## SPECTRAL ANALYSIS ON UPPER LIGHT CONE IN R ${ }^{3}$ AND THE RADON TRANSFORM

ANTONI WAWRZYÑCZYK

Introduction. The upper light cone $L$ in $\mathbf{R}^{3}$ is a homogeneous space of the 3dimensional Lorentz group $G$. It may be identified with the space of horocycles in the upper hyperboloide $H$ which is the symmetric space associated to $G$. There exists a duality between $H$ and $L$ (see [5] p. 144) and a general procedure leads to a generalized Radon transform:

$$
R: \mathcal{D}(H) \rightarrow \mathcal{D}(L)
$$

and the dual Radon transform

$$
B: \mathcal{E}(L) \rightarrow \mathcal{E}(H) .
$$

These operations commute with the natural action of the group $G$. It was proved in [12] that the spectral analysis holds in the space $\mathcal{E}(H)$ with respect to the system of functions given as follows:

$$
H \ni x \rightarrow\langle x \mid p\rangle^{\mu}, \quad \rho \in L, \mu \in \mathbf{C} .
$$

where $\langle\cdot \mid \cdot\rangle$ is the indefinite group invariant scalar product on $\mathbf{R}^{3}$.
In the present paper we study the problem of spectral analysis in the space $\mathcal{E}(L)$. One finds that the system of functions

$$
L \ni p \rightarrow\langle x \mid p\rangle^{\mu}, \quad x \in H
$$

is not sufficient for obtaining the spectral analysis on $L$, although the duality between $H$ and $L$ can suggest it. The reason is that in $\mathcal{E}(L)$ there appear finite dimensional and discrete series of irreducible representations of $G$.

Nevertheless, we obtain a spectral analysis theorem for a larger class of elementary functions (Theorems 5.3 and 5.4). It permits us to characterize in terms of spectral analysis the $G$-invariant space $\operatorname{ker} B$. In Section 6 we prove also a theorem of Pompeiu type, that is the necessary and sufficient condition for a compactly supported distribution on $L$ to span by translations and linear combinations a dense subset in $\mathcal{E}(L)$.

1. The group and its homogeneous spaces. Throughout what follows $G$ will denote the Lorentz group in three dimensions, that is the group of all linear mappings in $\mathbf{R}^{3}$ which preserve the form:

$$
\langle x \mid y\rangle:=-x_{1} y_{1}-x_{2} y_{2}+x_{3} y_{3}
$$

[^0]and the sign of $x_{3}$.
Let us distinguish in $G$ the following three 1-parameter subgroups:
\[

$$
\begin{aligned}
& A:=\left\{a(t) \mid a(t):=\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & \operatorname{ch} t & \operatorname{sh} t \\
0 & \operatorname{sh} t & \operatorname{ch} t
\end{array}\right), t \in \mathbf{R}\right\} \\
& K:=\left\{k(\theta) \mid k(\theta):=\left(\begin{array}{ccc}
\cos \theta & \sin \theta & 0 \\
-\sin \theta & \cos \theta & 0 \\
0 & 0 & 1
\end{array}\right), \theta \in \mathbf{R}\right\} \\
& N:=\left\{n(s) \mid n(s):=\left(\begin{array}{ccc}
1 & s & -s \\
-s & 1-\frac{s^{2}}{2} & \frac{s^{2}}{2} \\
-s & -\frac{s^{2}}{2} & 1+\frac{s^{2}}{2}
\end{array}\right), s \in \mathbf{R}\right\} .
\end{aligned}
$$
\]

Every element $g \in G$ can be represented in a unique way as

$$
g=g(\theta, t, s)=k(\theta) a(t) n(s)
$$

(the Iwasawa decomposition).
The group $N$ conserves in $\mathbf{R}^{3}$ the planes given by the equation $x_{2}-x_{3}=$ $\left\langle x \mid p_{0}\right\rangle=$ const., where $p_{0}:=(0,1,1)$. We denote also $x_{0}:=(0,0,1)$. The point $x_{0}$ is $K$-fixed and $p_{0}$ is $N$-fixed.

The orbit $H:=G x_{0}$ is the upper hyperboloide and as a homogeneous space is isomorphic to the quotient $G / K$. The upper light cone $L:=G p_{0}$ is in turn isomorphic to $G / N$.

For a given homogeneous space $(G, Y)$ and a function $f$ on $Y$ we denote by $L_{g}$ the translation
(1.2) $\quad L_{g} f(y):=f\left(g^{-1} y\right), \quad g \in G, y \in Y$.

The representation $G \ni g \rightarrow L_{g}$ will be referred to as the regular representation of $G$.

We are principally interested in the action of the group $G$ in the spaces $\mathcal{E}(H)$, $\mathcal{E}(L)$ and $\mathcal{D}(H), \mathcal{D}(L)$ of all smooth functions and all Schwartz test functions on $H$ and $L$, respectively. The topology in $E$ is the usual Fréchèt topology of the uniform convergence with all derivatives on compact sets. In the space $\mathcal{D}$ we have the topology of compact convergence with all derivatives. The dual space $\mathcal{D}^{\prime}$ is the space of all distribution and $\mathcal{E}^{\prime}$ can be identified with the space of compactly supported distributions. The group $G$ acts continuously on $\mathcal{E}$ and $\mathcal{D}$ by means of the operators $L_{g}$ and on the dual spaces $\mathcal{D}^{\prime}, \mathcal{E}^{\prime}$ by means of contragredient representation.

Let $T \in \mathcal{E}^{\prime}(G)$ and $S \in \mathcal{D}^{\prime}(Y)$. The convolution $T * S$ is the distribution on $Y$ defined by the formula

$$
\begin{equation*}
T * S(f):=T_{g} \otimes S_{y}(f(g y)), \quad f \in \mathcal{D}(Y) \tag{1.3}
\end{equation*}
$$

The invariant measures on $K, A, N$ respectively can be chosen in the following way:

$$
\begin{align*}
\int_{K} f d k & :=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(k(\theta)) d \theta  \tag{1.4}\\
\int_{A} \varphi d a & :=\int_{\mathbf{R}} \varphi(a(t)) d t \\
\int_{N} \varphi d n & :=\int_{\mathbf{R}} \varphi(n(s)) d s
\end{align*}
$$

In virtue of the Iwasawa decomposition the Haar measure on $G$ can be defined as follows:

$$
\begin{equation*}
\int_{G} f d g:=\int_{K} \int_{\mathbf{R}} \int_{N} f(k a(t) n) e^{t} d k d t d n . \tag{1.5}
\end{equation*}
$$

On the manifold $H$ we can define the following $G$-invariant measure:

$$
\int_{H} f d m:=\int_{K} \int_{N} f\left(a n x_{0}\right) d n d a .
$$

After determining a $G$-invariant measure $d y$ on a homogeneous space $y$ we are able to inject the space $\mathcal{E}(Y)$ into $\mathcal{D}^{\prime}(Y)$ in such a way that the translations in $\mathcal{E}(Y)$ coincide with the contragredient action of $G$ on $\mathcal{D}^{\prime}(Y)$. Namely, given $f \in \mathcal{E}(Y)$ we put:

$$
I(f)(\varphi):=\int_{K} f \varphi d y, \quad \varphi \in \mathcal{D}(Y)
$$

With the aid of the Haar measure on $K$ we can introduce the operator $P$ : $\mathcal{E}(G) \rightarrow \mathcal{E}(H)$ as follows

$$
P f(g K):=\int_{K} f(g k) d k
$$

The operator $P$ is surjective and commutes with the regular representation. A distribution $T \in \mathcal{E}^{\prime}(H)$ can be considered as a compactly supported distribution on $G$ according to the formula:

$$
\tilde{T}(f):=T(P f), \quad f \in \mathcal{D}(G)
$$

The space $\mathcal{E}^{\prime}(H)$ becomes now a convolution algebra with respect to the operation defined by the formula:

$$
T * S:=\tilde{T} * S, \quad T, S \in \mathcal{E}^{\prime}(H)
$$

the right hand side being defined by (1.3).
The space $\mathcal{D}(H)$ constitutes a subalgebra of $\mathcal{E}^{\prime}(H)$.
Every point $p \in L$ determines a subset $\xi_{p} \subset H$ called a horocycle and defined as:

$$
\xi_{p}:=\{x \in H \mid\langle x \mid p\rangle=1\} .
$$

Let us observe that $g \xi_{p}=\xi_{g p}$. In particular every horocycle is of the form $g \xi_{p_{0}}, g \in G$. Since the isotropy group of the point $\xi_{p_{0}}$ is just $N$ we can identify the space of all horocycles with $G / N=L$.

In a similar way there is a one-to-one correspondence between the points of $H$ and subsets of $L$ defined by

$$
\check{x}:=\{\rho \in L \mid\langle x \mid p\rangle=1\} \quad \text { for } x \in H .
$$

We observe that $\check{x}_{0}=K p_{0}$ and $(g x)^{\ulcorner }=g \check{x}$. The sets $\check{x} \subset L$ will be referred to as circles in $L$.
2. The Fourier, the Radon and the dual Radon transforms. Let us consider on $H \times L$ the family of functions given by:

$$
H \times L \ni(x, p) \rightarrow\langle x \mid p\rangle^{\mu}, \quad \mu \in \mathbf{C} .
$$

We shall use the notation:

$$
\begin{aligned}
& e_{\mu, \rho}(x):=\langle x \mid p\rangle^{\mu} \quad \text { and } \\
& e_{\mu, x}(\rho):=\langle x \mid p\rangle^{\mu}, \quad x \in H, p \in L .
\end{aligned}
$$

The invariance of the form $\langle\cdot \mid \cdot\rangle$ leads to relations:

$$
\begin{equation*}
e_{\mu, g x}=L_{g} e_{\mu, x} \quad \text { and } \quad e_{\mu, g p}=L_{g} e_{\mu, p} \tag{2.1}
\end{equation*}
$$

In particular the functions $e_{\mu, \rho_{0}}$ are $N$-fixed and the functions $e_{\mu, x_{0}}$ are $K$-fixed. We have:

$$
\begin{align*}
e_{\mu, x}\left(k a(t) p_{0}\right) & =e_{\mu, x}\left(k p_{0}\right) e^{\mu t} \quad \text { and }  \tag{2.2}\\
e_{\mu, p_{0}}\left(n a(t) x_{0}\right) & =e^{-\mu t} .
\end{align*}
$$

We shall refer to the functions $e_{\mu, x}$ on $L$ and to $e_{\mu, p}$ on $H$ as to the exponential functions on corresponding manifolds.

The Fourier transform of a function $f \in \mathcal{D}(H)$ is defined by the formula

$$
\begin{equation*}
\hat{f}(\lambda, \rho):=\int_{H} e_{-i \lambda-1 / 2, p}(x) f(x) d m(x), \quad \lambda \in \mathbf{C}, p \in L \tag{2.3}
\end{equation*}
$$

For $T \in \mathcal{E}^{\prime}(H)$ we put:

$$
\begin{equation*}
\hat{T}(\lambda, p):=T\left(e_{-i \lambda-1 / 2, p}\right) \tag{2.2}
\end{equation*}
$$

By the very definition one obtains:

$$
\begin{align*}
\left(L_{g} T\right)^{\wedge}(\lambda, \cdot) & =L_{g} \hat{T}(\lambda, \cdot) \quad \text { and }  \tag{2.5}\\
\hat{T}\left(\lambda, k a(t) p_{0}\right) & =e^{-(i \lambda+1 / 2) t} \hat{T}\left(\lambda, k p_{0}\right)
\end{align*}
$$

The Fourier transform originally defined on $C \times L$ is uniquely determined by its values on $\mathbf{C} \times K p_{0}=\mathbf{C} \times \mathbf{S}^{1}$. The restriction of the Fourier transformation to the space $\mathcal{E}^{\prime}(K \backslash H)$ of $K$-invariant and compactly supported distributions is called the spherical Fourier tansform on $H$.

We shall write

$$
\hat{T}(\lambda):=\hat{T}\left(\lambda, p_{0}\right)
$$

Theorem 1.1. [5] The spherical Fourier Transform on $H$ is an isomorphism of the convolution algebra $\mathcal{E}^{\prime}(K \backslash H)$ onto the space $A_{\text {s }}$ of all holomorphic functions $\varphi$ on the complex plane satisfying the following conditions:
$1^{\circ} \varphi(-z)=\varphi(z)$,
$2^{\circ}$ There exist constants $R, m, r>O$ such that

$$
|\varphi(z)| \leqq R(1+|z|)^{m} e^{r|\operatorname{lm} z|} .
$$

We observe that the image of the spherical Fourier transform on $H$ is just the subalgebra of all even elements in the algebra of the classical Fourier transforms of the elements of $\mathcal{E}^{\prime}(\mathbf{R})$.

The Radon transform on $H$ is the mapping which assigns to a function $f \in$ $\mathcal{D}(H)$ the function $R f \in \mathcal{D}(L)$ given by the formula:

$$
\begin{equation*}
R f\left(g \rho_{0}\right):=\int_{N} f\left(g n x_{0}\right) d n \tag{2.6}
\end{equation*}
$$

The value $R f\left(g p_{0}\right)$ can be interpreted as the mean value of the function $f$ on the horocycle $g \xi_{p_{0}}$.

The dual Radon transform maps $\mathcal{E}(L)$ into $\mathcal{E}(H)$ according to the formula:

$$
\begin{equation*}
B \psi\left(g x_{0}\right):=\int_{K} \psi\left(g k p_{0}\right) d k \tag{2.7}
\end{equation*}
$$

The value $B \psi\left(g x_{0}\right)$ is the mean value of $\psi$ on the circle $g k p_{0}$. Both operations commute with the translations acting on corresponding manifolds.

There is a simple relation between the Radon and the Fourier transforms on $H$. Let $f \in \mathcal{D}(H)$. We have

$$
\begin{aligned}
\hat{f}\left(\lambda, k p_{0}\right) & =\int_{H} f(x)\left\langle x \mid k p_{0}\right\rangle^{-i \lambda-1 / 2} d m \\
& \left.\left.=\int_{H} f(k x)\langle x| k p_{0}\right)\right\rangle^{-i \lambda-1 / 2} d m(x) \\
& =\int_{\mathbf{R}} \int_{N} f\left(k a(t) n x_{0}\right)\left\langle a(t) n x_{0} \mid p\right\rangle^{-i \lambda-1 / 2} d n d t \\
& =\int_{\mathbf{R}} e^{\langle i \lambda+1 / 2\rangle t} \int f\left(k a(t) n x_{0}\right) d n d t \\
& =\int_{\mathbf{R}} e^{\langle i \lambda+1 / 2\rangle t} R f\left(k a(t) p_{0}\right) d t .
\end{aligned}
$$

The symmetric space Fourier transform of $f \in \mathcal{D}(H)$ is then equal to the classical Fourier transform of the function

$$
\mathbf{R} \ni t \rightarrow e^{t / 2} R f\left(k a(t) p_{0}\right) .
$$

Let us consider on $\mathcal{D}(L)$ the operator given by the formula

$$
\tilde{\varphi}(\lambda, k):=\int_{\mathbf{R}} e^{\langle i \lambda+1 / 2\rangle t} \varphi\left(k a(t) p_{0}\right) d t
$$

Then we have

$$
\begin{equation*}
\hat{f}\left(\lambda, k p_{0}\right)=(R f)^{\sim}(\lambda, k) \tag{2.8}
\end{equation*}
$$

In particular for $f \in \mathcal{D}(K \backslash H)$ one obtains:

$$
\hat{f}(\lambda)=(R f)^{\sim}(\lambda, k) .
$$

Proposition (1.2.) Let $\psi \in \mathcal{D}(K \backslash L)$. Then $\psi \in R \mathcal{D}(K \backslash H)$ if and only if the function $\mathbf{R} \ni t \rightarrow e^{t / 2} \psi\left(a(t) p_{0}\right)$ is even.

Proof. In virtue of the formula (2.8) and the inversion formula for the classical Fourier transformation we obtain for any $f \in \mathcal{D}(K \backslash H)$ :

$$
\begin{equation*}
F(t):=e^{t / 2} R f\left(a(t) \rho_{0}\right)=\int_{\mathbf{R}} \hat{f}(\lambda) e^{-i \lambda t} d \lambda . \tag{2.9}
\end{equation*}
$$

Taking into account that $\hat{f}$ is an even function (Theorem 1.1) we observe that $F$ is even.

On the other hand if the function

$$
t \rightarrow e^{t / 2} \psi\left(a(t) p_{0}\right)
$$

is even its classic Fourier transform belongs to $\mathcal{D}(K \backslash H)^{\wedge}$, hence for some $\varphi \in$ $\mathcal{D}(K \backslash H)$ we have:

$$
\int_{\mathbf{R}} R \varphi\left(a(t) p_{0}\right) e^{(i \lambda+1 / 2) t} d t=\int_{\mathbf{R}} e^{t / 2} \psi\left(a(t) \rho_{0}\right) e^{i \lambda t} d t
$$

By the injectivity of the classical Fourier transform we obtain $R \varphi=\psi$.
3. Some properties of invariant subspaces in $\mathcal{E}(L)$. The manifold $L$ has an additional structure, namely that of the multiplication of elements by positive reals. This action obviously commutes with the group action. We denote:

$$
\begin{equation*}
\tau(t) f(p):=f\left(e^{t} p\right) \tag{3.1}
\end{equation*}
$$

For a given $\mu \in \mathbf{C}$ let us denote by $\mathcal{E}_{\mu}(L)$ the space of all elements of $\mathcal{E}(L)$ which satisfy:

$$
\begin{equation*}
f(r p)=r^{\mu} f(p), \quad r>0 \tag{3.2}
\end{equation*}
$$

The space $\mathcal{E}_{\mu}$ is invariant with respect to the left regular representation of $G$. The elements of $\mathcal{E}_{\mu}(L)$ are uniquely determined by their values of the orbit

$$
B:=K p_{0}=\left\{p \in L \mid p_{3}=1\right\} .
$$

For any $g \in G$ and $b \in B$ we have:

$$
\begin{aligned}
L_{g} f(b) & =f\left(g^{-1} b\right)=f\left(\frac{g^{-1} b}{\left\langle x_{0} \mid g^{-1} b\right\rangle}\left\langle x_{0} \mid g^{-1} b\right\rangle\right) \\
& =f\left(g^{-1} \cdot b\right)\left\langle g x_{0} \mid b\right\rangle^{\mu}=e_{\mu, b}\left(g x_{0}\right) f\left(g^{-1} \cdot b\right)
\end{aligned}
$$

where

$$
g \cdot b:=\frac{g b}{\left\langle g^{-1} x \mid b\right\rangle}
$$

defines an action of $G$ on $B$. We introduce a representation of $G$ on $\mathcal{E}(B)$ :

$$
U_{g}^{\mu} f(b):=\left\langle g x_{0} \mid b\right\rangle^{\mu} f\left(g^{-1} \cdot b\right)
$$

Then we have

$$
\begin{equation*}
L_{g} f \mid B=U_{g}^{\mu}(f \mid B) \quad \text { for } f \in \mathcal{E}_{\mu}(L) \tag{3.3}
\end{equation*}
$$

Let

$$
\int_{B} \varphi d b:=\int_{K} \varphi\left(k p_{0}\right) d k
$$

A direct calculation shows:

$$
\begin{equation*}
\int_{B}(g \circ b) d b=\int_{B} \varphi(b)\left\langle g x_{0} \mid b\right\rangle^{-1} d b . \tag{3.4}
\end{equation*}
$$

We denote

$$
(\varphi \mid \psi):=\int_{B} \varphi(b) \psi(b) d b
$$

Then we obtain by applying (3.4):

$$
\begin{equation*}
\left(U_{g}^{\mu} \varphi \mid \psi\right)=\left(\varphi \mid U_{g}^{-\mu-1} \psi\right) \tag{3.5}
\end{equation*}
$$

In particular

$$
\left(U_{g}^{\mu-1 / 2} \varphi \mid \psi\right)=\left(\varphi \mid U_{g^{-1}}^{-\mu-1 / 2} \psi\right)
$$

For an integral value $\mu=n$ the space $\mathbb{E}_{n}(L)$ contains an invariant subspace consisting of restrictions to $L$ of all homogeneous polynomials of order $n$. In order to describe all invariant subspaces in $\mathcal{E}_{\mu}(L)$ we introduce the following family of functions on $B$ :

$$
\begin{equation*}
\tilde{e}_{m}\left(k(\theta) \rho_{0}\right):=e^{i m \theta} . \tag{3.6}
\end{equation*}
$$

Proposition 3.1. $1^{\circ}$ If $\mu \in \mathbf{C}$ is not an integer then the representation $U_{\mu}$ is irreducible.
$2^{\circ}$ If $\mu=n \in \mathbf{N} \cup\{0\}$ then in the representation space $\mathcal{E}(B)$ there exist three invariant subspaces:

$$
\begin{aligned}
& \tilde{F}_{n}^{-}:=\text {closed linear span of }\left\{\tilde{e}_{k} \mid k \leqq n\right\}, \\
& \tilde{F}_{n}^{+}:=\text {closed linear span of }\left\{\tilde{e}_{k} \mid k \geqq-n\right\} \text { and } \\
& \tilde{E}_{n}:=\tilde{F}^{-} \cap \tilde{F}^{+} .
\end{aligned}
$$

$3^{\circ}$ If $\mu=-n, n \in \mathbf{N}$ then in the representation space there exist three invariant subspaces:

$$
\begin{aligned}
& \tilde{F}_{-n}^{-}:=\text {closed linear span of }\left\{\tilde{e}_{k} \mid k \leqq-n\right\}, \\
& \tilde{F}_{-n}^{+}:=\text {closed linear span of }\left\{\tilde{e}_{k} \mid k \geqq n\right\} \text { and } \\
& \tilde{E}_{-n}:=\tilde{F}_{-n}^{-}+\tilde{F}_{-n}^{+} .
\end{aligned}
$$

The space $\tilde{F}_{-n}^{ \pm}$is just the annihilator of the space $\tilde{F}_{n-1}^{ \pm}$with respect to the form $(\cdot \mid \cdot)$. We shall denote by $F_{n}^{ \pm}, n \in \mathbf{Z}$ the subspace of $\mathcal{E}_{\mu}$ of those elements
whose restrictions to $B$ belong to $\tilde{F}_{n}^{ \pm}$; similarly $E_{n}:=F_{n}^{-} \cap F_{n}^{+}, n \in \mathbf{N} \cup\{0\}$ and $E_{-n}:=F_{-n}^{-}+F_{-n}^{+}$. By virtue of (3.3) all these spaces are $G$-invariant.

Let us introduce:

$$
\begin{equation*}
e_{m}^{\mu}\left(k(\vartheta) a(t) \rho_{0}\right):=e^{i m \vartheta+\mu t} \tag{3.7}
\end{equation*}
$$

We shall also consider the restrictions of functions on $L$ to the orbit $A \rho_{0}$ which can be considered as a function on $\mathbf{R}$. Given $f \in \mathscr{E}(L)$ we write

$$
\begin{equation*}
f^{A}(t):=f\left(a(t) p_{0}\right) . \tag{3.8}
\end{equation*}
$$

Then we have

$$
(\tau(s) f)^{A}(t)=f^{A}(s+t) .
$$

Let $f \in \mathcal{E}(L)$. By the Fourier series of $f$ we mean the decomposition:

$$
\begin{align*}
& f(k(\theta) \rho)=\sum_{n=-\infty}^{\infty} e^{i n \theta} f_{n}(p), \text { where }  \tag{3.9}\\
& f_{n}(p):=\frac{1}{2 \pi} \int_{0}^{2 \pi} e^{-i n \theta} f(k(\theta) p) d \theta
\end{align*}
$$

The function $f_{n}$ belongs to $\mathcal{E}(L)$ and satisfies

$$
f_{n}(k(\theta) p)=e^{i n \theta} f(p)
$$

The series (3.9) converges in $\mathcal{E}(L)$ see [9]. In particular we have

$$
f\left(k(\theta) a(t) p_{0}\right)=\sum_{n=-\infty}^{\infty} e^{i n \theta} f_{n}^{A}(t) .
$$

For a given closed and $K$-invariant subspace $V \subset \mathcal{E}(L)$ we shall denote by $V_{n}$ the subspace of $V$ consisting of the functions $f_{n}, f \in V$ and by $V_{n}^{A}$ the space of all functions $f_{n}^{A}, f \in V$. In the sequel we are studying the invariance properties of the spaces $V_{n}^{A}$ for $V$ being $G$-invariant.

Proposition 3.2. Let $V \subset \mathcal{E}(L)$ be a closed and $G$-invariant subspace. Assume that $f \in V_{0}^{A}, \varphi \in \mathcal{D}(\mathbf{R})$ and suppose that $e^{t / 2} \varphi$ is even. Then $f * \varphi \in V_{0}^{A}$.

Proof. Let $u \in \mathcal{D}(K \backslash G / K)$ and $\psi \in V_{0}=V \cap \mathcal{E}(K \backslash L)$. The function

$$
L \ni p \rightarrow \int_{G} u(g) \psi(g p) d g
$$

also belongs to the space $V_{0}$ and then the function

$$
\mathbf{R} \ni t \rightarrow \int_{G} u(g) \psi\left(g a(t) p_{0}\right) d g
$$

is an element of $V_{0}^{A}$.
Representing the Haar measure on $G$ in the form (1.5) we obtain:

$$
\begin{aligned}
I(t) & :=\int_{G} u(g) \psi\left(g a(t) p_{0}\right) d g \\
& =\int_{K} \int_{\mathbf{R}} \int_{N} u(k a(s) n) \psi\left(k a(s) n a(t) p_{0}\right) e^{s} d k d s d n \\
& =\int_{\mathbf{R}} \int_{N} u(a(s) n) d n \psi\left(a(s) a(t) p_{0}\right) e^{s} d s \\
& =\int_{\mathbf{R}} R u\left(a(s) p_{0}\right) e^{s} \psi\left(a(s-t) p_{0}\right) d s .
\end{aligned}
$$

According to Proposition 1.2 the function

$$
s \rightarrow e^{s / 2} R u\left(a(s) p_{0}\right)
$$

is even hence

$$
e^{s} R u\left(a(s) p_{0}\right)=R u\left(a(-s) p_{0}\right) .
$$

We obtain

$$
I(t)=\int_{\mathbf{R}} R u\left(a(-s) p_{0}\right) \psi\left(a(s-t) p_{0}\right) d s=f * \varphi(t)
$$

where

$$
f(t):=\psi\left(a(t) p_{0}\right)=\psi^{A}(t) \quad \text { and } \quad \varphi(s):=\operatorname{Ru}\left(a(s) p_{0}\right) .
$$

The proof follows.
Corollary 3.3. Let $f \in V_{0}^{A}$ and let $T \in \mathcal{E}^{\prime}(\mathbf{R})$ be such that $e^{t / 2} T$ is even. Then $T * f \in V_{0}^{A}$.

Proof. Proposition 3.2 proves the statement for the distributions given by the test functions. By the density of $\mathcal{D}(\mathbf{R})$ in $\mathcal{E}^{\prime}(\mathbf{R})$ and the continuity of the mapping

$$
\mathcal{E}^{\prime}(\mathbf{R}) \times \mathcal{E}(\mathbf{R}) \ni(T, f) \rightarrow T * f \in \mathcal{E}(R)
$$

the proof follows.

Let us denote by $\mathcal{E}_{s}^{\prime}(\mathbf{R})$ the algebra of all even elements of $\mathcal{E}(\mathbf{R})$.
Corollary 3.4. If $V$ is closed and $G$-invariant then the space $e^{t / 2} V_{0}^{A}$ of all functions of the form $e^{t / 2} f, f \in V_{0}^{A}$ is a $\mathscr{E}_{s}^{\prime}(\mathbf{R})$-convolution module.

Now, we are going to deduce the same invariance property for functions which not necessarily belong to $V_{0}$ but satisfy a special condition of analicity.

Let us define

$$
D_{0} \varphi:=\varphi^{\prime}(0), \quad \varphi \in \mathcal{E}(\mathbf{R}) .
$$

Definition. A function $f \in \mathcal{E}(L)$ will be called $A$-analytic if

$$
\begin{equation*}
\tau(t) f\left(g p_{0}\right)=\sum_{m=0}^{\infty} \frac{t^{m}}{m!} D_{0}^{m}\left(l_{g^{-1}} f\right)^{A} \tag{3.10}
\end{equation*}
$$

and the series converges in $\mathcal{E}(\mathbf{R})$ for all $t \in \mathbf{R}$.
Given $f \in \mathcal{E}(L)$ we denote by $V(f)$ the closed linear span of all functions $L_{g} f, g \in G$.

We are going to prove:
Proposition 3.5. Let $f \in \mathcal{E}(L)$ be $A$-analytic. Then the function

$$
\psi:=e^{t / 2} \tau(t) f+e^{-t / 2} \tau(-t) f
$$

belongs to $V(f)$ for all $t \in \mathbf{R}$.
Proof. As proved in [5] every differential operator $D$ on $L$ which commutes with the regular representation is of the form:

$$
\begin{equation*}
D f\left(g p_{0}\right)=P\left(D_{0}\right)\left(L_{g^{-1}} f\right)^{A}, \tag{3.11}
\end{equation*}
$$

where $P(\cdot)$ is a polynomial with constant coefficients. Let us denote by $\mathrm{g}, \mathrm{a}, \mathrm{n}$ the Lie algebras of the groups $G, A, N$ respectively. By $\mathfrak{H}(\mathrm{g})$ we denote the enveloping algebra of g and by $Z(\mathrm{~g})$ its center. The differential of the regular representation of $G$ on $\mathcal{E}(L)$ give us a representation of $\mathfrak{U}(\mathrm{g})$ on $\mathcal{E}(L)$ by means of differential operators. This action of $\mathfrak{U}(\mathrm{g})$ conserves closed $G$-invariant subspaces. In particular the elements of the center $Z(\mathrm{~g})$ define differential operators on $\mathcal{E}(L)$ commuting with the regular representation, thus being of the form (3.11). The determination of the corresponding polynomial $P$ can be done with the aid of the Harish-Chandra isomorphism ([9] Lemma 2.3.4). If $Z \in Z(\mathrm{~g})$ we have

$$
\begin{equation*}
Z f\left(g p_{0}\right)=L_{g^{-1}} Z f\left(p_{0}\right)=Z\left(L_{g^{-1}} f\right)\left(p_{0}\right) \tag{3.12}
\end{equation*}
$$

On the other hand $Z=Y+Q$, where $Y \in \mathfrak{U}(a)$ and $Q$ belongs to the ideal $\mathfrak{U}(\mathrm{g}) \mathrm{n}$. Since the elements of the ideal vanish on $\mathcal{E}(L)(=\mathcal{E}(G / N))$ we obtain $Z f=Y f$.

Let us identify $\mathfrak{a} \cong \mathbf{R}$ by chosing $d / d t$ as a base in $a$. The element $Y$ considered as a polynomial on $a^{*} \cong \mathbf{R}$ is mapped into a symmetric polynomial under the application $\gamma$ defined by:

$$
\gamma p(x):=\rho\left(x-\frac{1}{2}\right), \quad x \in \mathbf{R} \quad \text { ([9] p. 168). }
$$

Let us consider $Y_{0}:=\gamma^{-1}\left(x^{2}\right)$ and let $Z_{0}=Y_{0}+Q \in Z(\mathrm{~g})$. Then

$$
Y_{0}=\left(\frac{d}{d t}+\frac{1}{2}\right)^{2}=: X^{2}
$$

The differential operator

$$
X=\frac{d}{d t}+\frac{1}{2}
$$

assigns to a function $\varphi \in \mathcal{E}(\mathbf{R})$ the element

$$
e^{-t / 2} \frac{d}{d t} e^{t / 2} \varphi
$$

In virtue of (3.12) and the last statement we obtain
Lemma 3.6. For any $f \in \mathcal{E}(L)$ the function

$$
\begin{align*}
L \ni g p_{0} \rightarrow Z_{0} f\left(g p_{0}\right) & =Y_{0} f\left(g p_{0}\right)=X^{2}\left(L_{g^{-1}} f\right)^{A}  \tag{3.13}\\
& =e^{t / 2} \frac{d^{2}}{d t^{2}} e^{t / 2}\left(L_{g^{-1}} f\right)^{A}(0)
\end{align*}
$$

belongs to $V(f)$.
Let us note that, assuming the convergence of the series in question we can write for $\varphi \in \mathcal{E}(\mathbf{R})$ :

$$
\begin{align*}
& \left(2 \sum_{k=0} \frac{t^{2 k}}{(2 k)!} e^{-t / 2} \frac{d^{2 k}}{d t^{2 k}} e^{t / 2} \varphi\right)(0)  \tag{3.14}\\
& \quad=e^{-t / 2}\left(\sum_{k=0} \frac{t^{k}}{k!} D_{0}^{k}+\sum_{k=0} \frac{(-t)^{k}}{k!} D_{0}^{k}\right) e^{t / 2} \varphi \\
& \quad=e^{-t / 2}\left((\tau(t)+\tau(-t)) e^{t / 2} \varphi\right)(0) \\
& \quad=e^{t / 2} \varphi(t)+e^{-t / 2} \varphi(-t)
\end{align*}
$$

If we apply the formula to the function $\left(L_{g^{-1}} f\right)^{A}$, (the convergence of the series is assured by the supposition of $f$ being $A$-analytic) the right hand side becomes the function $\varphi$ and the left hand side as a function of the variable $g p_{0}$ belongs to $V(f)$ in virtue of the Lemma 3.6. The proof follows.

As a corollary we obtain
Theorem 3.7. Let $f \in \mathcal{E}(L)$ be an A-analytic function. Then the space $\left(e^{t / 2} V(f)\right)^{A} \subset \mathcal{E}(\mathbf{R})$ is an $\mathcal{E}_{s}^{\prime}(\mathbf{R})$-convolution module.

Proof. If $T \in \mathcal{E}^{\prime}(\mathbf{R})$ is symmetric then

$$
T_{t}\left(e^{t / 2} f\left(g a(t) p_{0}\right)+e^{-t / 2} f\left(g a(-t) \rho_{0}\right)\right)=2 T\left(e^{t / 2}\left(L_{g^{-1}} f\right)^{A}\right)
$$

As a function of $\rho=g \rho_{0}$ this element belongs to $V(f)$ according to Proposition 3.5. The function

$$
R \ni s \rightarrow T_{t}\left(e^{t / 2} f\left(a(s+t) \rho_{0}\right)\right)=e^{-s / 2} T *\left(e^{t / 2} f\right)^{A}(s)
$$

belongs then to $V(f)^{A}$. This ends the proof.
Corollary 3.4 and Theorem 3.7 suggest the following
Conjecture. If $V \subset \mathcal{E}(L)$ is closed and $G$-invariant then $e^{t / 2} V^{A}$ is an $\mathcal{E}_{s}^{\prime}(\mathbf{R})$ convolution module.
4. Spectral analysis for $\mathcal{E}_{s}^{\prime}(\mathbf{R})$-modules. The spectral analysis theorem of $L$. Schwartz describes translation invariant closed subspaces in $\mathcal{E}(\mathbf{R})$ as the spaces generated by functions of the form $x^{m} e^{i \lambda x}, \lambda \in \mathbf{C}, m \in \mathbf{N}$. A closed subspace $V \subset \mathcal{E}(\mathbf{R})$ is translation invariant if and only if it is an $\mathcal{E}^{\prime}(\mathbf{R})$ convolution module.

Let us consider

$$
\begin{equation*}
V^{\perp}:=\left\{T \in \mathcal{E}^{\prime}(\mathbf{R}) \mid\langle T, \varphi\rangle=0, \varphi \in V\right\} . \tag{4.1}
\end{equation*}
$$

The annihilator space $V^{\perp}$ is an ideal in the convolution algebra $\mathcal{E}^{\prime}(\mathbf{R})$. By the Hahn-Banach theorem

$$
\begin{equation*}
V=\left(V^{\perp}\right)^{\perp}:=\{\varphi \in \mathcal{E}(\mathbf{R}) \mid\langle T, \varphi\rangle=0, T \in V\} . \tag{4.2}
\end{equation*}
$$

The problem of describing the closed $\mathcal{E}^{\prime}(\mathbf{R})$-modules reduces to the description of ideals of $\mathscr{E}^{\prime}(\mathbf{R})$.

In this section we shall denote by $\hat{f}$ the classical Fourier transform.
The space $J:=\left(V^{\perp}\right)^{\wedge}$ forms an ideal in the algebra $A=\left(\mathcal{E}^{\prime}(\mathbf{R})\right)^{\wedge}$. The theorem of L. Schwartz characterizes the ideals of $A$ in terms of its spectrum. Given an ideal $J \subset A$ we denote:

$$
\operatorname{Sp} J:=\{\lambda \in \mathbf{C} \mid \psi(\lambda)=0, \psi \in J\} .
$$

If $\lambda \in \operatorname{Sp} J$ we denote by $m(\lambda)$ the multiplicity of $\lambda$, that is the maximal natural number such that all elements of $J$ have in $\lambda$ zero of order $\leqq m(\lambda)$.

Theorem 4.1. [8] (Spectral synthesis theorem) Each ideal $J \subset A$ is uniquely determined by the set of pairs $(\lambda, m(\lambda)), \lambda \in \operatorname{Sp} J$.

If particular the theorem states that $V \neq 0$ if and only if $\operatorname{Sp} J \neq 0$ for $J:=\left(V^{\perp}\right)^{\wedge}, V$ being a closed $\mathcal{E}^{\prime}(\mathbf{R})$-module.

A simple calculation proves that $\lambda \in \operatorname{Sp} J$ with multiplicity $\geqq m(\lambda)$ if and only if the function $x^{m} e^{i \lambda x}$ belongs to $V$ for every $m \leqq m(\lambda)$. The spectrum of $V$ is defined as the spectrum of $\left(V^{\perp}\right)^{\wedge}$. In this way one obtains the version of the Schwartz theorem mentioned at the beginning of the section:

Theorem 4.1'. Every translation invariant and closed subspace $V \subset \mathcal{E}(\mathbf{R})$ is generated by all functions of the form

$$
x^{m} e^{i \lambda x}, \quad m \in \mathbf{N} \cup\{0\}, \quad \lambda \in \mathbf{C}
$$

contained in $V$. If $V \neq \mathcal{E}(\mathbf{R})$ then $\mathrm{Sp} J$ is a countable set without accumulation points.

The space $\mathscr{E}_{s}^{\prime}(\mathbf{R})$ of all even distributions forms a subalgebra of $\mathcal{E}^{\prime}(\mathbf{R})$ whose Fourier transforms is the algebra $A_{s}$ of all even elements in $A$. The spectral synthesis theorem for $A_{s}$ was proved in [4].

Theorem 4.2. Let $J$ be an ideal in $A_{s}$ and let $I$ be the ideal generated in $A$ by $J$. Then $J=I \cap A_{s}$. In particular $J$ is uniquely determined by the set of pairs $(\lambda, m(\lambda)), \lambda \in \operatorname{Sp} J=\operatorname{Sp} I$.

This theorem permits us to describe the $\mathcal{E}_{s}^{\prime}(\mathbf{R})$-modules in $\mathcal{E}_{s}(\mathbf{R})$. Let $V \subset$ $\mathcal{E}_{s}(\mathbf{R})$ be such a module. We define

$$
\begin{equation*}
V^{\perp}:=\left(V^{\perp}\right) \cap \mathcal{E}_{s}^{\prime}(\mathbf{R}) \quad \text { and } \quad\left(V^{\perp}\right)^{\perp}:=\left(V^{\perp}\right)^{\perp} \cap \mathcal{E}_{s}(\mathbf{R}) . \tag{4.3}
\end{equation*}
$$

Again, by the Hahn-Banach theorem we have

$$
\begin{equation*}
V=\left(V^{\perp}\right)^{\perp} \tag{4.4}
\end{equation*}
$$

Let $T \in V^{\perp}$ and let $\lambda_{0}$ be zero of order $m$ of the transform $\hat{T}$, that is let

$$
\frac{d^{k}}{d \lambda^{k}} \hat{T}\left(\lambda_{0}\right)=0 \quad \text { for all } k \leqq m
$$

Then we have

$$
0=\left.\frac{d^{k}}{d \lambda^{k}} T_{x}\left(e^{i \lambda x}\right)\right|_{\lambda=\lambda_{0}}=T_{x}\left((i x)^{k} e^{i \lambda_{0} x}\right) \quad \text { for all } k \leqq m
$$

and by the symmetry the same holds for $-\lambda_{0}$. In virtue of (4.4) we deduce that $\lambda \in \operatorname{Sp}\left(V^{\perp}\right)^{\wedge}$ with the multiplicity $m(\lambda)$ if and only if the symmetric parts of the functions $x^{k} e^{i \lambda x}$ and $x^{k} e^{-i \lambda x}$ belong to $V$ for all $k \leqq m(\lambda)$.

By applying Theorem 4.2 we get
Theorem 4.2'. Each closed $\mathfrak{E}_{s}^{\prime}(\mathbf{R})$-module $V \subset \mathcal{E}_{s}(\mathbf{R})$ is generated by all functions of the form

$$
\begin{aligned}
& x^{k} \operatorname{ch} \lambda x, \lambda \in \operatorname{Sp} V, k \text {-even and less or equal to } m(\lambda) \text { and } \\
& x^{k} \operatorname{sh} \lambda x, \lambda \in V, k \text {-odd and less or equal to } m(\lambda) .
\end{aligned}
$$

$\mathrm{Sp} V$ is countable and has no accumulation points.
The last result solves the problem of spectral synthesis in the space $\mathcal{E}(K \backslash H)$ (see [1] or [11]). In order to obtain at least the spectral analysis in the space $\mathcal{E}(L)$ we need information about the $\mathcal{E}_{s}^{\prime}$-modules in the whole space $\mathcal{E}(\mathbf{R})$.

In the sequel we denote by $A_{a}, \mathfrak{E}_{a}(\mathbf{R}), \mathcal{E}_{a}^{\prime}(\mathbf{R})$ the subspaces of all alternating elements in $A, \mathcal{E}(\mathbf{R})$ and $\mathcal{E}^{\prime}(\mathbf{R})$ respectively. The spaces $\mathcal{E}_{s}(\mathbf{R})$ and $\mathcal{E}_{a}(\mathbf{R})$ are $\mathcal{E}_{s}^{\prime}$-submodules of $\mathcal{E}(\mathbf{R})$ which contain no exponentials. We can suppose that on studying $\mathcal{E}_{s}^{\prime}$-modules we should rather consider the functions of the form

$$
a x^{k} e^{\lambda x}+b x^{m} e^{-\lambda x}
$$

as elementary functions on $\mathbf{R}$.
At the beginning let us assume that $V \subset \mathcal{E}_{a}(\mathbf{R})$ is a closed $\mathscr{E}_{s}^{\prime}(\mathbf{R})$-module. The operator of derivation $D: f \rightarrow f^{\prime}$ is continuous, commutes with the convolution and maps $\mathcal{E}_{s}(\mathbf{R})$ onto $\mathcal{E}_{a}(\mathbf{R})$ with kernel consisting of constant functions. The space

$$
W:=\left(D^{-1} V\right) \cap \mathcal{E}_{s}(\mathbf{R})
$$

is closed and is a $\mathscr{E}_{s}^{\prime}(\mathbf{R})$-submodule of $\mathscr{E}_{s}(\mathbf{R})$ hence is generated by the functions

$$
\begin{aligned}
x^{2 k} \operatorname{ch} \lambda x, & \mathrm{Sp} W, 2 k \leqq m(\lambda) \text { and } \\
x^{2 k+1} \operatorname{sh} \lambda x, & \mathrm{Sp} W, 2 k+1 \leqq m(\lambda) .
\end{aligned}
$$

The module $V$ is then generated by the derivatives of the above functions. If $\lambda \in \operatorname{Sp} W$ and $\lambda \neq 0$ the multiplicity of $\lambda$ in $V$ understood as the highest power of the variable $x$ appearing as a factor in this set of generating functions is equal to max $(0, m(\lambda))$. The multiplicity of $\lambda=0$ in $V$ is equal to $\max (0, m(0))$.

We have in this way a counterpart of Theorem 4.2':
Proposition 4.3. If $V \subset \mathcal{E}_{a}(\mathbf{R})$ is a closed $\mathcal{E}_{s}(\mathbf{R})$-module then $V$ is generated by all functions of the form

$$
x^{2} \operatorname{sh} \lambda x \text { and } x^{2 k-1} \operatorname{ch} x
$$

contained in $V$. If $V \neq \mathcal{E}_{a}(\mathbf{R})$ then the set of $\lambda$ 's for which some of above functions appear in $V$ is countable and has no accumulation points.

Passing to the general case, let $V$ be a nontrivial, closed $\mathcal{E}_{s}^{\prime}$-module in $\mathcal{E}(\mathbf{R})$ and $V^{\perp} \subset \mathscr{E}_{s}^{\prime}(\mathbf{R})$ its annihilator which is also an $\mathscr{E}_{s}^{\prime}$-module. The space $J:=$ $\left(V^{\perp}\right)^{\wedge} \subset A$ is then an $A_{s}$-module. We shall consider now the possible relations of $J$ to $A_{s}$ and $A_{a}$.

Case 1. If $J \subset A_{s}$ then by Theorem 4.2 it is determined uniquely by $\operatorname{Sp} J$ and the multiplicities. The elements of $V^{\perp}$ annihilate the whole space $\mathcal{E}_{a}(\mathbf{R})$ as well as the functions:

$$
\begin{aligned}
& x^{k} \operatorname{ch} \lambda x, \lambda \in \operatorname{Sp} V, k+2 m \leqq m(\lambda) \\
& x^{k} \operatorname{sh} \lambda x, \lambda \in \operatorname{Sp} V, k=2 m+1 \leqq m(\lambda)
\end{aligned}
$$

In this case $V=\mathcal{E}_{a}(\mathbf{R})+\left(V^{\perp}\right)^{\perp}=\mathcal{E}_{a}(\mathbf{R})+\left(V \cap \mathcal{E}_{s}(\mathbf{R})\right)$.
Case 2. Assume $J \subset A_{a}$. Then we have $V^{\perp} \subset \mathscr{E}_{a}^{\prime}(\mathbf{R})$ and

$$
V=\left(V^{\perp}\right)^{\perp}=\mathcal{E}_{s}(\mathbf{R})+\left(V \cap \mathcal{E}_{a}(\mathbf{R})\right)
$$

Applying Proposition 4.3 to the module $V \cap \mathcal{E}_{a}(\mathbf{R})$ we get:
The module $V$ is generated by $\mathcal{E}_{s}(\mathbf{R})$ and by all functions of the form $x^{2 k-1} \operatorname{ch} \lambda x$ and $x^{2 k} \operatorname{sh} \lambda x$ contained in $V$.

From now we can assume that neither $\left(J \subset \mathcal{A}_{s}\right)$ nor $J \subset A_{a}$. Let us consider the subspaces:

$$
\begin{array}{ll}
W_{s}:=\left\{f \in A_{s} \mid \exists \varphi \in J\right. & f(\lambda)=\varphi(\lambda)+\varphi(-\lambda)\}, \\
W_{a}:=\left\{f \in A_{a} \mid \exists \varphi \in J\right. & f(\lambda)=\varphi(\lambda)-\varphi(-\lambda)\} .
\end{array}
$$

The space $W_{s}$ is an ideal in $A_{s}$ and $W_{a}$ is an $A_{s^{-}}$submodule of $A_{a}$. According to our assumption $W_{s} \neq 0$ and $W_{a} \neq 0$.

Case 3. Suppose $W_{s} \neq A_{s}$. According to Theorem $4.2 \mathrm{Sp} W_{s}$ contains at least one point, say $\lambda_{0} \in \mathbf{C}$. Then for any $T \in V^{\perp}$ one has

$$
0=\hat{T}\left(\lambda_{0}\right)+\hat{T}\left(-\lambda_{0}\right)=T_{x}\left(\operatorname{ch} \lambda_{0} x\right)
$$

It means that the function ch $\lambda_{0} x$ belongs to $V$.
Case 4. Let $W_{a} \neq A_{a}$. There exist an ideal $I \subset A_{s}$ such that

$$
W_{a}:=\{\psi \mid \psi(\lambda)=\lambda \varphi(\lambda), \varphi \in I\} .
$$

Since $I$ is nontrivial we can choose $\lambda_{0} \in \operatorname{Sp} I$. Supposing $\lambda_{0} \neq 0$ we obtain for every $T \in V^{\perp}$ :

$$
0=\frac{1}{\lambda_{0}}\left(\hat{T}\left(\lambda_{0}\right)-\hat{T}\left(-\lambda_{0}\right)\right)=\frac{2}{\lambda_{0}} T_{x}\left(\operatorname{sh} \lambda_{0} x\right)
$$

which implies that $\operatorname{sh} \lambda_{0} x \in V$.
If $\operatorname{Sp} I=\{0\}$ and $m(0)=2 k>0$ then

$$
W_{a}=\left\{\lambda^{2 k+1} \varphi \mid \varphi \in A_{s}\right\} .
$$

This being the case we have for all $T \in V^{\perp}$ :

$$
0=\left.\frac{d^{2 k-1}}{d \lambda^{2 k-1}}(\hat{T}(\lambda)-\hat{T}(-\lambda))\right|_{\lambda=0}=2 T_{x}\left((i x)^{2 k-1}\right)
$$

which means that $x^{2 k-1} \in V$.
(Note that in cases 3 and 4 if zero is the unique element of the spectrum then $V$ is generated by polynomials.)

Case 5. It remains to consider the case $W_{a}=A_{a}$ and $W_{s}=A_{s}$. Let us define a relation $R$ in $A_{s} \times A_{a}$ by the formula:

$$
x R y \quad \text { if } \quad x+y \in J .
$$

If $x R y$ and $x R y^{\prime}$ then

$$
y-y^{\prime} \in J \cap A_{a}=: J_{a}
$$

We can lift the relation to $A_{s} \times\left(A_{a} / J_{a}\right)$ by putting

$$
x \tilde{R}[y] \text { if } x R y .
$$

This relation is an application commuting with the multiplication by elements of $A_{s}: \tilde{R}(f x)=f \tilde{R} x$.

Since the algebra $A_{s}$ contains the unity we obtain:

$$
\tilde{R} x=\tilde{R}(x \cdot 1)=x \tilde{R}(1)
$$

Let us fix a representant $y_{0}$ of the class $\tilde{R}(1)$. Then one has

$$
J=\left\{x+x y_{0}+J_{a} \mid x \in A_{s}\right\} .
$$

If in particular $y=x+x y_{0}+j, j \in J_{a}$ then

$$
x(\lambda)=\frac{1}{2}(y(\lambda)+y(-\lambda)) .
$$

Denoting $\check{y}(\lambda):=y(-\lambda)$ we have for every $y \in J$ :

$$
y-\frac{1}{2}(y+\check{y})\left(y_{0}+1\right)=\left(1+y_{0}\right) y+\left(1-y_{0}\right) \check{y} \in J_{a} .
$$

Case 5A. If $J_{a}=A_{a}$ then $J+A$ which corresponds to the case $V=0$ which was excluded.

Case 5B. Suppose $J_{a}=0$. This being the case the elements of $J$ are of the form $y=x\left(1+y_{0}\right), x \in A_{s}$. If $1+y_{0}$ is everywhere distinct from zero then the holomorphic and alternating function $y_{0}$ does not take the values 1 and -1 . In virtue of the Picard theorem $y_{0}$ is constant hence zero and we obtain $J=A_{s}$ once again.

Case 5C. Finally we suppose $A_{a} \neq J_{a} \neq 0$. Let $I$ be the ideal in $A_{s}$ such that all elements of $J_{a}$ are of the form $y(\lambda)=\lambda f(\lambda), f \in I$. For any $\lambda_{0} \in \mathrm{Sp} I$ and $T \in V^{\perp}$ we obtain:

$$
\begin{aligned}
0 & =\left(1+y_{0}\left(\lambda_{0}\right)\right) \hat{T}\left(\lambda_{0}\right)+\left(1-y_{0}\left(\lambda_{0}\right)\right) \hat{T}\left(-\lambda_{0}\right) \\
& =T_{x}\left(\left(1-y_{0}\left(\lambda_{0}\right)\right) e^{i \lambda_{0} x}\right)+\left(1+y_{0}\left(\lambda_{0}\right)\right) e^{-\lambda_{0} x} .
\end{aligned}
$$

This implies that the function

$$
a e^{i \lambda_{0} x}+b e^{-i \lambda_{0} x}
$$

belongs to $V$ for $a=1-y_{0}\left(\lambda_{0}\right)$ and $b=1+y_{0}\left(\lambda_{0}\right)$. The function is nontrivial for each $\lambda_{0}$.

All cases can be summarized in the following way:
Theorem 4.4 If $V$ is a nontrivial closed $\mathcal{E}_{s}(\mathbf{R})$-module in $\mathcal{E}(\mathbf{R})$ then there exists $\lambda \in \mathbf{C}$ such that the function

$$
a e^{i \lambda x}+b e^{-i \lambda x}
$$

is nontrivial and belongs to $V$, or the identity function $f(x)=x$ belongs to $V$. If the module $V$ does not contain any function of the first kind then $V$ is finite-dimensional and is generated by polynomials.
5. Spectral analysis theorems for $\mathcal{E}(L)$. In this section we apply Theorems 4.1-4 to obtain results about spectral analysis and in several cases even spectral synthesis in the space $\mathcal{E}(L)$.

First let us distinguish a particular invariant subspace $\mathcal{N} \subset \mathcal{E}(L)$. Let

$$
\mathcal{N}:=\left\{f \in \mathcal{E}(L) \mid \int_{K} f(g k \rho) d k=0, \quad g \in G, \quad p \in L\right\} .
$$

This is the maximal invariant subspace of $\mathcal{E}(L)$ which does not contain any $K$-fixed elements. Let us note that in the case of the manifold $H$ such a subspace is trivial. This fact is the reason of principal differences between the analysis on $H$ and $L$. By the very definition it's clear that $\mathcal{N}$ is closed, $G$-invariant and moreover invariant with respect to the operators $\tau(t), t \in \mathbf{R}$. Since the regular representation of $G$ commutes with $\tau$ the subspaces $\mathcal{N}_{q} \subset \mathcal{N}$ are also invariant. By Theorem 4.1' we obtain:

Lemma 5.1. The space $\mathcal{N}_{q}{ }^{A}$ is uniquely determined by its spectrum $\operatorname{Sp} \mathcal{N}_{q}{ }^{A}$ and the corresponding multiplicity function $m_{q}(\lambda) . \mathcal{N}_{q}{ }^{A}$ is lineary generated by the functions of the form

$$
t^{m} e^{i \lambda t}, \quad \lambda \in \operatorname{Sp} \mathcal{N}_{q}^{A}, \quad m \leqq m_{q}(\lambda)
$$

According to the lemma the functions $e_{q}^{i \lambda}$ belong to $\mathcal{N}$ for each $\lambda \in \operatorname{Sp} \mathcal{N}_{q}{ }^{A}$. Now, the integral equation defining $\mathcal{N}$ can be represented in the following way:

$$
0=\int_{K} L_{g^{-1}} e_{q}^{i \lambda}\left(k p_{0}\right) d k=\left(U_{g^{-1}}^{i \lambda} \tilde{e}_{q} \mid \tilde{e}_{0}\right)=\left(\tilde{e}_{q} \mid U_{g^{-1}}^{1 i \lambda-1} \tilde{e}_{0}\right)
$$

for all $g \in G$. The space generated by the vectors

$$
U_{g}^{-i \lambda-1} e_{0}, \quad g \in G
$$

differs from the whole space if and only if $-i \lambda-1=n \in \mathbf{N}$ (Proposition 3.1).
Thus $\mathcal{N}_{q} \cap \mathcal{E}_{i \lambda} \neq 0$ if and only if $i \lambda=-n-1$ and in this case $n+1 \leqq|q|$ according to Proposition 3.1.3. We have obtained

Proposition 5.2. The space $\mathcal{N}_{q}, q \in \mathbf{Z}$ is finite-dimensional and linearly generated by the functions of the form:

$$
\begin{aligned}
& \psi_{q, n, s}\left(k(\theta) a(t) p_{0}\right):=t^{s} e^{-n t+i q \theta} \quad \text { with } \\
& n \in \mathbf{N}, \quad n \leqq|q| \quad \text { and } \quad s \leqq m_{q}(-n) .
\end{aligned}
$$

In particular this means that the representation of $G$ in $\mathcal{N}$ is admissible and the character $e^{i q \theta}$ occurs in $\mathcal{N}$ with finite multiplicity which is less or equal to

$$
\sum_{n=1}^{q} m_{q}(-n)
$$

The calculation of the multiplicities $m_{q}(\lambda) \quad q \in \mathbf{Z}$ is an open problem.
Now, consider a closed and $G$-invariant subspace $V \subset \mathcal{N}$. According to Lemma 3.6 the finite-dimensional spaces $e^{t / 2} \mathcal{N}_{q}{ }^{A}$ are invariant with respect to the operator $d^{2} / d t^{2}$. The eigenvectors of this operator in the space $e^{t / 2} \mathcal{N}_{q}{ }^{A}$ are the functions

$$
e^{\langle-n+1 / 2\rangle t}, \quad n \leqq|q|
$$

and the elements of the Jordan base of the operator are of the form

$$
P(t) e^{(n+1 / 2) t}
$$

where $P$ is a polynomial of order less than or equal to $m_{q}(-n)$. This leads to

Proposition 5.3. Every closed and $G$-invariant subspace $V \subset \mathcal{N}$ is linearly generated by combinations of the functions $\psi_{q, n, s}$ contained in $V$.

Now, our purpose is to prove a spectral analysis theorem in $\mathcal{E}(L)$ hence, after obtaining even the spectral synthesis for subspaces of $\mathcal{N}$ it is sufficient to consider invariant subspaces in $\mathcal{E}(L)$ whose intersection with $\mathcal{N}$ are trivial. If $V \neq 0$ but $\mathcal{N} \cap V=0$ we have $V_{0} \neq 0$. In virtue of Corollary 3.4 the space $V:=e^{t / 2} V_{0}^{A}$ is an $\mathcal{E}_{s}^{\prime}(\mathbf{R})$-module. By Theorem 4.4 the space $V$ contains for some $\lambda$ the function

$$
a e^{i \lambda t}+b e^{-1 \lambda t} \quad \text { with } \quad|a|^{2}+|b|^{2} \neq 0
$$

or the function $f(t)=t$.
The space $V_{0}$ then contains the element

$$
\psi\left(k(\theta) a(t) \rho_{0}\right):=a e^{(i \lambda-1 / 2) t}+b e^{-(i \lambda+1 / 2) t}
$$

or the function

$$
\varphi\left(k(\theta) a(t) \rho_{0}\right):=t e^{-i / 2}
$$

The first function can be written as

$$
a e_{i \lambda-1 / 2, x_{0}}+b e_{-(i \lambda+1 / 2), x_{0}}
$$

and the second as

$$
e_{-1 / 2, x_{0}}^{\langle 1\rangle}:=\left.\frac{\partial}{\partial \lambda}\right|_{\lambda=-1 / 2} e_{\lambda, x_{0}} .
$$

Taking into account the $G$-invariance of $V$ we obtain:
Theorem 5.3. If $V \subset \mathcal{E}(L)$ is nontrivial, closed and $G$-invariant but $V \cap \mathcal{N}=$ 0 then either

$$
e_{-1 / 2, x}^{\langle 1\rangle} \in V \quad \text { for all } x \in H,
$$

or there exist $\lambda, a, b, \in \mathbf{C}$ such that $|a|^{2}+|b|^{2}>0$ and for all $x \in H$ the function

$$
a e_{i \lambda-1 / 2, x}+b e_{-(i \lambda+1 / 2), x}
$$

belongs to $V$.
Theorem 5.4. Each $G$-invariant and closed subspace $V \subset \mathcal{E}(L)$ such that $V \cap \mathcal{N} \neq 0$ contains for some $n \in \mathbf{N}$ at least one of the functions

$$
\psi^{ \pm}\left(k(\theta) a(t) \rho_{0}\right)=e^{ \pm i n \theta-n t}
$$

Proof. According to Proposition 5.2 we have in $V \cap \mathcal{N}$ a function of the form

$$
\psi_{q, n, 0}\left(k(\theta) a(t) \rho_{0}\right)=e^{i q \theta-n t}
$$

such that $n \leqq|q|$. This element belongs to $F_{-n}^{+}$or ${F_{-n}^{-}}^{-}$depending on the sign of $q$. Both spaces are carrier spaces of irreducible representations of $G$ hence together with the above element the whole space $F_{-n}^{+}$or $F_{-n}^{-}$belongs to $V$. In particular $\psi_{n}^{+}$or $\psi_{n}^{-}$belongs to $V$.

Theorems 5.3 and 5.4 mean that the spectral analysis holds in the space $\mathcal{E}(L)$ with respect to the elemental family given by the collection of all functions

$$
\psi_{n}^{ \pm}, \quad n \in \mathbf{N} ; \quad e_{-1 / 2, x}^{\langle 1\rangle}, \quad x \in H
$$

and the functions

$$
a e_{i \lambda-1 / 2, x}+b e_{--\langle i \lambda+1 / 2\rangle, x}, \quad x \in H,|a|^{2}+|b|^{2}>0 .
$$

Nevertheless, some additional information can be deduced with the aid of Theorems 3.7 and 4.4. Namely, we obtain

Proposition 5.5. Let $V \subset \mathcal{E}(L)$ be a closed and $G$-invariant subspace. Assume that for some $n \in \mathbf{Z}$ the space $V_{n}$ contains some $A$-analytic function. Then $V_{n}$ contains a function of the form

$$
\varphi_{n}\left(k(\theta) a(t) \rho_{0}\right):=t e^{i n \theta-t / 2}
$$

or the function

$$
a e_{n}^{\langle i \lambda-1 / 2\rangle}+b e_{n}^{-\langle i \lambda+1 / 2\rangle}, \quad|a|^{2}+|b|^{2} \neq 0 .
$$

6. Comments and applications. The space $\mathcal{E}(L)$ is mapped under the dual Radon transform (2.7) continuously into $\mathcal{E}(H)$. Since the mapping $B$ commutes with the group action, the space $\operatorname{ker} B$ is closed and $G$-invariant. By the very definition a function $f$ belongs to $\mathcal{N}$ if and only if for each $t \in \mathbf{R}$ one has $\tau(t) f \in \operatorname{ker} B$. In particular $\mathcal{N} \subset \operatorname{ker} B$ since $\mathcal{N}$ is $\tau$-invariant. In the sequel we shall see that $\mathcal{N} \neq \operatorname{ker} B$. This implies that $\operatorname{ker} B$ constitutes an example of a $G$-invariant subspace which is not $\tau$-invariant.

Let

$$
\varphi_{\lambda}(g k):=B e_{i \lambda-1 / 2, x_{0}}\left(g x_{0}\right)=\int_{K}\left\langle x_{0} \mid g k p_{0}\right\rangle^{i \lambda-1 / 2} d k
$$

The function $\varphi_{\lambda}$ is a zonal spherical function on $H$. It is known [5] that $\varphi_{-\lambda}=\varphi_{\lambda}$ for each $\lambda \in \mathbf{C}$. In particular it means that

$$
e_{i \lambda-1 / 2, x_{0}}-e_{-i \lambda-1 / 2, x_{0}} \in \operatorname{ker} B .
$$

By the $G$-invariance of $\operatorname{ker} B$ and the formula (2.1) the same is true for an arbitrary point $x \in H$. By applyting the derivation with respect to $\lambda$ one obtains

$$
e_{i \lambda-1 / 2, x}^{\langle 1\rangle} \in \operatorname{ker} B \quad \text { for all } \lambda \in \mathbf{C} \text { and } x \in H .
$$

All these functions do not belong to $\mathcal{N}$, hence $\mathcal{N} \neq \operatorname{ker} B$. At the same time we observe that in $\operatorname{ker} B$ appear elementary functions of all frequencies $\lambda \in \mathbf{C}$.

On the Pompeiu problem on $L$. Let $A \subset \mathbf{R}^{n}$ be a compact set and let $f \in L^{\infty}\left(\mathbf{R}^{n}\right)$. The Pompeiu problem consists in asking if the condition:

$$
\int_{g A} f(x) d x=0 \quad \text { for all rigid motions } g
$$

implies $f=0$ almost everywhere on $\mathbf{R}^{n}$. If the answer is positive one says that $A$ has the Pompeiu property. The problem is equivalent to the question if the translations of the characteristic function $1_{A}$ span a dense subset in $L^{1}\left(\mathbf{R}^{n}\right)$. By the Tauberian theorem of Wiener the set $A$ has the Pompeiu property if and only if the Fourier transform $\hat{1}_{A}$ is everywhere different from zero.

Various generalizations of the Pompeiu problem were considered on $\mathbf{R}$ and on symmetric spaces of rank one ([1], [2], [3], [6], [7].)
The results of the last section permit us to formulate a theorem of Pompeiu type for the manifold $L$.
Theorem 6.1. Let $T \in \mathcal{E}^{\prime}(L)$. The system of equations

$$
T\left(L_{g} f\right)=0, \quad \text { for each } g \in G
$$

has in $\mathcal{E}(L)$ only the trivial solution $f \equiv 0$ if and only if the following conditions are satisfied:
a) For every $\lambda \in \mathbf{C}$ the functions

$$
H \ni x \rightarrow T\left(e_{i \lambda-1 / 2, x}\right) \quad \text { and } \quad H \in x \rightarrow T\left(e_{-\langle i \lambda+1 / 2\rangle, x}\right)
$$

are linearly independent.
b) $T\left(e_{-1 / 2, x}^{\langle 1\rangle}\right) \neq 0$ for some $x \in H$.
c) For every $n \in \mathbf{N}$ there exists $Q \in \mathbf{Z}$ such that $|q| \geqq n$ and

$$
T\left(\psi_{q, n, 0}\right) \neq 0
$$

Proof. Let us denote

$$
V:=\left\{f \in \mathcal{E}(L) \mid T\left(L_{g} f\right)=0, g \in G\right\} .
$$

The space $V$ is closed and $G$-invariant. If $V \neq 0$ then according to Proposition 5.2 and Theorem 5.3 one of the following conditions is satisfied:

$$
\begin{aligned}
& \text { 1) } a e_{i \lambda-1 / 2, x}+b e_{-\langle i \lambda+1 / 2\rangle, x} \in V \quad \text { for some } \lambda \in \mathbf{C}, \\
& |a|^{2}+|b|^{2} \neq 0 \quad \text { and all } x \in H, \\
& \text { 2) } e_{-1 / 2, x}^{\langle 1\rangle} V, \quad \text { for all } x \in H,
\end{aligned}
$$

or
3) $V \cap \mathcal{N} \neq 0$ and consequently $F_{-n}^{+}$or ${F_{-n}^{-}}^{\text {b }}$ belongs to $V$ for $n \in \mathbf{N}$.

In the first case the condition $a$ ) is not satisfied: the case 2 ) contradicts $b$ ) and the case 3) contradicts c). Then the conditions a) b) c) are sufficient.

Now, if a) is not satisfied then for some $a, b \in \mathbf{C}$ and $\lambda \in \mathbf{C}$ we have $(a, b) \neq(0,0)$ and

$$
T\left(a e_{i \lambda-1 / 2, x}+b e_{-\langle i \lambda+1 / 2\rangle, x}\right)=0 \quad \text { for all } x \in H
$$

The nontrivial and $G$-invariant space spanned by the functions in parenthesis belong to $V$.

If $b$ ) is not satisfied then the function

$$
e_{-1 / 2, x}^{\langle 1\rangle}
$$

is a solution of the system in question.
If for some $n \in \mathbf{N}$ and all $q \in \mathbf{Z}$ such that $q \geqq n$ (or all $q$ such that $q \leqq-n$ ) we have

$$
T\left(\psi_{q, n, 0}\right)=0
$$

then the space $F_{-n}^{+}$(or $F_{-n}^{-}$) belongs to $V$. This ends the proof.

## References

1. S.C. Bagchi and A. Sitaram, Spherical mean-peridic functions on semisimple Lie groups, Pacific J. Math. 84 (1979), 241-250.
2. C.A. Berenstein and L. Zalcman, Pompeiu's problem on a symmetric space, Comment. Math Helvetici 55 (1980), 593-621.
3. L. Brown, B.M. Schreiber and B.A. Taylor, Spectral sythesis and the Pompeiu problem, Ann Inst. Fourier 23 (1973), 125-154.
4. L. Ehrenpreis and F.I. Mautner, Some properties of the Fourier transform on semi-simple Lie groups, II, Trans. Amer. Math. Soc. 84 (1957), 1-55.
5. S. Helgason, Groups and geometric analysis (Academic Press, New York, 1984).
6. L. Pysiak, On Pompeiu problem on symmetric spaces, preprint.
7. A. Sitaram, An analogue of the Wiener-Tauberian theorem for spherical transforms on semisimple Lie groups, Pacific J. Math. 89 (1980), 439-445.
8. L Schwartz, Théorie générale des functions moyenes-périodiques, Ann. of Math. 48 (1947), 857-929.
9. G. Warner, Harmonic analysis on semi-simple Lie groups, I, II (Springer Verlag, Berlin, 1972).
10. A. Wawrzyñczyk, Group representations and special functions (D. Reidel Comp.-PWN, Warsaw, 1984).
11. Spectral analysis and synthesis on symmetric spaces, J. of Math. Anal. and Appl. 127 (1987), 1-17.
12. Spectral analysis and mean-periodic functions on symmetric spaces of rank one, Bol. Soc. Mat. Mex, 30 (1985), 15-29.

Universidad Antónoma Metropolitana - Iztapalapa, México, México


[^0]:    Received March 30, 1988. This research was partially supported by Sistema Nacional de Investigadores de México.

