots+\min \left(m_{J}, \iota_{J}\right)=N-l
$$

and

$$
\frac{1}{N^{l}} \leqslant \frac{m_{1}!\cdot \ldots \cdot m_{J}!}{\iota_{1}!\cdot \ldots \cdot \iota_{J}!} \leqslant N^{l}
$$

and there are $c_{l} \leqslant J^{2 l}$ many such vectors.
Since at least one of the brackets in the sum is larger than $\beta^{J}$, the sum will be larger than $\frac{1}{2} \beta^{J} \cdot m_{1}!\cdot \ldots \cdot m_{J}!/ N!$ unless for some $\left|m_{1}-\iota_{1}\right|+\ldots+\left|m_{J}-\iota_{J}\right|=2 l \neq 0$,

$$
\left|\left[\frac{1}{K_{6}^{\iota_{1}\left(\iota_{1}!\right)^{2}}} \partial^{\iota_{1}} u_{1}(0) \cdot \ldots \cdot \frac{1}{K_{6}^{\iota_{J}\left(\iota_{J}!\right)^{2}}} \partial^{\iota_{J}} u_{J}(0)\right]\right| \geqslant \frac{1}{2\left(2 J^{2}\right)^{l}} \cdot \frac{m_{1}!\cdot \ldots \cdot m_{J}!}{\iota_{1}!\cdot \ldots \cdot \iota_{J}!} \beta^{J}
$$

But this implies that

$$
\left|\frac{1}{K_{6}^{\iota_{i}}\left(\iota_{i}!\right)^{2}} \partial^{\iota_{i}} u_{i}(0)\right| \geqslant\left(\frac{1}{4 C_{6}^{J} J^{2} N}\right)^{l} \beta^{J} \quad \forall i
$$

We now have an estimate for derivatives of smaller order, $\min \left(m_{i}, \iota_{i}\right)$, and the result follows by induction. (This induction is immediate for $N \geqslant 3$ and requires a little more details when $N=1$ and 2.)

## 4. Normal forms

Consider a symmetric matrix

$$
D: \mathbf{Z} \times \mathbf{Z} \times \mathbf{T} \rightarrow \mathbf{R}
$$

smoothly parametrized by $\theta \in \mathbf{T}$ and satisfying the shift condition

$$
D_{m+k, n+k}(\theta)=D_{m, n}(\theta+k \omega)
$$

where $\omega$ is a Diophantine number,

$$
\|k \omega\| \geqslant \frac{\varkappa}{|k|^{\tau}} \quad \forall k \in \mathbf{Z} \backslash\{0\}
$$

with $\varkappa>0, \tau>1$.
(A) Estimates.
( $\mathrm{A}_{1}$ )

$$
\left|D_{m, n}\right|_{\mathcal{C}^{\nu}} \leqslant \begin{cases}C e^{-|m-n| r} K^{\nu} & \forall \nu \geqslant 0  \tag{7}\\ 0, & |m-n| \geqslant N .\end{cases}
$$

(B) Block diagonalization. There exist a symmetric interval $I \ni 0$ and a disjoint decomposition $\bigcup \Lambda_{i}=\mathbf{Z}$ such that
$\left(\mathrm{B}_{1}\right) \# \Lambda_{i} \leqslant M \forall i ;$
and there exists a smooth orthogonal matrix $Q$ on I such that the conjugated matrix $\widetilde{D}=$ $Q^{*} D Q$ is a product of commuting blocks

$$
\prod \widetilde{D}_{\Lambda_{i}}(\theta) \quad \forall \theta \in I
$$

and
$\left(\mathrm{B}_{2}\right) Q_{m, n} \not \equiv 0$ only if $|m-n| \leqslant N$, and $\operatorname{diam}\left(\Lambda_{i}\right) \leqslant M N \forall i ;$
$\left(\mathrm{B}_{3}\right)$ for all $m, Q_{m, n} \neq 0$ for at most $M$ different $n$;
( $\mathrm{B}_{4}$ ) $|Q|_{\mathcal{C}^{\nu}} \leqslant K^{\nu} \quad \forall \nu \geqslant 0$.
The decomposition, which is independent of $\theta$, defines an equivalence relation $k \sim l$ on the integers and for each $k$ we denote its equivalence class by $\Lambda(k)$.

To each $\Lambda_{i}$ there is associated an invariant subspace $Q(\theta)\left(\mathbf{R}^{\Lambda_{i}}\right)$ of $D(\theta)$. ( $\left.\mathrm{B}_{1}\right)$ bounds the dimension of this space. By $\left(\mathrm{B}_{2}\right)$ its diameter is at most $(M+2) N$, and by $\left(\mathrm{B}_{3}\right)$ at most $M$ invariant subspaces $Q(\theta)\left(\mathbf{R}^{\Lambda_{i}}\right)$ occupy a fixed site $m$, i.e. the $m$ th component of $Q(\theta)\left(\mathbf{R}^{\Lambda_{i}}\right)$ is identically 0 except for at most $M$ many $\Lambda_{i}$ 's.

There are of course also one-dimensional invariant subspaces, but we have no control on their regularity. The regularity of $Q(\theta)\left(\mathbf{R}^{\Lambda_{i}}\right)$, however, is controlled by $\left(\mathrm{B}_{4}\right)$.

We can extend $Q$ to a neighborhood $I+k \omega$ of any point $k \omega$ by the shift condition. In this way we get a piecewise smooth $Q$ defined on all $\mathbf{T}$ and, in particular, $D(\theta)$ is pure point with a complete set of eigenvectors for all $\theta$. The most important assumptions will be related to these eigenvalues and their eigenspaces.
(C) Eigenvalues. There is a piecewise smooth function $E(\theta)$ such that

$$
\{E(k \omega)\}_{k \in \Lambda_{i}} \text { are the eigenvalues of } \widetilde{D}_{\Lambda_{i}}(0) \forall i,
$$

and there are sets $\Omega_{i} \supset \Lambda_{i}$ such that
( $\mathrm{C}_{1}$ ) for all $n$, if $\inf _{l \in \Lambda_{i}}|E(l \omega)-E(n \omega)|<\alpha$, then

$$
\begin{array}{ll}
n \omega \in m \omega+\frac{1}{2} I \quad \text { for some } m \in \Omega_{i}, \\
Q(\theta)\left(\mathbf{R}^{\Lambda(n)}\right) \subset \mathbf{R}^{\Omega_{i}+n-m} \quad \forall \theta \in I ;
\end{array}
$$

$\left(\mathrm{C}_{2}\right)$ for all $i$, the resultant

$$
u_{\Omega_{i}}(x, \theta)=\operatorname{Res}\left(\operatorname{det}\left(D(\theta+x)_{\Omega_{i}}-t I\right), \operatorname{det}\left(D(\theta)_{\Omega_{i}}-t I\right)\right)
$$

satisfies

$$
\begin{gathered}
\left|u_{\Omega_{i}}\right| C^{\nu}<(4 M C)^{2 M^{2}} L^{\nu} \quad \forall \nu \leqslant s M^{2}+1, \\
\max _{0 \leqslant \nu \leqslant s M^{2}}\left|\frac{1}{(\nu!)^{2} L^{\nu}} \partial_{x}^{\nu} u_{\Omega_{i}}(x, \theta)\right| \geqslant \beta \quad \forall x \forall \theta \in \mathbf{T} ;
\end{gathered}
$$

$\left(\mathrm{C}_{3}\right) \# \Omega_{i} \leqslant M \quad \forall i$;
$\left(\mathrm{C}_{4}\right)$ the intervals $\{n \omega+I\}_{\operatorname{dist}\left(n, \Omega_{i}\right)<N}$ are pairewise disjoint;
$\left(\mathrm{C}_{5}\right) \operatorname{diam}\left(\Omega_{i}\right) \leqslant(1 / \lambda)^{\tau+2} \forall i$;
$\left(\mathrm{C}_{6}\right)$ for all $i$ and for all $x \in I$,

$$
\begin{gathered}
\left|u_{\Omega_{i}}\right| \mathcal{C}^{\nu}<\left(4 M \cdot \frac{1}{2} C\right)^{2 M^{2}} L^{\nu} \quad \forall \nu \leqslant s M^{2}+1, \\
\max _{0 \leqslant \nu \leqslant M^{2}}\left|\frac{1}{(\nu!)^{2} L^{\nu}} \partial_{\theta}^{\nu} u_{\Omega_{i}}(x, \theta)\right| \geqslant \beta\left(\prod_{k, l \in \Omega_{i}}\|x+(k-l) \omega\|\right) \quad \forall \theta \in \mathbf{T} .
\end{gathered}
$$

The sets $\Omega_{i}$ play a double role: by the first condition of $\left(\mathrm{C}_{1}\right)$ they restrict the possible eigenvalues $E(n \omega)$ that are equal to $E(k \omega), k \in \Lambda_{i}$, up to an approximation $\alpha$ if $n \omega \notin \Omega_{i} \omega+\frac{1}{2} I$ then $|E(k \omega)-E(n \omega)| \geqslant \alpha$; by the second condition of $\left(\mathrm{C}_{1}\right)$ they also describe the location of the eigenvectors corresponding to such eigenvalues.
$u_{\Omega_{i}}$ measures the difference of the eigenvalues of the submatrices $D(\theta+x)_{\Omega_{i}}$ and $D(\theta)_{\Omega_{i}}$, and among these eigenvalues there are precisely certain eigenvalues of $D(\theta+x)$ and $D(\theta)$ respectively, when $\theta+x, \theta \in I$.

The norm of $u_{\Omega_{i}}$ is in $\left(\mathrm{C}_{2}\right)$ with respect to the variable $x$ and in $\left(\mathrm{C}_{6}\right)$ with respect to the variable $\theta$.

Remark 1. Hence, a normal form is an infinite-dimensional symmetric matrix, parametrized by $\theta \in \mathbf{T}$, satisfying a shift condition with respect to a Diophantine rotation of $\mathbf{T}$ and satisfying conditions (A), (B) and (C). The normal form depends on several parameters: $C, K, r, N$ give upper estimates on $D ; M, N, K$ give information about $Q$; $\alpha, C, M, L, \beta, N, \lambda$, which give lower estimates on the spectrum. It also depends on the parameters $s \geqslant 2$ and $\varkappa>0, \tau>1$ (through $\omega$ ) which will be kept fixed.

Remark 2. Capital letters can be increased and small letters decreased, but not independently because the parameters are interrelated. For example if we increase $M$ we must also increase $L$, and if we increase $L$ we must decrease $\beta$.

Remark 3. It follows from the generalized Young inequality [5] that ( $\mathrm{A}_{1}$ ) implies an estimate of $D$ in the operator norm on $l^{2}(\mathbf{Z}) \rightarrow l^{2}(\mathbf{Z})$ :

$$
|D(\theta)|_{\mathcal{C}^{\nu}}<C \frac{e^{r}+1}{e^{r}-1} K^{\nu}<C \frac{4}{r} K^{\nu},
$$

if $r \leqslant 1$. Notice that this provides an estimate of the eigenvalues of $D(\theta)$ and of their first derivatives with respect to $\theta$.

Remark 4. $u_{\Omega_{i}}(x, \theta)=\operatorname{det} P_{i}\left(\theta, D(x+\theta)_{\Omega_{i}}\right)$, where $P_{i}(\theta, t)=\operatorname{det}\left(D(\theta)_{\Omega_{i}}-t I\right)$. The $\operatorname{det}\left(D(\theta)_{\Omega_{i}}-t I\right)$ is a sum of $\leqslant M$ ! many monomials of degree $\# \Omega_{i}$ in the matrix elements of $D(\theta)_{\Omega_{i}}-t I$. Introducing $D(\theta+x)_{\Omega_{\mathrm{i}}}$ for $t$ we get a matrix whose elements are $\leqslant M!(M)^{M-1}$ many monomials of degree $\# \Omega_{i}$ in the matrix elements of these two matrices and in their differences. Taking the determinant gives this number to the power $\# \Omega_{i}$,
times $\left(\# \Omega_{i}\right)$ !. This shows that $u_{\Omega_{i}}(x, \theta)$ is a sum of less than $M^{2 M^{2}}$ many monomials of degree less than $\# \Omega_{i} \times \# \Omega_{i}$ in the components of $D(\theta)_{\Omega_{i}}$ and $D(\theta+x)_{\Omega_{i}}$ and their differences. In particular, if $C \geqslant 1$ it follows that

$$
\left|u_{\Omega_{i}}\right|_{\mathcal{C}^{\nu}}<(4 M C)^{2 M^{2}} K^{\nu} \quad \forall \nu \geqslant 0
$$

Idea of proof-continued. Consider a perturbation $D(\theta)+\varepsilon F_{1}(\theta)$ of a normal form. Condition (B) implies that $D(\theta)$ is pure point with finite-dimensional eigenvectors (i.e. eigenvectors that only occupy finitely many sites). Conditions (A) and (B) will enable us to solve the equation

$$
\left[X_{1}(\theta), D(\theta)\right]=F_{1}(\theta)-G_{1}(\theta)
$$

for some anti-symmetric matrix $X_{1}$ and some symmetric modification $G_{1}$-with reasonable estimates. (This will be done in Lemma 5.) Then, with $U(\theta, \varepsilon)=e^{\varepsilon X_{1}(\theta)}$ we have

$$
\begin{equation*}
U(\theta, \varepsilon)^{*}\left(D(\theta)+\varepsilon F_{1}(\theta)\right) U(\theta, \varepsilon)=D(\theta)+\varepsilon G_{1}(\theta)+\mathbf{O}^{2}(\varepsilon) \tag{8}
\end{equation*}
$$

Conditions $\left(\mathrm{C}_{1}\right)-\left(\mathrm{C}_{3}\right)$ will then guarantee that the main remaining part $D+\varepsilon G_{1}$ is pure point and satisfies conditions (A) and (B) (with different parameters). The conditions $\left(\mathrm{C}_{4}\right)-\left(\mathrm{C}_{5}\right)$ are added in order to assure that $D+\varepsilon G_{1}$ also satisfies (C) (with different parameters), and hence permits an inductive construction-condition ( $\mathrm{C}_{6}$ ) will play a role only when we take the limit of this construction. That $D+\varepsilon G_{1}$ will be in normal form will be proved in Lemma 7 and the iterative construction is carried out in Proposition 8.

For technical reasons we will in Lemma 7 consider the solution of (8) not only to the first order in $\varepsilon$ but to some much higher order-such a solution is derived in Corollary 6.

Consider now a symmetric matrix $F$, smoothly parametrized on $\mathbf{T}$, satisfying the shift condition and such that

$$
\begin{equation*}
\left|F_{m, n}\right|_{\mathcal{C}^{\nu}}<\varepsilon e^{-|m-n| \varrho} K_{9}^{\nu} \quad \forall \nu \geqslant 0 \tag{9}
\end{equation*}
$$

Let $D$ be in normal form and let $q^{k}(\theta)$ be an eigenvector of $D(\theta)$ corresponding to the eigenvalue $E(\theta+k \omega)$. The following result will not make any use of condition (C).

Lemma 5. Assume $r \leqslant 1 \leqslant C$. For all $0<\delta<1$, there are a smooth symmetric matrix $G$ and a smooth anti-symmetric matrix $X$ on $\mathbf{T}$, both satisfying the shift condition, such that

$$
\begin{gathered}
\left\langle q^{m}(\theta), G(\theta) q^{n}(\theta)\right\rangle=0 \quad \text { if }|E(\theta+m \omega)-E(\theta+n \omega)| \geqslant \delta, \\
\\
{[X, D]=F-G}
\end{gathered}
$$

and for all $\nu \geqslant 0$,

$$
\begin{gather*}
r \delta\left|X_{m, n}{\mid \mathcal{C}^{\nu}}+\left|G_{m, n}\right|_{\mathcal{C}^{\nu}}<\gamma_{3} N^{2} e^{6 M N \varrho} \varepsilon e^{-|m-n| \varrho} K_{10}^{\nu},\right. \\
K_{10}=\max \left(\gamma_{3}(C / r \delta)^{2 M^{2}\left(2 M^{2}+1\right)+1} K, \gamma_{3} N^{\tau}, \gamma_{3} /|I|, K_{9}\right), \tag{10}
\end{gather*}
$$

where $\gamma_{3}\left(\gamma_{0}, \varkappa, \tau, M\right)$ increases superexponentially in $M$. Moreover, if $F_{m, n} \equiv 0$ when $|m-n| \geqslant J^{2}-(2 M+4) N$, then both $X_{m, n}$ and $G_{m, n} \equiv 0$ for $|m-n| \geqslant J^{2}$.

Proof. Let us first consider the case when $F$ has only two non-zero "diagonals":

$$
F_{m, n} \equiv 0 \quad \text { unless } n-m= \pm k
$$

By a partition of unity we may arrange so that $F_{0, k}(\theta)$ is supported in a small interval $I^{\prime}$ of length $3 t$ where

$$
3 t=\min \left(\frac{\varkappa}{(8(M+2) N)^{r}},|I|\right)
$$

and there is no restriction in assuming that $I^{\prime}$ is contained in the middle of $I$. Such a localized matrix will then satisfy

$$
|F|_{\mathcal{C}^{\nu}}<2^{3} \varepsilon e^{-|k| e} \max \left(K_{9}, \gamma_{0} / t\right)^{\nu}
$$

where, we recall, the norm is the operator norm. (This follows from Lemmas A1-A2.)
Due to the shift condition and the Diophantine condition, only for few entries on each of the two non-zero "diagonals" will $F_{i, j} \not \equiv 0$. Indeed two entries on a "diagonal" for which $F_{i, j} \not \equiv 0$ will be more than $8(M+2) N$ sites apart.

Consider the matrix $Q$ given in condition (B) and let $\widetilde{F}=Q^{*} F Q$ and $\widetilde{D}=Q^{*} D Q$. By ( $\mathrm{B}_{4}$ ) and Lemma A2 we get

$$
\begin{aligned}
& |\widetilde{F}|_{\mathcal{C}^{\nu}}<2^{7} \varepsilon e^{-|k| \varrho} \max \left(K_{9}, \gamma_{0} / t, K\right)^{\nu}, \\
& |\widetilde{D}|_{\mathcal{C}^{\nu}}<\frac{2^{6}}{r} C K^{\nu} .
\end{aligned}
$$

Most components of $\widetilde{F}$ are identically 0 . In the strip

$$
S: k-4(M+2) N \leqslant m+n \leqslant k+4(M+2) N
$$

for example, $\widetilde{F}_{m, n} \equiv 0$ except for ( $m, n$ ) belonging to the two "squares"

$$
B((k, 0) ; N)=\{|m-k| \leqslant N,|n| \leqslant N\} \quad \text { and } \quad B((0, k) ; N)
$$

(This follows from the first part of $\left(\mathrm{B}_{2}\right)$.) Moreover, in these "squares" $\widetilde{F}_{m, n} \equiv 0$ unless

$$
(m, n) \in Z_{k} \times Z_{0} \cup Z_{0} \times Z_{k}, \quad \# Z_{i} \leqslant M, i=0, k
$$

(This follows from $\left(B_{3}\right)$.)
Assume now that $|k|>(2 M+2) N$ and let $\Lambda^{i}$ be the union of all $\Lambda_{\nu}$ such that $\Lambda_{\nu} \cap Z_{i} \neq 0, i=0, k$. Then for $i=0, k$,

$$
\# \Lambda^{i} \leqslant M^{2} \quad \text { and } \quad \Lambda^{i} \subset[i-(M+1) N, i+(M+1) N] .
$$

(This follows from ( $\mathrm{B}_{1}$ ) and the second part of $\left(\mathrm{B}_{2}\right)$.) In particular $\Lambda^{0} \cap \Lambda^{k}=\varnothing$.
By Lemma 2 we can solve on $I^{\prime}$

$$
\left[\bar{X}_{\Lambda}, \widetilde{D}_{\Lambda}\right]=\widetilde{F}_{\Lambda}-\bar{G}_{\Lambda}, \quad \Lambda=\Lambda^{0} \cup \Lambda^{k}
$$

with $d=2 M^{2}$ and with $\delta$ replaced by $\delta / 8 M^{2}$. This gives

$$
\begin{gathered}
r \delta\left|\bar{X}_{m, n}\right|_{\mathcal{C}^{\nu}}+\left|\bar{G}_{m, n}\right|_{\mathcal{C}^{\nu}<\gamma_{4} \varepsilon e^{-|k| \varrho}\left(K^{\prime}\right)^{\nu} \quad \forall \nu \geqslant 0,}^{K^{\prime}=\max \left(\gamma_{4}(C / r \delta)^{2 M^{2}\left(2 M^{2}+1\right)+1} K, \max \left(K, \gamma_{0} / t, K_{9}\right)\right),}
\end{gathered}
$$

and $\bar{X}_{\Lambda}$ and $\bar{G}_{\Lambda}$ has compact support in $I^{\prime}$.
Let us now define $\widetilde{X}_{m, n}=\bar{X}_{m, n}$ for $(m, n) \in \Lambda^{k} \times \Lambda^{0} \cup \Lambda^{0} \times \Lambda^{k}$ and $\widetilde{X}_{m, n}=0$ otherwise, and define $\widetilde{G}_{m, n}$ in the same way. Then

$$
([\widetilde{X}, \widetilde{D}])_{m, n}=\widetilde{F}_{m, n}-\widetilde{G}_{m, n}
$$

for all $(m, n) \in S$. If we now let $X^{\prime}=Q \widetilde{X} Q^{*}$ and $G^{\prime}=Q \widetilde{G} Q^{*}$, then $X_{m, n}^{\prime}$ and $G_{m, n}^{\prime}$ are $\equiv 0$ unless ( $m, n$ ) belongs to the larger "squares" $B((k, 0) ;(M+2) N)$ or $B((0, k) ;(M+2) N)$. Moreover, on $I^{\prime}$,

$$
\left(\left[X^{\prime}, D^{\prime}\right]\right)_{m, n}=F_{m, n}-G_{m, n}^{\prime}
$$

for all $(m, n)$ in the smaller strip

$$
S^{\prime}: k-(4 M+6) N \leqslant m+n \leqslant k+(4 M+6) N
$$

(Here we use again the first part of $\left(\mathrm{B}_{2}\right)$.)
In order to get a solution we now extend $X^{\prime}$ to a matrix $X$ defined for all $\theta$ satisfying the shift condition and such that

$$
X_{m, n}(\theta)=X_{m, n}^{\prime}(\theta), \quad(m, n) \in S, \theta \in I^{\prime}
$$

and we do the same for $G^{\prime}$. This gives the solution. In order to recover the exponential decay we must multiply the estimates by $e^{(2 M+4) N \varrho} \leqslant e^{6 M N \varrho}$.

In the case $|k| \leqslant(2 M+2) N, \Lambda^{0}$ and $\Lambda^{k}$ may not be disjoint. But then $\Lambda \times \Lambda \subset S$ and we define $\widetilde{X}_{m, n}$ to be $\bar{X}_{m, n}$ if $(m, n) \in \Lambda \times \Lambda$, and to be 0 otherwise-and the same for $\widetilde{G}_{m, n}$. With this difference the construction of the solution is the same in both cases.

Suppose now that $F$ has two non-zero "diagonals". Take a partition of unity $\left\{\phi_{i}\right\}$ supported in intervals $I_{i}^{\prime}$ of length $<3 t$ (and such that no three intervals have nonvoid common intersection). Then we apply the above result to each $\phi_{i} F$ obtaining $X_{i}$ and $G_{i}$. Then $\sum X_{i}$ and $\sum G_{i}$ will be a solution with estimates that are at most const $\cdot(2 M+4) N \leqslant$ const $\cdot M N$ larger than those of $X_{i}$ and $G_{i}$ respectively.

Suppose finally that $F$ is arbitrary and decompose it into $\sum F_{k}$, where

$$
\left(F_{k}\right)_{m, n} \equiv 0 \quad \text { unless } m-n= \pm k
$$

Then we construct $X_{k}$ and $G_{k}$ as above, and $\sum X_{k}$ and $\sum G_{k}$ will be solutions with estimates that are at most const $\cdot(2 M+4) N \leqslant$ const $\cdot M N$ larger than those of $X_{k}$ and $G_{k}$ respectively. This completes the proof of the lemma.

Notice that the solution looses no smoothness with respect to $F$ but only with respect to $D$. This is important in the next corollary, where we shall apply this construction several times.

Corollary 6. Assume $r, \varrho \leqslant 1 \leqslant C$ and

$$
\begin{gather*}
1>6 a+8 b,  \tag{11.1}\\
\varepsilon<\left(\frac{\varrho^{3}}{2^{16} \gamma_{3} N^{2} e^{6 M N \varrho C}}\right)^{8} \text { and } \varepsilon^{a}<\left(\frac{a \varrho}{2^{5} M N}\right)^{4},  \tag{11.2}\\
c>\left(8 M^{4}+4 M^{2}+2\right) b,  \tag{11.3}\\
\varepsilon^{c}<\min \left(\frac{1}{\gamma_{3}^{2}}\left(\frac{r}{C}\right)^{8 M^{4}+4 M^{2}+2}, \frac{1}{\gamma_{3}} N^{-\tau} K, \frac{1}{\gamma_{3}}|I| K, \frac{K}{K_{9}}\right) . \tag{11.4}
\end{gather*}
$$

Then there are a smooth orthogonal matrix $U$ and a smooth symmetric matrix $G$ on $\mathbf{T}$, both satisfying the shift condition, such that

$$
\begin{aligned}
U^{*}(D+F) U & =D+G+F^{\prime}, \\
\left\langle q^{m}(\theta), G(\theta) q^{n}(\theta)\right\rangle & =0 \quad \text { if }|E(\theta+m \omega)-E(\theta+n \omega)| \geqslant \varepsilon^{b}, \\
G_{m, n} & \equiv 0 \quad \text { if }|m-n| \geqslant(1 / \varepsilon)^{a},
\end{aligned}
$$

and for all $\nu \geqslant 0$,

$$
\begin{aligned}
\left|(U-I)_{m, n}\right|_{\mathcal{C}^{\nu}} & <\sqrt{\varepsilon} e^{-|m-n| \varrho / 2}\left((1 / \varepsilon)^{c} K\right)^{\nu} \\
\left|G_{m, n}\right|_{\mathcal{C}^{\nu}} & <\sqrt{\varepsilon} e^{-|m-n| \varrho / 2}\left((1 / \varepsilon)^{c} K\right)^{\nu} \\
\left|F_{m, n}^{\prime}\right| \mathcal{C}^{\nu} & <\varepsilon^{(1 / \varepsilon)^{a / 2} / 2} e^{-|m-n| \varrho / 2}\left((1 / \varepsilon)^{c} K\right)^{\nu}
\end{aligned}
$$

Proof. Let us first fix $\delta=\varepsilon^{b}$ and $J=(1 / \varepsilon)^{a / 2}$. Let us then define for $j \geqslant 1$,

$$
\begin{aligned}
{\left[X_{j}, D\right] } & =\widehat{F}_{j}-\operatorname{Proj}_{D, \delta}\left(\widehat{F}_{j}\right) \\
G_{j} & =\operatorname{Proj}_{D, \delta}\left(\widehat{F}_{j}\right) \\
F_{j+1} & =e^{-X_{j}} \ldots e^{-X_{1}}(D+F) e^{X_{1}} \ldots e^{X_{j}}-D-\left(G_{1}+\ldots+G_{j}\right)
\end{aligned}
$$

where $F_{1}=F, \widehat{F}_{j}$ is the truncation of $F_{j}$ at distance $(1 / \varepsilon)^{a}-(2 M+4) N$ from the main diagonal and $\operatorname{Proj}_{D, \delta}\left(\widehat{F}_{j}\right)$ is the $G$ defined in Lemma 5. We shall estimate these matrices using this lemma.

If $K_{10}$ is defined by (10) and if

$$
\varrho_{1}=\varrho, \quad \varrho_{2}=\varrho_{1}-\frac{\varrho}{2 J}, \quad \varepsilon_{1}=\varepsilon, \quad \varepsilon_{2}=\varepsilon^{3 / 2}
$$

then

$$
\begin{aligned}
& \left|\left(G_{1}\right)_{m, n}\right|_{\mathcal{C}^{\nu}}<A \varepsilon_{1} e^{-|m-n| \varrho_{1}} K_{10}^{\nu} \\
& \left|\left(X_{1}\right)_{m, n}\right|_{\mathcal{C}^{\nu}}<\frac{1}{r \delta} A \varepsilon_{1} e^{-|m-n| \varrho_{1}} K_{10}^{\nu}
\end{aligned}
$$

with $A=\gamma_{3} N^{2} e^{6 M N \varrho_{1}}$. In order to estimate $e^{ \pm X_{1}}-I$ we note that an $i$-fold product of $X_{1}$ 's satisfies

$$
\left|\left(X_{1} \cdot \ldots \cdot X_{1}\right)_{m, n}\right|_{C^{\nu}}<\frac{1}{4}\left(\frac{4^{2}}{\varrho_{1}-\varrho_{2}} \cdot \frac{1}{r \delta} A \varepsilon_{1}\right)^{i} e^{-|m-n| \varrho_{2}} K_{10}^{\nu}
$$

Using the power series expansion of $e^{ \pm X_{1}}$ we get

$$
\left|\left(e^{ \pm X_{1}}-I\right)_{m, n}\right|_{\mathcal{C}^{\nu}}<\frac{2^{5} J}{\varrho} \cdot \frac{1}{r \delta} A \varepsilon_{1} e^{-|m-n| \varrho_{2}} K_{10}^{\nu}
$$

under conditions (11.1)-(11.2). Using power series expansions and estimating products, we also get

$$
\left|\left(F_{2}\right)_{m, n}\right|_{\mathcal{C}^{\nu}}<\varepsilon_{2} e^{-|m-n| \varrho_{2}} K_{10}^{\nu}
$$

(Here we used the second part of (11.2) in order to estimate the difference $F_{1}-\widehat{F}_{1}$.)
Proceeding in the same way with

$$
\varrho_{j}=\varrho-\frac{\varrho}{2 J}(j-1) \quad \text { and } \quad \varepsilon_{j+1}=\varepsilon^{(j+2) / 2}
$$

we get, under conditions (11.1)-(11.2), that

$$
\begin{aligned}
\left|\left(G_{j}\right)_{m, n}\right| \mathcal{C}^{\nu} & <A \varepsilon_{j} e^{-|m-n| \varrho_{j}} K_{10}^{\nu} \\
\left|\left(X_{j}\right)_{m, n}\right| \mathcal{C}^{\nu} & <\frac{1}{r \delta} A \varepsilon_{j} e^{-|m-n| \varrho_{j}} K_{10}^{\nu} \\
\left|\left(e^{ \pm X_{j}}-I\right)_{m, n}\right| \mathcal{C}^{\nu} & <\frac{2^{5} J}{\varrho} \cdot \frac{1}{r \delta} A \varepsilon_{j} e^{-|m-n| \varrho_{j+1}} K_{10}^{\nu} \\
\left|\left(F_{j+1}\right)_{m, n}\right| \mathcal{C}^{\nu} & <\varepsilon_{j+1} e^{-|m-n| \varrho_{j+1}} K_{10}^{\nu}
\end{aligned}
$$

By (11.3)-(11.4) we get $K_{10}<(1 / \varepsilon)^{c} K$.
We now let $G=G_{1}+\ldots+G_{J}, U=e^{X_{1}} \ldots e^{X_{J}}$ and $F^{\prime}=F_{J+1}$. These matrices will satisfy the estimates of the corollary.

## 5. The inductive lemma

Lemma 7. Let $D$ be in normal form on an interval I with parameters $C, K, r, M, N, \alpha$, $L, \beta, \lambda$, and let $a<b<c$ be numbers restricted by

$$
\frac{1}{\tau M^{3 s M^{3}}} \leqslant a<\frac{b}{20 s \tau M^{4}}<\frac{c}{100 s^{2} \tau M^{8}}, \quad c \leqslant \frac{1}{5 s M^{2 s M^{3}}} .
$$

Assume, as simplification, that

$$
1 \leqslant L \leqslant K, \quad 8 \leqslant M, \quad 1<C<2, \quad r, \alpha, \beta \leqslant 1
$$

Let $F$ be a symmetric matrix satisfying the shift condition. Assume that

$$
\begin{gathered}
\lambda \leqslant|I| \leqslant \beta / L \\
\left|F_{m, n}\right|_{\mathcal{C}^{\nu}}<\varepsilon e^{-|m-n| r} K^{\nu} \quad \forall \nu \geqslant 0 .
\end{gathered}
$$

Then there is a constant $\Gamma=\Gamma\left(\varkappa, \tau, s, \gamma_{0}, M\right)$, superexponentially decaying in $M$, such that if

$$
|\varepsilon|<\Gamma\left[\frac{r^{\tau} \alpha \beta \lambda^{\tau^{2}}}{K N^{\tau}} e^{-N r}\right]^{e^{e^{s M^{4}}}}
$$

then there is a smooth orthogonal matrix $U$, satisfying the shift condition, such that

$$
\left|(U-I)_{m, n}\right|_{\mathcal{C}^{\nu}}<\sqrt{\varepsilon} e^{-|m-n| r^{\prime}}\left(K^{\prime}\right)^{\nu}
$$

and

$$
U^{*}(D+F) U=D^{\prime}+F^{\prime}
$$

with $D^{\prime}$ in normal form on an interval $I^{\prime}$, with parameters

$$
\begin{array}{llr}
C^{\prime}=\left(1+\varepsilon^{1 / 2}\right) C, & K^{\prime}=(1 / \varepsilon)^{c} K, & r^{\prime}=\frac{1}{2} r \\
\lambda^{\prime}=\left(\frac{1}{9}\right)^{M^{\prime}} \varepsilon^{b}, & M^{\prime}=M^{s M^{3}}, & N^{\prime}=(1 / \varepsilon)^{a} \\
\alpha^{\prime}=\varepsilon^{c}, & L^{\prime}=K, & \beta^{\prime}=\varepsilon^{b} K
\end{array}
$$

and with

$$
\begin{gathered}
2 \lambda^{\prime} \leqslant\left|I^{\prime}\right| \leqslant \varepsilon^{b} \\
\left|F_{m, n}^{\prime}\right| \mathcal{C}^{\nu}<\varepsilon^{(1 / \varepsilon)^{a / 2} / 2} e^{-|m-n| r^{\prime}}\left(K^{\prime}\right)^{\nu}
\end{gathered}
$$

In addition,

$$
\begin{gathered}
|E(l \omega)-E(k \omega)|<M^{\prime} \frac{K}{r} \varepsilon^{b} \quad \forall l \in \Lambda^{\prime}(k), \\
Q^{\prime}(\theta)\left(\mathbf{R}^{\Lambda^{\prime}(k)}\right) \subset \sum_{l \in \Lambda^{\prime}(k)} Q(\theta)\left(\mathbf{R}^{\Lambda(l)}\right) \quad \forall \theta \in I^{\prime},
\end{gathered}
$$

$D^{\prime}$ is in normal form with the same parameters also on $\frac{1}{2} I^{\prime}$.
Finally, if $M \geqslant 2 \tau$ then the closure of the sets

$$
\begin{gathered}
\left\{n \omega:|E(n \omega)-E(n \omega+l \omega)|<2 M^{\prime}(K / r) \varepsilon^{b}\right\} \quad \text { for all } 4(1 / \lambda)^{\tau+2}<|l| \leqslant M^{\prime} N^{\prime} \\
\left\{n \omega:|E(n \omega)-E(n \omega+l \omega)|<2 \varepsilon^{1 / 8}\right\} \quad \text { for all } M^{\prime} N^{\prime}<|l| \leqslant 4\left(1 / \lambda^{\prime}\right)^{\tau+2}
\end{gathered}
$$

are unions of, respectively, at most $\varepsilon^{-b / 5 s M^{2}}$ and $\varepsilon^{-M^{4} b}$ many components, each component being of length, respectively, at most $\varepsilon^{b / 4 s M^{2}}$ and $\varepsilon^{2 M^{4} b}$.

In order to avoid any confusion, let us point out that $Q, \Lambda_{i}, \Omega_{i}$ and $Q^{\prime}, \Lambda_{i}^{\prime}, \Omega_{i}^{\prime}$ are the orthogonal matrix and sets referred to in the definition of the normal forms of $D$ and $D^{\prime}$ respectively. $\varkappa, \tau, s$ are constants which are the same for both $D$ and $D^{\prime}$.

Proof. In the sequel we shall, without further mentioning, denote by $\gamma_{4}, \gamma_{5}, \ldots$ constants that only depend on $\varkappa, \tau, s, \gamma_{0}, M$, and that decay superexponentially in $M$.

In the proof we shall verify several inequalities involving $\varepsilon$ and the parameters $C, K, r, \ldots$. These inequalities will be fulfilled if $\Gamma$ is small enough and it is important that $\Gamma$ does not depend on the parameters but only on $\varkappa, \tau, s, \gamma_{0}$-we will not in general stress this any further.

Construction of $U$ and $D^{\prime}$. Condition (11) is fulfilled with $\varrho=r$ and $K_{9}=K$ since

$$
\varepsilon^{\tau a}<\gamma_{4}\left(\frac{r^{\tau} \lambda^{\tau}}{N^{\tau}}\right)^{4}
$$

(Here we used $r \leqslant 1<C<2$ and $\lambda \leqslant|I|$.) So we can apply Corollary 6. Clearly $U, D^{\prime}=$ $D+G$ and $F^{\prime}$ satisfy the required estimates and it remains to verify that $D^{\prime}$ satisfies conditions (B) and (C) for the appropriate parameters.

Construction of $\Lambda_{i}^{\prime}$. We define an equivalence relation $\sim^{\prime}$ on $\mathbf{Z}$ by declaring that two integers $k$ and $l$ are equivalent if $|k-l|<2 N^{\prime}$ and

$$
|E(k \omega)-E(l \omega)|<2^{9} M \frac{K}{r} \varepsilon^{b}
$$

or if there is a sequence $k=k_{1}, k_{2}, \ldots, k_{n+1}=l$ such that each consecutive pair $\left(k_{j}, k_{j+1}\right)$ is equivalent in the first sense. This equivalence relation defines the decomposition $\bigcup \Lambda_{i}^{\prime}=\mathbf{Z}$. We shall show that condition (B) is fulfilled for any interval $I^{\prime}$ of length $\leqslant \varepsilon^{b}$.

Verification of $\left(\mathrm{B}_{1}\right)$. Let $k \in \Lambda_{i}$ be given and consider a sequence of integers $k=$ $k_{1} \sim^{\prime} k_{2} \sim^{\prime} \ldots \sim^{\prime} k_{n}$, where $\left|k_{j}-k_{j+1}\right|<2 N^{\prime}$. We shall show that the number $n$ of such integers must be $\leqslant M^{\prime} / 2^{9} M \leqslant \frac{1}{2} M^{\prime}$, so assume that $n=M^{\prime} / 2^{9} M$.

We have that

$$
\eta:=M^{\prime} \frac{K}{r} \varepsilon^{b}<\alpha
$$

if

$$
\begin{equation*}
\varepsilon^{b}<\gamma_{5} \frac{\alpha r}{K} \tag{12}
\end{equation*}
$$

(For any given constant $\gamma_{5}$, (12) is fulfilled if $\Gamma$ is small enough. Of course, it is sufficient that $\gamma_{5}=1 / M^{\prime}$, but since we will refer to a similar condition several times we prefer to give (12) a more general form. This practice will be used in the sequel without explicit mention.)

Any $l=k_{j}$ must satisfy

$$
|E(k \omega)-E(l \omega)|<\eta
$$

so $l \omega \in m \omega+\frac{1}{2} I$ for some $m \in \Omega_{i}$ by $\left(\mathrm{C}_{1}\right)$. Condition $\left(\mathrm{C}_{1}\right)$ also implies that the eigenvector $q^{l}(\theta) \in \mathbf{R}^{\Omega_{i}+l-m}$ for all $\theta \in I$. In particular, $E(l \omega)$ is an eigenvalue of $D_{\Omega_{i}+l-m}(0)=$ $D_{\Omega_{i}}((l-m) \omega)$. Since $E(k \omega)$ is an eigenvalue of $D_{\Omega_{i}}(0)$ it follows that (l-m) $\omega$ must belong to the set

$$
\left\{x \in I:\left|u_{\Omega_{i}}(x, 0)\right|<\eta(16 / r)^{M^{2}}\right\}
$$

the eigenvalues of $D_{\Omega_{i}}$ are bounded by $C \cdot 4 / r<8 / r$ since $C<2$. By ( $\mathrm{C}_{2}$ ) and by Lemma 3 there are not more than

$$
M_{1}:=2^{s M^{2}}\left[\frac{2(8 M)^{2 M^{2}} L\left(s M^{2}+1\right)^{2}}{\beta}|I|+1\right]
$$

many components of this set, each component being of length at most

$$
\frac{2}{L}\left(\frac{16}{r}\right)^{1 / s}\left(\frac{2 \eta}{\beta}\right)^{1 / s M^{2}}
$$

Hence, if $l \sim^{\prime} k$ then $l \omega$ must belong to one out of at most $M_{1} M$ many intervals of this length—because $\# \Omega_{i} \leqslant M$ by $\left(\mathrm{C}_{3}\right)$.

Due to the size of the intervals and the distance between the $k_{j}$ 's, no such interval can contain more than one number $k_{j} \omega$ since

$$
\begin{gather*}
b>2 s \tau M^{2} a,  \tag{13.1}\\
\varepsilon^{b / 2}<\gamma_{6} \frac{\beta}{K} r^{M^{2}+1} \tag{13.2}
\end{gather*}
$$

This forces $n$ to be smaller than $M_{1} M$, which is smaller than $M^{\prime} / 2^{9} M$. (Here we use $L|I| \leqslant \beta$ and $M \geqslant 8$.)

Construction of $Q^{\prime}$. Consider now the block matrix (written symbolically)

$$
\widetilde{D}=Q^{*} D Q=\left(\begin{array}{ccccc}
\ddots & & & & \\
& \widetilde{D}_{\Lambda_{-1}} & & & \\
& & \widetilde{D}_{\Lambda_{0}} & & \\
& & & \widetilde{D}_{\Lambda_{1}} & \\
& & & & \ddots
\end{array}\right)
$$

where each block is defined on $I$ and of dimension $\leqslant M$. We now block diagonalize these blocks using Lemma 1 with $d \leqslant M$ and $\delta=\left(2^{7} K / r\right) \varepsilon^{b}$;

$$
\widetilde{\widetilde{D}}=\widetilde{Q}^{*} \widetilde{D} \widetilde{Q}=\left(\begin{array}{lllll}
\ddots & & & & \\
& \tilde{\widetilde{D}}_{\Lambda_{-1}} & & & \\
& & \tilde{\widetilde{D}}_{\Lambda_{0}} & & \\
& & & \widetilde{\tilde{D}}_{\Lambda_{1}} & \\
& & & & \ddots
\end{array}\right)
$$

where the blocks now decompose into subblocks

$$
\widetilde{\widetilde{D}}_{\Lambda_{i}}=\left(\begin{array}{ccccc}
\ddots & & & & \\
& \widetilde{\widetilde{D}}_{\Lambda_{i,-1}} & & & \\
& & \widetilde{\widetilde{D}}_{\Lambda_{i, 0}} & & \\
& & & \widetilde{\widetilde{D}}_{\Lambda_{i, 1}} & \\
& & & & \ddots
\end{array}\right)
$$

Then $\widetilde{Q}$ will be a block matrix over the decomposition $\bigcup \Lambda_{i}=\mathbf{Z}$.
Since

$$
|D|_{\mathcal{C}^{\nu}} \leqslant \frac{2^{3}}{r} K^{\nu} \quad \text { and } \quad|\widetilde{D}|_{\mathcal{C}^{\nu}} \leqslant \frac{2^{7}}{r} K^{\nu}=: C_{1} K_{1}^{\nu}
$$

the decomposition into subblocks is constant over intervals of length

$$
\leqslant \frac{\delta}{C_{1} K_{1}}=\varepsilon^{b}
$$

i.e. over the interval $I^{\prime}$. Hence, two eigenvalues $E(k \omega)$ and $E(n \omega)$ of a subblock $\tilde{\widetilde{D}}_{\Lambda_{i, j}}(0)$ will be separated by at most

$$
4 d \delta \leqslant 2^{9} M \frac{K}{r} \varepsilon^{b}
$$

Since the blocks have diameter less than $M N$ and since $M N<2 N^{\prime}$ if for example

$$
\begin{equation*}
\varepsilon^{a}<\gamma_{7} \frac{1}{N} \tag{14}
\end{equation*}
$$

it follows that $k \sim^{\prime} n$. Hence, $\widetilde{\widetilde{D}}(\theta)$ decomposes as blocks over the decomposition $\mathbf{Z}=\bigcup \Lambda_{i}^{\prime}$ for all $\theta \in I^{\prime}$.

Let now $E_{k}(\theta)$ be a continuous branch of an eigenvalue of $\widetilde{D}_{\Lambda(k)}(\theta)$ starting at $E(k \omega)$ for $\theta=0$. Suppose that $\left\langle q^{k}(\theta), G(\theta) q^{l}(\theta)\right\rangle \neq 0$ for some $\theta \in I^{\prime}$, where $q^{k}(\theta), q^{l}(\theta)$ are eigenvectors corresponding to $E_{k}(\theta)$ and $E_{l}(\theta)$ respectively. Then we know by Corollary 6 that

$$
\begin{gathered}
\left|E_{k}(\theta)-E_{l}(\theta)\right|<\varepsilon^{b}, \\
|k-l|<(1 / \varepsilon)^{a}+2(M+2) N .
\end{gathered}
$$

By (14) this implies-recall Remark 3-that

$$
\left|E_{k}(0)-E_{l}(0)\right|<\varepsilon^{b}+\frac{16}{r} K \varepsilon^{b} \quad \text { and } \quad|k-l|<2 N^{\prime}
$$

i.e. $k \sim^{\prime} l$. Hence, also $\tilde{\tilde{G}}(\theta)$ decomposes as blocks over the decomposition $\mathbf{Z}=\bigcup \Lambda_{i}^{\prime}$ for each $\theta \in I^{\prime}$, so with $Q^{\prime}=Q \widetilde{Q}$ we have

$$
\widetilde{D}^{\prime}=\left(Q^{\prime}\right)^{*} D^{\prime} Q^{\prime}=\prod \widetilde{D}_{\Lambda_{i}^{\prime}}^{\prime}
$$

Verification of $\left(\mathrm{B}_{2}\right)-\left(\mathrm{B}_{4}\right)$. If $Q_{m, l} \widetilde{Q}_{l, n} \neq 0$ then

$$
|n-m| \leqslant|n-l|+|l-m| \leqslant M N+N \leqslant N^{\prime}
$$

under condition (14). This proves the first part of $\left(\mathrm{B}_{2}\right)$-the second part is fulfilled by construction, because we have seen $\# \Lambda_{i}^{\prime} \leqslant \frac{1}{2} M^{\prime}$.

Assume $Q_{m, l} \widetilde{Q}_{l, n} \not \equiv 0$. For each $m$ there are at most $M$ different $l$ 's such that $Q_{m, l} \neq 0$ by $\left(\mathrm{B}_{3}\right)$, and for each $l$ there are at most $M$ different $n$ 's such that $\widetilde{Q}_{l, n} \not \equiv 0$ because $\widetilde{Q}$ is a block matrix with blocks of dimension $\leqslant M$. Hence, there are at most $M^{2}$ many $n$ 's-this proves ( $\mathrm{B}_{3}$ ).
$\left(B_{4}\right)$ follows easily from Lemma 1 if

$$
\begin{align*}
c & >2\left(M^{2}+M\right) b  \tag{15.1}\\
\varepsilon^{b} & <\gamma_{8} \tag{15.2}
\end{align*}
$$

We now turn to condition (C). Let us first define $E^{\prime}(\theta)$ from $E(\theta)$ by analyticity in the perturbation.

Construction of $\Omega_{i}^{\prime}$. The set $\Omega_{i}^{\prime}$ is certainly not unique and we shall make a construction that depends on a choice of $k \in \Lambda_{i}^{\prime}$-we denote by $\Omega(k)$ the particular $\Omega_{j}$ containing $\Lambda(k)$.

Since the eigenvalues of $D$ and $D^{\prime}$ differ by at most $\sqrt{\varepsilon} C \cdot 4 / r$ it follows that for any $l \sim k$,

$$
\left|E^{\prime}(l \omega)-E^{\prime}(n \omega)\right|<\alpha^{\prime} \Rightarrow\left\{\begin{array}{l}
|E(k \omega)-E(n \omega)|<\eta+\alpha^{\prime}+\sqrt{\varepsilon} \cdot 16 / r<2 \eta<\alpha \\
\text { and } \\
|E(l \omega)-E(n \omega)|<\alpha^{\prime}+\sqrt{\varepsilon} \cdot 16 / r<2 \alpha^{\prime}
\end{array}\right.
$$

if (12) holds and if for example $2 b<c<\frac{1}{4}$. Then $n \omega \in m_{n} \omega+\frac{1}{2} I$ and $l \omega \in m_{l} \omega+\frac{1}{2} I$ for some $m_{n}, m_{l} \in \Omega(k)$, and, as in the proof of $\left(\mathrm{B}_{1}\right), n \omega-\left(l-m_{l}+m_{n}\right) \omega$ will belong to the set

$$
\left\{x \in I:\left|u_{\Omega(k)}\left(x, l-m_{l}\right)\right|<2 \alpha^{\prime}(16 / r)^{M^{2}}\right\} .
$$

This set has at most $M_{1}$ (defined in the proof of $\left(\mathrm{B}_{1}\right)$ ) many components, each of length less than

$$
\frac{2}{L}\left(\frac{16}{r}\right)^{1 / s}\left(\frac{4 \alpha^{\prime}}{\beta}\right)^{1 / s M^{2}}
$$

This bound is $<\lambda^{\prime}$ if

$$
\begin{align*}
c & >2 s M^{2} b  \tag{16.1}\\
\varepsilon^{c / 2} & <\gamma_{9} r^{M^{2}} \beta \tag{16.2}
\end{align*}
$$

Hence, we can cover $X:=\bigcup_{l \in \Lambda_{i}^{\prime}}\left\{n \omega:\left|E^{\prime}(l \omega)-E^{\prime}(n \omega)\right|<\alpha^{\prime}\right\}$ by $\leqslant M^{2} M_{1}^{2}$ many intervals-because $\# \Omega(k) \leqslant M$ and $\# \Lambda_{i}^{\prime} \leqslant M M_{1}$-each one of length $\lambda^{\prime}$ and intersecting $\Omega(k) \omega+\frac{1}{2} I$. If $X_{1}, X_{2}, \ldots$ are these intervals, then the first condition of $\left(\mathrm{C}_{1}\right)$ is fulfilled if $\Omega_{i}^{\prime} \omega+\frac{1}{2} I^{\prime} \supset \bigcup X_{j}$. The additional statement even says that $\Omega_{i}^{\prime} \omega+\frac{1}{4} I^{\prime} \supset \bigcup X_{j}$, which will require no extra work.

Let now

$$
\Lambda^{\prime}\left(X_{j}\right)=\left\{n: n+m \in \Lambda^{\prime}(m), \text { for some } m \omega \in X_{j} \text { with } \inf _{l \in \Lambda_{i}^{\prime}}\left|E^{\prime}(l \omega)-E^{\prime}(m \omega)\right|<\alpha^{\prime}\right\}
$$

It is easy to see, as in the proof of $\left(\mathrm{B}_{1}\right)$, that $\# \Lambda^{\prime}\left(X_{j}\right) \leqslant M_{1} M$. In fact, if $n, n^{\prime} \in \Lambda^{\prime}\left(X_{j}\right)$ then $|E(m \omega+n \omega)-E(k \omega)|<2 \eta$ and $\left|E\left(m^{\prime} \omega+n^{\prime} \omega\right)-E(k \omega)\right|<2 \eta$ for some $m \omega, m^{\prime} \omega \in X_{j}$. By (12) it then follows that $(m+n) \omega$ and $\left(m^{\prime}+n^{\prime}\right) \omega$ will belong to a union of at most $M M_{1}$ many intervals, which are of length $<\lambda^{\prime}$ if (16) holds. Now if $n \neq n^{\prime}$, then

$$
\left|(m+n) \omega-\left(m^{\prime}+n^{\prime}\right) \omega\right| \geqslant \frac{\varkappa}{\left(2 M^{\prime} N^{\prime}\right)^{\tau}}-\varepsilon^{b}
$$

which is $>\varepsilon^{b}$ if

$$
\begin{align*}
b & >2 \tau a,  \tag{17.1}\\
\varepsilon^{b} & <\gamma_{10} . \tag{17.2}
\end{align*}
$$

Hence, $(m+n) \omega$ and $\left(m^{\prime}+n^{\prime}\right) \omega$ will belong to different intervals.
Choose now $k_{j}$ such that $X_{j} \subset k_{j} \omega+\frac{1}{4} I^{\prime}$ for each $j$-the distance of $k_{j} \omega$ from $X_{j}$ must not exceed, say, $2^{-4} \varepsilon^{b}$ since we want $\left|I^{\prime}\right|$ to be less than $\varepsilon^{b}$-we don't require that $k_{j} \omega \in X_{j}$ but if it is then we can choose $\left|I^{\prime}\right|=8 \lambda^{\prime}$.

If $n \in \Lambda^{\prime}\left(X_{j}\right)$ then there is some $m \omega \in X_{j}$ such that $m+n \sim^{\prime} m$ and

$$
\left|E^{\prime}(l \omega)-E^{\prime}(m \omega)\right|<\alpha^{\prime} \quad \text { for some } l \in \Lambda_{i}^{\prime} .
$$

Then

$$
\begin{aligned}
|E(m \omega+n \omega)-E(k \omega)| \leqslant \mid & |E(m \omega+n \omega)-E(m \omega)|+\left|E(m \omega)-E^{\prime}(m \omega)\right| \\
& +\left|E^{\prime}(m \omega)-E^{\prime}(l \omega)\right|+\left|E^{\prime}(l \omega)-E(l \omega)\right| \\
& +|E(l \omega)-E(k \omega)| \leqslant \eta+\frac{8}{r} \sqrt{\varepsilon}+\alpha^{\prime}+\frac{8}{r} \sqrt{\varepsilon}+\eta,
\end{aligned}
$$

which is less than $\alpha$ under condition (12). By ( $\mathrm{C}_{1}$ ) it follows that

$$
(m+n) \omega \in p \omega+\frac{1}{2} I \quad \text { for some } p \in \Omega(k) .
$$

Since $\left\|\left(k_{j}-m\right) \omega\right\|<\varepsilon^{b}$ and $\lambda \leqslant|I|$ it follows that $\left(k_{j}+n\right) \omega \in p \omega+\frac{3}{4} I$ if

$$
\begin{equation*}
\varepsilon^{b}<\gamma_{11} \lambda . \tag{18}
\end{equation*}
$$

Hence, if we define $\left[k_{j}+n\right]_{\Omega(k)}$ to be this integer $p-$ notice that it is unique by $\left(\mathrm{C}_{4}\right)$ condition ( $\mathrm{C}_{1}$ ) implies that

$$
Q(\theta)\left(\mathbf{R}^{\Lambda(m+n)}\right) \subset \mathbf{R}^{\Omega(k)+m+n-\left[k_{j}+n\right]_{\Omega(k)}} \quad \forall \theta \in I .
$$

By the construction of $Q^{\prime}=Q \widetilde{Q}$ we have

$$
\begin{aligned}
Q^{\prime}(\theta)\left(\mathbf{R}^{\Lambda^{\prime}(m)}\right) & \subset \sum_{m+n \sim^{\prime} m} Q^{\prime}(\theta)\left(\mathbf{R}^{\Lambda(m+n)}\right) \subset \sum_{m+n \sim^{\prime} m} Q(\theta)\left(\mathbf{R}^{\Lambda(m+n)}\right) \\
& \subset \sum_{m+n \sim^{\prime} m} \mathbf{R}^{\Omega(k)+m+n-\left[k_{j}+n\right]_{\Omega(k)}} \\
& \subset \sum_{m+n \sim^{\prime} m} \mathbf{R}^{\Omega(k)+k_{j}+n-\left[k_{j}+n\right]_{\Omega(k)}+\left(m-k_{j}\right)} \\
& \subset \mathbf{R}^{\Omega_{i}^{\prime}+\left(m-k_{j}\right)},
\end{aligned}
$$

if

$$
\Omega_{i}^{\prime} \supset \bigcup_{n \in \Lambda^{\prime}\left(X_{j}\right)}\left[\Omega(k)+k_{j}+n-\left[k_{j}+n\right]_{\Omega(k)}\right] .
$$

So if we choose

$$
\Omega_{i}^{\prime}=\Omega(k)+\bigcup_{j}\left\{k_{j}+n-\left[k_{j}+n\right]_{\Omega(k)}: n \in \Lambda^{\prime}\left(X_{j}\right)\right\}=\Omega(k)+\Delta(k)
$$

then condition $\left(\mathrm{C}_{1}\right)$ will hold.
Notice again that $\Delta(k) \omega \subset \frac{3}{4} I$-this will be used in the verification of $\left(\mathrm{C}_{6}\right)$.
Since

$$
\# \Omega_{i}^{\prime} \leqslant \# \Omega(k) \cdot \# \Delta(k) \leqslant M \cdot M^{2} M_{1}^{2} \cdot M M_{1}<M^{\prime}
$$

also $\left(\mathrm{C}_{3}\right)$ will hold. (Here we used that $M \geqslant 8$.)
For $l_{1} \neq l_{2} \in \Lambda_{i}^{\prime}$,

$$
\left\|\left(l_{1}-l_{2}\right) \omega\right\| \geqslant \frac{\varkappa}{\left|l_{1}-l_{2}\right|^{\tau}} \geqslant \frac{\varkappa}{\left(2 M^{\prime} N^{\prime}\right)^{\tau}}
$$

which is $>\varepsilon^{b}$ if for example (17) holds. Hence, each $l \omega, l \in \Lambda_{i}^{\prime}$, belongs to one and only one $X_{j}$. We can therefore include among the $k_{j}$ 's all elements of $\Lambda_{i}^{\prime}$ assuring that

$$
\Omega_{i}^{\prime} \supset \Lambda_{i}^{\prime}
$$

Verification of $\left(\mathrm{C}_{5}\right)$. In order to verify $\left(\mathrm{C}_{5}\right)$ we must restrict the choice of the $k_{j}$ 's. If $X_{j} \cap \Lambda_{i}^{\prime} \omega=\{l \omega\}$, then we are forced to take $k_{j}=l$. If $X_{j} \cap \Lambda_{i}^{\prime} \omega=\varnothing$, then we have some freedom in the choice and we decide to choose $k_{j} \omega \in X_{j}$ in such a way that $k_{j}$ is as close as possible to $\Lambda_{i}^{\prime}$. There are one or two such choices. (Indeed, suppose that there are at least two such choices, $k_{1}$ and $k_{2}$ say. Then

$$
\varepsilon^{b} \geqslant\left\|\left(k_{1}-k_{2}\right) \omega\right\| \geqslant \frac{\varkappa}{\left|k_{1}-k_{2}\right|^{\tau}},
$$

so by (17),

$$
\left|k_{1}-k_{2}\right| \geqslant\left(\frac{\varkappa}{\varepsilon^{b}}\right)^{1 / \tau}>M^{\prime} \frac{1}{\varepsilon^{a}}=\operatorname{diam}\left(\Lambda_{i}^{\prime}\right)
$$

This means that $\Lambda_{i}^{\prime}$ must lie in between $k_{1}$ and $k_{2}$, which excludes a third choice.) If there are two, choose any one of them. It is an easy exercise, using the Diophantine property of $\omega$ to show that for any interval $I^{\prime}$ of length $\geqslant \lambda^{\prime}$ and any sequence of integers $l, l+1, \ldots, l+n$ of length

$$
\geqslant \frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{\lambda^{\prime}}\right)^{\tau+1}
$$

there is (at least) one integer $k$ in this sequence such that $k \omega \in I^{\prime}$. Hence, we can choose the $k_{j}$ 's so that

$$
\operatorname{dist}\left(k_{j}, \Lambda_{i}^{\prime}\right) \leqslant \frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{\lambda^{\prime}}\right)^{\tau+1}
$$

Then

$$
\operatorname{diam}\left(\Omega_{i}^{\prime}\right) \leqslant \operatorname{diam}\left(\Lambda_{i}^{\prime}\right)+2 \frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{\lambda^{\prime}}\right)^{\tau+1}+2 M^{\prime} N^{\prime}+2 \operatorname{diam}(\Omega(k))<\left(\frac{1}{\lambda^{\prime}}\right)^{\tau+2}
$$

under conditions (17)-(18).
Verification of $\left(\mathrm{C}_{4}\right)$. This is somewhat delicate and we will have to change the choice of the $k_{j}$ 's-in order to guarantee that $\Omega_{i}^{\prime} \supset \Lambda_{i}^{\prime}$ we are not allowed to change those $k_{j}$ 's that belong to $\Lambda_{i}^{\prime}$-in order to cover $X_{j}$ by $k_{j} \omega+\frac{1}{4} I^{\prime}$ with $\left|I^{\prime}\right| \leqslant \varepsilon^{b}$, we must keep $k_{j} \omega$ at distance less than $2^{-4} \varepsilon^{b}$ from $X_{j}$-in order to keep the bound ( $\mathrm{C}_{5}$ ) we cannot increase the $k_{j}$ 's too much.

Suppose that the intervals $\left\{m \omega+I^{\prime}:\left|m-k_{j}\right| \leqslant 2 M^{\prime} N^{\prime}, j \geqslant 1\right\}$ are not pairwise disjoint. Then there are two $k_{j}$ 's, $k_{1}$ and $k_{2}$ say, such that $\left(k_{1}+n_{1}\right) \omega+I^{\prime}$ and $\left(k_{2}+n_{2}\right) \omega+I^{\prime}$ intersect for some $\left|n_{1}\right|,\left|n_{2}\right| \leqslant 2 M^{\prime} N^{\prime}$, i.e.

$$
\left\|k_{1} \omega-\left(k_{2}+n\right) \omega\right\|<\left|I^{\prime}\right|, \quad|n|=\left|n_{2}-n_{1}\right| \leqslant 4 M^{\prime} N^{\prime}
$$

This implies that

$$
\left|k_{1}-k_{2}\right| \geqslant\left(\frac{\varkappa}{\varepsilon^{b}}\right)^{1 / \tau}-4 M^{\prime} N^{\prime}
$$

which is $>2 M^{\prime} N^{\prime}$ under condition (17). In particular, both $k_{1}$ and $k_{2}$ cannot belong to $\Lambda_{i}^{\prime}$-if one does, let it be $k_{2}$ say. Moreover,

$$
\left(k_{2}+n\right) \omega+\frac{1}{4} \cdot 9 I^{\prime} \supset k_{1} \omega+\frac{1}{4} I^{\prime} \supset X_{1}
$$

and for all $|m|<4 M^{\prime} N^{\prime}, m \neq n$,

$$
\left\|k_{1} \omega-\left(k_{2}+m\right) \omega\right\| \geqslant\|(m-n) \omega\|-\lambda^{\prime} \geqslant \frac{\varkappa}{\left(8 M^{\prime} N^{\prime}\right)^{\tau}}-\lambda^{\prime} \geqslant \frac{1}{2} \cdot \frac{\varkappa}{\left(8 M^{\prime}\right)^{\tau}} \varepsilon^{a \tau} \gg \varepsilon^{b}
$$

under (17).
Hence, if we replace $k_{1}$ by $k_{2}+n$-still denoted $k_{1}$--and increase $I^{\prime}$ to $9 I^{\prime}$-still denoted $I^{\prime}$-we can assume that

$$
\left\{m \omega+I^{\prime}:\left|m-k_{j}\right| \leqslant 2 M^{\prime} N^{\prime}, j=1,2\right\} \quad \text { are pairwise disjoint, }
$$

and

$$
k_{1} \omega+\frac{1}{4} I^{\prime} \cup k_{2} \omega+\frac{1}{4} I^{\prime} \supset X_{1} \cup X_{2}
$$

It is now sufficient that the intervals $\left\{m \omega+I^{\prime}:\left|m-k_{j}\right| \leqslant 4 M^{\prime} N^{\prime}, j \geqslant 2\right\}$ are pairwise disjoint. If they are not, then there are two $k_{j}$ 's, $k_{2}$ and $k_{3}$ say, such that $\left(k_{2}+n_{2}\right) \omega+I^{\prime}$ and $\left(k_{3}+n_{3}\right) \omega+I^{\prime}$ intersect for some $\left|n_{2}\right|,\left|n_{3}\right| \leqslant 4 M^{\prime} N^{\prime}$, i.e. $\left\|k_{1} \omega-\left(k_{2}+n\right) \omega\right\|<\left|I^{\prime}\right|$ for $|n|=\left|n_{3}-n_{2}\right| \leqslant 8 M^{\prime} N^{\prime}$. Now we proceed inductively. Hence, by changing those $k_{j}$ 's that do not belong to $\Lambda_{i}^{\prime}$, and increasing the size of $I^{\prime}$, we can achieve that

$$
\left\{m \omega+I^{\prime}\right\}_{\left|m-k_{j}\right| \leqslant 2 M^{\prime} N^{\prime}, j \geqslant 1} \quad \text { are pairwise disjoint, }
$$

and

$$
\bigcup_{j \geqslant 1} k_{j} \omega+\frac{1}{4} I^{\prime} \supset \bigcup_{j \geqslant 1} X_{j}
$$

Since the number of $k_{j}$ 's is less than $M^{2} M_{1}^{2} \leqslant M^{\prime}-1$, the $k_{j}$ 's increase by at most

$$
4 M^{\prime} N^{\prime}+2 \cdot 4 M^{\prime} N^{\prime}+2^{2} \cdot 4 M^{\prime} N^{\prime}+\ldots+2^{M^{\prime}-1} \cdot 4 M^{\prime} N^{\prime} \leqslant 2^{M^{\prime}+2} \cdot M^{\prime} N^{\prime}
$$

and therefore

$$
\operatorname{dist}\left(k_{j}, \Lambda_{i}^{\prime}\right) \leqslant 2 \frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{\lambda^{\prime}}\right)^{\tau+1}
$$

under condition (17)—hence, $\left(\mathrm{C}_{5}\right)$ still holds.
Since we can start the induction by an interval $I^{\prime}$ of size $8 \lambda^{\prime}$ and since the interval increases at most $9^{M^{\prime}-1}$ times, it will remain smaller than $\varepsilon^{b}$.

Finally, if $\operatorname{dist}\left(n, \Omega_{i}^{\prime}\right)<N^{\prime}$ it follows that

$$
\operatorname{dist}\left(n,\left\{k_{j}: j \geqslant 1\right\}\right)<N^{\prime}+M^{\prime} N^{\prime}+\operatorname{diam}(\Omega(k)) \leqslant 2 M^{\prime} N^{\prime}
$$

if

$$
\begin{equation*}
\varepsilon^{a}<\gamma_{12} \lambda \tag{19}
\end{equation*}
$$

This proves $\left(\mathrm{C}_{4}\right)$.
Verification of $\left(\mathrm{C}_{2}\right)$. Let us denote $\Omega(k)$ and $\Delta(k)$ by $\Omega_{i}$ and $\Delta_{i}$ respectively, so that $\Omega_{i}^{\prime}=\Omega_{i}+\Delta_{i}$. The construction of $\Omega_{i}$ together with condition $\left(\mathrm{C}_{4}\right)$ imply that $\Omega_{i}+k$ and $\Omega_{i}+l$ are separated by at least $N$ sites, unless $k=l$. Hence, by condition (A),

$$
D(\theta)_{\Omega_{i}^{\prime}}=\prod_{k \in \Delta_{i}} D(\theta)_{\Omega_{i}+k}
$$

for all $\theta$. Hence,

$$
\begin{aligned}
u_{\Omega_{i}^{\prime}}(x, \theta) & =\operatorname{Res}\left(\operatorname{det}\left(D(\theta+x)_{\Omega_{i}^{\prime}}-t I\right), \operatorname{det}\left(D(\theta)_{\Omega_{i}^{\prime}}-t I\right)\right) \\
& =\prod_{k, l \in \Delta_{i}} \operatorname{Res}\left(\operatorname{det}\left(D(\theta+x)_{\Omega_{i}+k}-t I\right), \operatorname{det}\left(D(\theta)_{\Omega_{i}+l}-t I\right)\right) \\
& =\prod_{k, l \in \Delta_{i}} u_{\Omega_{i}}(x+(k-l) \omega, \theta+l \omega)
\end{aligned}
$$

By Remark 4 we have, for $\theta$ fixed, that

$$
\left|u_{\Omega_{i}}\right| \mathcal{C}^{\nu}<(4 C M)^{2 M^{2}}\left(L^{\prime}\right)^{\nu} \quad \forall \nu \geqslant 0
$$

and applying Lemma 4 gives

$$
\max _{0 \leqslant \nu \leqslant s\left(M^{\prime}\right)^{2}}\left|\frac{1}{(\nu!)^{2}\left(L^{\prime}\right)^{\nu}} \partial_{x}^{\nu} u_{\Omega_{i}^{\prime}}(x, \theta)\right| \geqslant 2 \beta^{\prime},
$$

since

$$
\begin{equation*}
\varepsilon^{b}<\gamma_{13} \frac{1}{K}\left(\beta\left(\frac{L}{K}\right)^{s M^{2}}\right)^{\left(M^{\prime} / M\right)^{2\left(s\left(M^{\prime}\right)^{2}+1\right)}} \tag{20}
\end{equation*}
$$

What we want is an estimate for $u_{\Omega_{i}^{\prime}}^{\prime}$, which is defined as $u_{\Omega_{i}^{\prime}}$ but with the matrix $D^{\prime}$ instead of $D$. For this we only observe that, for $\theta$ fixed,

$$
\left|u_{\Omega_{i}^{\prime}}^{\prime}-u_{\Omega_{i}^{\prime}}\right|_{C^{\nu}}<\left(4 C M^{\prime}\right)^{2\left(M^{\prime}\right)^{2}} \sqrt{\varepsilon}\left(K^{\prime}\right)^{\nu}
$$

which is $<\varepsilon^{b} K^{\nu}$ for all $\nu \leqslant s\left(M^{\prime}\right)^{2}+1$, if

$$
\begin{gather*}
\frac{1}{4}>b+\left(s\left(M^{\prime}\right)^{2}+1\right) c  \tag{21.1}\\
\varepsilon^{1 / 4}<\gamma_{14} \tag{21.2}
\end{gather*}
$$

This implies the second part of condition $\left(\mathrm{C}_{2}\right)$.
Using the estimate of $\left|u_{\Omega_{i}}\right|_{\mathcal{C}^{\nu}}$ obtained from Remark 4, Lemma 4 gives an estimate of $\left|u_{\Omega_{i}^{\prime}}\right|_{\mathcal{C}^{\nu}}$. The estimate of $\left|u_{\Omega_{i}^{\prime}}^{\prime}-u_{\Omega_{i}^{\prime}}\right|_{\mathcal{C}^{\nu}}$ above now gives the first part of $\left(\mathrm{C}_{2}\right)$ if for example (21.1) holds.

Verification of $\left(\mathrm{C}_{6}\right)$. The first part follows exactly as in $\left(\mathrm{C}_{2}\right)$. As for the second part we let

$$
v_{\Omega_{i}}(y, \theta)=u_{\Omega_{i}}(y, \theta)\left(\prod_{m, n \in \Omega_{i}}\|y+(m-n) \omega\|\right)^{-1}
$$

If we notice that $u_{\Omega_{i}}(y, \theta) \in \mathbf{O}\left(|y|^{\# \Omega_{i}}\right)$ then we have, for $|y|<\frac{4}{5}|I|$ fixed, that

$$
\begin{aligned}
\left|v_{\Omega_{i}}\right|_{C^{\nu}} & <(8 M)^{2 M^{2}}\left(\frac{5}{\lambda}\right)^{M^{2}}\left(\frac{(\nu+M)!}{\nu!}\right)^{2}\left(L^{\prime}\right)^{M}\left(L^{\prime}\right)^{\nu} \\
& <(8 M)^{2 M^{2}}\left(\frac{5}{\lambda}\right)^{M^{2}}\left(s M^{\prime}\right)^{4 M}\left(L^{\prime}\right)^{M}\left(L^{\prime}\right)^{\nu} \quad \forall \nu \leqslant s\left(M^{\prime}\right)^{2}+1
\end{aligned}
$$

where we have used $\left(\mathrm{C}_{4}\right)$ to estimate $\|y-(m-n) \omega\|$ from below by $\frac{1}{5}|I| \geqslant \frac{1}{5} \lambda$ when $m \neq n$.

Let now

$$
v_{\Omega_{i}^{\prime}}(x, \theta)=\prod_{k, l \in \Delta_{i}} v_{\Omega_{i}}(x+(k-l) \omega, \theta+l \omega)
$$

and observe that if $x \in I^{\prime}$, then

$$
|x+(k-l) \omega|<\left|I^{\prime}\right|+\frac{3}{4}|I|<\frac{4}{5}|I|
$$

if (18) holds-recall that $\Delta_{i} \omega \subset \frac{3}{4} I$. By Lemma 4 we get that

$$
\max _{0 \leqslant \nu \leqslant s\left(M^{\prime}\right)^{2}}\left|\frac{1}{(\nu!)^{2}\left(L^{\prime}\right)^{\nu}} \partial_{\theta}^{\nu} v_{\Omega_{i}^{\prime}}(x, \theta)\right| \geqslant 2 \beta^{\prime}
$$

since

$$
\varepsilon^{b}<\gamma_{15} \frac{1}{K}\left(\beta \frac{\lambda^{2 M^{2}}}{K^{2 M}}\left(\frac{L}{K}\right)^{s M^{2}}\right)^{\left(M^{\prime} / M\right)^{2\left(s\left(M^{\prime}\right)^{2}+1\right)}}
$$

Hence

$$
\max _{0 \leqslant \nu \leqslant s\left(M^{\prime}\right)^{2}}\left|\frac{1}{(\nu!)^{2}\left(L^{\prime}\right)^{\nu}} \partial_{\theta}^{\nu} u_{\Omega_{i}^{\prime}}(x, \theta)\right| \geqslant 2 \beta^{\prime}\left(\prod_{k, l \in \Omega_{i}^{\prime}}\|x+(k-l) \omega\|\right)
$$

If we notice that $u_{\Omega_{i}^{\prime}}^{\prime}(x, \theta) \in \mathbf{O}\left(|x|^{\# \Omega_{i}^{\prime}}\right)$, then we have, for $x$ fixed, that

$$
\left|u_{\Omega_{i}^{\prime}}^{\prime}-u_{\Omega_{i}^{\prime}}\right| \mathcal{C}^{\nu}<\left(4 C M^{\prime}\right)^{2\left(M^{\prime}\right)^{2}}\left(\frac{\left(\nu+M^{\prime}\right)!}{\nu!}\right)^{2}\left(K^{\prime}\right)^{M^{\prime}}\|x\|^{\# \Omega_{i}^{\prime}} \sqrt{\varepsilon}\left(K^{\prime}\right)^{\nu}
$$

for all $\nu \geqslant 0$. Now this is bounded by

$$
\left(4 C M^{\prime}\right)^{2\left(M^{\prime}\right)^{2}}\left(s M^{\prime}\right)^{4 M^{\prime}}\left(K^{\prime}\right)^{M^{\prime}}\|x\|^{\# \Omega_{i}^{\prime}} \sqrt{\varepsilon}\left(K^{\prime}\right)^{\nu} \quad \forall \nu \leqslant s\left(M^{\prime}\right)^{2}+1
$$

and since

$$
\min _{k \neq l \in \Omega_{i}^{\prime}}\|(k-l) \omega\| \geqslant\left|I^{\prime}\right| \geqslant\left(\frac{1}{9}\right)^{M^{\prime}} \varepsilon^{b}
$$

this is less than

$$
\left(\prod_{k, l \in \Omega_{i}^{\prime}}\|x+(k-l) \omega\|\right) \varepsilon^{b} K^{\nu} \quad \forall \nu \leqslant s\left(M^{\prime}\right)^{2}+1
$$

if

$$
\begin{gather*}
\frac{1}{4}>b\left(\left(M^{\prime}\right)^{2}+1\right)+\left(s\left(M^{\prime}\right)^{2}+1+M^{\prime}\right) c \\
\varepsilon^{1 / 4}<\gamma_{16}(1 / K)^{M^{\prime}}
\end{gather*}
$$

This gives the second part of $\left(\mathrm{C}_{6}\right)$.

The additional statement. This is just a formulation of the construction made in the proof.

The final statement. This is the only part where we use condition $\left(\mathrm{C}_{6}\right)$. Consider

$$
\left\{\begin{array}{l}
\left\{n \omega:|E(n \omega)-E(n \omega+l \omega)|<2 \eta=2 M^{\prime}(K / r) \varepsilon^{b}\right\} \\
4(1 / \lambda)^{\tau+2}<|l| \leqslant M^{\prime} N^{\prime}
\end{array}\right.
$$

Divide $\mathbf{T}$ into intervals of length $r \alpha / 16 K$ and choose a $k \omega$ in each interval. For any $n$ the minimum of $|E(n \omega)-E(k \omega)|$ over all such $k$ is $<\frac{1}{2} \alpha$. If now

$$
|E(n \omega)-E(n \omega+l \omega)|<2 \eta
$$

then by $\left(\mathrm{C}_{1}\right), n \omega \in m \omega+\frac{1}{2} I$ and $(n+l) \omega \in m^{\prime} \omega+\frac{1}{2} I$ for some $m, m^{\prime} \in \Omega(k)$. Hence, $\theta=$ $(n-m) \omega \in I$ belongs to the set

$$
Z:=\left\{\theta \in I:\left|u_{\Omega(k)}(x, \theta)\right|<2 \eta(16 / r)^{M^{2}}\right\}
$$

where we have put $x=\left(l-m^{\prime}+m\right) \omega$.
For any $m_{1}, m_{2} \in \Omega(k)$ we get by $\left(\mathrm{C}_{5}\right)$ that

$$
0<\left|l-m^{\prime}+m+\left(m_{1}-m_{2}\right)\right|<M^{\prime} N^{\prime}+4(1 / \lambda)^{\tau+2}
$$

which is less than $2 M^{\prime} N^{\prime}$ under (19). Hence, we get that

$$
\left\|x+\left(m_{1}-m_{2}\right) \omega\right\| \geqslant \frac{\varkappa}{\left(2 M^{\prime} N^{\prime}\right)^{\tau}}
$$

Using now $\left(\mathrm{C}_{6}\right)$ together with Lemma 3 we find that $Z$ has at most

$$
M^{\prime}\left(\frac{L}{\beta}|I|\left(\frac{\left(2 M^{\prime} N^{\prime}\right)^{\tau}}{\varkappa}\right)^{M^{2}}+1\right)
$$

many components, each of length less than

$$
\frac{2}{L}\left(\frac{16}{r}\right)^{1 / s}\left(\frac{2 \eta}{\beta}\right)^{1 / s M^{2}}\left(\frac{\left(2 M^{\prime} N^{\prime}\right)^{\tau}}{\varkappa}\right)^{1 / s}
$$

which is $<\varepsilon^{b / 4 s M^{2}}$ under condition (13).
Since $\# \Omega(k) \leqslant M$, since the number of $l$ 's is less than $2 M^{\prime} N^{\prime}$ and the number of $k$ 's is less than const $\cdot K / r \alpha$, we get the result if

$$
\begin{align*}
b & >20 s \tau M^{4} a  \tag{22.1}\\
\varepsilon^{b} & <\gamma_{17}(r \alpha / K)^{10 s M^{2}} \tag{22.2}
\end{align*}
$$

The estimate of the second set is completely analogous-it is here we use that $M \geqslant 2 \tau$.

## 6. Proof of theorem

Proposition 8. Let $D$ be in normal form on an interval I with parameters $C, K, r, M$, $N, \alpha, L, \beta, \lambda$, satisfying the assumptions

$$
1<L \leqslant K, \quad \max (8,2 \tau) \leqslant M, \quad \frac{4}{3}<C<\frac{5}{3}, \quad r, \alpha, \beta \leqslant 1
$$

and $r N=1$.
Let $F$ be a smooth symmetric matrix satisfying the shift condition. Assume that

$$
\begin{gathered}
\lambda \leqslant|I| \leqslant \beta / L, \\
\left\{F_{m, n} \mid \mathcal{C}^{\nu}<\varepsilon e^{-|m-n| r} K^{\nu} \quad \forall \nu \geqslant 0 .\right.
\end{gathered}
$$

Then there exists a constant $\Gamma^{\prime}=\Gamma^{\prime}\left(\varkappa, \tau, s, \gamma_{0}, M\right)$, superexponentially decaying in $M$, such that if

$$
|\varepsilon|<\Gamma^{\prime}\left(\frac{\alpha \beta r^{\tau} \lambda^{\tau^{2}}}{K N^{\tau}} e^{-N r}\right)^{e^{e^{s M^{4}}}}
$$

then there is a smooth orthogonal matrix $U$, satisfying the shift condition and such that

$$
U^{*}(D+F) U=D_{\infty}
$$

where $D_{\infty}(\theta)$ is a norm limit of matrices with pure point spectrum. Moreover, $D_{\infty}(\theta)$ is pure point with finite-dimensional eigenvectors for a.e. $\theta$, and the measure of $\sigma(D) \backslash \sigma(D+F)$ goes to 0 as $\varepsilon \rightarrow 0$.

Proof. Let us choose

$$
\begin{aligned}
C_{j+1} & =\left(1+\varepsilon_{j}^{1 / 2}\right) C_{j}, & K_{j+1}=\left(1 / \varepsilon_{j}\right)^{c_{j}} K_{j}, & r_{j+1}=\varepsilon_{j}^{a_{j}}, \\
\lambda_{j+1} & =\left(\frac{1}{9}\right)^{M_{j+1}} \varepsilon_{j}^{b_{j}}, & M_{j+1}=M_{j}^{s M_{j}^{3}}, & N_{j+1}=\left(1 / \varepsilon_{j}\right)^{a_{j}}, \\
\alpha_{j+1} & =\varepsilon_{j}^{c_{j}}, & L_{j+1}=K_{j}, & \beta_{j+1}=\varepsilon_{j}^{b_{j}} K_{j}, \\
\varepsilon_{j+1} & =\varepsilon_{j}^{\frac{1}{2}\left(\frac{1}{\left.\varepsilon_{j}\right)^{a_{j} / 2}}\right.} & &
\end{aligned}
$$

where we let $C_{1}=C, K_{1}=K, \ldots$ be the parameters of the normal form $D_{1}=D$, where $\varepsilon_{1}=\varepsilon$ and

$$
a_{j}=\frac{1}{\tau}\left(\frac{1}{M_{j}}\right)^{3 s M_{j}^{3}}, \quad b_{j}=20 s \tau M_{j}^{4} a_{j}, \quad c_{j}=\frac{1}{5 s}\left(\frac{1}{M_{j}}\right)^{2 s M_{j}^{3}} .
$$

We choose

$$
\frac{1}{\Gamma^{\prime}(M)}=\exp \circ \ldots \circ \exp \left(A M^{4}\right), \quad A=A\left(\varkappa, \tau, s, \gamma_{0}\right)
$$

in such a way that for all $j \geqslant 0$,

$$
\begin{aligned}
\Gamma^{\prime}\left(M_{j+1}\right) & \geqslant(1 / e)^{\left(1 / \Gamma^{\prime}\left(M_{j}\right)\right)^{a_{j} / 2}} \\
\Gamma^{\prime}\left(M_{j}\right)^{a_{j} / 2} & \leqslant(1 / e)^{e^{s M_{j+1}^{4}}} \\
\Gamma^{\prime}\left(M_{j}\right) & \leqslant \Gamma\left(M_{j}\right)
\end{aligned}
$$

where $\Gamma$ is the constant in Lemma 7.
We shall show that for all $j \geqslant 1$,

$$
\varepsilon_{j}<\Gamma^{\prime}\left(M_{j}\right)\left(\frac{\alpha_{j} \beta_{j} r_{j}^{\tau} \lambda_{j}^{\tau^{2}}}{K_{j} N_{j}^{\tau}} e^{-N_{j} r_{j}}\right)^{e^{s M_{j}^{4}}}
$$

This implies that

$$
\varepsilon_{j}<\Gamma^{\prime}\left(M_{j}\right)\left(\left(\frac{1}{9}\right)^{M_{j} \tau^{2}} e^{-1}\right)^{e^{e^{M_{j}^{4}}}}
$$

so $\varepsilon_{j}$ decays very rapidly.
Let us assume that the inequality holds for $j$ and then prove it for $j+1$. If it holds for $j$ then

$$
\begin{gathered}
\Gamma^{\prime}\left(M_{j+1}\right)\left(\frac{\alpha_{j+1} \beta_{j+1} r_{j+1}^{\tau} \lambda_{j+1}^{\tau^{2}}}{K_{j+1} N_{j+1}^{\tau}} e^{-1}\right)^{e^{e M_{j+1}^{4}}} \\
\geqslant\left(\frac{1}{e}\right)^{\left(1 / \Gamma^{\prime}\left(M_{j}\right)\right)^{a_{j} / 2}}\left(\varepsilon_{j}^{c_{j}+b_{j}+a_{j} \tau+\tau^{2} b_{j}+c_{j}+\tau a_{j}}\left(\frac{1}{9}\right)^{M_{j+1} \tau^{2}} e^{-1}\right)^{e^{s M_{j+1}^{4}}} \\
\geqslant \varepsilon_{j}^{\frac{1}{4}\left(\frac{1}{\varepsilon_{j}}\right)^{a_{j} / 2}}\left(\varepsilon_{j}^{1 / 8}\left(\frac{1}{9}\right)^{M_{j+1} \tau^{2}}\right)^{e^{e^{s M_{j+1}^{4}}} \geqslant \varepsilon_{j}^{\frac{1}{4}\left(\frac{1}{\varepsilon_{j}}\right)^{a_{j} / 2}} \varepsilon_{j}^{\frac{1}{4} e^{s M_{j+1}^{4}}}},
\end{gathered}
$$

which is greater than $\varepsilon_{j+1}$. Hence, the basic smallness assumption of Lemma 7 is fulfilled-all other assumptions of that lemma are immediate.

By Lemma 7 there is for each $j \geqslant 1$ an orthogonal matrix $U_{j}$ satisfying the shift condition and such that

$$
\left|\left(U_{j}-I\right)_{m, n}\right|_{\mathcal{C}^{\nu}}<\sqrt{\varepsilon_{j}} e^{-|m-n| r_{j+1}} K_{j+1}^{\nu}
$$

and

$$
\left(U_{1} \cdot \ldots \cdot U_{j}\right)^{*}(D+F)\left(U_{1} \cdot \ldots \cdot U_{j}\right)=D_{j+1}+F_{j+1}
$$

where $D_{j+1}$ is in normal form with parameters $C_{j+1}, K_{j+1}, r_{j+1}, M_{j+1}, N_{j+1}, \alpha_{j+1}, L_{j+1}$, $\beta_{j+1}, \lambda_{j+1}$, and where

$$
\left|\left(F_{j+1}\right)_{m, n}\right|_{C^{\nu}}<\varepsilon_{j+1} e^{-|m-n| r_{j+1}} K_{j+1}^{\nu}
$$

Hence

$$
U_{1} \cdot \ldots \cdot U_{j} \rightarrow U, \quad F_{j} \rightarrow 0 \quad \text { and } \quad D_{j} \rightarrow D_{\infty}
$$

in norm, and it remains to check that $D_{\infty}$ is pure point. For this we need the additional and final statements of Lemma 7.

Clearly there is a uniform limit

$$
E^{j}(\theta) \rightarrow E^{\infty}(\theta)
$$

which describes the spectrum of $D_{\infty}(\theta)$-it is the closure of the image of $E^{\infty}$. Consider now the closure $Z^{j}$ of the set of all $\theta$ such that

$$
\left|E^{\infty}(\theta)-E^{\infty}(\theta+l \omega)\right|<\frac{3}{2} \eta_{j} \quad \text { for some } 4\left(1 / \lambda_{j}\right)^{\tau+2}<|l| \leqslant M_{j+1} N_{j+1}
$$

or

$$
\left.\mid E^{\infty}(\theta)-E^{\infty}(\theta+l \omega)\right) \left\lvert\,<\frac{3}{2} \varepsilon_{j}^{1 / 8} \quad\right. \text { for some } M_{j+1} N_{j+1}<|l| \leqslant 4\left(1 / \lambda_{j+1}\right)^{\tau+2}
$$

where $\eta_{j}=M_{j+1}\left(K_{j} / r_{j}\right) \varepsilon_{j}^{b_{j}}$. According to the final statement of Lemma 7 this set is of measure less than

$$
\text { const } \cdot \varepsilon_{j}^{b_{j} / 20 s M_{j}^{2}}
$$

From this we conclude that for a.e. $\theta$ each $\theta+k \omega$ will belong to only finitely many $Z^{j}$ 's. Suppose $\theta=0$ is of this sort, i.e. for all $k$ there is a $j_{0}(k)$ such that $k \omega \notin Z^{j}$ for $j \geqslant j_{0}(k)$. Hence, for such $j$ 's,

$$
\left|E^{j}(k \omega)-E^{j}(k \omega+l \omega)\right| \geqslant 2 \eta_{j} \quad \text { for all } 4\left(1 / \lambda_{j}\right)^{\tau+2}<|l| \leqslant M_{j+} N_{j+1}
$$

and

$$
\left|E^{j}(k \omega)-E^{j}(k \omega+l \omega)\right| \geqslant 2 \varepsilon_{j}^{1 / 8} \quad \text { for all } M_{j+1} N_{j+1}<|l| \leqslant 4\left(1 / \lambda_{j+1}\right)^{\tau+2}
$$

This implies that $\Lambda^{j}(k) \subset\left[k-4\left(1 / \lambda_{j 0}(k)\right)^{\tau+2}, k+4\left(1 / \lambda_{j_{0}(k)}\right)^{\tau+2}\right]$ for all $j \geqslant j_{0}(k)$. The blocks $\Lambda^{j}(k)$ therefore become eventually stationary:

$$
\Lambda^{j+1}(k)=\Lambda^{j}(k) \quad \forall j \geqslant j_{1}(k)
$$

and $D(0)$ is pure point. (Here we used the first two additional statements of Lemma 7.)
The same argument works for an arbitrary $\theta$ and shows that $D(\theta)$ is pure point for a.e. $\theta$. One only needs to convince oneself that when we shift base point on T from 0 to
$\theta_{0}$ then $Z_{\theta_{0}}^{j}=Z^{j}-\theta_{0}$. In order to see this, observe first that if $D$ is in normal form on an interval $I$ with orthogonal matrix $Q$, block decomposition $\mathbf{Z}=\bigcup \Lambda_{i}$, spectral function $E$ and sets $\left\{\Omega_{i}\right\}$, then $\widetilde{D}(\theta)=D(\theta+k \omega)$ will be in normal form with corresponding objects

$$
I, \quad T_{k}^{-1} Q T_{k}, \quad \bigcup\left(\Lambda_{i}-k\right), \quad \widetilde{E}(\theta)=E(\theta+k \omega), \quad\left\{\Omega_{i}-k\right\}
$$

where $T_{k}$ is the shift by $k$ sites. (But $\widetilde{D}(\theta)=D\left(\theta+\theta_{0}\right)$ may not be in normal form for arbitrary $\theta_{0}$.)

The $D^{\prime}$ obtained by Lemma 7 is special however. If $\theta_{0}$ is given, choose $k$ so that $\left\|k \omega-\theta_{0}\right\|<\frac{1}{16}\left|I^{\prime}\right|$. Then $D_{k \omega}^{\prime}(\theta)=D^{\prime}(\theta+k \omega)$ is in normal form on $I^{\prime}$ and on $\frac{1}{2} I^{\prime}$-this is the third additional statement-with

$$
T_{k}^{-1} Q T_{k}, \quad \bigcup\left(\Lambda_{i}-k\right), \quad E_{k \omega}^{\prime}(\theta)=E^{\prime}(\theta+k \omega), \quad\left\{\Omega_{i}-k\right\}
$$

Then it follows that $D_{\theta_{0}}^{\prime}(\theta)=D^{\prime}\left(\theta+\theta_{0}\right)$ is in normal form on $\frac{3}{4} I^{\prime}$ since

$$
\left(k \omega-\theta_{0}\right)+\frac{1}{2}\left(\frac{3}{4} I^{\prime}\right) \supset \frac{1}{4} I^{\prime} \quad \text { and } \quad\left(k \omega-\theta_{0}\right)+\left(\frac{3}{4} I^{\prime}\right) \subset I^{\prime}
$$

Moreover, $E^{\prime}\left(\theta+\theta_{0}\right)$ is a spectral function for $D_{\theta_{0}}^{\prime}$.
Hence, we can choose a sequence $k_{j} \omega$ such that $\left\|k_{j} \omega-\theta_{0}\right\|<\frac{1}{16}\left|I_{j}\right|$. Then $D_{\theta_{0}}^{j}(\theta)=$ $D^{j}\left(\theta+\theta_{0}\right)$ will be in normal form on $\frac{3}{4} I_{j}$ with $E^{j}\left(\theta+\theta_{0}\right)$ as spectral function, and

$$
D_{\theta_{0}}^{j}(\theta) \rightarrow D_{\theta_{0}}^{\infty}(\theta)=D^{\infty}\left(\theta+\theta_{0}\right) \quad \text { and } \quad E^{j}\left(\theta+\theta_{0}\right) \rightarrow E^{\infty}\left(\theta+\theta_{0}\right)
$$

Hence $E_{\theta_{0}}^{\infty}(\theta)=E^{\infty}\left(\theta+\theta_{0}\right)$ and $Z_{\theta_{0}}^{j}=Z^{j}-\theta_{0}$.
Finally, the resolvent set of $D_{j+1}$ consists first of all of a modification $\Sigma_{j+1}^{\prime}$ of all the gaps of $D_{j}$ under the perturbation $F_{j}$, and secondly of newly created gaps which are contained in the image under $E^{j}$ of the set of $\theta$ such that

$$
\left|E^{j}(\theta)-E^{j}(\theta+l \omega)\right|<2 \eta_{j}
$$

for some $|l| \leqslant M_{j+1} N_{j+1}$. These contributions are easy to estimate and the estimate converges as $j \rightarrow \infty$ to something that goes to 0 with $\varepsilon$.

Proof of theorem. In order to prove the theorem we first observe that we can assume that $|E|<1$, so we can take $\frac{4}{3}<C<\frac{5}{3}$.

Now we let $D(\theta)=\operatorname{diag}(E(\theta+n \omega))$. It satisfies condition (A) for $K$ of the theorem, for $N=1$, and for $r=1$.

If we let $I$ be any symmetric interval around $0, \Lambda(k)=\{k\}$ for all $k$, and if we let $Q$ be the identity matrix, then also condition (B) is fulfilled for $N=1$ and any $M \geqslant 1$.

Let now

$$
M=\max \left(2^{s+4} C \frac{K^{s+1}((s+1)!)^{2}}{\xi}, 2 \tau, 8\right)
$$

By Lemma 3,

$$
\mid E(\theta)-E(k \omega)) \mid \geqslant \alpha
$$

outside a set of at most $M$ many intervals, each of length at most

$$
l(\alpha):=\frac{2}{K}\left(\frac{2 \alpha(s!)^{2} K^{s}}{\xi}\right)^{1 / s}
$$

By taking $\alpha$ small enough we can assume that each interval is separated from any other by a distance $\geqslant 4 l\left(\alpha_{0}\right)$. Choose then $\alpha$ even smaller so that

$$
\frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{l(\alpha)}\right)^{\tau+1} \leqslant\left(\frac{1}{4 l(\alpha)}\right)^{\tau+2}
$$

and

$$
4 l(\alpha) \leqslant \frac{\beta}{L}
$$

where $\beta, L$ will be defined below (independent of $\alpha$ ).
We choose a $m \omega$ in each such interval with the integer $m$ as close to $k$ as possible,

$$
|m-k| \leqslant \frac{2 \pi}{\varkappa}\left(\frac{2 \pi}{l(\alpha)}\right)^{\tau+1}
$$

and we let $\Omega(k)$ be the set of such $m$ 's. Then $\left(\mathrm{C}_{1}\right)$ and $\left(\mathrm{C}_{3}\right)-\left(\mathrm{C}_{5}\right)$ are fulfilled for $M$ as above and for $\lambda=|I| \leqslant 4 l(\alpha)$.

The first part of $\left(\mathrm{C}_{2}\right)$ and $\left(\mathrm{C}_{6}\right)$ holds with $L=K$-by Remark 4. By Lemma 4, the second part of $\left(\mathrm{C}_{2}\right)$ and $\left(\mathrm{C}_{6}\right)$ will be fulfilled for

$$
\beta=\left(\frac{1}{e}\right)^{M^{4 s M^{2}}}\left(\frac{l\left(\alpha_{0}\right) \xi}{\left(C K\left(s M^{2}+1\right)^{2}\right)^{2} K^{s}(s!)^{2}}\right)^{M^{2\left(s M^{2}+1\right)}}
$$

as in the verification of $\left(\mathrm{C}_{2}\right)$ and $\left(\mathrm{C}_{6}\right)$ in Lemma 7.
With this choice of parameters, $D(\theta)$ will be of normal form and satisfy the assumptions of Proposition 8. The perturbation

$$
F(\theta)=\left(\begin{array}{cccccc}
\ddots & \vdots & \vdots & \vdots & \vdots & \\
\ldots & -\varepsilon & 0 & -\varepsilon & 0 & \ldots \\
\cdots & 0 & -\varepsilon & 0 & -\varepsilon & \ldots \\
\ldots & \vdots & \vdots & \vdots & \vdots & \ddots
\end{array}\right)
$$

satisfies

$$
\left|F_{m, n}\right| \mathcal{C}^{\nu}<(e \varepsilon) e^{-|m-n| r} K^{\nu}
$$

Now the theorem follows from Proposition 8.

## 7. Some further properties

We have not studied the spectrum as a set, nor the occurrence of gaps-these play no role in the proof. One would, however, expect that gaps occur (at least for analytic potentials) and that the spectrum is a Cantor set.

We want to discuss the continuity of the eigenvalues as functions of $\theta$. Since $D_{j}(\theta)$ is smoothly conjugated to a block matrix with finite-dimensional blocks, there are continuous choices $E_{k}^{j}(\theta)$ of eigenvalues of $D_{j}(\theta)$ on $\mathbf{T} \backslash\{$ point \}-we don't claim that they are continuous on the whole of $T$. We can choose the continuous branches in such a way that $E_{k}^{j}(0)=E^{j}(k \omega)$ and

$$
\left|E_{k}^{j}(\theta)-E_{k}^{j+1}(\theta)\right| \leqslant\left|D_{j}(\theta)-D_{j+1}(\theta)\right|<\text { const } \cdot \frac{\sqrt{\varepsilon_{j}}}{r_{j}^{2}}
$$

It follows that there are continuous choices of the eigenvalues of $D_{\infty}(\theta)$, and hence of $D(\theta)+F(\theta)$. This analysis, however, does not permit us to conclude anything about smoothness.

Another question we like to address is the multiplicity of the spectrum. The "almost multiplicities" we have been dealing with are unbounded, but in the end the true multiplicity is always one. Indeed, if $D(0)+F(0)$ had a double eigenvalue $E_{k}^{\infty}(0)=E_{l}^{\infty}(0)$ there would also be two orthogonal eigenvectors, $q_{k}^{\infty}(0)$ and $q_{l}^{\infty}(0)$. This means that

$$
\begin{equation*}
-\varepsilon\left(u_{n+1}+u_{n-1}\right)+V_{n} u_{n}=z u_{n} \tag{z}
\end{equation*}
$$

has two independent solutions in $l^{2}(\mathbf{Z})$ when $z=E_{k}^{\infty}(0)=E_{l}^{\infty}(0)$. But if $\left(*_{z}\right)$ has two independent solutions in $l^{2}(Z)$ for one $z$ then this will be the case for all $z$-see for example [3, p. 245], where the proof in the continuous case translates directly to the discrete case. But for sufficiently large $|z|$ the equation is reducible to a constant-coefficient equation so there are no solutions in $l^{2}(\mathbf{Z})$ at all. Hence, the spectrum of $D(0)+F(0)$ must be simple.

Since our approach permits us to analyze the system for all $\theta$ to any degree of approximation, one can use it to try to understand the equation also for $\theta$ in the exceptional measure-zero set. We expect that, for generic $\theta$ and for generic potentials (in the Gevrey class we have studied, or perhaps in some weaker Gevrey class), the blocks do not become stationary but continue to increase and give rise to eigenvectors in $l^{\infty}(\mathbf{Z})$ and a singular continuous spectrum for $D(\theta)+F(\theta)$. (A result of this type has been announced in [9].)

As for the exponential decay of the eigenvectors, we have not provided any proof of this. The reason why we don't immediately get such decay is that the parameter $r_{j}$, with which we measure the exponential decay, decreases in order to control the factor $e^{N_{j} r_{j}}$
appearing in Lemma 5. This cannot be avoided since $r$ is a global parameter, measuring the minimal exponential decay of all the eigenvalues-therefore it must decrease to zero. However, if we measured the exponential decay of eigenvectors associated to a given block $\Lambda^{j}(k)$ we would probably see that the decay stops once the blocks become stationaryonly a very small decay, which does not go to zero, will be required by the construction in Corollary 6.

Considered as a dynamical system on $\mathbf{T} \times \operatorname{SL}(2, \mathbf{R})$ the theorem implies, by Kotani's theorem [1, Proposition VII.3.3], that the skew product

$$
\left(\begin{array}{c}
\theta \\
u \\
u^{\prime}
\end{array}\right) \rightarrow\left(\begin{array}{c}
\theta+\omega \\
u^{\prime} \\
-u+(E(\theta+\omega)-\lambda) u^{\prime} / \varepsilon
\end{array}\right)
$$

has non-zero Lyapunov exponents for almost every $\lambda$.

## Appendix

Lemma A1. There is a $\mathcal{C}^{\infty}$-function $\phi \geqslant 0$ on $\mathbf{R}$, supported in $\left[-\frac{1}{2}, \frac{1}{2}\right]$, such that

$$
|\phi|_{\mathcal{C}^{\nu}}<\gamma_{0} \gamma_{0}^{\nu} \quad \forall \nu \geqslant 0
$$

for some $\gamma_{0}>0$.
Given two intervals $I_{2} \subset I_{1}$ with $\operatorname{dist}\left(I_{2}, \partial I_{1}\right) \geqslant \delta$. Then there is a smooth function $\phi \geqslant 0$ of compact support in $I_{1}$ which is $\equiv 1$ on $I_{2}$ and such that

$$
|\phi|_{\mathcal{C}^{\nu}}<\left(\frac{\gamma_{0}}{\delta}\right)^{\nu} \quad \forall \nu \geqslant 0
$$

Proof. The class of all $\mathcal{C}^{\infty}$-functions satisfying such a bound is not quasi-analytic [13, 19.11]. Hence it contains a non-zero function $\psi$ which satisfies this bound for some $\gamma_{0}>0$ and which is compactly supported [13, 19.10]-by a scaling we can obviously assume that it is supported in $\left[-\frac{1}{2}, \frac{1}{2}\right]$. Then if we take $\phi=\psi^{2}$ and replace $\gamma_{0}$ by a suitable larger constant we have our function.

By multiplying the function by an appropriate constant we can assume that its integral equals one. We localize then our function on the left component of $I_{1} \backslash I_{2}$. Since such a component is of length at least $\delta$, it will satisfy

$$
|\phi|_{\mathcal{C}^{\nu}}<\frac{\gamma_{0}}{\delta}\left(\frac{\gamma_{0}}{\delta}\right)^{\nu} \quad \forall \nu \geqslant 0
$$

We now define $\phi$ on the right component in the same way, but negative, and then we take the primitive of this function.

Let now $u_{j}$ be a sequence of smooth functions defined on an open interval $I$.

Lemma A2. Assume that $\left|u_{j}\right|_{\mathcal{C}^{\nu}}<C_{j} K^{\nu} \forall \nu \geqslant 0$. Then for all $\nu \geqslant 0$,

$$
\begin{aligned}
\left|u_{1} \cdot \ldots \cdot u_{J}\right|_{\mathcal{C}^{\nu}} & <4^{J-1}\left(C_{1} \cdot \ldots \cdot C_{j}\right) K^{\nu}, \\
\left|e^{u_{1}}\right|_{\mathcal{C}^{\nu}} & <e^{4 C_{1}} K^{\nu}, \\
\left|\sqrt{1+u_{1}}\right|_{\mathcal{C}^{\nu}} & <\sqrt{1+C_{1}} K^{\nu} \quad \text { if } C_{1}<\frac{1}{4} \\
\left|\frac{1}{u_{1}}\right|_{\mathcal{C}^{\nu}} & <\frac{1}{\delta}\left(\frac{4}{\delta} C_{1} K\right)^{\nu} \quad \text { if }\left|u_{1}\right|>\delta .
\end{aligned}
$$

Proof.

$$
\left|\partial^{\nu}\left(u_{1} u_{2}\right)\right|=\left|\sum_{i=0}^{\nu}\binom{\nu}{i} \partial^{\nu-i} u_{1} \partial^{i} u_{2}\right|<\left(C_{1} C_{2}\right) K^{\nu}(\nu!)^{2} \sum_{i=0}^{\nu}\binom{\nu}{i} \frac{(i!)^{2}((\nu-i)!)^{2}}{(\nu!)^{2}}
$$

and the result for $J=2$ follows since $\sum_{i=0}^{\nu}\binom{\nu}{i}^{-1}<4$ for all $\nu$-just notice that $\binom{\nu}{i} \geqslant 2^{i}$ if $i \leqslant \frac{1}{2} \nu$. For $J \geqslant 3$ the result follows by induction.

We use the power series expansions of $e^{u}$ and $\sqrt{1+u}$ which converge absolutely in $|u|<\infty$ and in $|u|<1$ respectively, to prove the second and third estimates.

Notice that $\left|u_{1}\right|>\delta$ implies that $C_{1}>\delta$. By a scaling we can suppose that $\delta=1$ and $C_{1} \geqslant 1$. The fourth statement is obvious for $\nu=0$ so we proceed by induction. Hence

$$
\begin{aligned}
\left|\partial^{\nu}\left(\frac{1}{u_{1}}\right)\right| & =\left|-\frac{1}{u_{1}} \sum_{i=1}^{\nu}\binom{\nu}{i} \partial^{\nu-i}\left(\frac{1}{u_{1}}\right) \partial^{i} u_{1}\right| \\
& <4^{\nu-1}(\nu!)^{2}\left(C_{1} K\right)^{\nu} \sum_{i=1}^{\nu}\binom{\nu}{i} \frac{(i!)^{2}((\nu-i)!)^{2}}{(\nu!)^{2}} \\
& <(\nu!)^{2}\left(4 C_{1} K\right)^{\nu} .
\end{aligned}
$$

Remark. The preceding lemma is valid also for matrices if we understand by $1 / u$ the inverse of $u$, and if we use the operator norm satisfying

$$
|u v| \leqslant|u| \cdot|v| .
$$

Lemma A3. Let $v^{1}, \ldots, v^{d}$ be a basis of $\mathbf{R}^{d}$ for $\theta \in I \subset \mathbf{R}, d \geqslant 2$, such that

$$
\begin{aligned}
&\left|v^{m}\right|_{\mathcal{C}^{\nu}}<C K^{\nu} \quad \forall \nu \geqslant 0, \\
&\left|\left\langle v^{m}, v^{n}\right\rangle\right|_{\mathcal{C}^{0}}<\frac{1}{4 d^{2} 3^{d+4}}, \quad m \neq n, \\
& 1-\frac{1}{4 d^{2} 3^{d+4}}<\left|v^{m}\right|_{\mathcal{C}^{0}}<1+\frac{1}{4 d^{2} 3^{d+4}},
\end{aligned}
$$

for all $m, n$ and for $C \geqslant 1$. Then the orthonormal basis $\hat{v}^{1}, \ldots, \hat{v}^{d}$ obtained by GramSchmidt orthogonalization verifies

$$
\left|\hat{v}^{m}\right|_{\mathcal{C}^{\nu}} \leqslant\left(d^{8^{d}} C K\right)^{\nu} \quad \forall \nu \geqslant 0
$$

Proof. We have

$$
\left|v^{m}\right|_{\mathcal{C}^{\nu}}<\left(1+\frac{1}{4 d^{2} 3^{d+4}}\right)(C K)^{\nu}
$$

By Gram-Schmidt we first obtain an orthogonal basis $\tilde{v}^{1}, \ldots, \tilde{v}^{d}$, and if we define

$$
\left\{\begin{array}{l}
w^{1}=\tilde{v}^{1}=v^{1} \\
w^{m}=\left|w^{1}\right|^{2} \cdot \ldots \cdot\left|w^{m-1}\right|^{2} \tilde{v}^{m}
\end{array}\right.
$$

then these vectors will satisfy

$$
w^{m}=\left|w^{1}\right|^{2} \cdot \ldots \cdot\left|w^{m-1}\right|^{2}\left(v^{m}-\sum_{j=1}^{m-1}\left\langle v^{m}, w^{j}\right\rangle \frac{1}{\left|w^{j}\right|^{2}} w^{j}\right)
$$

(Notice that there is no division in this formula.) From this we get by induction, using Lemma A2,

$$
\begin{gathered}
\left|w^{m}\right|_{\mathcal{C}^{\nu}}<C_{m}(C K)^{\nu} \\
C_{m}=m d^{m} 4^{2 m-2}\left(C_{1} \cdot \ldots \cdot C_{m-1}\right)^{2} C_{1}, \quad C_{1}=1+\frac{1}{4 d^{2} 3^{d+4}}<2
\end{gathered}
$$

which gives

$$
C_{m} \leqslant\left(4^{2 m-2} m d^{m} C_{1}\right)^{3^{m-2}} C_{1}^{2 \cdot 3^{m-2}} \leqslant\left(4^{2 m} m d^{m}\right)^{3^{m-2}}
$$

It follows also that, for each $m$,

$$
\sqrt{1-\frac{1}{4}} \leqslant\left(1-\frac{1}{3^{d+2}}\right)^{3^{m-1}} \leqslant\left|w^{m}\right| \leqslant\left(1+\frac{1}{3^{d+2}}\right)^{3^{m-1}} \leqslant \sqrt{1+\frac{1}{4}}
$$

and from Lemma A2 that

$$
\left|\frac{w^{m}}{\sqrt{\left\langle w^{m}, w^{m}\right\rangle}}\right|_{\mathcal{C}^{\nu}} \leqslant\left(4^{6} d C_{d}^{3} C K\right)^{\nu}
$$

Polynomials. Let $P(\lambda, \theta)$ be a polynomial in $\lambda$ of degree $d$ with leading coefficient 1 , smoothly parametrized by $\theta \in I$, and satisfying

$$
|P(\lambda, \cdot)|_{\mathcal{C}^{\nu}}<C K^{\nu} \quad \forall \nu \geqslant 0
$$

in a disk $|\lambda|<R$ of radius $R$. Assume that the roots $\left\{E_{m}(\theta)\right\}_{1}^{d}$ are real and lie in $|\lambda|<R-1$.

Lemma A4. If the roots belong to two groups,

$$
\left\{\begin{array}{l}
E_{1}(\theta), \ldots, E_{k}(\theta) \\
E_{k+1}(\theta), \ldots, E_{d}(\theta)
\end{array}\right.
$$

such that any root of one group is separated from any root of the other group by at least $\delta<1$, then the polynomial

$$
\prod_{m=1}^{k}\left(\lambda-E_{m}(\theta)\right)=\sum_{j=0}^{k} e_{k-j}(\theta) \lambda^{j}
$$

satisfies

$$
\left|e_{j}\right|_{\mathcal{C}^{\nu}}<\left(4^{d+2} R C \frac{1}{\delta^{d-1}}\right)^{j}\left(2^{d+2} C \frac{1}{\delta^{d}} K\right)^{\nu}
$$

Proof. Choose a curve $\Delta(\theta)$ in $|\lambda|<R-\frac{1}{2}$, piecewise constant in $\theta$, keeping a distance $\geqslant \frac{1}{2} \delta$ to all the roots $E_{1}(\theta), \ldots, E_{d}(\theta)$ of $P(\lambda, \theta)$ and surrounding the first $k$ of these roots. $\Delta(\theta)$ may consist of several components so we can choose it to be of length at most $d \pi \delta$. Then we have for all $\lambda \in \Delta(\theta)$,

$$
\begin{gathered}
|P(\lambda, \theta)| \geqslant\left(\frac{1}{2} \delta\right)^{d}=\varrho \\
\left|\frac{1}{P(\lambda, \cdot)}\right|_{\mathcal{C}^{\nu}}<\frac{1}{\varrho}\left(\frac{4}{\varrho} C K\right)^{\nu}, \\
\left|\frac{P_{\lambda}(\lambda, \cdot)}{P(\lambda, \cdot)}\right|_{\mathcal{C}^{\nu}}<4 \cdot 2 C \frac{1}{\varrho}\left(\frac{4}{\varrho} C K\right)^{\nu},
\end{gathered}
$$

where the second and third estimates follow from Lemma A2-we have used the Cauchy formula to estimate $\left|P_{\lambda}(\lambda, \cdot)\right|_{\mathcal{C}^{\nu}}$ on $|\lambda|<R-\frac{1}{2}$.

Consider now the power symmetric functions in the first $k$ roots:

$$
p_{j}(\theta)=E_{1}(\theta)^{j}+\ldots+E_{k}(\theta)^{j} .
$$

We have the integral representation

$$
p_{j}(\theta)=\frac{1}{2 \pi} \oint_{\Delta(\theta)} \lambda^{j} \frac{P_{\lambda}(\lambda, \theta)}{P(\lambda, \theta)} d \lambda
$$

from which we get

$$
\left|p_{j}\right|_{\mathcal{C}^{\nu}}<4 d \delta \frac{C}{\varrho}\left(\frac{4 C}{\varrho} K\right)^{\nu} R^{j}
$$

The $e_{j}$ 's are the elementary symmetric functions in the first $k$ roots and we have the relation

$$
e_{j}=\frac{(-1)^{j}}{j!} \operatorname{det}\left(\begin{array}{ccccc}
p_{1} & 1 & 0 & 0 & \ldots \\
p_{2} & p_{1} & 2 & 0 & \ldots \\
p_{3} & p_{2} & p_{1} & 3 & \ldots \\
\vdots & \vdots & \vdots & \vdots & \ddots \\
p_{j} & p_{j-1} & p_{j-2} & p_{j-3} & \ldots
\end{array}\right)
$$

for $j \geqslant 1$ [11, p. 20]-of course $e_{0}=1$. So $j!e_{j}$ is a sum of at most $j!$ many products

$$
p_{\iota_{1}} \cdot \ldots \cdot p_{\iota_{l}}, \quad \iota_{1}+\ldots+\iota_{l}=j \leqslant k<d
$$

each with a coefficient that is at most $(j-1)$ !. From Lemma A2 we now get the estimate we want.

Symmetric matrices. Let $D(\theta)$ be a symmetric ( $d \times d$ )-matrix, smooth in the parameter $\theta \in I \subset \mathbf{R}$, and let $\left\{E_{m}(\theta)\right\}_{1}^{d}$ be its eigenvalues and $\left\{q^{m}(\theta)\right\}_{1}^{d}$ the corresponding eigenvectors. The eigenvalues and eigenvectors are smooth in $\theta$ when all eigenvalues are simple. The eigenvalues can in fact always be chosen to be continuous, and even $\mathcal{C}^{1}$ [10, p. 122], in $\theta$, but this is not the case for the eigenvectors. A noteworthy exception is when $D$ is analytic in the parameter because then the eigenvalues and the eigenvectors are analytic [6]. For example, if $D=D_{0}+\varepsilon D_{1}$ with $D_{0}$ diagonal, then the analyticity in $\varepsilon$ gives a unique choice of the eigenvalues.

We always have estimates of the first derivative $\partial^{1} E_{m}$ in terms of $D$.
Lemma A5.

$$
\left|E_{m}(\theta)\right| \leqslant|D(\theta)| \quad \text { and } \quad\left|\partial E_{m}(\theta)\right| \leqslant|\partial D(\theta)|
$$

Proof. Let $q^{m}(\theta)$ be the eigenvector corresponding to $E_{m}(\theta)$. Then

$$
\left(D(\theta)-E_{1}(\theta) I\right) q^{1}(\theta)=0
$$

and if we take the scalar product with $q^{1}(\theta)$ then we get an estimate of $E_{1}$.
If we differentiate the relation and take the scalar product with $q^{1}(\theta)$, and use that the eigenvectors are orthogonal, then we get an estimate of $\partial E_{1}$.

For higher-order derivatives we have no such estimates. For example, if

$$
D(\theta)=\left(\begin{array}{cc}
\theta & \varepsilon \\
\varepsilon & -\theta
\end{array}\right)
$$

then $E_{1}(\theta)= \pm \sqrt{\theta^{2}+\varepsilon^{2}}$ and, hence, $\partial^{2} E_{1}= \pm 1 / \varepsilon$.

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