# Medium amplitude limit cycles of some classes of generalized Liénard systems 

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## 1. Introduction and statement of the results

The bifurcation of limit cycles by perturbing a planar system which has a continuous family of cycles, i.e. periodic orbits, has been an intensively studied phenomenon; see for instance $[13,16,2]$ and references therein. The simplest planar system having a continuous family of cycles is the linear center, and a special family of its perturbations is given by the generalized polynomial Liénard systems:

$$
\dot{x}=y+\sum_{i=1}^{\mu} \varepsilon^{i} F_{i}(x), \quad \dot{y}=-x+\sum_{i=1}^{\nu} \varepsilon^{i} g_{i}(x)
$$

where $\mu, \nu \in \mathbb{N}, g_{i}(x)$ and $F_{i}(x)$ are polynomials for $i \geq 1$, and $\varepsilon$ is a small parameter.
The classical and generalized Liénard systems appear very often in several branches of science and engineering, as biology, chemistry, mechanics, electronics, etc., see for instance [20] and references therein. In particular Liénard systems are frequent specially in physiological processes, see for instance [10]. Further, some planar systems can be transformed into (generalized) Liénard systems, see for example [5, 15]. In addition, the generalized polynomial Liénard systems is one of the most considered families in the study of limit cycles, see [18].

We assume $F_{\mu}(x) \not \equiv 0$ and $g_{\nu}(x) \not \equiv 0$, then we define

$$
m=\max _{1 \leq i \leq \mu}\left\{\operatorname{deg} F_{i}(x)\right\}
$$

and

$$
n=\max _{1 \leq i \leq \nu}\left\{\operatorname{deg} g_{i}(x)\right\} .
$$

For a small enough $\varepsilon$, let $\mathcal{F}_{\nu}^{\mu}(m, n)$ be the maximum number of limit cycles of $\left(1_{\varepsilon}\right)$ that bifurcate from cycles of the linear center $\left(1_{0}\right)$, i.e. the maximum number of medium amplitude limit cycles which can bifurcate from $\left(1_{0}\right)$ under the perturbation $\left(1_{\varepsilon}\right)$, in short

$$
\mathcal{H}_{\nu}^{\mu}(m, n)=\left\{\begin{array}{l}
\text { Maximum number of medium } \\
\text { amplitude limit cycles of }\left(1_{\varepsilon}\right)
\end{array}\right\} .
$$

The main problem concerning $\mathcal{H}_{\nu}^{\mu}(m, n)$ is finding its exact value.

We are interested in $\nu \geq 1$ because when $g_{i}(x) \equiv 0$ for all $i \geq 1$ the maximum number of medium amplitude limit cycles of $\left(1_{\varepsilon}\right)$ is well-known. Indeed, if we denote by $\mathcal{H}_{0}^{\mu}(m)$ the maximum number of medium amplitude limit cycles of $\left(1_{\varepsilon}\right)$ in such a case, then we know from [17] that $\mathcal{H}_{0}^{1}(m) \geq[(m-1) / 2]$, where [.] denotes the integer part function. Moreover, by following [7, Theorem 3.1] we can prove that $\mathcal{H}_{0}^{\mu}(m)=[(m-1) / 2]$ for $\mu \geq 1$. An explicit proof of this statement is provided in [1, Section 3.2.2] because it is known that the cyclicity of a non-degenerated center (as in our case) coincide with the cyclicity of the open period annulus surrounding it. See for instance [9]. Theorem 1 (below) is a generalization of this result, and it also improves the results of Section 4.3 .2 in [1] which prove that $\mathcal{H}_{\nu}^{\mu}(m, n)=[(m-1) / 2]$ for some families of generalized Liénard systems. A rewiev about the results conserning small and medium amplitude limit cycles of $\left(1_{\varepsilon}\right)$ is given in [19], where is also proved that

$$
\begin{aligned}
& \mathcal{H}_{1}^{1}(m, n) \geq\left[\frac{m-1}{2}\right] \\
& \mathcal{H}_{2}^{2}(m, n) \geq \max \left\{\left[\frac{m-1}{2}\right],\left[\frac{m}{2}\right]+\left[\frac{n}{2}\right]-1\right\} \\
& \mathcal{H}_{3}^{3}(m, n) \geq\left[\frac{m+n-1}{2}\right]
\end{aligned}
$$

However, the exact values of $\mathcal{H}_{1}^{1}(m, n), \mathcal{H}_{2}^{2}(m, n)$, and $\mathcal{H}_{3}^{3}(m, n)$ were not reported there.
In this paper we give the exact value of $\mathcal{H}_{\nu}^{\mu}(m, n)$ for two subfamilies of $\left(1_{\varepsilon}\right)$. More precisely, we consider the families:

$$
\mathcal{G} \mathcal{L} 1:=\left\{\begin{array}{c}
\text { Systems }\left(1_{\varepsilon}\right) \text { assuming that } \\
g_{i}(x) \text { is odd for } 1 \leq i \leq \nu
\end{array}\right\}
$$

and

$$
\mathcal{G} \mathcal{L} 2:=\left\{\begin{array}{l}
\text { Systems }\left(1_{\varepsilon}\right) \text { assuming that } \\
F_{i}(x) \text { is even for } \mu_{0}<i \leq \mu
\end{array}\right\},
$$

where $\mu_{0}$ is the smallest integer with $1 \leq \mu_{0} \leq \mu$ such that $F_{\mu_{0}}(x) \not \equiv 0$.
We will give the exact values of $\tilde{\mathcal{F}}_{\nu}^{\mu}(m, n)$ and $\overline{\mathcal{F}}_{\nu}^{\mu}(m, n)$ the maximum number of medium amplitude limit cycles of systems in $\mathcal{G} \mathcal{L} 1$ and $\mathcal{G} \mathcal{L} 2$, respectively. We note that if $\mu_{0}=\mu$, then $\overline{\mathcal{H}}_{\nu}^{\mu}(m, n)=\mathcal{H}_{\nu}^{\mu}(m, n)$.

Our main result is the following.
Theorem 1. The following statements hold.
(a) The exact value of $\tilde{\mathcal{H}}_{\nu}^{\mu}(m, n)$ is $\left[\frac{m-1}{2}\right]$. Moreover, for each $s$ with $0 \leq s \leq\left[\frac{m-1}{2}\right]$ there exist systems in GL1 having exactly s hyperbolic limit cycles.
(b) The exact value of $\overline{\mathcal{H}}_{\nu}^{\mu}(m, n)$ is either $\left[\frac{m-1}{2}\right]$ if $m$ is odd or $\left[\frac{m}{2}\right]+\left[\frac{n}{2}\right]-1$ if $m$ is even. Moreover, for each $s$ with $0 \leq s \leq\left[\frac{m}{2}\right]+\left[\frac{n}{2}\right]-1$ there exist systems in $\mathcal{G} \mathcal{L} 2$ having exactly $s$ hyperbolic limit cycles.

The assumptions on $g_{i}(x)$ and $F_{i}(x)$ in definitions of $\mathcal{G} \mathcal{L} 1$ and $\mathcal{G} \mathcal{L} 2$, respectively, are necessary. Otherwise, we can construct systems $\left(1_{\varepsilon}\right)$ having more medium amplitude limit cycles, see Remark 1 in Section 3.

Theorem 1 is a generalization of Theorem 1.1 in [22], where the case $\mu=\nu=1$ was considered. We note that in such a case $\overline{\mathcal{H}}_{1}^{1}(m, n)=\mathcal{H}_{1}^{1}(m, n)$. Hence Theorem 1.(b) gives the exact value of $\mathcal{H}_{1}^{1}(m, n)$.

The proof of Theorem 1 is based on computing the maximum number of isolated zeros of the first non-vanishing Poincaré-Pontryagin-Melnikov function of the displacement function of $\left(1_{\varepsilon}\right)$, by taking into account the restrictions: $g_{i}(x)$ odd for $1 \leq i \leq \nu$ and $F_{i}(x)$ even for $\mu_{0}<i \leq \mu$, respectively.

The paper is organized as follows. In Section 2 we recall the definition of the displacement function of $\left(1_{\varepsilon}\right)$, as well as the algorithm to compute the Poincaré-Pontryagin-Melnikov functions. Preliminary results that allow us to provide elementary proofs of the main result are given in Section 3. Finally, in Section 4 we will prove Theorem 1.

## 2. Poincaré-Pontryagin-Melnikov functions

The linear center $\left(1_{0}\right)$ is the Hamiltonian system associated to the polynomial $H=\left(x^{2}+y^{2}\right) / 2$; hence its cycles are the circles $\gamma_{c}=\{H-c=0\}$ with $c>0$. By using $c$ as a parameter, the first return map of $\left(1_{\varepsilon}\right)$ can be expressed in terms of $\varepsilon$ and $c: \mathcal{P}(\varepsilon, c)$. Therefore the corresponding displacement function $L(\varepsilon, c)=\mathcal{P}(\varepsilon, c)-c$ is analytic for small enough $\varepsilon$ and can be written as the power series in $\varepsilon$

$$
\begin{equation*}
L(\varepsilon, c)=\varepsilon L_{1}(c)+\varepsilon^{2} L_{2}(c)+O\left(\varepsilon^{3}\right) \tag{2}
\end{equation*}
$$

where $L_{i}(c)$ with $i \geq 1$ is the Poincaré-Pontryagin-Melnikov function of order $i$, which is defined for $c \geq 0$.
Let $L_{k}(c)$ with $k \geq 1$ be the first non-vanishing coefficient in (2). The zeros of $L_{k}(c)$ are important in the study of medium amplitude limit cycles of $\left(1_{\varepsilon}\right)$ because of the Poincaré-Pontryagin-Andronov criterion: The maximum number of isolated zeros, counting multiplicities, of $L_{k}(c)$ is an upper bound for $\mathcal{H}_{\nu}^{\mu}(m, n)$. Furthermore each simple zero $c_{0}$ of $L_{k}(c)$ corresponds to one and only one limit cycle of $\left(1_{\varepsilon}\right)$ with $\varepsilon$ small enough bifurcating from the cycle $\gamma_{c_{0}}$.

Now, we will recall the algorithm to compute the functions $L_{i}(c)$. System $\left(1_{\varepsilon}\right)$ can be written as

$$
\dot{x}=y, \quad \dot{y}=-x+\sum_{i \geq 1} \varepsilon^{i}\left(g_{i}(x)+f_{i}(x) y\right)
$$

where $f_{i}(x)=F_{i}^{\prime}(x)$, or equivalently as

$$
d H-\varepsilon \omega_{1}-\varepsilon^{2} \omega_{2}-\cdots=0
$$

with $\omega_{i}=\left(g_{i}(x)+f_{i}(x) y\right) d x$ and $\omega_{i} \equiv 0$ for $i>\max \{\mu, \nu\}$.
As we know, $L_{1}(c)$ is given by the classical Poincaré-Pontryagin formula $L_{1}(c)=\int_{\gamma_{c}} \omega_{1}$. A construction to compute the second order Poincaré-Pontryagin-Melnikov function of a perturbed system of the form $d H-\varepsilon \omega_{1}$ with $\omega_{1}$ an arbitrary polynomial 1-form was given by Yakovenko [1995]. After, Françoise [1996] gave the algorithm to know the Poincaré-Pontryagin-Melnikov function of any order of $d H-\varepsilon \omega_{1}$. Finally, Iliev [1999] gave the result for computing the higher order Poincaré-Pontryagin-Melnikov functions of a perturbed system of the form $d H-\varepsilon \omega_{1}-\varepsilon^{2} \omega_{2}-\cdots=0$, where $\omega_{i}$ for $i \geq 1$ are arbitrary polynomial 1-forms. His result is the following.

Theorem 2. [11]. If $k \geq 2$ and $L_{1}(c) \equiv \cdots \equiv L_{k-1}(c) \equiv 0$, then there are polynomials $q_{1}, \ldots, q_{k-1}$ and $Q_{1}, \ldots, Q_{k-1}$ such that $\Omega_{1}=d Q_{1}+q_{1} d H, \ldots, \Omega_{k-1}=d Q_{k-1}+q_{k-1} d H$, and

$$
L_{k}(c)=\int_{\gamma_{c}} \Omega_{k}
$$

where

$$
\begin{equation*}
\Omega_{1}=\omega_{1}, \Omega_{l}=\omega_{l}+\sum_{i+j=l} q_{i} \omega_{j}, \text { and } 2 \leq l \leq k \tag{4}
\end{equation*}
$$

The proof of this result easily follows from the Poincaré-Pontryagin formula, and the Ilyashenko-Gavrilov theorem ([12], [8]): If $\int_{\gamma_{c}} \omega=0$ for all $c \geq 0$, then $\omega=d Q+q d H$, where $Q$ and $q$ are polynomials, and by applying an induction argument. For a detailed proof, see for instance [11], [14].

On the other hand, we know from [11] that $L_{k}(c)$ has at most $[k(\max \{n, m\}-1) / 2]$ positive zeros, counting multiplicities. However, this result does not give the value of $\mathcal{H}_{\nu}^{\mu}(m, n)$ because the upper bound for $k$ depending on $\mu, \nu, m$, and $n$ is unknown. In addition, the number of isolated zeros of the first nonvanishing Poincaré-Pontryagin-Melnikov function does not provide the number of limit cycles of ( $1_{\varepsilon}$ ) with $\varepsilon$ small enough as shows next example.

Example 1. Consider the Liénard system

$$
\dot{x}=y+\varepsilon x-\varepsilon^{2} x^{3}, \quad \dot{y}=-x
$$

or equivalently $d H-\varepsilon \omega_{1}-\varepsilon^{2} \omega_{2}=0$ with $\omega_{1}=y d x$ and $\omega_{2}=3 x^{2} y d x$, where $\varepsilon$ is a small parameter.
A simple computation gives $L_{1}(c)=\int_{\gamma_{c}} \omega_{1}=-2 \pi c$. Hence system ( $5_{\varepsilon}$ ) does not have limit cycles bifurcating from the cycles of the linear center. However, for any $\varepsilon>0$ small enough the system ( $5_{\varepsilon}$ ) has a limit cycle bifurcating from the infinity; more precisely, if we consider $\left(5_{\varepsilon}\right)$ on the Poincaré sphere $\mathbb{S}^{2}$, then the limit cycle bifurcates from the equator of $\mathbb{S}^{2}$ which is known as "the circle at infinity" or "points at infinity" of $\mathbb{R}^{2}$ [21]. Indeed, by using the function

$$
V_{\varepsilon}(x, y)=4 y^{2}+4 \varepsilon x\left(1-\varepsilon x^{2}\right) y+4 x^{2}-\frac{3}{\varepsilon^{2}}
$$

it is not difficult to prove that the $1 / V_{\varepsilon}(x, y)$ is a Dulac function for $\left(5_{\varepsilon}\right)$ in $\mathbb{R}^{2} \backslash\left\{V_{\varepsilon}(x, y)=0\right\}$; moreover, it is easy to see that for $\varepsilon \in(0,1)$ the curve $\left\{V_{\varepsilon}(x, y)=0\right\}$ has an oval surrounding the origin (the unique singularity of $\left.\left(5_{\varepsilon}\right)\right)$. Hence, $\mathbb{R}^{2} \backslash\left\{V_{\varepsilon}(x, y)=0\right\}$ has a 1-connected component $\widetilde{U}_{\varepsilon}$, then the generalized Bendixon-Dulac theorem [6] ensures that ( $5 \varepsilon_{\varepsilon}$ ) has a hyperbolic limit cycle $\Gamma_{\varepsilon}$ in $\widetilde{U}_{\varepsilon}$ for each $\varepsilon \in(0,1)$. Thus, $\Gamma_{\varepsilon}$ contains the oval of $\left\{V_{\varepsilon}(x, y)=0\right\}$. See Section 4 of [4] for more details. Finally, a straightforward computation shows that the circle $x^{2}+y^{2}=1 /(2 \varepsilon)^{2}$ is contained in the bounded region limited by the oval of $\left\{V_{\varepsilon}(x, y)=0\right\}$. This implies that $\Gamma_{\varepsilon}$ bifurcates from the "the circle at infinity" of $\mathbb{R}^{2}$.

In next section we will give some properties on $\omega_{i}$ which will allow us to simplify the computation of the Poincaré-Pontryagin-Melnikov functions

## 3. Preliminary results

For computing $L_{k}(c)$ for $\left(1_{\varepsilon}\right)$ we will use the following two elementary lemmas whose proof is omitted.
Lemma 3. Let $P$ be a polynomial in the ring $\mathbb{R}\left[x^{2}, H\right]$. We define $\operatorname{deg}_{2} P$ to be the degree of $P$ in $\mathbb{R}\left[x^{2}, H\right]$.
(a) For $i, j \geq 0$ there are homogeneous polynomials $Q_{i j}, q_{i j} \in \mathbb{R}\left[x^{2}, H\right]$ with $\operatorname{deg}_{2} Q_{i j}=i+j$ and $\operatorname{deg}_{2} q_{i j}=$ $i+j-1$, such that

$$
H^{i} x^{2 j} d x=d\left(x Q_{i j}\right)+\left(x q_{i j}\right) d H
$$

or

$$
H^{i} x^{2 j+1} d x=d\left(x^{2} Q_{i j}\right)+\left(x^{2} q_{i j}\right) d H
$$

If $i=0$, then $q_{i j} \equiv 0$.
(b) For $i, j \geq 0$ there are homogeneous polynomials $Q_{i j}, q_{i j} \in \mathbb{R}\left[x^{2}, H\right]$ with $\operatorname{deg}_{2} Q_{i j}=i+j+1$ and $\operatorname{deg}_{2} q_{i j}=i+j$, such that

$$
H^{i} x^{2 j+1} y d x=d\left(y Q_{i j}\right)+\left(y q_{i j}\right) d H
$$

(c) For $i, j \geq 0$ we have

$$
\int_{\gamma_{c}} H^{i} x^{2 j} y d x=\frac{-\pi c}{2^{j}(2 j+1)}\binom{2(j+1)}{j+1} c^{i+j}
$$

Lemma 4. If $\omega, \eta \in \mathcal{A}$ and $q \in \mathcal{S}$ where

$$
\mathcal{A}:=\left\{(x A+x y B) d x \mid A, B \in \mathbb{R}\left[x^{2}, H\right]\right\}
$$

and

$$
\mathcal{S}:=\left\{x^{2} q_{1}+y q_{2} \mid q_{1}, q_{2} \in \mathbb{R}\left[x^{2}, H\right]\right\}
$$

then $\omega+\eta \in \mathcal{A}$ and $q \omega \in \mathcal{A}$.
The next two results are straightforward consequences of these two previous lemmas.
Corollary 5. If $\omega \in \mathcal{A}$, then $\int_{\gamma_{c}} \omega \equiv 0, \omega=d Q+q d H$ with $q \in \mathcal{S}$, and $q \omega \in \mathcal{A}$.
Corollary 6. If $P\left(x^{2}\right)=\sum_{r=0}^{d} p_{r} x^{2 r} \in \mathbb{R}\left[x^{2}\right]$, then

$$
\int_{\gamma_{c}} P\left(x^{2}\right) y d x=-\pi c \sum_{r=0}^{d}\binom{2(r+1)}{r+1} \frac{p_{r}}{2^{r}(2 r+1)} c^{r}
$$

We now will prove two lemmas which will be useful in the proof of Theorem 1.
Lemma 7. Suppose $k \geq 2$. Then the following statements hold.
(a) $\omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$ if and only if $\Omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$, where $\Omega_{l}$ is defined as in (4).
(b) If $\Omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$, then $L_{1}(c) \equiv \cdots \equiv L_{k-1}(c) \equiv 0$ and $L_{k}(c)=\int_{\gamma_{c}} \omega_{k}$.

Proof. (a) We proceed by induction on $k$. If $k=2$, then statement $(a)$ is true because $\Omega_{1}=\omega_{1} \in \mathcal{A}$. We now assume that the statement is true for $k-1$, and we will prove it for $k$.

From the induction hypothesis it follows that $\omega_{l}, \Omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-2$. Thus, by Corollary 5 , $\Omega_{l}=d Q_{l}+q_{l} d H$ with $q_{l} \in \mathcal{S}$ for all $1 \leq l \leq k-2$, and by using Lemma 4 we conclude that $\bar{\Omega}_{k-1}:=$ $\sum_{i+j=l} q_{i} \omega_{j} \in \mathcal{A}$. Hence, since $\Omega_{k-1}=\omega_{k-1}+\bar{\Omega}_{k-1}$, Lemma 4 implies that $\omega_{k-1} \in \mathcal{A}$ if and only if $\Omega_{k-1} \in \mathcal{A}$.
(b) By Corollary $5, \Omega_{l}=d Q_{l}+q_{l} d H$ with $q_{l} \in \mathcal{S}$ for all $1 \leq l \leq k-1$, and $L_{1}(c) \equiv \cdots \equiv L_{k-1}(c) \equiv 0$. In addition, by the statement $(a), \omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$. Thus, $\bar{\Omega}_{k}:=\sum_{i+j=k} q_{i} \omega_{j} \in \mathcal{A}$ because of Lemma 4, which implies that $\int_{\gamma_{c}} \bar{\Omega}_{k} \equiv 0$ by Corollary 5. Finally, from Theorem 2 we have $L_{k}(c)=\int_{\gamma_{c}} \omega_{k}+\int_{\gamma_{c}} \bar{\Omega}_{k}$. Therefore $L_{k}(c)=\int_{\gamma_{c}} \omega_{k}$.

Before announce next lemma, we note that each polynomial $h(x)=\sum_{r=0}^{m-1} a_{r} x^{r}$ of degree $m-1$ can be written as

$$
\begin{equation*}
h(x)=\hat{h}\left(x^{2}\right)+x \tilde{h}\left(x^{2}\right) \tag{6}
\end{equation*}
$$

where

$$
\begin{equation*}
\hat{h}\left(x^{2}\right)=\sum_{r=0}^{\left[\frac{m-1}{2}\right]} a_{2 r+1} x^{2 r}, \quad \tilde{h}\left(x^{2}\right)=\sum_{r=0}^{\left[\frac{m-2}{2}\right]} a_{2 r+2} x^{2 r} . \tag{7}
\end{equation*}
$$

Lemma 8. Let $\omega=(g(x)+f(x) y) d x$, where $f(x)=\sum_{r=0}^{m-1} a_{r} x^{r}$ and $g(x)=\sum_{s=0}^{n} b_{s} x^{s}$.
(a) $\int_{\gamma_{c}} \omega=\int_{\gamma_{c}} \hat{f}\left(x^{2}\right) y d x$ and its value is

$$
-\pi c \sum_{r=0}^{\left[\frac{m-1}{2}\right]}\binom{2(r+1)}{r+1} \frac{a_{2 r+1}}{2^{r}(2 r+1)} c^{r}
$$

(b) If $\int_{\gamma_{c}} \omega \equiv 0$, then $\omega=d Q+(y \bar{q}) d H$ with $\bar{q} \in \mathbb{R}\left[x^{2}, H\right]$ of degree $\operatorname{deg}_{2} \bar{q}=[(m-2) / 2]$, and $\int_{\gamma_{c}}(y \bar{q}) \omega=$ $\int_{\gamma_{c}} \bar{q} \hat{g}\left(x^{2}\right) y d x$ whose value is

$$
-\pi c \sum_{s=0}^{\left[\frac{n}{2}\right]} \sum_{r=0}^{\left[\frac{m-2}{2}\right]}\binom{2(s+r+1)}{s+r+1} \frac{\left(b_{2 s}\right)\left(a_{2 r+2}\right)}{2^{s+r}(2 s+1)} c^{s+r}
$$

(c) $\int_{\gamma_{c}}(y \bar{q}) \omega \equiv 0$ if and only if $\bar{q} \equiv 0$ or $\hat{g}\left(x^{2}\right) \equiv 0$.

Proof. (a) By statements (a) and (b) of Lemma 3, $\int_{\gamma_{c}} \omega=\int_{\gamma_{c}} \hat{f}\left(x^{2}\right) y d x$, and the value of this integral follows from Corollary 6.
(b) If $\int_{\gamma_{c}} \omega \equiv 0$, then $\hat{f}\left(x^{2}\right) \equiv 0$ by $(a)$. This implies that $\omega=g(x) d x+x \tilde{f}\left(x^{2}\right) y d x$ and by (7) we have

$$
\omega=d\left(\int g(x) d x\right)+\sum_{r=0}^{\left[\frac{m-2}{2}\right]} a_{2 r+2} x^{2 r+1} y d x
$$

From Lemma 3.(b) we obtain $x^{2 r+1} y d x=d\left(y \bar{Q}_{r}\right)+\left(y \bar{q}_{r}\right) d H$ for some homogeneous polynomials $\bar{Q}_{r}, \bar{q}_{r} \in$ $\mathbb{R}\left[x^{2}, H\right]$ with $\operatorname{deg}_{2} \bar{Q}_{r}=r+1$ and $\operatorname{deg}_{2} \bar{q}_{r}=r$, respectively. Hence

$$
\omega=d Q+(y \bar{q}) d H
$$

with

$$
Q=\int g(x) d x+y \sum_{r=0}^{\left[\frac{m-2}{2}\right]} a_{2 r+2} \bar{Q}_{r} \quad \text { and } \quad \bar{q}=\sum_{r=0}^{\left[\frac{m-2}{2}\right]} a_{2 r+2} \bar{q}_{r}
$$

Thus $\bar{q} \in \mathbb{R}\left[x^{2}, H\right]$ is homogeneous and $\operatorname{deg}_{2} \bar{q}=\left[\frac{m-2}{2}\right]$. Moreover, a simple computation shows that

$$
\begin{equation*}
\bar{q}_{r}=2 \sum_{i=0}^{r}\binom{r+1}{i}\left(\frac{r+1-i}{2 i+1}\right)(2 H)^{r-i}\left(x^{2}-2 H\right)^{i} \tag{8}
\end{equation*}
$$

As $(y \bar{q}) \omega=\bar{q} \hat{g}\left(x^{2}\right) y d x+\bar{q} \tilde{g}\left(x^{2}\right) x y d x+\bar{q} \tilde{f}\left(x^{2}\right) x y^{2} d x$ and $\bar{q} \tilde{f}\left(x^{2}\right) x y^{2} d x=\bar{q} \tilde{f}\left(x^{2}\right) x\left(2 H-x^{2}\right) d x$, it follows that $(y \bar{q}) \omega=\bar{q} \hat{g}\left(x^{2}\right) y d x+d Q_{2}+q_{2} d H$ because of statements $(a)$ and ( $b$ ) of Lemma 3. Hence we obtain $\int_{\gamma_{c}}(y \bar{q}) \omega=\int_{\gamma_{c}} \bar{q} \hat{g}\left(x^{2}\right) y d x$. That is,

$$
\int_{\gamma_{c}}(y \bar{q}) \omega=\int_{\gamma_{c}}\left(\sum_{r=0}^{\left[\frac{m}{2}\right]-1} a_{2 r+2} \bar{q}_{r}\right)\left(\sum_{s=0}^{\left[\frac{n}{2}\right]} b_{2 s} x^{2 s}\right) y d x .
$$

By using expression (8) of $\bar{q}_{r}$, a straightforward computation, and Lemma $3(c)$ we obtain the formula given in the statement. Finally, statement $(c)$ follows from the formula given in statement $(b)$.

Remark 1. System ( $1_{\varepsilon}$ ) with $\mu=\nu=1, F_{1}(x)=-x^{2}$, and $g_{1}(x)=1-x^{2}$ does not satisfy the hypothesis in definition of $\mathcal{G} \mathcal{L} 1$ because $g_{1}(x)$ is not an odd function. Here $m=n=2$ and from Theorem 1.(a) it follows that $\tilde{\mathcal{H}}_{1}^{1}(2,2)=0$; however, for $\varepsilon$ small enough, this system has one medium amplitude limit cycle. Indeed, we need only to prove that the first non-vanishing coefficient of the displacement function (2), associated to this system, has a simple positive zero. For that we write system in the form $\left(3_{\varepsilon}\right)$ as $d H-\varepsilon \omega=0$ with $\omega=\left(1-x^{2}-2 x y\right) d x$. By Lemma 8. $(a), L_{1}(c) \equiv 0$, and by Theorem 2 and Lemma 8. $(b), L_{2}(c)=-\pi c(4-2 c)$.

Now, system $\left(1_{\varepsilon}\right)$ with $\mu=\nu=2, F_{1}(x)=-3 x^{2}, F_{2}(x)=-2 x^{3}, g_{1}(x)=x^{2}+x^{3}$, and $g_{2}(x)=$ $\left(-5+25 x^{2}\right) / 6$ does not satisfy the hypothesis in definition of $\mathcal{G} \mathcal{L} 2$ because $F_{2}(x)$ is not an even function. In this case $m=n=3$ and by Theorem $1 .(b), \mathcal{H}_{2}^{2}(3,3)=1$; however, for $\varepsilon$ small enough, the resulting system has two medium amplitude limit cycles. Indeed, following previous ideas, and using Theorem 2 and Lemma 8 it is easy to see that $L_{1}(c) \equiv 0, L_{2}(c) \equiv 0$, and $L_{3}(c)=-\pi c(c-1)(c-2)$.

## 4. Proof of the Theorem 1

We can assume, after a linear change of variables if necessary, that $F_{i}(0)=0$ for all $1 \leq i \leq \mu$. Suppose that $F_{i}(x)=\sum_{r=1}^{m}\left(a_{i(r-1)} / r\right) x^{r}$ and $g_{i}(x)=\sum_{s=0}^{n} b_{i s} x^{s}$. Thus, $f_{i}(x)=F_{i}^{\prime}(x)=\sum_{r=0}^{m-1} a_{i r} x^{r}$ and $g_{i}(x)$ can be written as $f_{i}(x)=\hat{f}_{i}\left(x^{2}\right)+x \tilde{f}_{i}\left(x^{2}\right)$ and $g_{i}(x)=\hat{g}_{i}\left(x^{2}\right)+x \tilde{g}_{i}\left(x^{2}\right)$, respectively, according to (6).

Proof of Theorem 1. (a) By hypothesis, $g_{i}(x)$ is odd for $1 \leq i \leq \nu$, which means that $g_{i}(x)=x \tilde{g}_{i}\left(x^{2}\right)$ for $1 \leq i \leq \nu$. Let $L_{k}(c)$ be the first non-vanishing Poincaré-Pontryagin-Melnikov function in (2). For proving the statement we will prove first that $L_{k}(c)$ has at most $[(m-1) / 2]$ positive zeros, and then that for each $s$ with $0 \leq s \leq[(m-1) / 2]$ we can choose systems in $\mathcal{G} \mathcal{L} 1$ in such a way that $L_{k}(c)$ has exactly $s$ simple positive zeros.

If $k=1$, then the assertion is true. Indeed, we have

$$
L_{1}(c)=\int_{\gamma_{c}} x \tilde{g}_{i}\left(x^{2}\right) d x+\hat{f}_{1}\left(x^{2}\right) y d x+\tilde{f}_{1}\left(x^{2}\right) x y d x
$$

Thus, as $\int_{\gamma_{c}} x \tilde{g}_{i}\left(x^{2}\right) d x \equiv 0$ and $\int_{\gamma_{c}} \tilde{f}_{1}\left(x^{2}\right) x y d x \equiv 0$ by Corollary 5 , we obtain $L_{1}(c)=\int_{\gamma_{c}} \hat{f}_{1}\left(x^{2}\right) y d x$. From (7) we know that $\operatorname{deg}_{2} \hat{f}_{1}\left(x^{2}\right)=[(m-1) / 2]$, which implies that $L_{1}(c)$ has at most $[(m-1) / 2]$ positive zeros because of Corollary 6 . In addition, since $\hat{f}_{1}\left(x^{2}\right)$ has $[(m-1) / 2]+1$ independent coefficients, for each $s$ with $0 \leq s \leq[(m-1) / 2]$ we can choose suitable coefficients of $\hat{f}_{1}\left(x^{2}\right)$ in such a way that $L_{k}(c)$ has exactly $s$ simple positive zeros. Therefore, by applying the Poincaré-Pontryagin-Andronov criterion it follows that
the corresponding system $\left(1_{\varepsilon}\right)$, which belongs to $\mathcal{G} \mathcal{L} 1$, has exactly $s$ hyperbolic limit cycles. In particular we have proved that $\tilde{\mathcal{H}}_{\nu}^{\mu}(m, n)=[(m-1) / 2]$.

Suppose then that $k \geq 2$. If we prove that $\Omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$, then $L_{k}(c)=\int_{\gamma_{c}} \omega_{k}$ by Lemma 7 , and by applying the same idea as in previous paragraph the assertion follows. Therefore, it remains to show that $\Omega_{l} \in \mathcal{A}$ for $1 \leq l \leq k-1$.

We proceed by induction on $k$. If $k=2$, then $L_{1}(c) \equiv 0$, which implies that

$$
\Omega_{1}=\left(x \tilde{g}_{1}\left(x^{2}\right)+x y \tilde{f}_{1}\left(x^{2}\right)\right) d x \in \mathcal{A}
$$

Hence the assertion is true for $k=2$.
We now assume that the assertion is true for $k-2$, and we will prove it for $k-1$. By induction hypothesis, $\Omega_{i} \in \mathcal{A}$ for $1 \leq i \leq k-2$, which implies that $\Omega_{i}=d Q_{i}+q_{i} d H$ with $q_{i} \in \mathcal{S}$ for $1 \leq i \leq k-2$ by Corollary 5 . Furthermore, by Lemma $7, \omega_{j} \in \mathcal{A}$ for $1 \leq j \leq k-2$. Hence $\bar{\Omega}_{k-1}:=\sum_{i+j=k-1} q_{i} \omega_{j}($ with $1 \leq i, j \leq k-2)$ is an element of $\mathcal{A}$ because of Lemma 4. Since $\Omega_{k-1}=\omega_{k-1}+\bar{\Omega}_{k-1}$,

$$
L_{k-1}(c)=\int_{\gamma_{c}} \Omega_{k-1}=\int_{\gamma_{c}} \omega_{k-1}=\int_{\gamma_{c}} \hat{f}_{k-1}\left(x^{2}\right) y d x
$$

which vanishes identically. Therefore, $\omega_{k-1}=\left(x \tilde{g}_{k-1}\left(x^{2}\right)+x y \tilde{f}_{k-1}\left(x^{2}\right)\right) d x \in \mathcal{A}$. Thus $\Omega_{k-1} \in \mathcal{A}$.
(b) Firstly we will show two properties concerning $\omega_{i}$ and $\int_{\gamma_{c}} \omega_{i}$ which we will use along the proof of the statement. Then, we will split the proof into two cases: $m$ odd and $m$ even.

For $1 \leq i<\mu_{0}$ the 1 -form $\omega_{i}$ has the form $g_{i}(x) d x$ that is exact: $\omega_{i}=d Q_{i}+q_{i} d H$ with $q_{i} \equiv 0$. Hence, from (4) it follows that $\Omega_{i}=\omega_{i}$ for $1 \leq i \leq \mu_{0}$. Thus, $L_{i}(c)=\int_{\gamma_{c}} \Omega_{i} \equiv 0$ for $1 \leq i<\mu_{0}$ and $L_{\mu_{0}}(c)=$ $\int_{\gamma_{c}} \Omega_{\mu_{0}}=\int_{\gamma_{c}} \omega_{\mu_{0}}$. On the other hand, since $F_{i}(x)$ is even for $\mu_{0}<i \leq \mu, f_{i}(x)=x \tilde{f}_{i}\left(x^{2}\right)$ for $\mu_{0}<i \leq \mu$. Thus, for $i>\mu_{0}$ we have $\omega_{i}=d\left(\int g_{i}(x) d x\right)+x \tilde{f}_{i}\left(x^{2}\right) y d x$. Moreover, $x^{2 r+1} y d x=d\left(y \bar{Q}_{0 r}\right)+\left(y \bar{q}_{0 r}\right) d H$ because of Lemma 3.(b). Hence,

$$
\begin{equation*}
\omega_{i}=d\left(\bar{Q}_{i}\right)+\left(y \bar{q}_{i}\right) d H \tag{9}
\end{equation*}
$$

of course $\bar{q}_{i} \equiv 0$ for $i>\mu$. Therefore we obtain

$$
\begin{equation*}
\int_{\gamma_{c}} \omega_{i} \equiv 0 \quad \text { for } i>\mu_{0} \tag{10}
\end{equation*}
$$

Case $m$ odd. In this case $\operatorname{deg} F_{\mu_{0}}(x)=m$ because $F_{i}(x)$ is an even polynomial for $\mu_{0}<i \leq \mu$. Since $F_{\mu_{0}}^{\prime}(x)=f_{\mu_{0}}(x)=\hat{f}_{\mu_{0}}\left(x^{2}\right)+x \tilde{f}_{\mu_{0}}\left(x^{2}\right)$ has an even degree, $\hat{f}_{\mu_{0}}\left(x^{2}\right) \not \equiv 0$. Hence, from Lemma 8.(a) it follows that $L_{\mu_{0}}(c)=\int_{\gamma_{c}} \omega_{\mu_{0}}=\int_{\gamma_{c}} \hat{f}_{\mu_{0}}\left(x^{2}\right) y d x \not \equiv 0$, and it has at most $[(m-1) / 2]$ positive zeros, counting multiplicities; moreover, we can choose suitable coefficients of $F_{\mu_{0}}(x)$ in such a way that $L_{\mu_{0}}(c)$ has exactly $[(m-1) / 2]$ simple positive zeros. Therefore by the Poincaré-Pontryagin-Andronov criterion, $\overline{\mathcal{F}}_{\nu}^{\mu}(m, n)=[(m-1) / 2]$.

Case $m$ even. Let $L_{k}(c)$ be the first non-vanishing Poincaré-Pontryagin-Melnikov function of (2). If $k=\mu_{0}$, then $L_{\mu_{0}}(c)=\int_{\gamma_{c}} \omega_{\mu_{0}}$ has at most $[(m-1) / 2]$ positive zeros, counting multiplicities, because of Lemma 8. $(a)$. Since $m$ is even, $[(m-1) / 2] \leq[m / 2]+[n / 2]-1$. Hence $L_{\mu_{0}}(c)$ has at most $[m / 2]+[n / 2]-1$ positive zeros, counting multiplicities.

We claim that if $k \geq \mu_{0}+1$, then $\omega_{1}, \ldots, \omega_{k-1-\mu_{0}} \in \mathcal{A}, \Omega_{i}=d Q_{i}+q_{i} d H$ with $q_{i} \in \mathcal{S}$ for $\mu_{0} \leq i \leq$ $k-1$, and $L_{k}(c)=\int_{\gamma_{c}}\left(y \bar{q}_{\mu_{0}}\right) \omega_{k-\mu_{0}}=\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{k-\mu_{0}}\left(x^{2}\right) y d x$. By assuming that this assertion is true and
by applying Lemma 8 .(b) we conclude that $L_{k}(c)$ has at most $[m / 2]+[n / 2]-1$ positive zeros, counting multiplicities; moreover, for each $s$ with $0 \leq s \leq[m / 2]+[n / 2]-1$ we can choose suitable coefficients of $\bar{q}_{\mu_{0}}$ and $\hat{g}_{k-\mu_{0}}\left(x^{2}\right)$ in such a way that $L_{k}(c)$ has exactly $s$ simple positive zeros. Thus, by the Poincaré-Pontryagin-Andronov criterion the corresponding system $\left(1_{\varepsilon}\right)$ has exactly $s$ hyperbolic limit cycles; in particular we have $\overline{\mathcal{H}}_{\nu}^{\mu}(m, n)=[m / 2]+[n / 2]-1$. Therefore, to finish the proof of statement (b) we need only to confirm our assertion, which we prove next by proceeding by induction on $k$.

If $k=\mu_{0}+1$, then we will prove that $\Omega_{\mu_{0}}=d Q_{\mu_{0}}+q_{\mu_{0}} d H$ with $q_{\mu_{0}} \in \mathcal{S}$, and that $L_{\mu_{0}+1}(c)=$ $\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{1}\left(x^{2}\right) y d x$. We know that $\Omega_{\mu_{0}}=\omega_{\mu_{0}}$ and since $k=\mu_{0}+1, L_{\mu_{0}}(c)=\int_{\gamma_{c}} \Omega_{\mu_{0}} \equiv 0$. Thus, from Lemma 8.(b) it follows that $\Omega_{\mu_{0}}=\omega_{\mu_{0}}=d Q_{\mu_{0}}+q_{\mu_{0}} d H$, where $q_{\mu_{0}}=y \bar{q}_{\mu_{0}} \not \equiv 0 \in \mathcal{S}$. On the other hand, by Theorem 2, $L_{\mu_{0}+1}(c)=\int_{\gamma_{c}} \Omega_{\mu_{0}+1}$, where $\Omega_{\mu_{0}+1}=\omega_{\mu_{0}+1}+\sum_{i+j=\mu_{0}+1} q_{i} \omega_{j}$. Since $q_{i} \equiv 0$ for $1 \leq i<\mu_{0}$, $\Omega_{\mu_{0}+1}=\omega_{\mu_{0}+1}+q_{\mu_{0}} \omega_{1}$. Hence, by (10) we obtain $L_{\mu_{0}+1}(c)=\int_{\gamma_{c}}\left(y \bar{q}_{\mu_{0}}\right) \omega_{1}=\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{1}\left(x^{2}\right) y d x$.

If $k=\mu_{0}+2$, then $\Omega_{\mu_{0}}=\omega_{\mu_{0}}=d Q_{\mu_{0}}+q_{\mu_{0}} d H$, where $q_{\mu_{0}}=y \bar{q}_{\mu_{0}} \not \equiv 0 \in \mathcal{S}$ and $L_{\mu_{0}+1}(c) \equiv 0$. Since $\bar{q}_{\mu_{0}} \not \equiv 0, \hat{g}_{1}\left(x^{2}\right) \equiv 0$ by Lemma 8.(c). Thus, $\Omega_{1}=\omega_{1} \in \mathcal{A}$, and Corollary 5 implies that $\Omega_{1}=d Q_{1}+q_{1} d H$ with $q_{1} \in \mathcal{S}$. Moreover, from (9) $\omega_{\mu_{0}+1}=d\left(\bar{Q}_{\mu_{0}+1}\right)+\left(y \bar{q}_{\mu_{0}+1}\right) d H$, whence

$$
\Omega_{\mu_{0}+1}=\omega_{\mu_{0}+1}+q_{\mu_{0}} \omega_{1}=d Q_{\mu_{0}+1}+q_{\mu_{0}+1} d H
$$

with $q_{\mu_{0}+1} \in \mathcal{S}$ because of Corollary 5 .
On the other hand, from Theorem 2 we have

$$
L_{\mu_{0}+2}(c)=\int_{\gamma_{c}} \omega_{\mu_{0}+2}+\sum_{i+j=\mu_{0}+2} \int_{\gamma_{c}} q_{i} \omega_{j}
$$

As $\omega_{1} \in \mathcal{A}$ and $q_{\mu_{0}+1} \in \mathcal{S}$, then $q_{\mu_{0}+1} \omega_{1} \in \mathcal{A}$ following Lemma 4 and $\int_{\gamma_{c}} q_{\mu_{0}+1} \omega_{1} \equiv 0$ by Corollary 5. In addition, we know that $q_{i} \equiv 0$ for $1 \leq i<\mu_{0}$ and $\int_{\gamma_{c}} \omega_{\mu_{0}+2} \equiv 0$ by (10). Hence $L_{\mu_{0}+2}(c)=\int_{\gamma_{c}}\left(y \bar{q}_{\mu_{0}}\right) \omega_{2}=$ $\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{2}\left(x^{2}\right) y d x$.

We now assume that the assertion holds for $k-1$, and we will prove it for $k$. By Theorem $2, L_{k}(c)=$ $\int_{\gamma_{c}} \Omega_{k}$, where

$$
\Omega_{k}=\omega_{k}+q_{1} \omega_{k-1}+\cdots+q_{\mu_{0}-1} \omega_{k+1-\mu_{0}}+q_{\mu_{0}} \omega_{k-\mu_{0}}+q_{\mu_{0}+1} \omega_{k-1-\mu_{0}}+\cdots+q_{k-2} \omega_{2}+q_{k-1} \omega_{1}
$$

Since $q_{i} \equiv 0$ for $1 \leq i<\mu_{0}, \Omega_{k}=\omega_{k}+q_{\mu_{0}} \omega_{k-\mu_{0}}+q_{\mu_{0}+1} \omega_{k-1-\mu_{0}}+\cdots+q_{k-2} \omega_{2}+q_{k-1} \omega_{1}$.
On the other hand, from the induction hypothesis it follows that $\omega_{1}, \ldots, \omega_{k-2-\mu_{0}} \in \mathcal{A}, \Omega_{i}=d Q_{i}+q_{i} d H$ with $q_{i} \in \mathcal{S}$ for $\mu_{0} \leq i \leq k-2$, and $L_{k-1}(c)=\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{k-1-\mu_{0}}\left(x^{2}\right) y d x$. Since $L_{k-1}(c) \equiv 0, \hat{g}_{k-1-\mu_{0}}\left(x^{2}\right) \equiv 0$ because of Lemma 8. $(c)$, which implies that $\omega_{k-1-\mu_{0}} \in \mathcal{A}$. Therefore, $q_{\mu_{0}} \omega_{k-1-\mu_{0}}+\cdots+q_{k-3} \omega_{2}+q_{k-2} \omega_{1} \in \mathcal{A}$ by Lemma 4 . Moreover, from (10) $\omega_{k-1}=d\left(\bar{Q}_{k-1}\right)+\left(y \bar{q}_{k-1}\right) d H$, and by applying Corollary 5 we obtain

$$
\Omega_{k-1}=d Q_{k-1}+q_{k-1} d H, \quad \text { with } q_{k-1} \in \mathcal{S}
$$

Hence $q_{\mu_{0}+1} \omega_{k-1-\mu_{0}}+\cdots+q_{k-2} \omega_{2}+q_{k-1} \omega_{1} \in \mathcal{A}$ by Lemma 4. In addition, $\omega_{k}=d\left(\bar{Q}_{k}\right)+\left(y \bar{q}_{k}\right) d H$ by (10). Thus, we obtain $L_{k}(c)=\int_{\gamma_{c}} q_{\mu_{0}} \omega_{k-\mu_{0}}=\int_{\gamma_{c}} \bar{q}_{\mu_{0}} \hat{g}_{k-\mu_{0}}\left(x^{2}\right) y d x$.

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