# Central Limit Theorems for Quadratic Forms in Random Variables Having Long-Range Dependence 

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## 1. Introduction

Let $f(x)$ and $g(x)$ be integrable real symmetric functions on $[-\pi, \pi]$ that are bounded on subintervals that exclude the origin. Let $X_{1}, X_{2}, \ldots$ be a mean zero stationary Gaussian sequence with spectral density $f(x)$, and let $\ldots,-a_{1}, a_{0}, a_{1}, \ldots$ be the Fourier coefficients of $g(x)$. We prove that the distribution of the normalized quadratic form

$$
Z_{N}=\frac{1}{\sqrt{N}}\left\{\sum_{i=1}^{N} \sum_{j=1}^{N} a_{i-j} X_{i} X_{j}-E \sum_{i=1}^{N} \sum_{j=1}^{N} a_{i-j} X_{i} X_{j}\right\}
$$

converges to a normal distribution if there exist constants $\alpha<1$ and $\beta<1$ with $\alpha+\beta<1 / 2$ such that for each $\delta>0, f(x)=O\left(|x|^{-\alpha-\delta}\right)$ and $g(x)=O\left(|x|^{-\beta-\delta}\right)$ as $x \rightarrow 0$.

Of particular interest are the cases where $f(x) \sim x^{-\alpha} L_{1}(x)$ and $g(x) \sim x^{-\beta} L_{2}(x)$ as $x \rightarrow 0$ with $L_{1}$ and $L_{2}$ slowly varying. The exponents $\alpha$ and $\beta$ are allowed to be positive, zero or negative. The sequence $\left\{X_{j}\right\}$ is said to exhibit a long-range dependence when $\alpha>0$. When $\alpha<0$, the covariances $r_{k}$ $=E X_{j} X_{j+k}$ satisfy $\sum_{k=-\infty}^{+\infty} r_{k}=0$.

Suppose $f(x) \sim \sim^{k=-\infty} x_{1}(x)$ and $g(x) \sim x^{-\beta} L_{2}(x)$ as $x \rightarrow 0$. Rosenblatt (1961) showed that in the special case $1 / 2<\alpha<1$ and $a_{i-j}=\delta_{i j}$, the quadratic form $\sum_{i=1}^{N} \sum_{j=1}^{N} a_{i-j} X_{i} X_{j}$, adequately normalized, converges to a non-normal distribution. The assumption $a_{i-j}=\delta_{i j}$ implies $g(x)$ constant and thus $\beta=0$. Our result shows that the normalized quadratic form $Z_{N}$ converges to a normal distribu-

[^0]tion when $1 / 2<\alpha<1$ and $\beta<1 / 2-\alpha<0$. If $\alpha \leqq 1 / 2$, it is even possible to choose $\beta>0$ as long as $\beta<\min (1 / 2-\alpha, 1)$.

These results are used in the study of the asymptotic behavior of maximum likelihood type estimators related to the sequence $\left\{X_{j}\right\}$ (Fox and Taqqu 1986). Examples of sequences $\left\{X_{j}\right\}$ satisfying $f(x) \sim x^{-\alpha} L_{1}(x)$ that are of special interest include fractional Gaussian noise and fractional ARMA.

A sequence $\left\{X_{j}\right\}$ is fractional Gaussian noise (Mandelbrot and Van Ness 1968) if its covariance satisfies

$$
r(k)=E X_{j} X_{j+k}=\frac{\sigma^{2}}{2}\left\{| | k|-1|^{2 H}-2|k|^{2 H}+(|k|+1)^{2 H}\right\}
$$

for $0<H<1$. In that case (Sinai 1976)

$$
f(x)=\frac{\sigma^{2}}{\int_{-\infty}^{+\infty}(1-\cos y)|y|^{-1-2 H} d y}(1-\cos x) \sum_{k=-\infty}^{+\infty}|x+2 k \pi|^{-1-2 H},
$$

so that $\alpha=2 H-1 \in(-1,1)$.
A sequence $\left\{X_{j}\right\}$ is fractional ARMA (Hoskings 1981) if its spectral density is

$$
f(x)=\left|e^{i x}-1\right|^{-d} \frac{\left|\varphi\left(e^{i x}\right)\right|^{2}}{\left|\psi\left(e^{i x}\right)\right|^{2}}
$$

where $\varphi$ and $\Psi$ are polynomials having no zeroes on the unit circle and $d<1$. In that case $\alpha=d$. Heuristically, fractional ARMA is the sequence, which, when differenced $d / 2$ times, yields an autoregressive-moving average (ARMA) sequence with spectral density $\left|\varphi\left(e^{i x}\right)\right|^{2} /\left|\Psi\left(e^{i x}\right)\right|^{2}$.

Our main results are in Sect. 2. Sections 3 through 7 are devoted to the proof of Theorem 1. That proof uses "power counting" arguments in the sense of mathematical physics. In Sect. 3 we introduce the power counting set-up and state an extension of a power counting theorem of Lowenstein and Zimmerman (1975). Preliminary lemmas are proven in Sect. 4 and, together with the results of Sect. 5, they are used to establish Propositions 6.1 and 6.2 of Sect. 6. These propositions describe the asymptotic behavior of certain multiple integrals. Section 7 contains the proof of Theorem 1 . Theorem 4 is proven in Sect. 8.

## 2. Main Results

Let $f(x)$ and $g(x)$ be integrable real symmetric functions on $[-\pi, \pi]$, not necessarily non-negative. Define the Fourier coefficients

$$
r_{n}=\int_{-\pi}^{\pi} e^{i n x} f(x) d x
$$

and

$$
a_{n}=\int_{-\pi}^{\pi} e^{i n x} g(x) d x
$$

Let $R_{N}$ and $A_{N}$ be the $N \times N$ matrices with entries $\left(R_{N}\right)_{j, k}=r_{j-k}$ and $\left(A_{N}\right)_{j, k}$ $=a_{j-k}, 0 \leqq j, k \leqq N-1$. Let $\operatorname{Tr} M$ denote the trace of a matrix $M$.

We say that $f$ satisfies the regularity condition if the discontinuities of $f$ have Lebesgue measure 0 and $f$ is bounded on the interval $[\delta, \pi]$ for all $\delta>0$.

Theorem 1. Suppose that $f$ and $g$ each satisfy the regularity condition. Suppose in addition that there exist $\alpha<1$ and $\beta<1$ such that for each $\delta>0$

$$
|f(x)|=O\left(|x|^{-\alpha-\delta}\right) \quad \text { as } x \rightarrow 0
$$

and

$$
|g(x)|=O\left(|x|^{-\beta-\delta}\right) \quad \text { as } x \rightarrow 0
$$

Then
a) If $p(\alpha+\beta)<1$,

$$
\lim _{N \rightarrow \infty} \frac{\operatorname{Tr}\left(R_{N} A_{N}\right)^{p}}{N}=(2 \pi)^{2 p-1} \int_{-\pi}^{\pi}[f(x) g(x)]^{p} d x .
$$

b) If $p(\alpha+\beta) \geqq 1$,

$$
\operatorname{Tr}\left(R_{N} A_{N}\right)^{p}=o\left(N^{p(\alpha+\beta)+\varepsilon}\right) \text { for every } \varepsilon>0
$$

The theorem is proven in Sect. 7. The proof of Part a) amounts to showing that
where

$$
\lim _{N \rightarrow \infty} \int_{[-\pi, \pi]^{2 p}} Q(y) d \mu_{N}(y)=\int_{[-\pi, \pi]^{2 p}} Q(y) d \mu(y)
$$

$$
\begin{aligned}
Q(y)= & f\left(y_{1}\right) g\left(y_{2}\right) f\left(y_{3}\right) g\left(y_{4}\right) \ldots b\left(y_{2 p-1}\right) g\left(y_{2 p}\right), \\
d \mu_{N}(y)= & \frac{1}{(2 \pi)^{2 p-1}} N \sum_{j_{1}=0}^{N-1} \ldots \sum_{j_{2}=0}^{N-1} e^{i\left(j_{1}-j_{2}\right) y_{1}} e^{i\left(j_{2}-j_{3}\right) y_{2}} \\
& \ldots e^{i\left(j_{2 p-1}-j_{2 p}\right) y_{2 p-1}} e^{i\left(j_{2 p}-j_{1}\right) y_{2 p}} d y_{1} \ldots d y_{2_{p}},
\end{aligned}
$$

and where $\mu$ is Lebesgue measure concentrated on the diagonal of $[-\pi, \pi]^{2 p}$.
Introduce now a stationary Gaussian sequence $X_{j}, j \geqq 1$ with mean 0 and spectral density $f(x) \geqq 0$, so that

$$
E X_{j} X_{j+k}=r_{k}=\int_{-\pi}^{\pi} e^{i k x} f(x) d x
$$

Let $x_{N}$ denote the random vector $\left(X_{1}, X_{2}, \ldots, X_{N}\right)$. Put $\mu_{N}=E x_{N}^{\prime} A_{N} x_{N}$.
Theorem 2. Suppose that $f$ and $g$ each satisfy the regularity condition. Suppose in addition that there exist $\alpha<1$ and $\beta<1$ such that $\alpha+\beta<1 / 2$ and such that for each $\delta>0$

$$
\begin{array}{ll}
f(x)=O\left(|x|^{-\alpha-\delta}\right) & \text { as } x \rightarrow 0 \\
g(x)=O\left(|x|^{-\beta-\delta}\right) & \text { as } x \rightarrow 0
\end{array}
$$

Then

$$
\frac{x_{N}^{\prime} A_{N} x_{N}-\mu_{N}}{\sqrt{N}}
$$

tends in distribution to a normal random variable with mean 0 and variance $16 \pi^{3} \int_{-\pi}^{\pi}[f(x) g(x)]^{2} d x$.

Proof. Since the sequence $X_{j}$ is Gaussian, the $p^{\text {th }}$ cumulant of $x_{N}^{\prime} A_{N} x_{N}$ is equal to $2^{p-1}(p-1)!\operatorname{Tr}\left(R_{N} A_{N}\right)^{p}$. (See, for example, Grenander and Szego 1958, p. 218). Thus the $p^{\text {th }}$ cumulant of

$$
\frac{x_{N}^{\prime} A_{N} x_{N}-\mu_{N}}{\sqrt{N}}
$$

is given by

$$
c_{p}(N)= \begin{cases}0 & \text { if } p=1 \\ 2^{p-1}(p-1)!\frac{\operatorname{Tr}\left(R_{N} A_{N}\right)^{p}}{N^{p / 2}} & \text { if } p \geqq 2\end{cases}
$$

An application of Theorem 1 yields

$$
\lim c_{P}(N)= \begin{cases}0 & \text { if } p \neq 2 \\ 16 \pi^{3} \int_{-\pi}^{\pi}[f(x) g(x)]^{2} d x & \text { if } p=2\end{cases}
$$

This implies the conclusion of Theorem 2.
The following is an immediate consequence of Theorem 2.
Theorem 3. Suppose that $f$ and $g$ each satisfy the regularity condition. Suppose in addition that there exist $\alpha<1$ and $\beta<1$ such that $\alpha+\beta<1 / 2$,

$$
f(x) \sim|x|^{-\alpha} L_{1}(x) \quad \text { as } x \rightarrow 0
$$

and

$$
g(x) \sim|x|^{-\beta} L_{2}(x) \quad \text { as } x \rightarrow 0
$$

where $L_{1}$ and $L_{2}$ are slowly varying at 0 . Then the conclusion of Theorem 2 holds.

The next theorem, which is used in Fox and Taqqu (1986), is proven in Sect. 8. Define $\bar{X}_{N}=(1 / N) \sum_{j=1}^{N} X_{j}$ and the random vector $\tilde{x}_{N}=\left(X_{1}-\bar{X}_{N}, \ldots, X_{N}\right.$
$\left.-\bar{X}_{N}\right)$.

Theorem 4. If the conditions of Theorem 2 are satisfied, then

$$
\frac{\tilde{x}_{N}^{\prime} A_{N} \tilde{x}_{N}-E\left\{\tilde{x}_{N}^{\prime} A_{N} \tilde{x}_{N}\right\}}{\sqrt{N}}
$$

tends in distribution to a normal random vector with mean 0 and variance $16 \pi^{3} \int_{-\pi}^{\pi}[f(x) g(x)]^{2} d x$.

## 3. Power Counting Theorems

Power counting methods can be used to verify the convergence of multiple integrals whose integrands are products of powers of affine functionals. Let $b_{1}, \ldots, b_{m}$ and $\theta_{1}, \ldots, \theta_{m}$ be real constants and let $M_{1}(x), \ldots, M_{m}(x)$ be $m$ linear functionals on $\mathbb{R}^{n}$. Put $L_{j}(x)=M_{j}(x)+\theta_{j}, j=1, \ldots, m$. Define the function $P$ : $\mathbb{R}^{n} \rightarrow \mathbb{R} \cup\{\infty\}$ by

$$
P(x)=\left|L_{1}(x)\right|^{b_{1}}\left|L_{2}(x)\right|^{b_{2}} \ldots\left|L_{m}(x)\right|^{b_{m}}
$$

Define $T=\left\{L_{1}, \ldots, L_{m}\right\}$ and let $W \subset T$. Let $\operatorname{span}\{W\}$ denote the set of linear combinations of elements of $W$ and $s(W)$ denote those linear combinations which coincide with elements of $T$. Thus

$$
s(W)=T \cap \operatorname{span}\{W\}
$$

For each $W \subset T$ we define the quantity

$$
d(P, W)=|W|+\sum_{\left\{j: L_{j} \in S(W)\right\}} b_{j},
$$

where $|W|$ denotes the cardinality of $W$. We refer to $d(P, W)$ as the dimension of $P$ with respect to $W$. We say that $W=\left\{L_{i_{1}}, \ldots, L_{i_{k}}\right\}$ is strongly independent if $M_{i_{1}}, \ldots, M_{i_{k}}$ are linearly independent. Let $S$ be the set of those $L_{j}$ in $T$ that have exponents $b_{j}<0$. Finally, for each $t>0$, let

$$
U_{t}=[-t, t]^{n}=\left\{x \in \mathbb{R}^{n}:\left|x_{i}\right| \leqq t, i=1, \ldots, n\right\} .
$$

The next theorem extends a basic result of Lowenstein and Zimmermann (1975). It is proved at the end of Sect. 4.

Theorem 3.1. Suppose that $d(P, W)>0$ for every strongly independent set $W \subset S$. Then $\int_{U_{t}} P(x) d x<\infty$ for all $t>0$.

To illustrate the application of the theorem, let $n=3$ and define $P(x)$ : $\mathbb{R}^{3} \rightarrow \mathbb{R} \cup\{\infty\}$ by

$$
P(x)=\left|x_{1}+x_{2}+2\right|^{b_{1}}\left|x_{1}+x_{2}+x_{3}-1\right|^{b_{2}}\left|x_{3}-3\right|^{b_{3}}
$$

where $b_{1}, b_{2}, b_{3}<0$. Define $L_{1}(x)=x_{1}+x_{2}+2, L_{2}(x)=x_{1}+x_{2}+x_{3}-1$ and $L_{3}(x)$ $=x_{3}-3$. Then $S=T=\left\{L_{1}, L_{2}, L_{3}\right\}$. The strongly independent subsets of $S$ are $\left\{L_{1}\right\},\left\{L_{2}\right\},\left\{L_{3}\right\},\left\{L_{1}, L_{2}\right\},\left\{L_{1}, L_{3}\right\}$ and $\left\{L_{2}, L_{3}\right\}$. We have $d\left(P,\left\{L_{j}\right\}\right)=1+b_{j}, j$ $=1,2,3$. The other three dimensions are all equal to $2+b_{1}+b_{2}+b_{3}$ because for example $s\left(\left\{L_{1}, L_{2}\right\}\right)=\left\{L_{1}, L_{2}, L_{3}\right\}$. Therefore $\int_{U_{t}} P(x) d x$ will be finite provided
that $b_{1}+b_{2}+b_{3}>-2$ and $b_{1}, b_{2}, b_{3}>-1$. that $b_{1}+b_{2}+b_{3}>-2$ and $b_{1}, b_{2}, b_{3}>-1$.
Remark. Suppose the condition of Theorem 3.1 is satisfied. Then in fact $d(P, W)>0$ for every $W \subset T$. To see this, suppose first that $W \subset T$ contains only one element not in $S$, say $L$. Then $W=W_{0} \cup\{L\}$, where $W_{0} \subset S$ and $d\left(P, W_{0}\right)>0$. If $d(P, W) \leqq 0$ then $S(W)$ must contain some element of $S$ which is not in $W_{0}$, say $L^{\prime}$. Then $W_{1}=W_{0} \cup\left\{L^{\prime}\right\}$ satisfies $W_{1} \subset S$ and $d\left(P, W_{1}\right)$ $=d(P, W)<0$, since $S\left(W_{1}\right)=S\left(W_{2}\right)$ and $\left|W_{1}\right|=\left|W_{2}\right|$. This contradicts the assump-
tion. Hence any subset of $T$ which differs from a subset of $S$ by one element has positive dimension. The same method can be used inductively to show that all subsets of $T$ have positive dimension.

## 4. Preliminary Lemmas

Retain the notation introduced in Sect. 3. Fix a permutation $\sigma=\left(\sigma_{1}, \ldots, \sigma_{m}\right)$ of $\{1, \ldots, m\}$ and let

$$
E_{\sigma}^{t}=\left\{x \in U_{t}:\left|L_{\sigma_{1}}(x)\right| \leqq\left|L_{\sigma_{2}}(x)\right| \leqq \ldots \leqq\left|L_{\sigma_{m}}(x)\right|\right\}
$$

We use the greedy algorithm to construct a basis $B_{\sigma}$ for $T$. The greedy algorithm proceeds as follows. We put $L_{\sigma_{1}} \in B_{\sigma}$. We put $L_{\sigma_{2}} \in B_{\sigma}$ if $L_{\sigma_{2}}$ is not in the span of $\left\{L_{\sigma_{\sigma}}\right\}$. On the $j^{\text {th }}$ step we put $L_{\sigma_{j}} \in B_{\sigma}$ if $L_{\sigma_{j}}$ is not in the span of $\left\{L_{\sigma_{1}}, \ldots, L_{\sigma_{j-1}}\right\}$. It is well known that in this way we obtain a basis $B_{\sigma}$ $=\left\{L_{\tau_{1}}, \ldots, L_{\tau_{r}}\right\}$ for $T$, where $r$ is the rank of $T$. We then have

$$
\begin{equation*}
\left|L_{\tau_{1}}\right| \leqq\left|L_{\tau_{2}}\right| \leqq \ldots \leqq\left|L_{\tau_{r}}\right|, \quad x \in E_{\sigma}^{t} \tag{4.1}
\end{equation*}
$$

The functions $L_{\tau_{1}}, \ldots, L_{\tau_{r}}$ are linearly independent but not necessarily strongly independent.

We use $B_{\sigma}$ to construct the partition of $T$ given by
and

$$
T_{1}=s\left\{L_{\tau_{1}}\right\}
$$

$$
T_{k}=s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\} / s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k-1}}\right\}, \quad k=2, \ldots, r .
$$

Lemma 4.1. For each permutation $\sigma$ there is a constant $C_{\sigma}$ (independent of $x$ and t) such that if $L \in T_{k}$ then
a) $|L| \leqq C_{\sigma}\left|L_{\tau_{k}}\right|, x \in E_{\sigma}^{t}$,
and
b) $\left|L_{\tau_{k}}\right| \leqq|L|, x \in E_{\sigma}^{t}$.

Proof. a) If $L \in T_{k}$ then $L=a_{1} L_{\tau_{1}}+\ldots+a_{k} L_{\tau_{k}}$ for some constants $a_{1}, \ldots, a_{k}$. Therefore

$$
|L| \leqq\left|a_{1}\right|\left|L_{\tau_{1}}\right|+\ldots+\left|a_{k}\right|\left|L_{\tau_{k}}\right|, \quad x \in \mathbb{R}^{n} .
$$

Relation (4.1) implies that for $x \in E_{\sigma}^{t}$ the right hand side is less than $\left(\left|a_{1}\right|+\ldots\right.$ $\left.+\left|a_{k}\right|\right)\left|L_{\mathrm{t}_{k}}\right|$.
b) Suppose that $L \in T_{k}$. We must have either $L=L_{\tau_{k}}$ or else $L$ was rejected by the greedy algorithm. In proving b) we can thus assume that $L$ was rejected by the greedy algorithm. Since $L \in T_{k}$ it follows that $L \notin s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k-1}}\right\}$. Therefore it must be that $L$ was considered by the greedy algorithm after $L_{\tau_{k}}$. But the greedy algorithm considers candidates in order of increasing absolute value on $E_{\sigma}^{t}$. Thus we must have $\left|L_{\tau_{k}}\right| \leqq|L|, x \in E_{\sigma}$. This completes the proof of Lemma 4.1.

The next lemma provides a majorant for $P(x)$ involving only elements of $B_{\sigma}$.

Lemma 4.2. For each permutation $\sigma$ there is a constant $C_{1}$ (independent of $x$ and t) such that

$$
P(x) \leqq C_{1}\left|L_{\tau_{1}}\right|^{A_{1}} \ldots\left|L_{\tau_{r}}\right|^{4_{r}}, \quad x \in E_{\sigma}^{t},
$$

where

$$
A_{1}=d\left(P,\left\{L_{\tau_{1}}\right\}\right)-1
$$

and

$$
\Delta_{k}=d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)-d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}-1}\right\}\right)-1, \quad k=2, \ldots, r .
$$

Proof. We have

$$
P(x)=\prod_{k=1}^{r} F_{k}(x)
$$

where

$$
F_{k}(x)=\prod_{\left\{j: L_{j} \in T_{k}\right\}}\left|L_{j}\right|^{b_{j}}=\left(\prod_{\left\{: L_{j} \in T_{k} \backslash S\right\}}\left|L_{j}\right|^{b_{j}}\right)\left(\prod_{\left\{j: L_{j} \in T_{k} \cap S\right\}}\left|L_{j}\right|^{b_{j}}\right)
$$

Fix $k \leqq r$ and consider the two products on the right hand side. In the first product all of the exponents are non-negative because the $L_{j}$ 's do not belong to $S$. Therefore Lemma 4.1a implies that the first product is majorized on $E_{\sigma}^{t}$ by

$$
\prod_{\left\{: L_{j \in T_{k}} \backslash S\right\}} C_{\sigma}^{b_{j}}\left|L_{\tau_{k}}\right|^{b_{j}}
$$

In the second product all of the exponents are negative. Thus Lemma 4.1 b implies that the second product is majorized on $E_{\sigma}^{t}$ by

$$
\prod_{\left(j: L_{j} \in T_{k} \cap S\right\}}\left|L_{\tau_{k}}\right|^{b_{j}} .
$$

Combining these facts we conclude that there is a constant $C_{2}$ such that

$$
F_{k}(x) \leqq C_{2}\left|L_{\tau_{k}}\right|^{p_{k}}, \quad x \in E_{\sigma}^{t}, k \leqq r
$$

where

$$
p_{k}=\sum_{\left\{j: L_{j} \in T_{k}\right\}} b_{j}
$$

Lemma 4.2 will follow from this inequality if we show that $\Delta_{k}=p_{k}, k$ $=1, \ldots, r$. We have

$$
d\left(P,\left\{L_{\tau_{1}}\right\}\right)=1+\sum_{\left\{j: L_{j} \in s\left(L_{\tau_{2}}\right)\right\}} b_{j}=1+\sum_{\left\{j: L_{y} \in T_{1}\right\}} b_{j}=1+p_{1}
$$

Thus

$$
\Delta_{1}=d\left(P,\left\{L_{\tau_{1}}\right\}\right)-1=p_{1} .
$$

If $k \geqq 2$ then

$$
\begin{aligned}
d\left(P,\left(L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right) & =k+\sum_{\left\{j: L_{j \in s}\left(L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right\}} b_{j} \\
& =1+\left((k-1)+\sum_{\left\{: L_{j} \in s\left(L_{\tau_{1}}, \ldots, L_{\nu_{k}-1}\right)\right.} b_{j}\right)+\sum_{\left\{: L_{j} \in T_{k}\right\}} b_{j} \\
& =1+d\left(P\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}-1}\right\}\right)+p_{k} .
\end{aligned}
$$

Thus

$$
\Delta_{k}=d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)-d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k-1}}\right\}\right)-1=p_{k}
$$

This completes the proof of Lemma 4.2.
Lemma 4.3. Let $\varphi_{1}, \varphi_{2}, \ldots, \varphi_{n}$ be given real numbers. Then for all $t>0$

$$
\int_{\left|x_{1}\right| \leqq\left|x_{2}\right| \leqq \ldots \leqq\left|x_{n}\right| \leqq t}\left|x_{1}\right|^{\varphi_{1}}\left|x_{2}\right|^{\varphi_{2}} \ldots\left|x_{n}\right|^{\varphi_{n}} d x_{1} d x_{2} \ldots d x_{n}<\infty
$$

if $d_{k}=k+\sum_{j=1}^{k} \varphi_{j}>0$ for $k=1, \ldots, n$.
Proof. It clearly suffices to consider the case $t=1$. We proceed by induction on $n$. The lemma is obviously true for $n=1$. Now suppose that the lemma holds for $n-1$ and that we are given $\varphi_{1}, \ldots, \varphi_{n}$ satisfying the hypotheses of the lemma. Choose $\delta \geqq 0$ such that $d_{n}-\delta>0$ and $\varphi_{n}-\delta \neq-1$. (If $\varphi_{n} \neq-1$ we can take $\delta=0$ ). Then the above integral (with $t=1$ ) is less than

$$
\begin{aligned}
& \int_{\left|x_{1}\right| \leqq\left|x_{2}\right| \leqq \ldots \leqq\left|x_{n}\right| \leqq 1}\left|x_{1}\right|^{\varphi_{1}} \ldots\left|x_{n-1}\right|^{\varphi_{n-1} \mid}\left|x_{n}\right|^{\varphi_{n}-\delta} d x_{1} \ldots d x_{n} \\
& =\int_{\left|x_{1}\right| \leqq\left|x_{2}\right| \leqq \ldots \leqq\left|x_{n-1}\right|}\left|x_{1}\right|^{\mid \varphi_{1}} \ldots\left|x_{n-1}\right|^{\mid \varphi_{n-1}} \int_{\left|x_{n-1}\right| \leqq\left|x_{n}\right| \leqq 1}\left|x_{n}\right|^{\varphi_{n}-\delta} d x_{n} d x_{1} \ldots d x_{n-1}
\end{aligned}
$$

After evaluating the integral over $x_{n}$, we obtain

$$
\begin{aligned}
& \frac{2}{\varphi_{n}-\delta+1}\left\{\int_{\left|x_{1}\right| \leqq\left|x_{2}\right| \leqq \ldots \leqq\left|x_{n-1}\right| \leqq 1}\left|x_{1}\right|^{\varphi_{1}} \ldots\left|x_{n-1}\right|^{\varphi_{n-1}} d x_{1} \ldots d x_{n-1}\right. \\
& \quad-\int_{\left|x_{1}\right| \leqq\left|x_{2}\right| \leqq \ldots \leqq\left|x_{n-1}\right| \leqq 1}\left|x_{1}\right|^{\left.\varphi_{1} \ldots\left|x_{n-2}\right|^{\varphi_{n-2}}\left|x_{n-1}\right|^{\varphi_{n-1}+\varphi_{n}-\delta+1} d x_{1} \ldots d x_{n}\right\}}
\end{aligned}
$$

The induction hypothesis implies that the first integral in the braces is finite. To apply the induction hypothesis to the second integral, note that

$$
(n-1)+\varphi_{1}+\ldots+\varphi_{n-2}+\left(\varphi_{n-1}+\varphi_{n}-\delta+1\right)=n+\varphi_{1}+\ldots+\varphi_{n}-\delta=d_{n}-\delta>0
$$

Thus the second integral is finite, which completes the proof of Lemma 4.3.

Lemma 4.4. Let $\sigma$ be a permutation of $\{1, \ldots, m\}$ and let $I$ be the largest index such that $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{I}}\right\}$ is strongly independent. If

$$
d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)>0, \quad k=1, \ldots, I
$$

then

$$
\int_{E_{\Phi}^{t}} P(x) d x<\infty
$$

Proof. According to Lemma 4.2 it suffices to show

$$
\int_{E_{\sigma}}\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{r}}\right|^{\Lambda_{r}} d x<\infty,
$$

where $\Delta_{1}, \ldots, \Delta_{r}$ are as defined in Lemma 4.2,

Case 1. $I=r$. Let $C_{3}=\max \left\{\left|L_{j}(x)\right|: x \in U_{t}, 1 \leqq j \leqq m\right\}$. The last integral is majorized by

$$
C_{4} \int_{\left|y_{1}\right| \leqq\left|y_{2}\right| \leqq \ldots \leqq\left|y_{r}\right| \leqq C_{3}}\left|y_{1}\right|^{\Lambda_{1}} \ldots\left|y_{r}\right|^{4_{r}} d y_{1} \ldots d y_{r},
$$

where $C_{4}$ is a constant obtained by integrating over $n-r$ extraneous variables. Note that $\Delta_{1}, \ldots, \Delta_{r}$ satisfy

$$
k+\sum_{i=1}^{k} \Lambda_{i}=d\left(P,\left\{L_{\tau 1}, \ldots, L_{\tau_{k}}\right\}\right)>0, \quad k=1, \ldots, r
$$

Hence Lemma 4.4 implies the conclusion in this case.
Case 2. $I<r$. In this case there are constants $a_{1}, \ldots, a_{I}$ so that

$$
M_{\tau_{I+1}}=a_{1} M_{\tau_{1}}+\ldots+a_{I} M_{\tau_{I}}
$$

Then

$$
\begin{aligned}
L_{\tau_{I+1}} & =M_{\tau_{I+1}}+\theta_{\tau_{I+1}} \\
& =a_{1} M_{\tau_{1}}+\ldots+a_{I} M_{\tau_{I}}+\theta_{\tau_{I+1}} \\
& =a_{1} L_{\tau_{1}}+\ldots+a_{I} L_{\tau_{I}}+w,
\end{aligned}
$$

 $L_{\tau_{1}}, \ldots, L_{\tau_{I}}$, if follows that $w \neq 0$. Thus we can choose a constant $\lambda$ so that

$$
\left|L_{\tau_{1+1}}\right| \leqq\left|\frac{w}{2}\right| \quad \text { whenever }\left|L_{\tau_{1}}\right| \leqq \ldots \leqq\left|L_{\tau_{\tau}}\right| \leqq \lambda
$$

Since $\left|L_{\tau_{I}}\right| \leqq\left|L_{\tau_{I+1}}\right| \leqq \ldots \leqq\left|L_{\tau_{r}}\right|$ for $x \in E_{\sigma}^{t}$, there is a constant $C_{5}$ depending on $\lambda$ and $w$ so that

$$
\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{r}}\right|^{4_{r}} \leqq C_{5}\left|L_{\tau_{1}}\right|^{\mid \Lambda_{1}} \ldots\left|L_{\tau_{I}}\right|^{\Lambda_{I}} \quad \text { if } x \in E_{\sigma}^{t}, \quad\left|L_{\tau_{I}}\right| \leqq \lambda
$$

and also a constant $C_{6}$ depending on $\lambda$ so that

$$
\left|L_{\tau_{1}}\right|^{\Delta_{1}} \ldots\left|L_{\tau_{r}}\right|^{A_{r}} \leqq C_{6}\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{I}-1}\right|^{\Lambda_{I-1}} \quad \text { if } x \in E_{\sigma}^{l}, \quad\left|L_{\tau_{I}}\right|>\lambda .
$$

Since $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{1}}\right\}$ and $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{I-1}}\right\}$ are strongly independent, the proof can be completed as in Case 1.

Proof of Theorem 3.1. Suppose that the conditions of Theorem 3.1 hold. Let $\sigma$ be a permutation of $\{1, \ldots, m\}$ and define $I$ as in Lemma 4.4. The remark following Theorem 3.1 implies that $d\left(P,\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)>0, k=1, \ldots, I$. Thus we can use Lemma 4.4 to conclude that $\int_{E_{\sigma}^{\varepsilon}} P(x) d x<\infty$. Theorem 3.1 follows because $U_{t}$ is the union over $\sigma$ of the sets $\stackrel{E_{\sigma}^{t}}{E_{\sigma}^{t}}$.

## 5. Counting Powers

This section is devoted to "counting powers" in the function $p_{\eta}: \mathbb{R}^{2 p_{p}} \rightarrow \mathbb{R}$ given by

$$
\begin{aligned}
p_{\eta}(x)= & \left|x_{2}+\ldots+x_{2 p}\right|^{\eta-1}\left|x_{2}\right|^{n-1}\left|x_{3}\right|^{n-1} \ldots\left|x_{2 p}\right|^{\eta-1}\left|x_{1}\right|^{-\alpha}\left|x_{1}+x_{2}\right|^{-\beta}\left|x_{1}+x_{2}+x_{3}\right|^{-\alpha} \\
& \cdot\left|x_{1}+x_{2}+x_{3}+x_{4}\right|^{-\beta} \ldots\left|x_{1}+\ldots+x_{2 p-1}\right|^{-\alpha}\left|x_{1}+\ldots+x_{2 p}\right|^{-\beta},
\end{aligned}
$$

where $\alpha<1, \beta<1$ and $0<\eta<1$. The results are stated in Propositions 5.1 and 5.2. Introduce the set of linear functionals on $\mathbb{R}^{2 p}$

$$
T=\left\{x_{2}+\ldots+x_{2 p}, x_{2}, x_{3}, \ldots, x_{2 p}, x_{1}, x_{1}+x_{2}, \ldots, x_{1}+\ldots+x_{2 p}\right\}
$$

For each $W \subset T$ we define the set $s\{W\}$ and the quantity $d\left(P_{\eta}, W\right)$ as in Sect. 3 and we say that $W$ is an independent set if it is strongly independent. (Here $W$ does not involve additive constants.)
Proposition 5.1. Let $\alpha<1, \beta<1$ and let $\eta$ satisfy $0<\eta<1$ and $n>(a+\beta) / 2$. If $W$ $\subset T$ is an independent set such that $|W|=2 p-1$ and $W \subset\left\{x_{2}+\ldots\right.$ $\left.+x_{2 p}, x_{2}, x_{3}, \ldots, x_{2 p}\right\}$, then $d\left(P_{n}, W\right)=2 p \eta-1$.
Proof. It is clear that if $W$ satisfies the conditions of Proposition 5.1 then $s\{W\}$ $=\left\{x_{2}+\ldots+x_{2 p}, x_{2}, x_{3}, \ldots, x_{2 p}\right\}$. Therefore

$$
d\left(P_{\eta}, W\right)=(2 p-1)+2 p(\eta-1)=2 p \eta-1 .
$$

Proposition 5.2. Let $\alpha<1, \beta<1$ and let $\eta$ satisfy $0<\eta<1$ and $n>(\alpha+\beta) / 2$. If $W$ $\subset T$ is an independent set such that either $|W| \neq 2 p-1$ or $W \nsubseteq\left\{x_{2}+\ldots\right.$ $\left.+x_{2_{p}}, x_{2}, x_{3}, \ldots, x_{2 p}\right\}$, then $d\left(P_{\eta}, W\right)>0$.

The rest of this section is devoted to the proof of Proposition 5.2.
In proving that proposition we can restrict ourselves to considering sets $W$ $\subset T$ which do not contain $x_{2}+\ldots+x_{2 p}$. To see this, assume that $x_{2}+\ldots$ $+x_{2 p} \in W$. Suppose first that the set $s\{W\} \backslash s\left\{W \backslash x_{2}+\ldots+x_{2 p}\right\}$ contains some functional $L$ other than $x_{2}+\ldots+x_{2 p}$. Then we consider the set $W^{\prime}$ which is $W$ with $x_{2}+\ldots+x_{2 p}$ replaced by $L$, that is $W^{\prime}=W \cup\{L\} \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$. Clearly, $x_{2}+\ldots+x_{2 p} \notin W^{\prime}$. Furthermore, $W^{\prime}$ has the same span and cardinality as $W$. Therefore $d\left(P_{\eta}, W^{\prime}\right)=d\left(P_{\eta}, W\right)$. On the other hand, suppose that there is no such $L$. In this case we put $W^{\prime}=W \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$. We have $\left|W^{\prime}\right|=|W|-1$ and $s\left\{W^{\prime}\right\}=s\{W\} \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$. Hence

$$
d\left(P_{\eta}, W^{\prime}\right)=d\left(P_{\eta}, W\right)-1-(\eta-1)=d\left(P_{\eta}, W\right)-\eta<d\left(P_{\eta}, W\right)
$$

Thus in either case there is a set $W^{\prime}$ which does not contain $x_{2}+\ldots+x_{2 p}$ and satisfies $d\left(P_{\eta}, W^{\prime}\right) \leqq d\left(P_{\eta}, W\right)$. Hence we can assume that $W$ does not contain $x_{2}$ $+\ldots+x_{2 p}$.

In proving Proposition 5.2 we can also restrict ourselves to sets $W \subset T$ which satisfy $\left\{x_{k}, x_{1}+\ldots+x_{k}\right\} \notin W, k=2, \ldots, 2 p$. For suppose that $T$ does not satisfy this restriction. Let $j$ be the largest $k$ for which $\left\{x_{k}, x_{1}+\ldots+x_{k}\right\} \subset W$. Let $W^{\prime}=W \cup\left\{x_{1}+\ldots+x_{j-1}\right\} \backslash\left\{x_{1}+\ldots+x_{j}\right\}$. Since the sets $\left\{x_{j}, x_{1}+\ldots+x_{j-1}\right\}$ and $\left\{x_{j}, x_{1}+\ldots+x_{j}\right\}$ have the same span and cardinality, it follows that $d\left(P_{n}, W^{\prime}\right)$ $=d\left(P_{\eta}, W\right)$. It is clear that the largest value of $k$ for which $\left\{x_{k}, x_{1}+\ldots+x_{k}\right\}$ $\subset W^{\prime}$ is at most $j-1$. After repeating this process at most $j-2$ more times we obtain a set $W^{\prime \prime}$ satisfying $d\left(P_{\eta}, W^{\prime \prime}\right)=d\left(P_{\eta}, W\right)$ and $\left\{x_{k}, x_{1}+\ldots+x_{k}\right\} \nleftarrow W^{\prime}, k$ $=2, \ldots, 2 p$. Thus we can restrict ourselves to sets $W$ which do not contain both $x_{k}$ and $x_{1}+\ldots+x_{k}$.

We will assume from now on that $W \subset T$ satisfies both of the above restrictions. To describe the sets $W$ which we will be considering, it is helpful to think of the elements of $T \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$ arranged in columns as follows:

$$
x_{1}\left|\begin{array}{c|c}
x_{2} & x_{3} \\
x_{1}+x_{2} & x_{1}+x_{2}+x_{3}
\end{array}\right| \ldots\left|\begin{array}{c}
x_{2 p} \\
x_{1}+\ldots+x_{2 p}
\end{array}\right|
$$

In the rest of this section we consider sets $W$ which contain at most one element from each column. For any set $T^{\prime} \subset T$ we say that $T^{\prime}$ contains the $k^{\text {th }}$ column if $x_{k} \in T^{\prime}$ or $x_{1}+\ldots+x_{k} \in T^{\prime}$.

The proof of Proposition 5.2 involves three lemmas.
Lemma 5.3. Suppose that $W$ does not contain the $k^{\text {th }}$ column. Then $s\{W\}$ does not contain the $k^{\text {th }}$ column.
Proof. We prove that neither $x_{k}$ nor $x_{1}+\ldots+x_{k}$ is in $s\{W\}$. We distinguish two cases.

Case I. There is no $j>k$ such that $x_{1}+\ldots+x_{j} \in W$. In this case the conclusion of the lemma is clear since no element of $W$ contains the summand $x_{k}$.
Case II. There exists $j>k$ such that $x_{1}+\ldots+x_{j} \in W$.
Suppose that $j$ is the smallest index with this property. Then the only elements of $W$ which contain the summand $x_{k}$ are among $\left\{x_{1}+\ldots+x_{j}, x_{1}+\ldots\right.$ $\left.+x_{j+1}, \ldots, x_{1}+\ldots+x_{2 p}\right\}$. Since $x_{j} \notin W$ these are also the only elements of $W$ which contain the summand $x_{j}$. Thus in any linear combination of the elements of $W$ the summands $x_{k}$ and $x_{j}$ appear with the same coefficient. Hence neither $x_{k}$ nor $x_{1}+\ldots+x_{k}$ can be linear combinations of elements of $W$. This completes the proof of Lemma 5.3.

We now partition $W$ into blocks of contiguous columns. Any two blocks are separated by at least one column not in $W$. Formally, we will say that a set $B \subset W$ is a block of columns, if there exist $l_{B}<r_{B}$ such that

1) $W$ contains neither column $l_{B}-1$ nor column $r_{B}+1$.
2) $B$ contains column $l_{B}$ through $r_{B}$ and no other columns.

With this definition we obtain a partition $W=\bigcup_{j=1}^{n} B_{j}$, where each $B_{j}$ is a block of columns. We will assume that $B_{j}$ is to the left of $B_{j+1}$ for each $j$.

Define the function $Q_{\eta}(x)=P_{n}(x) \cdot\left|x_{2}+\ldots+x_{2 p}\right|^{1-\eta}$. It is clear that

$$
d\left(P_{n}, W\right)= \begin{cases}d\left(Q_{n}, W\right), & \text { if } x_{2}+\ldots+x_{2 p} \notin s\{W\} \\ n-1+d\left(Q_{\eta}, W\right) & \text { if } x_{2}+\ldots+x_{2 p} \in S\{W\} .\end{cases}
$$

Furthermore Lemma 5.3 implies that $d\left(Q_{\eta}, W\right)=\sum_{j=1}^{n} d\left(Q_{\eta}, B_{j}\right)$. Thus we have

$$
d\left(P_{\eta}, W\right)= \begin{cases}\sum_{j=1}^{n} d\left(Q_{n}, B_{j}\right) & \text { if } x_{2}+\ldots+x_{2 p} \notin s(W)  \tag{5.1}\\ n-1+\sum_{j=1}^{n} d\left(Q_{\eta}, B_{j}\right) & \text { if } x_{2}+\ldots+x_{2 p} \in s(W)\end{cases}
$$

The next lemma is useful in determining the quantities $d\left(Q_{\eta}, B_{j}\right)$. A block of columns will be called nonsimple if it contains $x_{1}+\ldots+x_{k}$ for some $k \geqq 1$.
Lemma 5.4. Let $B$ be a nonsimple block of columns. Put $l=l_{B}$ and $r=r_{B}$. Let $m$ be the smallest $k$ satisfying $x_{1}+\ldots+x_{k} \in B$.

1) If $l \leqq j<m$, then $x_{j} \in s(B)$ and $x_{1}+\ldots+x_{j} \notin s(B)$.
2) $x_{m} \notin s(B)$ and $x_{1}+\ldots+x_{m} \in s(B)$.
3) If $m<j \leqq r$, then $x_{j} \in s(B)$ and $x_{1}+\ldots+x_{j} \in s(B)$.

Proof. 1) Let $l \leqq j<m$. Since $j<m$ we have $\left\{x_{l}, x_{l+1}, \ldots, x_{j}\right\} \subset B$. Suppose that $x_{1}$ $+\ldots+x_{j} \in s(B)$. The identity $x_{1}+\ldots+x_{l-1}=\left(x_{1}+\ldots+x_{j}\right)-x_{l}-x_{l+1}-\ldots-x_{j}$ implies that $x_{1}+\ldots+x_{l-1} \in s(B)$. This contradicts Lemma 5.3. We conclude that $x_{1}+\ldots+x_{j} \notin s(B)$.
2) The definition of $m$ implies that $x_{1}+\ldots+x_{m} \in B$. Suppose that $x_{m} \in S(B)$. We have

$$
x_{1}+\ldots+x_{l-1}=\left(x_{1}+\ldots+x_{m}\right)-x_{l}-x_{l+1}-\ldots-x_{m}
$$

again contradicting Lemma 5.3.
3) This is proven by induction. It is clear that if $x_{1}+\ldots+x_{j} \in s(B)$ and $B$ contains column $j+1$, then $\left\{x_{j+1}, x_{1}+\ldots+x_{j+1}\right\} \subset s(B)$. To start the induction off, note that $x_{1}+\ldots+x_{m} \in s(B)$ and $B$ contains column $m+1$. This completes the proof of Lemma 5.4.

If $B$ is a simple block of columns, then $B \subset\left\{x_{2}, x_{3}, \ldots, x_{2 p}\right\}$ and therefore

$$
\begin{equation*}
d\left(Q_{\eta}, B\right)=|B|+|B|(\eta-1)=|B| \eta>0 \tag{5.2}
\end{equation*}
$$

To determine $d\left(Q_{\eta}, B\right)$ for a nonsimple block, we need to take into account the parities of the integers $m$ and $r$ introduced in the statement of Lemma 5.4. This is done in the next lemma. First define

$$
\gamma_{1}=(m-l) \eta+(r-m)\left[\eta-\frac{(\alpha+\beta)}{2}\right]
$$

and

$$
\gamma_{2}=(m-l) \eta+(r-m+1)\left[\eta-\frac{(\alpha+\beta)}{2}\right]
$$

Note that under the conditions of Proposition 5.2 we have $\gamma_{1} \geqq 0$ and $\gamma_{2}>0$.
Lemma 5.5. Suppose that the conditions of Proposition 5.2 hold. Let $B$ be a nonsimple block of columns.

1) If $m$ and $r$ are both odd, then

$$
d\left(Q_{n}, B\right)=(1-\alpha)+\gamma_{1} \geqq 1-\alpha>0
$$

2) If $m$ and $r$ are both even, then

$$
d\left(Q_{\eta}, B\right)=(1-\beta)+\gamma_{1} \geqq 1-\beta>0
$$

3) If $m$ and $r$ have different parities, then

$$
d\left(Q_{\eta}, B\right)=(1-\eta)+\gamma_{2}>1-\eta>0
$$

Proof. Note that $d\left(Q_{\eta}, B\right)$ is equal to the cardinality of $B$ plus the sum of the powers of all the elements of $s(B) \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$. The cardinality of $B$ contributes $(r-l)+1$ to $d\left(Q_{\eta}, B\right)$.

According to Lemma 5.4, the set $s(B) \backslash\left\{x_{2}+\ldots+x_{2 p}\right\}$ is equal to $W_{1} \cup W_{2}$, where

$$
W_{1}=\left\{x_{l}, x_{l+1}, \ldots, x_{m-1}, x_{m+1}, x_{m+2}, \ldots, x_{r}\right\}
$$

and

$$
W_{2}=\left\{x_{1}+\ldots+x_{m}, x_{1}+\ldots+x_{m+1}, \ldots, x_{1}+\ldots+x_{r}\right\} .
$$

(When $m=1$ we let $W_{1}=\left\{x_{2}, \ldots, x_{r}\right\}$.)
Counting the powers associated with $W_{1}$ we obtain a contribution $(r-l)(\eta$ $-1)=-(r-l)+(m-l) \eta+(r-m) \eta$.

Counting the powers associated with $W_{2}$ we obtain a contribution

$$
\begin{array}{ll}
-\alpha-\frac{(r-m)}{2}(\alpha+\beta) & \text { if } m, r \text { are both odd } \\
-\beta-\frac{(r-m)}{2}(\alpha+\beta) & \text { if } m, r \text { are both even } \\
-\frac{(r-m+1)}{2}(\alpha+\beta) & \text { if } m, r \text { have different parities. }
\end{array}
$$

Summing the appropriate contributions and using the inequalities $\alpha<1, \beta<1$, $\gamma_{1} \geqq 0$ and $\gamma_{2} \geqq 0$ we obtain the results of Lemma 5.5.
Proof of Proposition 5.2. Suppose that the conditions of Proposition 5.2 hold and that the independent subset $W$ of $T$ also satisfies the restrictions described above. (Namely, $W$ does not contain $x_{2}+\ldots+x_{2 p}$ and $\left\{x_{k}, x_{1}+\ldots+x_{k}\right\} \notin W, k$ $=2, \ldots, 2 p$.) Relation (5.1), relation (5.2) and Lemma 5.5 imply that $d\left(P_{\eta}, W\right)>0$ if $x_{2}+\ldots+x_{2 p} \notin s(W)$. To complete the proof, assume that $x_{2}+\ldots+x_{2 p} \in s(W)$. This implies that $r_{B_{n}}=2 p$ (where $B_{n}$ is the rightmost block of $W$ ), because the summand $x_{2_{p}}$ appears only in the $2 p^{\text {th }}$ column.

First we will show that $B_{n}$ is nonsimple, that is, it contains $x_{1}+\ldots+x_{k}$ for some $k \geqq 1$. Put $l=l_{B_{n}}$. Put $l=1$, then $x_{1} \in B_{n}$ and so $B_{n}$ is nonsimple. If $l=2$ and $B_{n}$ is simple, then $W=B_{n}=\left\{x_{2}, \ldots, x_{2 p}\right\}$, contradicting the assumptions of the proposition. If $l>2$ and $B_{n}$ is simple, then no element of $W$ contains the summand $x_{1-1}$, contradicting the assumption that $x_{2}+\ldots+x_{2 p} \in S(W)$. Thus $B_{n}$ must be nonsimple.

Next we will show that $l_{B_{1}}=1$. Since $B_{n}$ is nonsimple, Lemma 5.4 shows that $x_{1}+\ldots+x_{2 p} \in s(W)$. Since we have assumed that $x_{2}+\ldots+x_{2 p} \in s(W)$, it follows that $x_{1} \in S(W)$. Thus we must have $l_{B_{1}}=1$ in order to avoid contradicting Lemma 5.3.

To complete the proof, we distinguish two cases, according to whether $W$ consists of a single block or more than one block.
Case I. $n=1$. In this case we have only one block $B_{1}$ satisfying $l_{B_{1}}=m_{B_{1}}=1$ and $r_{B_{1}}=2 p$. Lemma 5.5 implies that $d\left(Q_{\eta}, B_{1}\right)=1-\eta+\gamma_{2}$. According to (5.1),

$$
d\left(P_{\eta}, W\right)=d\left(Q_{\eta}, B_{1}\right)+\eta-1=\gamma_{2}>0 .
$$

Case II. $n>1$. We again have $l_{B_{1}}=m_{B_{1}}=1$. Thus either Part 1 or Part 3 of Lemma 5.5 applies. Hence $d\left(Q_{\eta}, B_{1}\right) \geqq 1-\alpha$ or $d\left(Q_{\eta}, B_{1}\right) \geqq 1-\eta$.

Since $r_{B_{n}}=2 p$ and $B_{n}$ is nonsingular, either Part 2 or Part 3 of Lemma 5.5 applies to $B_{n}$. Thus $d\left(Q_{\eta}, B_{n}\right) \geqq 1-\beta$ or $d\left(Q_{\eta}, B_{n}\right) \geqq 1-\eta$.

The proof can now be completed as follows. According to (5.2) and Lemma 5.5, we have $d\left(Q_{\eta}, B_{j}\right)>0, j=1, \ldots, n$. Thus by (5.1),

$$
\begin{aligned}
d\left(P_{\eta}, W\right) & =\eta-1+\sum_{j=1}^{n} d\left(Q_{\eta}, B_{j}\right) \\
& \geqq n-1+d\left(Q_{\eta}, B_{1}\right)+d\left(Q_{\eta}, B_{n}\right) .
\end{aligned}
$$

If $d\left(Q_{\eta}, B_{1}\right) \geqq 1-\eta$, then $d\left(P_{\eta}, W\right) \geqq d\left(Q_{\eta}, B_{n}\right)>0$. Similarly, $d\left(P_{\eta}, W\right)>0 \quad$ if $d\left(Q_{\eta}, B_{n}\right) \geqq 1-\eta$. Therefore we can assume that $d\left(Q_{\eta}, B_{1}\right) \geqq 1-\alpha$ and $d\left(Q_{\eta}, B_{n}\right) \geqq 1$ $-\beta$. Then $d\left(P_{\eta}, W\right) \geqq \eta-1+(1-\alpha)+(1-\beta)=1-\eta+2[\eta-(\alpha+\beta) / 2]>1-\eta>0$. This completes the proof of Proposition 5.2.

## 6. Applications of Power Counting

In this section, we establish Propositions 6.1 and 6.2 , which will be used in the proof of Theorem 1.

For each integer $N \geqq 1$ define the function

$$
h_{N}(z)= \begin{cases}\min \left(\frac{1}{|z+2 \pi|}, N\right) & -2 \pi \leqq z \leqq-\pi \\ \min \left(\frac{1}{|z|}, N\right) & -\pi \leqq z \leqq \pi \\ \min \left(\frac{1}{|z-2 \pi|}, N\right) & \pi \leqq z \leqq 2 \pi\end{cases}
$$

and the function $f_{N}: \mathbb{R}^{2 p} \rightarrow \mathbb{R}$ by

$$
\begin{aligned}
f_{N}(y)= & h_{N}\left(y_{1}-y_{2 p}\right) h_{N}\left(y_{2}-y_{1}\right) h_{N}\left(y_{3}-y_{2}\right) \ldots h_{N}\left(y_{2 p}-y_{2 p-1}\right) \\
& \cdot\left|y_{1}\right|^{-\alpha}\left|y_{2}\right|^{-\beta}\left|y_{3}\right|^{-\alpha} \ldots\left|y_{2 p}\right|^{-\beta}
\end{aligned}
$$

where $\alpha<1$ and $\beta<1$. Given $t>0$ put $U_{t}=[-t, t]^{2 p}$ and $V$ $=\left\{y \in \mathbb{R}^{2 p}:\left|y_{1}\right| \leqq \frac{1}{2}\left|y_{2}\right|\right\}$.

The following results are useful in studying the behavior of $\int_{U_{\pi}} f_{N}(y) d y$ as $N \rightarrow \infty$.

Proposition 6.1. Let $\alpha<1$ and $\beta<1$.
a) If $\alpha+\beta \leqq 0$, then as $N \rightarrow \infty$,

$$
\int_{U_{\pi \cap V}} f_{N}(y) d y=O\left(N^{\varepsilon}\right)
$$

for every $\varepsilon>0$.
b) If $\alpha+\beta>0$, then as $N \rightarrow \infty$,

$$
\int_{U_{\pi} \cap V} f_{N}(y) d y=O\left(N^{p(\alpha+\beta)+\varepsilon}\right)
$$

for every $\varepsilon>0$.

Proposition 6.2. Let $\alpha<1$ and $\beta<1$.
a) If $p(\alpha+\beta)<1$, then

$$
\lim _{t \rightarrow 0} \limsup _{N \rightarrow \infty} \frac{\int_{U_{t}} f_{N}(y) d y}{N}=0
$$

b) If $p(\alpha+\beta) \geqq 1$, then as $N \rightarrow \infty$

$$
\int_{U_{\pi c}} f_{N}(y) d y=O\left(N^{p(\alpha+\beta)+\varepsilon}\right)
$$

for every $\varepsilon>0$.
In order to prove Propositions 6.1 and 6.2 we need to put the problem into the framework described in Sect. 4. Choose $\eta$ satisfying $0<\eta<1$. If $1 / N \leqq|z| \leqq \pi$ then we have

$$
h_{N}(z)=\frac{1}{|z|} \leqq \frac{1}{|z|} N^{n}|z|^{n}=N^{n}|z|^{n-1} .
$$

If $|z|<1 / N$ then

$$
h_{N}(z)=N^{\eta} N^{1-\eta} \leqq N^{\eta}|z|^{\eta-1}
$$

Thus

$$
h_{N}(z) \leqq N^{n} \mid z^{\mid n-1}, \quad-\pi \leqq z \leqq \pi, 0<\eta<1 .
$$

This implies $f_{N}(y) \leqq f_{N, \eta}(y)$, where $f_{N, \eta}(y)$ is defined as $f_{N}(y)$ with $h_{N}$ replaced by

$$
h_{N, \eta}(z)= \begin{cases}N^{\eta}|z+2 \pi|^{\eta-1} & -2 \pi \leqq z \leqq-\pi \\ N^{\eta}|z|^{\eta-1} & -\pi \leqq z \leqq \pi \\ N^{\eta}|z-2 \pi|^{\eta-1} & \pi \leqq z \leqq 2 \pi\end{cases}
$$

To study $\int_{U_{t}} f_{N, n}(y) d y$ we make the change of variable $x_{1}=y_{1}, x_{k}=y_{k}-y_{k-1}$, $k=2, \ldots, 2 p$. Thus we define

$$
\begin{aligned}
f_{N}^{\prime}(x)= & h_{N}\left(x_{2}+\ldots+x_{2 p}\right) h_{N}\left(x_{2}\right) \ldots h_{N}\left(x_{2 p}\right) \\
& \cdot\left|x_{1}\right|^{-\alpha}\left|x_{1}+x_{2}\right|^{-\beta}\left|x_{1}+x_{2}+x_{3}\right|^{-\alpha} \ldots\left|x_{1}+\ldots+x_{2 p}\right|^{-\beta}
\end{aligned}
$$

and $f_{N, n}^{\prime}(x)$ in the same way, with $h_{N}$ replaced by $h_{N, n}$. Define the set $U_{t}^{\prime}$ so that $\int_{U_{i}} f_{N}^{\prime}(x) d x=\int_{U_{t}} f_{N}(y) d y$ and let $V^{\prime}=\left\{x:\left|x_{1}\right| \leqq \frac{1}{2}\left|x_{1}+x_{2}\right|\right\}$.

Note that if $y \in U_{\pi}$ and $\left|y_{k}-y_{k-1}\right| \geqq \frac{3 \pi}{2}$, then $\left|y_{k-1}\right| \geqq \pi / 2$ and $\left|y_{k}\right| \geqq \pi / 2$. Hence for $x \in U_{\pi}^{\prime}$

$$
\begin{align*}
\left|x_{1}+\ldots+x_{k-1}\right| \geqq \frac{\pi}{2} & \text { if }\left|x_{k}+2 \pi\right| \leqq \frac{\pi}{2}  \tag{6.1a}\\
\left|x_{1}+\ldots+x_{k-1}\right| \geqq \frac{\pi}{2} & \text { if }\left|x_{k}-2 \pi\right| \leqq \frac{\pi}{2}  \tag{6.1~b}\\
\left|x_{1}+\ldots+x_{k}\right| \geqq \frac{\pi}{2} & \text { if }\left|x_{k}+2 \pi\right| \leqq \frac{\pi}{2} \tag{6.2a}
\end{align*}
$$

and

$$
\begin{equation*}
\left|x_{1}+\ldots+x_{k}\right| \geqq \frac{\pi}{2} \quad \text { if }\left|x_{k}-2 \pi\right| \leqq \frac{\pi}{2} \tag{6.2b}
\end{equation*}
$$

It is clear that if $y \in U_{\pi} \cap V$ then $\left|y_{2}-y_{1}\right| \leqq \frac{3}{2}\left|y_{2}\right| \leqq 3 \pi / 2$. Thus

$$
\begin{equation*}
\left|x_{2}\right| \leqq \frac{3 \pi}{2} ; \quad x \in U_{\pi}^{\prime} \cap V^{\prime} . \tag{6.3}
\end{equation*}
$$

In order to apply the lemmas of Sect. 4 introduce the functionals

$$
\begin{aligned}
M_{1}(x) & =x_{2}+\ldots+x_{2 p} \\
M_{2}(x) & =x_{2} \\
& \vdots \\
M_{2 p}(x) & =x_{2 p} \\
M_{2 p+1}(x) & =x_{1} \\
M_{2 p+2}(x) & =x_{1}+x_{2} \\
& \vdots \\
M_{4 p}(x) & =x_{1}+\ldots+x_{2 p} .
\end{aligned}
$$

Choose $\left\{\theta_{1}, \ldots, \theta_{2 p}\right\} \subset\{-2 \pi, 0,2 \pi\}$ and put $\theta_{2 p+1}=\theta_{2 p+2}=\ldots=\theta_{4 p}=0$. Define $L_{j}(x)=M_{j}(x)+\theta_{j}, j=1, \ldots, 4 p$. Let $U_{t}^{\prime \prime}$ be the subset of $U_{t}^{\prime}$ on which $f_{N, \eta}^{\prime}(x)$ $=N^{2 p \eta} P_{n}^{\prime}(x)$, where

$$
\begin{aligned}
P_{\eta}^{\prime}(x)= & \left|L_{1}\right|^{n-1}\left|L_{2}\right|^{\mid-1} \ldots\left|L_{2 p}\right|^{\eta-1} \\
& \cdot\left|L_{2 p+1}\right|^{-\alpha}\left|L_{2 p+2}\right|^{-\beta} \ldots\left|L_{4 p}\right|^{-\beta} .
\end{aligned}
$$

In the proofs of Propositions 6.1 and 6.2 we use the notation introduced in Sect. 3. For example $T=\left\{L_{1}, \ldots, L_{4 p}\right\}$ and $r$ is the rank of $T$.

Fixing a permutation $\sigma=\left(\sigma_{1}, \ldots, \sigma_{4 p}\right)$ of $\{1, \ldots, 4 p\}$ we define

$$
E_{\sigma}^{t}=\left\{x \in U_{t}^{\prime \prime}:\left|L_{\sigma_{1}}(x)\right| \leqq \ldots \leqq\left|L_{\sigma_{4 p}}(x)\right|\right\}
$$

and, as in Sect. 4 construct a basis $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{r}}\right\}$ for $T$ satisfying

$$
\left|L_{\tau_{1}}\right| \leqq\left|L_{\tau_{2}}\right| \leqq \ldots \leqq\left|L_{\tau_{r}}\right|, \quad x \in E_{\sigma}^{t}
$$

In proving Propositions 6.1 and 6.2 it suffices to show that the conclusions hold with $U_{t}$ replaced by $E_{\sigma}^{t}, V$ replaced by $V^{\prime}$ and $f_{N}(y)$ replaced by $f_{N}^{\prime}(x)$ or $f_{N, \eta}^{\prime}(x)$.
Proof of Proposition 6.1. Fix $\eta$ satisfying $0<\eta<1$ and $\eta>(\alpha+\beta) / 2$. Since $f_{N, \eta}^{\prime}(x)$ $=N^{2 p \eta} P_{\eta}^{\prime}(x)$, both parts of Proposition 6.1 will follow if we show

$$
\begin{equation*}
\int_{E_{\sigma}^{\pi} \cap V^{\prime}} P_{\eta}^{\prime}(x) d x<\infty . \tag{6.4}
\end{equation*}
$$

To show (6.4) we distinguish two cases.
Case I. $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\} \notin\left\{L_{1}, \ldots, L_{2 p}\right\}$.
In this case we will show that in fact $\int_{E_{\sigma}^{\pi}} P_{\eta}^{\prime}(x) d x<\infty$.
Subcase I.1. $\theta_{1}=\ldots=\theta_{2 p}=0$.
In this case $P_{\eta}^{\prime}(x)=P_{\eta}(x)$, where $P_{\eta}(x)$ is as defined in Sect. 5. Thus Proposition 5.2 and Lemma 4.4 imply $\int_{E_{\vec{\sigma}}} P_{\eta}^{\prime}(x) d x<\infty$.

Subcase I.2. $\theta_{j} \neq 0$ for some $j$.
Here some of the dimensions may be non-positive, but relations (6.1)-(6.3) will allow us to deal with that situation. If $d\left(P_{\eta}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)>0, k=1, \ldots, 2 p$, then Lemma 4.4 implies that $\int_{E_{\sigma}^{\pi}} P_{\eta}^{\prime}(x) d x<\infty$. Otherwise put $W=\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}$, where $k$ is the smallest index satisfying $d\left(P_{n}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right) \leqq 0$. As in Sect. 5, $s(W)$ consists of a collection of blocks of columns, plus possibly $L_{1}$. Let $B$ be a block whose contribution to $d\left(P_{\eta}^{\prime}, W\right)$ is non-positive. Then it is clear that $\theta_{j} \neq 0$ for some $j$ satisfying $l_{B} \leqq j \leqq r_{B}$ and that $x_{1}+\ldots+x_{m} \in s(B)$ for some $m$ satisfying $l_{B} \leqq m \leqq r_{B}$. (For $d\left(P_{\eta}^{\prime}, B\right)$ would be positive if there were no such $j$ (Lemma 5.5) or if there were no such $m$ (direct computation).)

Let $m$ be the smallest index with this property. We distinguish two cases, in both of which we will show that $P_{\eta}^{\prime}(x)$ is at most a constant times $\left|L_{\tau_{1}}\right|^{4_{i}} \ldots\left|L_{\tau_{k-1}}\right|^{\Delta_{k-1}}$, with $\Delta_{i}$ defined as in Lemma 4.2.
Subcase 1.2.i. Some $j$ satisfies $m<j \leqq r_{B}, L_{j} \in S(B)$ and $\theta_{j} \neq 0$.
Let $j$ be the smallest index with this property. Then it is clear that $x_{1}+\ldots$ $+x_{j-1} \in s(B)$. By Lemma 4.2, $P_{\eta}^{\prime}(x)$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau r}\right|^{\Lambda_{r}}$. Since $L_{j}$ and $x_{1}+\ldots+x_{j-1}$ are in $s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}$ and $\left|L_{\tau_{1}}\right| \leqq \ldots \leqq\left|L_{\tau_{r}}\right|$ on $E_{\sigma}^{\pi}$, Lemma 4.1 implies that there is a constant $C$ such that if $x \in E_{\sigma}^{\pi}$

$$
C\left|x_{1}+\ldots+x_{j-1}\right| \leqq\left|L_{\tau_{k}}\right|
$$

and

$$
C\left|L_{j}\right| \leqq\left|L_{\tau_{k}}\right|
$$

Thus if $x \in E_{\sigma}^{\pi}$ and $\left|L_{j}\right| \geqq \pi / 2, L_{\tau_{k}}$ is bounded away from 0 . If $\left|L_{j}\right| \leqq \pi / 2$, relation (6.1) implies that $L_{\tau_{k}}$ is bounded away from 0 , since $L_{j}=x_{2} \pm 2 \pi$. So $L_{\tau_{k}}$ is bounded away from 0 on $E_{\sigma}^{\pi}$, which implies that $L_{\tau_{k+1}}, \ldots, L_{\tau_{r}}$ are bounded away from 0 also. It follows that $P_{\eta}^{\prime}(x)$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{k-1}}\right|^{4_{k-1}}$ on $E_{\sigma}^{\pi}$.
Subcase I.2.ii. No $j$ satisfies $m<j \leqq r_{B}, L_{j} \in s(B)$ and $\theta_{j} \neq 0$.
Note that $x_{m} \notin s(B)$, for otherwise we would have $x_{1}+\ldots+x_{m-1} \in s(B)$, contracting our choice of $m$. Since $L_{m}=x_{m}+\theta_{m}$ it follows that either $L_{m} \notin s(B)$ or $\theta_{m}$ $\neq 0$. If $L_{m} \notin s(B)$, then

$$
\begin{aligned}
s(B)=\{ & L_{l_{B}}, \ldots, L_{m-1}, x_{m+1}, \ldots, x_{r_{B}}, x_{1}+\ldots+x_{m}, x_{1}+ \\
& \left.\ldots+x_{m+1}, \ldots, x_{1}+\ldots+x_{r B}\right\} .
\end{aligned}
$$

We see that $d\left(P_{\eta}^{\prime}, B\right)$ would not change if $L_{l_{B}}=x_{l_{B}}, \ldots, L_{m-1}=x_{m-1}$, but with this change it becomes the dimension of the block described in Lemma 5.4, which is positive by Lemma 5.5, contradicting our assumption. Hence $L_{m} \in s(B)$ and $\theta_{m} \neq 0$. But then relation (6.2) allows us to argue as in Subcase I.2.i that $\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{k-1}}\right|^{\Delta_{k-1}}$ provides a majorant for $P_{\eta}^{\prime}(x)$.

To complete the proof in Subcase I. 2 it suffices to show that $\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{k-1}}\right|^{\Lambda_{k-1}}$ is integrable on $E_{\sigma}^{\pi}$. But this follows from Lemma 4.4, since $d\left(P_{\eta}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{j}}\right\}\right)>0, j=1, \ldots, k-1$. This establishes that $\int_{\boldsymbol{E}_{\sigma}^{\pi}} P_{\eta}^{\prime}(x) d x<\infty$ in Subcase I.2.

Case II. $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\} \subset\left\{L_{1}, \ldots, L_{2 p}\right\}$.
Subcase II.1. $\theta_{1}=\theta_{2}+\ldots+\theta_{2 p}$.
In this case we have $s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}=\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}, k=1, \ldots, 2 p-2$, and $s\left\{L_{\tau_{1}}, \ldots, L_{i_{2 p-1}}\right\}=\left\{L_{1}, \ldots, L_{2_{2}}\right\}$.

Suppose first that $\theta_{2} \neq 0$, so that $L_{2}=x_{2} \pm 2 \pi$. Then (6.3) implies that $\left|L_{2}\right| \geqq \pi / 2$ on $E_{\sigma}^{\pi} \cap V^{\prime}$. According to Lemma 4.2, $P_{\eta}^{\prime}(x)$ is at most a constant times $\left|L_{\tau_{1}}\right|^{4_{1}} \ldots\left|L_{\tau_{2 p}}\right|^{A_{22}}$. Let $j<2 p$ be the integer satisfying $L_{2} \in T_{j}$. Then Lemma 4.1 implies that $\left|L_{\tau_{2} p}\right| \geqq \ldots \geqq\left|L_{j}\right| \geqq \frac{1}{C_{g}}\left|L_{2}\right| \geqq \frac{\pi}{2 C_{\sigma}}$ for $x \in E_{\sigma}^{\pi}$. Hence $P_{\eta}^{\prime}(x)$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\Lambda_{i}} \ldots\left|L_{\tau_{j}-1}\right|^{\Delta_{j-1}}$ for $x \in E_{\sigma}^{\pi} \cap V^{\prime}$. Since $j-1 \leqq 2 p-2$, $d\left(P_{\eta}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)>0, k=1, \ldots, j-1$. Therefore Lemma 4.4 implies that the integral of this product over $E_{\sigma}^{\pi} \cap V^{\prime}$ is finite, establishing (6.4) when $\theta_{2} \neq 0$.

Now suppose $\theta_{2}=0$, so that $L_{2}=x_{2}$. Define

Then

$$
L_{q}=\left\{L_{1}, \ldots, L_{2 p}\right\} \backslash\left\{L_{\tau_{1}}, \ldots, L_{\tau_{z p-1}}\right\}
$$

$$
T_{k}=\left\{L_{\tau_{k}}\right\}, \quad k=1, \ldots, 2 p-2
$$

and

$$
T_{2 p-1}=\left\{L_{1}, \ldots, L_{2_{p}}\right\} \backslash\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-2}}\right\}=\left\{L_{\tau_{2 p-1}}, L_{q}\right\}
$$

The next step is to use this to establish

$$
\begin{equation*}
\left|L_{2 p+1}\right| \leqq\left|L_{q}\right|, \quad x \in E_{\sigma}^{t} \cap V^{\prime} \tag{6.5}
\end{equation*}
$$

Since $L_{q} \in T_{2 p-1}$, Lemma 4.1b implies that $\left|L_{\tau_{2 p-1}}\right| \leqq\left|L_{q}\right|$ for $x \in E_{\sigma}^{\pi}$. Hence $\left|L_{\tau_{1}}\right| \leqq\left|L_{\tau_{2}}\right| \leqq \ldots \leqq\left|L_{\tau_{2 p-1}}\right| \leqq\left|L_{q}\right|$ for $x \in E_{\sigma}^{\pi}$. Therefore on $E_{\sigma}^{\pi}$ we have $L_{q}=$ $\max \left\{\left|L_{1}\right|,\left|L_{2}\right|, \ldots,\left|L_{2 p}\right|\right\}$. In particular $\left|L_{2}\right| \leqq\left|L_{q}\right|$ on $E_{\sigma}^{\pi}$. For $x \in V^{\prime}$ we have $\left|L_{2 p+1}\right|=\left|x_{1}\right| \leqq\left|x_{2}\right|=\left|L_{2}\right|$. These last two inequalities imply (6.5).

Since

$$
P_{\eta}^{\prime}(x)=\left|L_{\tau_{1}}\right|^{\eta-1} \ldots\left|L_{\tau_{2 p-1}}\right|^{\eta-1}\left|L_{q}\right|^{\eta-1}\left|L_{2 p+1}\right|^{-\alpha}\left|L_{2 p+2}\right|^{-\beta} \ldots\left|L_{4 p}\right|^{-\beta},
$$

relation (6.5) implies that $P_{\eta}^{\prime}(x) \leqq P_{\eta}^{\prime \prime}(x), x \in E_{\sigma}^{\pi} \cap V^{\prime}$, where

$$
P_{n}^{\prime \prime}(x)=\left|L_{\tau_{1}}\right|^{n-1} \ldots\left|L_{\tau_{2 p-1}-1}\right|^{n-1}\left|L_{2 p+1}\right|^{-\alpha+n-1}\left|L_{2 p+2}\right|^{-\beta} \ldots\left|L_{4 p}\right|^{-\beta}
$$

Hence (6.4) will follow if we show

$$
\begin{equation*}
\int_{E=\pi} P_{n}^{\prime \prime}(x) d x<\infty . \tag{6.6}
\end{equation*}
$$

To show (6.6) we use Lemma 4.4. For $k \leqq 2 p-2$ we have $s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}$ $=\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}$, from which it follows that

$$
d\left(P_{\eta}^{\prime \prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)=k+k(\eta-1)=k \eta>0, \quad k \leqq 2 p-2
$$

Since $s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\}=\left\{L_{1}, \ldots, L_{2_{p}}\right\}$ we have

$$
d\left(P_{\eta}^{\prime \prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\}\right)=(2 p-1)+(2 p-1)(\eta-1)=(2 p-1) \eta>0
$$

Finally, if $\theta_{1}=\ldots=\theta_{2 p}=0$, then

$$
d\left(P_{\eta}^{\prime \prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p}}\right\}\right)=2 p+2 p(\eta-1)-p \alpha-p \beta=2 p\left[\eta-\frac{(\alpha+\beta)}{2}\right]>0
$$

On the other hand, if some $\theta_{j} \neq 0$ then $\left\{x_{1}+\ldots+x_{m}, L_{1}, \ldots, L_{2 p}\right\}$ $\subset s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p}}\right\}$ for some $m$ and we can argue as in Subcase I. 2 that $P_{n}^{\prime \prime}$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\eta-1} \ldots\left|L_{\tau_{2 p-1}}\right|^{\eta-1}$. Either way, (6.6) follows from Lemma 4.4, completing the proof in Subcase II.1.

Subcase II.2. $\theta_{1} \neq \theta_{2}+\ldots+\theta_{2 p}$.
In this case we will show that $\int_{E^{\pi}} P_{\eta}^{\prime}(x) d x<\infty$. We have

$$
s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}=\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}, \quad k=1, \ldots, 2 p-1
$$

Put $L_{q}=\left\{L_{1}, \ldots, L_{2 p}\right\} \backslash\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\}$.
Suppose first that $L_{\tau_{2 p}}=L_{q}$. Then $s\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2} p}\right\}=\left\{L_{1}, \ldots, L_{2_{p}}\right\}$, so that $d\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}=k(\eta-1)+k=k \eta>0, k=1, \ldots, 2 p$. Hence Lemma 4.4 implies that $\int_{E_{\boldsymbol{\pi}}^{\pi}} P_{\eta}^{\prime}(x) d x<\infty$.

Now suppose that $L_{\tau_{2 p}} \neq L_{q}$ so $L_{\tau_{2 p}}=x_{1}+\ldots+x_{m}$ for some $1 \leqq m \leqq 2 p$. Since $\theta_{j} \neq 0$ for some $j$, we can argue as in Subcase I. 2 that $P_{\eta}^{\prime}$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\eta-1} \ldots\left|L_{\tau_{2 p-1}}\right|^{\eta-1}$. Since $d\left(P_{\eta}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right)>0, k=1, \ldots, 2 p-1$, this product is integrable over $E_{\sigma}^{\pi}$, completing the proof of Proposition 6.1.
Proof of Proposition 6.2. Let $\sigma$ be a permutation of $\{1, \ldots, 4 p\}$.
Case I. $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\} \neq\left\{L_{1}, \ldots, L_{2 p}\right\}$.
As in Case I of the proof of Proposition 6.1, we show that $\int_{E_{\mathcal{F}}} P_{\eta}^{\prime}(x) d x<\infty$ if $\eta>(\alpha+\beta) / 2$. Since $f_{N, \eta}^{\prime}(x)=N^{2 p \eta} P_{\eta}^{\prime}(x)$, both parts of Proposition 6.2 follow by choosing $\eta$ appropriately.
Case II. $\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\} \subset\left\{L_{1}, \ldots, L_{2 p}\right\}$.
Subcase II.1. $\theta_{1}=\theta_{2}+\ldots+\theta_{2 p}$.
Proof of Part a in Subcase II.1. We saw in the proof of Proposition 6.1 that
where

$$
\begin{equation*}
\left|L_{\tau_{1}}\right| \leqq\left|L_{\tau_{2}}\right| \leqq \ldots \leqq\left|L_{\tau_{2 p-1}}\right| \leqq\left|L_{q}\right|, \quad x \in E_{\sigma}^{t} \tag{6.7}
\end{equation*}
$$

$$
L_{q}=\left\{L_{1}, \ldots, L_{2_{p}}\right\} \backslash\left\{L_{\tau_{1}}, \ldots, L_{\tau_{2 p-1}}\right\}
$$

Since $h_{N}(z) \leqq N$, we have

$$
f_{N}^{\prime}(x) \leqq N^{2 p}\left|L_{2 p+1}\right|^{-\alpha}\left|L_{2 p+2}\right|^{-\beta} \ldots\left|L_{4 p}\right|^{-\beta}, \quad x \in \mathbb{R}^{2 p}
$$

If $\theta_{1}=\ldots=\theta_{2 p}=0$ then $T_{2 p}=\left\{L_{2 p+1}, \ldots, L_{4 p}\right\}$ and Lemma 4.1 implies that $\left|L_{2_{p+1}}\right|^{-\alpha}\left|L_{2 p+2}\right|^{-\beta} \ldots\left|L_{4 p}\right|^{-\beta}$ is at most a constant times $\left|L_{\tau_{2 p}}\right|^{-p(\alpha+\beta)}$ for $x \in E_{\sigma}^{\pi}$. On the other hand if some $\theta_{j} \neq 0$ the fact that $L_{\tau_{2 p}}=x_{1}+\ldots+x_{m}$ for some $m$ allows us to argue as in Subcase I.2.i that $L_{\tau_{2 p}}$ is bounded away from 0 on $E_{\sigma}^{\pi}$, which implies that $\left|L_{2 p+1}\right|^{-\alpha} \ldots\left|L_{4 p}\right|^{-\beta}$ is bounded on $E_{\sigma}^{\pi}$. Thus
this product is at most a constant times $M\left(L_{\tau_{2 p}}\right)$, where

$$
M(z)=\max \left(|z|^{-p(\alpha+\beta)}, 1\right)
$$

Hence there is a constant $C$ so that

$$
\begin{equation*}
f_{N}^{\prime}(x) \leqq C N^{2 p} M\left(L_{\tau_{2 p}}\right), \quad x \in E_{\sigma}^{\pi} \tag{6.8}
\end{equation*}
$$

Define the sets

$$
\begin{gathered}
G_{N, 0}=E_{\sigma}^{t} \cap\left\{\frac{1}{N} \leqq\left|L_{\tau_{1}}\right|\right\}, \\
G_{N, j}=E_{\sigma}^{\imath} \cap\left\{\left|L_{\tau_{j}}\right| \leqq \frac{1}{N} \leqq\left|L_{\tau_{j+1}}\right|\right\}, \quad j=1, \ldots, 2 p-2
\end{gathered}
$$

and

$$
G_{N, 2 p-1}=E_{\sigma}^{t} \cap\left\{\left|L_{\tau_{2 p-1}}\right| \leqq \frac{1}{N}\right\}
$$

Because of (6.7) it is clear that

$$
E_{\sigma}^{t}=G_{N, 0} \cup G_{N, 1} \cup \ldots \cup G_{N, 2 p-1}
$$

Thus it suffices to show that the conclusion of Part a holds with $U_{t}$ replaced by $G_{N, j}, j=0, \ldots, 2 p-1$. Define also the sets

$$
\begin{aligned}
K_{N, j}= & \left\{\left|L_{\tau_{k}}\right| \leqq \frac{1}{N}, k=1, \ldots, j\right\} \\
& \cap\left\{\frac{1}{N} \leqq\left|L_{\tau_{k}}\right|, L_{\tau_{k}} \in A_{t}, k=j+1, \ldots, 2 p-1\right\} \cap\left\{L_{\tau_{\lambda_{p}}} \in A_{t}\right\}
\end{aligned}
$$

where $A_{t}=\bigcup_{k=1}^{4 p}\left\{L_{k}(x): x \in E_{\sigma}^{t}\right\}$. Note that the measure of $A_{t}$ tends to 0 as $t \rightarrow 0$. In view of (6.7) we have $G_{N, j} \subset K_{N, j}$. To prove Part a we distinguish two subcases according to whether $j=2 p-1$ or not.
Subcase II.1.i. $j=2 p-1$. From (6.8) we conclude that $\int_{G_{N}, 2 p-1} f_{N}^{\prime}(x) d x$ is at most
a constant times

$$
N^{2 p} \int_{G_{N}, 2 p-1} M\left(L_{\tau_{2 p}}\right) \leqq N^{2 p} \int_{K_{N}, 2 p-1} M\left(L_{\tau_{2 p}}\right)
$$

where on $K_{N, 2 p-1}$, we have $\left|L_{\tau_{k}}\right| \leqq 1 / N$ for $k=1, \ldots, 2 p-1$ and $L_{\tau_{2 p}} \in A_{t}$. We see, on making the appropriate change of variable, that the right hand side is majorized by

$$
N^{2 p}\left[\int_{-1 / N}^{1 / N} d w\right]^{2 p-1} \int_{A_{t}} M(z) d z
$$

Therefore

$$
\limsup _{N} \frac{1}{N} \int_{G_{N}, 2 p-1} f_{N}^{\prime}(x) d x
$$

is at most a constant times $\int_{A_{t}} M(z) d z$, which implies the conclusion of Part a.

Subcase II.1.ii. $j<2 p-1$. For $x \in G_{N, j}$ we have

$$
\begin{aligned}
& h_{N}\left(L_{\tau_{k}}\right)=N, \quad k=1, \ldots, j, \\
& h_{N}\left(L_{\tau_{k}}\right)=\left|L_{\tau_{k}}\right|^{-1}, \quad k=j+1, \ldots, 2 p-1, \\
& h_{N}\left(L_{q}\right)=\left|L_{q}\right|^{-1} \leqq\left|L_{\tau_{2 p-1}}\right|^{-1} .
\end{aligned}
$$

and, by (6.7)

These facts in combination with (6.8) yield

$$
f_{N}^{\prime}(x) \leqq C N^{j}\left|L_{\tau_{j+1}}\right|^{-1} \ldots\left|L_{\tau_{2 p-2}}\right|^{-1}\left|L_{\tau_{2 p-1}}\right|^{-2} M\left(L_{\tau_{2 p}}\right), \quad x \in G_{N, j}
$$

According to (6.7), $\left|L_{\tau_{k}}\right| \leqq\left|L_{\tau_{2 p-1}}\right|, k=j+1, \ldots, 2 p-1$, and thus $f_{N}^{\prime}(x)$ is at most

$$
C N^{j}\left\{\prod_{k=j+1}^{2 p-1}\left|L_{\tau_{k}}\right|^{-1-\frac{1}{2 p-1-j}}\right\} M\left(L_{\tau_{2 p}}\right)
$$

for $x \in G_{N, j}$. Integrating this expression over $K_{N, j}$ we have at most a constant times

$$
N^{j}\left(\int_{-1 / N}^{1 / N} d v\right)^{j}\left(\int_{\frac{1}{N} \leqq w, w \in A_{c}} w^{-\left\{1+\frac{1}{2 p-1-j}\right\}} d w\right)^{2 p-1-j} \int_{A_{t}} M(z) d z
$$

The first integral in brackets is $2 N^{-1}$. The second is $O\left(N^{1 /(2 p-1-j)}\right)$. So the whole expression is $O(N) \int M(z) d z$. This concludes the proof of Part a in Subcase II.1.

Proof of Part $b$ in Subcase II.1. Fix $\varepsilon>0$. Under the conditions of Part $b$ we can choose $\eta$ satisfying $0<\eta<1, \eta>(\alpha+\beta) / 2$ and $1<2 p \eta<p(\alpha+\beta)+\varepsilon$. Thus it suffices to show that $\int_{E \pi} P_{n}^{\prime}(x) d x<\infty$ under these conditions.

First suppose that $\theta_{1}=\theta_{2}=\ldots=\theta_{2 p}=0$. Then Propositions 5.1 and 5.2 imply that $d\left(P_{\eta}^{\prime}, W\right)>\min (2 p \eta-1,0)$ for every strongly independent $W \subset T$. Since $2 p \eta>1$, Theorem 3.1 implies the desired conclusion in this case.

On the other hand if some $\theta_{j} \neq 0$, then some dimensions may be negative. However, since $2 p \eta>1, d\left(P_{\eta}^{\prime}, W\right)>0$ whenever $W \subset\left\{L_{1}, \ldots, L_{2 p}\right\}$. Hence if $k$ is the smallest index satisfying $d\left(P_{n}^{\prime},\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}\right) \leqq 0$, then we must have $x_{1}+\ldots$ $+x_{m} \in S\left\{L_{\tau_{1}}, \ldots, L_{\tau_{k}}\right\}$ for some $m$, and thus we can argue as Subcase $I .2$ of the proof of Proposition 6.1 that $P_{\eta}^{\prime}$ is at most a constant times $\left|L_{\tau_{1}}\right|^{\Lambda_{1}} \ldots\left|L_{\tau_{k-1}}\right|^{A_{k-1}}$, so that Lemma 4.4 can be used to complete the proof of Part b in Subcase II.1.
Subcase II.2. $\theta_{1} \neq \theta_{2}+\ldots+\theta_{2 p}$.
As in Subcase II. 2 of the proof of Proposition 6.1, we have

$$
\int_{E_{\mathcal{F}}^{\pi}} P_{\eta}^{\prime}(x) d x<\infty \quad \text { if } \eta>(\alpha+\beta) / 2
$$

Since $f_{N, \eta}^{\prime}(x)=N^{2 p \eta} P_{\eta}^{\prime}(x)$, both parts of Proposition 6.2 follow by choosing $\eta$ appropriately.

## 7. Proof of Theorem 1

The proof requires a lemma in addition to Propositions 6.1 and 6.2. We use the notation introduced in Sect. 2 prior to the statement of Theorem 1. Fix $p \geqq 1$ and note that

$$
\begin{align*}
\operatorname{Tr}\left(R_{N} A_{N}\right)^{p}= & \sum_{j_{1}=0}^{N-1} \cdots \sum_{j_{2}=0}^{N-1} r_{j_{1}-j_{2}} a_{j_{2}-j_{3}} r_{j_{3}-j_{4}} \ldots a_{j_{2 p}-j_{1}} \\
= & \sum_{j_{1}=0}^{N-1} \cdots \sum_{j_{2 p}=0}^{N-1}\left(\int_{-\pi}^{\pi} \ldots \int_{-\pi}^{\pi} e^{i\left(j_{1}-j_{2}\right) y_{1}} e^{i\left(j_{2}-j_{3}\right) y_{2}} \ldots e^{i\left(j_{2 p}-j_{1}\right) y_{2 p}}\right. \\
& \left.\cdot f\left(y_{1}\right) g\left(y_{2}\right) f\left(y_{3}\right) \ldots g\left(y_{2_{p}}\right) d y_{1} \ldots d y_{2_{p}}\right) \\
= & \int_{U_{\pi}} P_{N}(y) Q(y) d y \tag{7.1}
\end{align*}
$$

where

$$
\begin{gathered}
P_{N}(y)=\sum_{j_{1}=0}^{N-1} \cdots \sum_{j_{2}=0}^{N-1} e^{i\left(i_{1}-j_{2}\right) y_{1}} e^{i\left(j_{2}-j_{3}\right) y_{2}} \ldots e^{i\left(j_{2 p}-j_{1}\right) y_{2} p}, \\
Q(y)=f\left(y_{1}\right) g\left(y_{2}\right) f\left(y_{3}\right) \ldots g\left(y_{2 p}\right)
\end{gathered}
$$

and

$$
U_{\mathrm{t}}=[-t, t]^{2 p} .
$$

To state the lemma, introduce the diagonal

$$
D=\left\{y \in U_{\pi}: y_{1}=y_{2}=\ldots=y_{2 p}\right\} .
$$

Let $\mu$ be the measure on $U_{\pi}$ which is concentrated on $D$ and satisfies $\mu\left\{y: a \leqq y_{1}=y_{2}=\ldots=y_{2 p} \leqq b\right\}=b-a$ for all $-\pi \leqq a \leqq b \leqq \pi$. Thus $\mu$ is Lebesgue measure on $D$, normalized so that $\mu(D)=2 \pi$.
Lemma 7.1. Define the measure $\mu_{N}$ on $U_{\pi}$ by

$$
\mu_{N}(E)=\frac{1}{(2 \pi)^{2 p-1} N} \int_{E} P_{N}(y) d y, \quad E \subset U_{\pi}
$$

Then $\mu_{N}$ converges weakly to $\mu$ as $N \rightarrow \infty$.
Proof. Since $U_{\pi}$ is compact, it suffices to show that the Fourier coefficients of $\mu_{N}$ converge to those of $\mu$. Fixing integers $n_{1}, n_{2}, \ldots, n_{2 p}$, the corresponding Fourier coefficient of $\mu$ is

$$
\begin{aligned}
\int_{U_{\pi}} e^{i\left(n_{1} y_{1}+\ldots+n_{2 p} y_{2 p}\right)} d \mu(y) & =\int_{-\pi}^{\pi} e^{i\left(n_{1}+\ldots+n_{2 p}\right) x} d x \\
& = \begin{cases}0 & \text { if } \sum_{j=1}^{2 p} n_{j} \neq 0 \\
2 \pi & \text { if } \sum_{j=1}^{2 p} n_{j}=0\end{cases}
\end{aligned}
$$

The corresponding Fourier coefficient of $\mu_{N}$ is

$$
\begin{align*}
C_{N}= & C_{N}\left(n_{1}, n_{2}, \ldots, n_{2 p}\right) \\
= & \int_{U_{\pi}} e^{i\left[n_{1} y_{1}+\ldots+n_{2 p} y_{2 p}\right]} d \mu_{N}(y) \\
= & \frac{1}{(2 \pi)^{2 p-1} N} \sum_{j_{1}=0}^{N-1} \cdots \sum_{j_{2 p}=0}^{N-1}\left\{\int_{-\pi}^{\pi} e^{i\left[n_{1}+j_{1}-j_{2}\right] y_{1}} d y_{1}\right. \\
& \left.\cdot \int_{-\pi}^{\pi} e^{i\left[n_{2}+j_{2}-j_{3}\right] y_{2}} d y_{2} \ldots \int_{-\pi}^{\pi} e^{i\left[n_{2 p}+j_{2 p}-j_{1}\right] y_{2 p}} d y_{2 p}\right\} . \tag{7.2}
\end{align*}
$$

Fix $j_{1}, \ldots, j_{2 p}$. In order for the expression in braces to be nonzero we must have

$$
\begin{gather*}
n_{1}=-\left(j_{1}-j_{2}\right) \\
n_{2}=-\left(j_{2}-j_{3}\right)  \tag{7.3}\\
\vdots \\
n_{2 p-1}=-\left(j_{2 p-1}-j_{2 p}\right) \\
n_{2 p}=-\left(j_{2 p}-j_{1}\right) .
\end{gather*}
$$

But then $n_{1}+\ldots+n_{2 p}=0$. Thus if $n_{1}+\ldots+n_{2 p} \neq 0$ each of the summands in (7.2) is equal to 0 . Therefore $C_{N}=0$ if $n_{1}+\ldots+n_{2 p} \neq 0$.

Suppose $n_{1}+\ldots+n_{2 p}=0$. Then each summand in (7.2) equals 0 or $(2 \pi)^{2 p}$. When the summand equals $(2 \pi)^{2 p}$, the indices $j_{1}, \ldots, j_{2_{p}}$ satisfy (7.3), which implies

$$
\begin{align*}
& j_{2}=j_{1}+\left(n_{1}\right) \\
& j_{3}=j_{1}+\left(n_{1}+n_{2}\right)  \tag{7.4}\\
& \quad \vdots \\
& j_{2 p}=j_{1}+\left(n_{1}+\ldots+n_{2 p-1}\right) .
\end{align*}
$$

Define

$$
\begin{gathered}
M=\max \left\{n_{1}+\ldots+n_{k}: k=1, \ldots, 2 p-1\right\}, \quad M^{+}=\max (M, 0), \\
m=\min \left\{n_{1}+\ldots+n_{k}: k=1, \ldots, 2 p-1\right\}, \quad \text { and } \quad m^{+}=\max (-m, 0) .
\end{gathered}
$$

Fix $j_{1}$ satisfying $0 \leqq j_{1} \leqq N-1$ and determine $j_{2}, \ldots, j_{2 p}$ according to (7.4). In order for the inequalities $0 \leqq j_{k} \leqq N-1, k=2, \ldots, 2 p$ to be satisfied we must have $j_{1} \leqq N-1-M$ and $j_{1} \geqq m$. Thus the sum in (7.2) is equal to

$$
\sum_{j_{1}=m^{+}}^{N-1-M^{+}}(2 \pi)^{2 p}=\left(N-M^{+}-m^{+}\right)(2 \pi)^{2 p}
$$

Therefore

$$
C_{N}=\frac{2 \pi\left(N-M^{+}-m^{+}\right)}{N}
$$

which tends to $2 \pi$ as $N \rightarrow \infty$. This completes the proof of Lemma 7.1.
Proof of Theorem 1. We must evaluate the asymptotic behavior of

$$
\int_{U_{\pi}} P_{N}(y) Q(y) d y
$$

Introduce the sets

$$
\begin{aligned}
W_{k} & =\left\{y \in \mathbb{R}^{2 p}:\left|y_{k}\right| \leqq \frac{\left|y_{k+1}\right|}{2}\right\}, \quad k=1, \ldots, 2 p-1, \\
W_{2 p} & =\left\{y \in \mathbb{R}^{2 p}:\left|y_{2 p}\right| \leqq \frac{\left|y_{1}\right|}{2}\right\},
\end{aligned}
$$

and

$$
W=W_{1} \cup W_{2} \cup \ldots \cup W_{2 p}
$$

We shall divide the domain of integration $U_{\pi}$ into three parts as follows. Let

$$
\begin{aligned}
E_{t} & =U_{\pi} \backslash\left\{W \cup U_{t}\right\}, \\
F_{t} & =U_{t} \backslash W,
\end{aligned}
$$

and

$$
G=U_{\pi} \cap W
$$

For each $0<t \leqq \pi$, the sets $E_{t}, F_{t}$ and $G$ are disjoint and satisfy $U_{\pi}=E_{t} \cup F_{t} \cup G$.
According to (7.1), Part a of the theorem will be proven if we show that $p(\alpha$ $+\beta)<1$ implies

$$
\begin{gather*}
\lim _{N \rightarrow \infty} \frac{\int_{E_{t}} P_{N} Q}{N}=(2 \pi)^{2 p-1} \int_{t \leqq|z| \leqq \pi}[f(z) g(z)]^{p} d z, \quad 0<t \leqq 1,  \tag{7.5}\\
\lim _{t \rightarrow 0} \limsup _{N \rightarrow \infty} \frac{\int_{F_{t}} P_{N} Q}{N}=0, \tag{7.6}
\end{gather*}
$$

and

$$
\begin{equation*}
\lim _{N \rightarrow \infty} \frac{\int_{G} P_{N} Q}{N}=0 \tag{7.7}
\end{equation*}
$$

To prove (7.6) it is enough to show that when $p(\alpha+\beta)<1$

$$
\begin{equation*}
\lim _{t \rightarrow 0} \limsup _{N \rightarrow \infty} \frac{\int_{U_{t}}\left|P_{N} Q\right|}{N}=0 \tag{7.8}
\end{equation*}
$$

Since $G=\bigcup_{k=1}^{2 p}\left[W_{k} \cap U_{\pi}\right]$, relation (7.7) will hold, if $p(\alpha+\beta)<1$ implies

$$
\begin{equation*}
\lim _{N \rightarrow \infty} \frac{\int_{U_{t} W_{k}}\left|P_{N} Q\right|}{N}=0, \quad k=1, \ldots, 2 p \tag{7.9}
\end{equation*}
$$

From the definitions of $P_{N}$ and $Q$ it is clear that
and

$$
\int_{U_{\pi} \cap W_{1}}\left|P_{N} Q\right|=\int_{U_{\pi} \cap W_{3}}\left|P_{N} Q\right|=\ldots=\int_{U_{\pi} \cap W_{2 p-1}}\left|P_{N} Q\right|
$$

$$
\int_{U_{\pi} \cap W_{2}}\left|P_{N} Q\right|=\int_{U_{\pi} \cap W_{4}}\left|P_{N} Q\right|=\ldots=\int_{U_{\pi} \cap W_{2 p}}\left|P_{N} Q\right|
$$

Because of the symmetry between $\alpha$ and $\beta$ in the hypotheses of the theorem, it is clear that if we prove that $p(\alpha+\beta)<1$ implies

$$
\begin{equation*}
\lim _{N \rightarrow \infty} \frac{\int_{U_{\pi} \cap W_{1}}\left|P_{N} Q\right|}{N}=0 \tag{7.10}
\end{equation*}
$$

we will have also established

$$
\lim _{N \rightarrow \infty} \frac{\int_{U_{x} \cap W_{2}}\left|P_{N} Q\right|}{N}=0
$$

Thus (7.9) will follow from (7.10).
In conclusion, Part a of the theorem will be proven if we show that $p(\alpha$ $+\beta)<1$ implies (7.5), (7.8) and (7.10).

To prove Part $b$, we must show that for $p(\alpha+\beta) \geqq 1$

$$
\begin{equation*}
\int_{U_{\pi}}\left|P_{N} Q\right|=o\left(N^{p(\alpha+\beta)+\varepsilon}\right) \quad \text { for every } \varepsilon>0 \tag{7.11}
\end{equation*}
$$

We start with relation (7.5) and show that it holds in fact for all real values of $\alpha$ and $\beta$. We begin by showing that $Q$ is bounded on $E_{t}$. Let $y \in E_{t}$. Since $E_{t}$ is in the complement of $U_{t}$, there is some $k$ such that $\left|y_{k}\right|>t$. Since $E_{i}$ is also in the complement of $W_{j}, j=1, \ldots, 2 p$, we have $\left|y_{j}\right|>\left|y_{j+1} / 2\right|, j=1, \ldots, 2 p-1$ and $y_{2 p}>y_{1} / 2$. Thus we have

$$
\left|y_{k+1}\right|>\frac{\left|y_{k+2}\right|}{2}>\frac{\left|y_{k+3}\right|}{4}>\ldots>\frac{\left|y_{2}\right|}{2^{2 p-k}}>\frac{\left|y_{1}\right|}{2^{2 p-k+1}}>\ldots>\frac{\left|y_{k}\right|}{2^{2 p-1}}>\frac{t}{2^{2 p-1}}
$$

Therefore $\left|y_{j}\right|>t / 2^{2 p-1}, j=1, \ldots, 2 p$, for $y \in E_{t}$. Hence $Q$ is bounded on $E_{t}$. Since $E_{t} \cap D=\left\{y: y_{1}=\ldots=y_{2 p}, t \leqq\left|y_{1}\right| \leqq \pi\right\}$, relation (7.5) follows from Lemma 7.1.

Before proving (7.8), (7.10) and (7.11) we need to obtain majorants for $P_{N}$ and $Q$. We have

$$
\begin{aligned}
P_{N}(y) & =\left(\sum_{j_{1}=0}^{N-1} e^{i\left(y_{1}-y_{2 p}\right) j_{1}}\right)\left(\sum_{j_{2}=0}^{N-1} e^{i\left(y_{2}-y_{1}\right) j_{2}}\right) \ldots\left(\sum_{j_{2}=0}^{N-1} e^{i\left(y_{2 p}-y_{2 p-1}\right) j_{2} p}\right) \\
& =h_{N}^{*}\left(y_{1}-y_{2_{p}}\right) h_{N}^{*}\left(y_{2}-y_{1}\right) \ldots h_{N}^{*}\left(y_{2 p}-y_{2 p-1}\right),
\end{aligned}
$$

where

$$
h_{N}^{*}(z)=\sum_{j=0}^{N-1} e^{i z j}
$$

Since $h_{N}^{*}(z)=\left(1-e^{i N x}\right) /\left(1-e^{i x}\right)$ for $z \neq 0,\left|1-e^{i N z}\right| \leqq 2$ and $\left|1-e^{i z}\right| \geqq|z| / 2$ for $|z| \leqq \pi$, we obtain $h_{N}^{*}(z) \leqq 4|z|^{-1}$ for $|z| \leqq \pi$. For $\pi \leqq z \leqq 2 \pi$ this implies $h_{N}^{*}(z)$ $=h_{N}^{*}(z-2 \pi) \leqq 4|z-2 \pi|^{-1}$. For $-2 \pi \leqq z \leqq-\pi$ we have $h_{N}^{*}(z)=h_{N}^{*}(z+2 \pi) \leqq$ $4|z+2 \pi|^{-1}$. Since $\left|h_{N}^{*}(z)\right| \leqq N$, these inequalities imply that $\left|h_{N}^{*}(z)\right| \leqq 4 h_{N}(z)$, $-2 \pi \leqq z \leqq 2 \pi$, where $h_{N}(z)$ is as defined at the beginning of Sect. 6 .

Thus $\left|P_{N}(y)\right|$ is at most $4^{2 p}$ times

$$
h_{N}\left(y_{1}-y_{2_{p}}\right) h_{N}\left(y_{2}-y_{1}\right) \ldots h_{N}\left(y_{2 p}-y_{2 p-1}\right)
$$

For fixed $\delta>0$ let $\alpha_{0}=\alpha+\delta$ and $\beta_{0}=\beta+\delta$. It is clear that under the hypotheses of the theorem, $Q(y)$ is at most a constant times

$$
Q^{\prime}(y)=\left|y_{1}\right|^{-\alpha_{0}}\left|y_{2}\right|^{-\beta_{0}}\left|y_{3}\right|^{-\alpha_{0}} \ldots\left|y_{2 p}\right|^{\beta_{0}} .
$$

Thus, the proof of the theorem can be completed by showing that (7.8), (7.10) and (7.11) hold with the integrand $P_{N} Q$ replaced by $f_{N}(y)$, defined as in Sect. 6. We can now apply Propositions 6.1 and 6.2. Assume first that $p(\alpha+\beta) \geqq 1$. Choose $\delta>0$. Then $p\left(\alpha_{0}+\beta_{0}\right)>1$. Therefore Part b of Proposition 6.2 implies that

$$
\int_{U_{\pi}} f(y) d y=O\left(N^{p\left(\alpha_{0}+\beta_{0}\right)+\varepsilon}\right)=O\left(N^{p(\alpha+\beta)+2 p \delta+\varepsilon}\right)
$$

Since $\delta$ can be made arbitrarily small, (7.11) follows.
Now suppose $p(\alpha+\beta)<1$. To prove (7.8), choose $\delta>0$ such that $p\left(\alpha_{0}+\beta_{0}\right)<1$. Then (7.8) follows from Part a of Proposition 6.2. To prove (7.10) we consider two cases. If $\alpha+\beta<0$ choose $\delta>0$ such that $\alpha_{0}+\beta_{0}<0$. Then (7.10) is a consequence of Part a of Proposition 6.1. If $\alpha+\beta \geqq 0$, choose $\delta$ such that $p\left(\alpha_{0}+\beta_{0}\right)<1$ and use Part b of Proposition 6.1. This completes the proof of Theorem 1 .

## 8. Proof of Theorem 4

Lemma 8.1. If the conditions of Theorem 2 are satisfied, then

$$
\lim _{N \rightarrow \infty} \frac{1}{\sqrt{N}} E\left|x_{N}^{\prime} A_{N} x_{N}-\tilde{x}_{N}^{\prime} A_{N} \tilde{x}_{N}\right|=0
$$

Proof. The beginning of the proof follows Walker (1964). Note

$$
\begin{aligned}
x_{N}^{\prime} A_{N} x_{N}-\tilde{x}_{N}^{\prime} A_{N} \tilde{x}_{N}= & \sum_{j=1}^{N} \sum_{k=1}^{N} a_{j-k}\left(\bar{X}_{N}^{2}-X_{j} \bar{X}_{N}-X_{k} \bar{X}_{N}\right) \\
= & \bar{X}_{N}^{2} \int_{-\pi}^{\pi} g(x) \sum_{j=1}^{N} \sum_{k=1}^{N} e^{i(j-k) x} d x \\
& -2 \bar{X}_{N} \int_{-\pi}^{\pi} g(x) \sum_{j=1}^{N} \sum_{k=1}^{N} X_{j} e^{i(j-k) x} d x \\
= & =F_{N}-G_{N} .
\end{aligned}
$$

We consider $E\left|F_{N}\right|$ first. We have

$$
\begin{aligned}
E \bar{X}_{N}^{2} & =\frac{1}{N^{2}} \sum_{j=1}^{N} \sum_{k=1}^{N} r_{j-k}=\frac{1}{N^{2}} \int_{-\pi}^{\pi} f(x) \sum_{j=1}^{N} \sum_{k=1}^{N} e^{i(j-k) x} d x \\
& =\frac{1}{N^{2}} \int_{-\pi}^{\pi} f(x) h_{N}^{*}(x) h_{N}^{*}(-x) d x
\end{aligned}
$$

which is at most a constant times

$$
\frac{1}{N^{2}} \int_{-\pi}^{\pi}|x|^{-\alpha-\delta} h_{N}(x) h_{N}(-x) d x \leqq N^{2 \eta-2} \int_{-\pi}^{\pi}|x|^{-\alpha-\delta+2 \eta-2} d x
$$

where $\delta>0,0<\eta<1, h_{N}^{*}(x)=\sum_{j=1}^{N-1} e^{i x j}$ and $h_{N}(x) \leqq h_{N, \eta}(x)$ as at the beginning of Sect. 6. Choose $\delta$ so that $\alpha+2 \delta<1$ and put $\eta=(1+\alpha+2 \delta) / 2$. Then the last integral is finite, so $E \vec{X}_{N}^{2}=O\left(N^{\alpha-1+2 \delta}\right)$.

A similar argument shows that

$$
\int_{-\pi}^{\pi} g(x) \sum_{j=1}^{N} \sum_{k=1}^{N} e^{i(j-k) x} d x=O\left(N^{\beta+i+2 \delta}\right)
$$

Hence

$$
E\left|F_{N}\right|=o\left(N^{\frac{1}{2}}\right)
$$

Next we consider

$$
\left(E\left|G_{N}\right|\right)^{2} \leqq 4 E \bar{X}_{N}^{2} E\left\{\int_{-\pi}^{\pi} g(x) \sum_{j=1}^{N} \sum_{k=1}^{N} X_{j} e^{i(j-k) x} d x\right\}^{2}
$$

The second expectation above is equal to

$$
\begin{aligned}
& \int_{-\pi}^{\pi} \int_{-\pi}^{\pi} \int_{-\pi}^{\pi} g(x) g(y) f(z) \sum_{k_{1}=0}^{N-1} \cdots \sum_{k_{4}=0}^{N-1} \\
& \cdot e^{i\left(k_{1}-k_{2}\right) x} e^{i\left(k_{3}-k_{4}\right) y} e^{i\left(k_{1}-k_{3}\right) z} d x d y d z
\end{aligned}
$$

which is at most a constant times

$$
\int_{-\pi}^{\pi} \int_{-\pi}^{\pi} \int_{-\pi}^{\pi}|x|^{-\beta-\delta}|y|^{-\beta-\delta}|z|^{-\alpha-\delta} h_{N}(x+z) h_{N}(-x) h_{N}(y-z) h_{N}(-y) d x d y d z
$$

Put $\eta_{1}=\beta+2 \delta$ and $\eta_{2}=(1+\alpha) / 2$. We have $h_{N}(-x) \leqq N^{\eta_{1}} h_{N, \eta_{1}}(-x)$,

$$
h_{N}(-y) \leqq N^{\eta_{1}} h_{N, \eta_{1}}(-y), \quad h_{N}(x+z) \leqq N^{\eta_{2}} h_{N, \eta_{2}}(x+z)
$$

and $h_{N}(y-z) \leqq N^{\eta_{2}} h_{N, \eta_{2}}(y-z)$.
In order for $|x+z|$ to exceed $3 \pi / 2$ we must have $|x| \geqq \pi / 2$ and $|z| \geqq \pi / 2$, in which case the integrand is majorized by $N^{2 \eta_{1}+2 \eta_{2}}$ times

$$
|y|^{-\beta-\delta+\eta_{1}-1}\left|x+z+\theta_{1}\right|^{\eta_{2}-1}\left|y-z+\theta_{2}\right|^{\eta_{2}-1}
$$

where $\left\{\theta_{1}, \theta_{2}\right\} \subset\{-2 \pi, 0,2 \pi\}$. This product is clearly integrable for any choice of $\theta_{1}$ and $\theta_{2}$.

In order for $|y-z|$ to exceed $3 \pi / 2$ we must have $|y| \geqq \pi / 2$ and $|z| \geqq \pi / 2$ in which case the integrand is majorized by $N^{2 \eta_{1}+2 \eta_{2}}$ times

$$
|x|^{-\beta-\delta+n_{1}-1}\left|x+z+\theta_{3}\right|^{n_{2}-1}\left|y-z+\theta_{4}\right|^{n_{2}-1}
$$

for some $\theta_{3}, \theta_{4} \subset\{-2 \pi, 0,2 \pi\}$. This product is also integrable.

If neither of the above cases holds, then the integrand is majorized by $N^{2 \eta_{1}+2 \eta_{2}}$ times

$$
|x|^{-\beta-\delta+\eta_{1}-1}|y|^{-\beta-\delta+\eta_{1}-1}|z|^{-\alpha-\delta}|x+z|^{\eta_{2}-1}|y-z|^{\eta_{2}-1} .
$$

Using Theorem 3.1, it is easily checked that this product is integrable. Thus we conclude that

$$
\left(E\left|G_{N}\right| \theta^{2} \leqq 4 E \bar{X}_{N}^{2} O\left(N^{2 \eta_{1}+2 \eta_{2}}\right)=O\left(N^{2 \alpha+2 \beta+4 \delta}\right)\right.
$$

Since $\alpha+\beta<1 / 2$, we see that $E\left|G_{N}\right|=o\left(N^{\frac{1}{2}}\right)$, completing the proof of Lemma 8.1.

## References

1. Fox, R., Taqqu, M.S.: Large-sample properties of parameter estimates for strongly dependent stationary time series. Ann. Stat. 14, 517-532 (1986)
2. Grenander, U., Szegö, G.: Toeplitz forms and their applications. University of California Press, Berkeley: University of California Press (1958)
3. Hoskings, J.R.M.: Fractional differencing. Biometrika 68, 165-176 (1981)
4. Lowenstein, J.H., Zimmermann, W.: The power counting theorem for Feynman integrals with massless propagators. Commun. Math. Phys. 44, 73-86 (1975)
5. Mandelbrot, B.B., Van Ness, J.W.: Fractional Brownian motions, fractional noises and applications. SIAM Review 10, 422-437 (1968)
6. Rosenblatt, M.: Independence and dependence. Proc. 4th Berkeley Symp. Math. Stat. Probab., pp. 431-443. Berkeley: University of California Press: Berkeley (1961)
7. Sinai, Ya.G.: Self-similar probability distributions. Theory Probab. Appl. 21, 64-80 (1976)
8. Walker, A.N.: Asymptotic properties of least squares estimators of parameters of the spectrum of a stationary nondeterministic time series. J. Aust. Math. Soc. 4, 363-384 (1964)

Received November 18, 1983; in revised form March 17, 1986


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    * Research supported by the National Science Foundation grant ECS-80-15585

